Algirdas Vaclovas Valiulis

A HISTORY OF MATERIALS AND TECHNOLOGIES DEVELOPMENT

TEXTBOOK

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VILNIAUS GEDIMINO TECHNIKOS UNIVERSITETAS

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Contents

Acknowledgements

This textbook can be exclusively helpful for engineering students to take a course on history at Technical University. The work gives valuable information taken from the textbooks or other books published by my colleagues in this particular field. Inevitably, as the textbook of such a broad scope drew on the artefacts or expertise of many scholars, I gratefully thank them personally. One of the pleasures of writing this textbook are learning from other people, frequently from the other field of interests you are involved in. There are dozens of people who would have been able to write every chapter of this textbook in a more appropriate way than I did. However, the only relief is that a significant part of immensely rich materials of technology development is exemplified by the artifacts of the development of local or regional technology. Such an approach can be justified by a wish to get familiar with local and incoming students of engineering and technology development in the area of the Eastern part of the Baltic Sea. My job was to combine materials revealed and displayed by other authors or scientists into a single compilation. Therefore, I am truly grateful to everyone the scientific research areas of which I have covered as well as to the personnel of Vilnius Gediminas Technical University Press *Technika* that contributed to this textbook. Finally, I would also like to acknowledge Science Museums, companies and individuals for permission to reproduce illustrations contained in this book.

Preface

The textbook presents the history of inventions from a stone axe to a spacecraft and from flint tools to silicon chips in computer. The book focuses on simplifying the study on the History of Technology by putting into the hands of the reader a single volume telling the whole story presented in twenty-eight chapters and illustrated by various pictures. Technology refers to the collection of tools, including machinery, arrangements and procedures used by human societies. The book begins with a general historical survey of ancient civilizations, then, considers such topics as metalworking, early building construction and sources of power, the development of the steam engine, machine tools, modern transport, mining coal and ores, the rise of modern chemical industry, the internal combustion engine, electricity, etc. From the very beginning, human has existed in a hostile environment and has had to use his wits in the fierce struggle for survival. Created or applied technologies affect human ability to adapt to their natural environments. The use of technology began with the conversion of natural resources (stone, bone, animal leather) into simple tools. Up to now, it has been believed that the development of technology was restricted only to human beings. However, recent scientific studies have indicated that other primates, not in the same scale as humans, developed simple tools and learned to pass their knowledge to other generations. The distinction between science, engineering and technology is not always clear. *Science* is reasoned investigation into the phenomena aimed at discovering enduring principles among the elements of the phenomenal world by employing formal techniques like the scientific method. *Engineering* is the goal-oriented process of designing and making tools and systems to exploit natural phenomena for practical human needs. *Technologies* are often a consequence of science and engineering, because sometimes they have to satisfy contradictory requirements comprising utility, usability, durability, safety, etc. In a simple way, technology can be defined as the entities, both material and immaterial, created by the application of mental and physical effort in order to achieve some value. In this usage, technology refers to tools and machines that may be used for solving real-world problems.

It is unaccountable that the available studies of engineering have paid so little attention to the history of technology development. What pharaohs, emperors, generals and statesmen did in the past and how they won their fighting was largely dependent on the state of their possessed technologies. By studying the past, we are able to observe their successes while perceiving their mistakes. Normally, in the process of creating engineering design, the inventor or designer starts his preparatory quest with an exhaustive look at the present and past. Typically, the inventor seeks for improving design or technology based on past and present experience, and this is the first step in the process of moving towards a new solution. Thus, the history of technology and the history of invention are very much the same.

Technology is all around us: we live in the world where everything that exists can be classified as either the work of nature or man. Technology has affected society in a number of ways. In many societies, technology has helped with developing more advanced economies, but plenty of technological processes produce pollution as well as deplete natural resources and cause detriment for the Earth's environment. Philosophical debates over the present and future use of technology have arisen in society and have evoked disagreements whether technology improves the human condition or worsens it. Unfortunately, we are not remote observers of history. We are the passengers of the fast moving coach able to observe the surrounding from inside rather than from outside.

CHAPTER 1 The Place of Technology in History

This chapter will help in

- acquiring a broad knowledge of the dawn of human history;
- explaining the measures of human progress;
- computing cultural evolution in terms of the amount of energy harnessed per capita per year;
- describing the earliest tools and technologies;
- defining the reasons for the spread of Palaeolithic people across Europe.

Dawn of Human History

Human history began in the Pleistocene – the geological epoch that lasted from about 2,588,000 to 11,700 years ago, spanning the world's recent period of repeated glaciations. Up to 30% of the Earth's surface was glaciated periodically during the Pleistocene in the beginning of which, *Paranthropus* species and early human ancestors were still present. However, during the Lower Palaeolithic Period, they disappeared, and the only hominid species found in fossil records is *Homo erectus* for much of the Pleistocene. The Middle and late Palaeolithic saw the appearance of the new types of humans and the development of more elaborate tools than those discovered in the previous eras. According to mitochondrial timing techniques, modern humans migrated from Africa in the middle Palaeolithic, spreading all over the ice-free world during the late Pleistocene. Human history most importantly influenced a drop in the sea level aided in humanoid movement from Asia into North America, as the landmass connecting the two areas in the Bering Strait in Alaska (Beringia) surfaced to act as a bridge between them. The corridor came into existence towards the end of the late Wisconsin and could have served to accommodate two-way traffic.

Large portions of Europe, North America (including Greenland) and South America, all of Antarctica and small sections of Asia were entirely covered by ice. In North America, during the peak of the Wisconsin glaciation approximately 18,000 years ago, there were two massive yet independent ice sheets. Both the eastern Laurentide and the western Cordilleran ice sheets were over 3900 meters thick. The Laurentide ice sheet was a massive sheet of ice that covered hundreds of thousands of square miles of North America 95,000–20,000 years before the present day (Figure 1.1) (A. S. Dyke, J. T. Andrews *et al.* 2002). It now appears increasingly likely that humans occupied North America before the last glacial expansion of the late Wisconsin Ice Age, ca 20 to 25 thousand years before present, in which case the much diminished ice would not have acted as an impassable barrier. The corridor, such as it was, came into existence and could have served to accommodate two-way traffic.

The ice sheet disintegrated and created at its front an immense proglacial Lake Agassiz formed from its meltwater (Figure 1.2). Lake Agassiz was likely the largest of these proglacial lakes occupying, at various times in its 4000-year history, large portions of North America. Around 13,000 years ago, it may have covered as much as 440,000 km^2, which is approximately the size of the Black Sea.

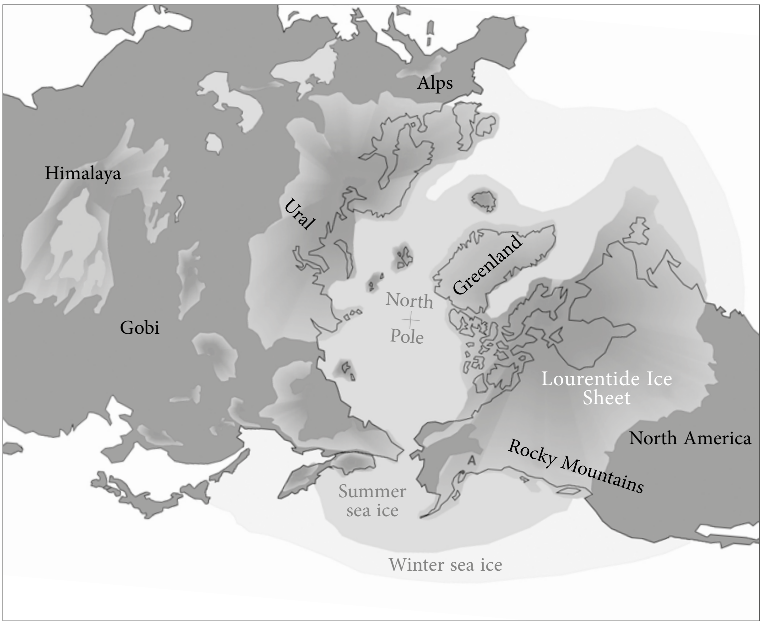

Figure 1.1. The Northern hemisphere glaciation

Public domain

Figure 1.2. Lake Agassiz in North America

Thermohaline circulation, the so-called "ocean conveyor belt," would be radically altered by melting ice sheets. The ocean conveyor belt circulates nutrient-rich water from polar regions throughout the world's oceans in a long, slow, continual loop. Circulation relies on the relationship between water having different densities. Cold, saline water from polar regions gradually rises to the surface in the tropics. Melting ice sheets would increase the amount of warm and freshwater in polar marine ecosystems. This would slow "deep water formation" and the development of cold, saline, nutrient-rich water on which the entire marine ecosystems depend. The major drainage events taking place around Lake Agassiz were of such magnitudes that they had a significant impact on climate, sea level and possibly early human civilization. Lake freshwater released into the Arctic Ocean is considered to be able to disrupt oceanic circulation and cause temporary cooling. The failure of North Atlantic thermohaline circulation is used for explaining a rapid change in the climate, and the freezing of Europe after the Gulf Stream was disrupted. Today warm water is conveyed across to Europe near the surface of the Atlantic, and at its northernmost point, it cools and sinks because the water from the tropics has higher salt content making it denser and therefore heavier than the water near the Arctic. Cold dense water then flows south close to the ocean floor and back towards central and south America to complete the cycle. Around 10,000 years ago, Lake

Agassiz refilled. The last major shift in drainage occurred around 8,200 years ago. The melting of the remaining Hudson Bay ice caused Lake Agassiz to drain nearly completely. This final drainage of Lake Agassiz is associated with an estimated 0.8 to 2.8 m rise in the global sea level.

In Europe, between 25,000 to 13,000 BP, the ice sheets were at their maximum size for only a short period. During the glacial maximum in Scandinavia, only the western parts of Jutland were ice-free, and a large part of what is today the North Sea was dry land connecting Jutland with Britain (Figure 1.3). The Baltic Sea is a result of meltwater from the *Weichsel glaciation* combining with saltwater from the North Sea when the straits between Sweden and Denmark opened. Initially, when ice began melting about 10,300 B.C., seawater filled the isostatically depressed area, a temporary marine incursion that geologists dub the Yoldia Sea. Then, as post-glacial isostatic rebound lifted the region about 9500 B.C., the deepest basin of the Baltic became a freshwater lake, in palaeological contexts referred to as Ancylus Lake. The lake was filled by glacial runoff, but as the worldwide sea level continued rising, saltwater again breached the sill about 8000 BP thus forming a marine Littorina Sea that was followed by another freshwater phase before the present brackish marine system was established. At its present state of development, the marine life of the Baltic Sea is less than about 4000 years old. Overlying ice had exerted pressure on the Earth's surface. As a result of melting ice, the land has continued to rise yearly in Scandinavia, mostly in northern Sweden and Finland where the land is rising at a rate of as much as 8–9 mm per year, or 1 meter in 100 years.

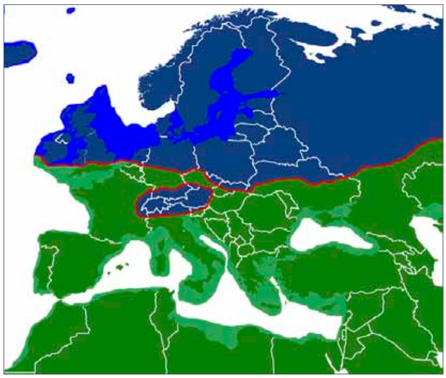

Figure 1.3. Europe during the last ice age (20.000 years ago)

How Can the Progress of Humankind be Measured?

Many sociologists and anthropologists have created social theories dealing with social and cultural evolution. Scientists are searching for the primary factor driving the development of human civilization. Over 10,000 years ago, there were fewer than 10 million people on the planet. At the beginning of 21st century, the number reached more than 7 billion, 99 per cent of whom are better fed, better sheltered, better entertained and better protected against disease than their Stone Age ancestors. The availability of almost everything a person could want or need has been going erratically upwards: calories, vitamins, clean water, machines means of travelling faster than we can run, the ability to communicate over longer distances we can shout, etc.

In accordance with Matt Ridley (Matt Ridley, 2010), the history of progress is based on a simple but unpopular idea: *specialization and markets* are the prime movers of progress. The ability to specialize and trade makes humans unique. Instead of catching our own food, making our own shelter, etc. (as other animals do), humans have created a system where everyone can specialize and trade with others engaged in other things. This means that those best at making houses make houses, those best at making food make food, etc. As regards trading, we can each benefit from that others do and vice versa. *Erectus hominids*, in other words, had almost everything we might call human: two legs, two hands, big brains, opposable thumbs, fire, cooking, tools, technology, cooperation, long childhoods and kindly demeanour. And yet, there was no sign of cultural take-off, little progress in technology, little expansion of range or niche. Then there appeared upon the earth a new kind of hominid which refused to play by the rules. Without any changes in its body, and without any succession of species, it just kept changing its habits. For the first time, its technology changed faster than its anatomy. This is not true for other creatures, not even brainy ones like chimpanzees, bottlenose dolphins, parrots and octopi. They may occasionally use tools, they may occasionally shift their ecological niche, but they do not "*raise their standard of living*", or experience "*economic growth*". After all, late Neanderthals had on average bigger brains than we have, yet did not experience this headlong cultural change. Language, cognitive reasoning, fire, cooking, tool making, self-awareness, deception, imitation, art, religion, throwing weapons, upright stance, and grandparental care – the list of features suggested as unique to human beings is long indeed. Big brains and language may be necessary for human beings to cope with a life of technological modernity. However, big brains and language are not themselves the explanation of prosperity, progress or poverty. They do not themselves deliver a changing standard of living. Neanderthals had all of these: huge brains, probably complex languages, lots of technology. Nevertheless, they never burst out of their niche. Probably, when looking inside our heads, we would be looking in the wrong place to explain this extraordinary capacity for change in the species. It was not something that happened within a brain. It was something that happened *between brains*. It was a collective phenomenon. Chimpanzees may teach each other how to spear bush babies with sharpened sticks, and killer whales may teach each other how to snatch sea lions off beaches, but only human beings have the cumulative culture that goes into the design of a loaf of bread or coat. There was a point in human pre-history when big-brained, cultural, learning people for the first time began to exchange things with each other and that once they started doing so, culture suddenly became cumulative, and the great headlong experiment of human economic "progress" began. By exchanging, human beings discovered "the division of labour" and the specialisation of efforts and talents for mutual gain. But some ape-men had begun exchanging food or tools with others in such a way that both partners to the exchange were better off, and both were becoming more specialised. The more human beings diversified as consumers and specialized as producers, and the more they then exchanged, the better off they have been, are and will be.

By the age of 15, chimpanzees have produced about 40% and consumed about 40% of the calories they will need during their entire lives. By the same age, human hunter-gatherers have consumed about 20% of their lifetime calories, but produced just 4%. More than any other animal, human beings borrow against their future capabilities by depending on others in their early years. A big reason for this is that hunter-gatherers have always specialized in foods that need *extraction* and *processing* – roots that need to be dug and cooked, clams that need to be opened, nuts that need to be cracked, carcasses that need to be butchered – whereas chimpanzees eat things that simply need to be found and gathered, like fruit or termites. Learning to do this extraction and processing takes time, practice and a big brain, but once a human being has learnt, he or she can produce a huge surplus of calories to share with the other humans.

By the middle of the 21^{st} century the human race will have expanded in ten thousand years from less than ten million to nearly ten billion people. Some of the billions alive today still live in misery and dearth even worse than the worst experienced in the Stone Age. Some are worse off than they were just a few months or years before. But the vast majority of people are much better fed, much better sheltered, much better entertained, much better protected against disease and much more likely to live to old age than their ancestors have ever been. The cumulative accretion of knowledge by specialists that allows us each to consume more and more different things by each producing fewer and fewer is, probably, the central story of humanity.

Some scientists declare *technological progress* to be the primary factor driving the development of human civilization. The basis of the three-age system emphasizing archaeological periods is work done by Danish archaeologist Christian J. Thomsen and published in 1819. He related the classification of artifacts to technology, i.e. according to the materials in which they were made (stone, bronze, and iron), thereby defining the Stone Age, the Bronze Age and the Iron Age. The scheme was not universally accepted until the International Congress of Archaeology in Budapest in 1876. The Copper Age that lasted between the Stone and Bronze Ages was added to the three-age system.

One way is to take a look at the extraordinary evolution of the man-made materials and how these materials have been developed and used over time, including their range, variety, quality and performance. Even at a cursory glance – from the bronze and iron ages to the silicon age that dominated the end of the 20th century, the emerging picture is quite clear: progress has been phenomenal. Materials can enable industrial and commercial success in both the existing and not-yet existing products and processes: they may introduce new functionalities and improved properties adding value to the existing products and process thus representing an invisible revolution. At the same time, the engineered production of materials by design might allow the development of products and processes under a really sustainable systemic approach.

Material research can be presented starting from the materials themselves (e.g. metals, biomaterials, polymers), the industrial sector (e.g. metallurgy, chemistry) or their applications (e.g. energy, health, transport) as well through other approaches. Materials profit from a wide range of scientific disciplines such as chemistry, physics, biology and engineering as well as from all available technologies and multidisciplinary approaches like nanotechnology and biotechnology.

The Australian archaeologist Vere Gordon Childe was one who was convinced that we should look upon pre-history primarily as the history of technology. His work *Prehistory of European Society; Piecing Together the Past: The Interpretation of Archaeological Data*) appeared in 1925–2008 (Vere Gordon Childe, 2008). A rise in and fall of technologies for hunting and weapon-making, herding and domesticating animals, crop-growing and agriculture, pottery and metal working were the factors in contributing to the beginning of the pre-history of technologies. V.G. Childe held that technologies enabled one tribe to overcome another in battle and technologies that enabled people to produce such a surplus of food that the great states (Egypt, Mesopotamia) could be set up. When studying the history of mankind from the point of view of technological development, it is possible to distinguish seven to some extent overlapping ages:

1. The era of nomadic hunter-gatherers using tools and weapons fashioned from easily available wood, bone or stone and able to induce and control fire;
2. The Metal Ages when an increase in the specialization of tasks encouraged changes in social structures;
3. The First Machine Age, that of the first clocks and printing press, when knowledge began to be standardized and widely disseminated;
4. The beginning of quantity production when, with the early application of steam power, the factory system began irreversibly to displace craft-based manufacture;

5. The full flowering of the Steam Age affecting all areas of economic and social life;
6. The rapid spread of the internal combustion engine, which, within 50 years, virtually ousted steam as the primary source of power;
7. The present age of electrical, electronic and information technologies promising to change human life more swiftly and radically than any of its predecessors.

Other scientists and anthropologists declare a concept of three major stages of social evolution (savagery, barbarism and civilization) that can be divided by technological milestones, like fire, a bow and pottery in the savage era, the domestication of animals, agriculture and metalworking in the barbarian era and the alphabet and writing in the civilization era.

Others (Leslie White) decided that the measure to judge the evolution of culture was energy, and "the primary function of culture" is to "harness and control energy" (Leslie White, 1959). He differentiates between five stages of human development:

- at the first, people use energy of their own muscles;
- at the second, people use the energy of domesticated animals;
- at the third, people use the energy of plants (agricultural revolution);
- at the fourth, people learn to use the energy of natural resources such as coal, oil, gas, etc.;
- at the fifth, people harness nuclear energy.

In his own words, "culture evolves as the amount of energy harnessed per capita per year is increased, or as the efficiency of the instrumental means of putting energy to work is increased". Societies that capture more energy and use it more efficiently have an advantage over other societies. White introduced a formula

$$P = ET,$$

where P represents the degree of cultural development in terms of the produced product, E is a measure of energy consumed per capita per year, T is a measure of efficiency in utilising energy harnessed.

Gerhard Lenski focuses on information (Gerhard Lenski, 2005). The more information and knowledge (especially allowing the shaping of a natural environment) a given society has, the more advanced it is. He identifies four stages of human development based on advances in the history of communication:

- at the first stage, information is passed by genes;
- at the second, humans gain sentence and therefore can learn and pass information through with reference to the gained experience;
- at the third, humans start using signs and develop logic;
- at the fourth, humans can create symbols, develop language and writing.

Advancements in technology for communication translate into advancements in economic and political systems, the distribution of wealth, social inequality and other spheres of social life. He also differentiates societies based on their level of technology, communication and economy:

- hunters and gatherers,
- simple agricultural and herding,
- advanced agricultural,
- industrial,
- special such as fishing, maritime societies, etc.

Finally, from the late 1970s, sociologists and anthropologists have suggested the theories of post-industrial societies arguing that the current era of industrial society is coming to the end, and *services and information* are becoming more important than industry and goods.

The Earliest Tools and Technologies

The earliest tools allowed humans to create means that helped with leaping into the top of the food chain and competing for surviving along with all other live beings.

Oldowan chopping tools are the earliest record tools. The oldest examples of record are dated to be from around 2.6 million years ago. These primitive stone tools were largely single faced chipped stone implements with a sharp edge produced by banging a heavier rock into a smaller one until the sharp edge was formed when a part of the smaller stone flaked off.

A hand-axe is really old. This tool separated men from apes. The hand axe is basically a rock with sides that have been planed to a point. These were not as refined as the blades produced in later times through more advanced processes, but merely crudely sharpened stones. It seems that most hand axes have a sharp border all around. Experts hypothesize that they could have been used as cutting and chopping tools, digging implements, flake cores, traps or even for a purely ritual significance.

Archaeological record dates a *javelin* back to around 400,000 years. These tools could be the oldest ones known to mankind as some primate species have the capacity to make them. *Blades* started out as stone cylinders or knife-shaped rocks that were pounded with other rocks to make them sharper through more refined stone-tool making processes such as knapping. Blades are considered a more modern "middle" stone age tool and the earliest ones of record date back to somewhere between two and five hundred thousand years ago. Blades can be defined as being stone flakes that are at least twice as long as they are wide, with fairly parallel sides and at least two ridges on the outer side.

A sling appeared around 64,000 years ago, the same time as a bow and arrow. The most notable use of the sling in recorded history is the biblical recording of the boy David defeating the mighty warrior Goliath. As mentioned above though, the sling is an inferior tool in comparison to the bow. The sling shot apparently capable of penetrating body armour.

The bow and *arrow* appeared around 64,000 years ago and allowed man to hunt animals and conquer nations at a distance with much greater range and accuracy than other earlier projectile tools like slings and spears. Far superior to the sling in most respects, this tool had been used for millennia and remained the predominant long distance weapon of choice until it was surpassed by more modern guns developed in the last three hundred years or so.

Even a small thing like a hand sewing *needle* has had a role in the history of humans. They were originally made out of bone or ivory and were used for sewing. Ancient tailors used flax thread for sewing more form fitting clothes out of fur, leather or even leaves, allowing for increased dexterity and warmth on the move. The oldest known bone ones were found in what is now south-western France and has been estimated to be over 25,000 years old. Needles made from copper, silver and bronze were used in ancient Egypt. The oldest iron needle known was found in what is now Germany and dates back to the 3rd century B.C. Native Americans used porcupine quills and the pointed end of agave leaves for sewing needles. The fibbers of the agave leaf were also used for thread. Metal needle making was perfected by Muslims in Spain in the 11th century. Europe learned the art of needle making from Arab needle makers, and it came to England in the 17th century. Before this time, metal needles were made in Europe by the local blacksmith and resulted in very crude needles. Metal needles were handcrafted before the industrial age. The process began with cutting wire long enough to make two needles. Then, points were ground on either end of the wire that was flattened in the middle with the eyes punched out. Later, the needles were separated.

Thread is a tightly twisted strand of two or more ply of yarn used for hand and machine sewing. The first 'thread' to be used in sewing was made from animal sinew and plant fibbers. The difference between thread and yarn is that thread is used for sewing together garments and other products, whereas yarn is a collection of fibbers that is woven or knitted into textiles. All threads are made from yarn which is not made of threads. The three types of thread include *animal, plant* and *synthetic* and are based on the materials they are made from. Silk is an example of thread made from animal products. The silk caterpillar weaves a cocoon made from silk it produces. Cotton is an example of plant fibbers used for making thread. The fibbers of cotton are spun into fine yarn; two or more strands of yarn are twisted together to make the thread. Nylon and polyester thread are the examples made from synthetic materials.

An awl, as a needle, is really old. Awls were used for poking holes in animal hide, after which the awl was removed, and sinew or other binding materials could then be forced through the holes to make some sweet new buck-skin pants. Awls were made by taking bones or sticks and pounding or shaving them into finer points.

The exact time of *fire application* (on purpose) in which humans learned to control fire is still open to debate. Most experts in the archaeological community are willing to agree that man has controlled fire for at least 400,000 years, though some evidence suggests that it may have been as much as 1.5 million years ago that man began to use fire (cave of Alistrati, Greece).

Archaeology provides our earliest information on those works of humanity that could be considered as the examples of the earliest engineers (Table 1.1). During the long evolution of early man, technology, as a whole, is best characterized as conservative. Innovations were singular and often separated by millennia. The development of early stone tools is an example of this innovation. Surely this is, as anthropologists believe, the result of biological factors such as the evolution of the human brain. The development of technological thought, and thus the genesis of "engineered" works, is directly tied to the evolution of the human brain. This development can be inferred from two related sources:

- human skeleton remains of the Pleistocene;
- human cultural remains.

One of the first human innovations were related with stone tool manufacture. The first appearance of identifiable tools occurred at about 2.25 million years ago. The oldest stone tools have been found in ancient (2–2.25 mill. years) sites in Africa along the margins of Pleistocene lakes or streams. The remains of the associated skeleton have been termed *Homo habilis* by L. B. Leakey and were first identified as the Kenyan site of Olduvai Gorge along the Central African Rift Valley chain. Camp sites have been identified as associated with the remains of hominid tools. Only by 500.000 years (before present), the camp sites of extensive occupation with the demonstrated use of controlled fire appeared. Table 1.1. The origins of humanity and the earliest evidence

In anthropology, the origins of humanity are sought in the material evidence of creative intellect. The earliest evidence is in the form of stone tools and the control of fire. At a place in China, known as Chou Koutien, is the type site for *Homo erectus sinanthropus* or "Peking Man". The tools typical of this stage of human development are called core tools. The other members of *Homo erectus*, outside Asia, had a separate tradition of lithic technology in which the core tools predominated. Combined with bone points and clubs, these stone tools allowed *Homo erectus* to occupy large areas of Africa, Europe and Asia. These groups were the hunters of Pleistocene megafauna such as mammoth, giant sloth, camel, horse and bison.

Table 1.1. The origins of humanity and the earliest evidence

Sort of human activity	Years before present
Farming and food production	10–12,000
Mechanical devices like bows and spear throwers	30,000
Representational art	30,000
Hafted tools appear	50–100,000
Humans occupy the cold, temperate zone	300–500,000
Definite signs of controlled fire	300–500,000
Last definite signs of multiple species of coexistent hominids	1 million
Oldest definite "camp sites"	2 million
Oldest known definite stone artifacts (pebble tools)	2–2.25 million
Oldest fossil skull with a brain case much larger than that of any ape	2–2.5 million
Oldest known definite indications of a fully bipedal mode of locomotion	3–4 million
Oldest known traces of hominoids with teeth and jaws like those of later hominoids (at least three species, including *Ramapithecus, Sivapithecus* and *Giganthopithecus*)	9–14 million

Excavations in Nice, France, have yielded the first evidence of man-made structures. These shelters, over 13 meters in length, were made of joined along the ridgeline and upheld by four vertical supports. They were probably covered with leaves, although other materials such as skins could have been utilized. The perimeter of stones braced the lean-to walls on windward sides. Throughout the Pleistocene epoch spanning approximately three million years before the modern era (termed the Holocene), the climate was the engine of changes.

The inhabitants of a simple camp in Nice were mid-Pleistocene (400–500,000 B.P.) hunters living in a Europe enjoying a respite from the cold of the glacial ages. During the Pleistocene, at least four major glaciations have been verified, with numerous warmer or colder sub-stages of partial retreats or advances. These hunters and their simple stone axe technology ranged over three continents – Africa, Europe and Asia – during the middle Pleistocene until roughly the end of the third glaciation, termed the *Riss*.

About 100,000 years B.P., *Homo erectus* disappeared in the fossil record and was replaced by more advanced forms such as *Homo sapiens neanderthalensis*. This early man used much different stone technology called the Mousterian or prepared-core tools.

What is important to us in this discussion of early humans, is to form an appreciation of the long prehistory of the slow technological development of humanity. Human technology was simple and direct. We can speculate that this implies a direct relationship between man's needs and his intellectual capacity to find ways to meet these needs. Even a small thing like a hand *sewing needle* had a tremendous role in the early history of humans. A variety of archaeological findings illustrate sewing was present for thousands of years. The first sewing needles were made from bone and used for sewing animal hides together, which allowed prehistoric people move from natural caves into open terrain and follow animal herds. The sewed garment and tents protected humans from snow and fierce winds in winter. With such artificially made protection, humans could survive in the pre-Arctic areas left by the retreat of the glaciers. The oldest known bones found in Europe have been estimated to be over 25,000 years old. The earliest bone needle dates back to 61,000 B.C. and has been discovered in South Africa cave. The ivory needles discovered in Russia have been dated back to 30,000 years ago.

The Earliest Builders

We do know that after 10,000 years B.C., we see engineering thought evident in expanded technology that includes the construction of structures and boats, construction of water reservoirs and irrigation canals, the domestication and use of animals for motive power and the development of metallurgy. This "new stone age" is termed the *Neolithic* period and marked the end of the glacial age – a hard and difficult lime – with vast areas being uninhabitable for animals and humans. The upper Palaeolithic man did develop a rich way of life centred on hunting and gathering, but it was still technology employing stone and bone tools (Figure 1.4; Figure 1.5).

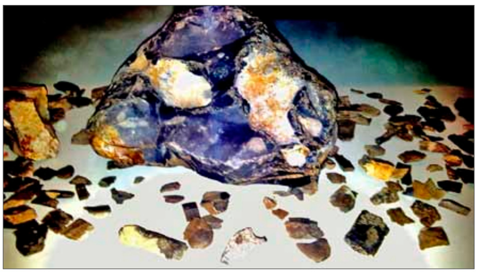

Figure 1.4. Tool making from a piece of flint (Kernave excavation, Lithuania)

Figure 1.5. Stone axes from Kernave excavation, Lithuania

With a retreat from glaciers, the environment warmed, spelling the doom of Pleistocene animals such as the mammoth, cave bear, saber-tooth tiger, American horse and numerous other extinct species. The life ways built upon the pur-

Figure 1.6. The ruins of ancient Jericho

suit of these animals ended as well. With the Neolithic, for the first time, humans had to build for permanency. The development of primitive agriculture is one of the major definitive characteristics of the Neolithic period. In the early spread of Neolithic agriculture (13,000–8,000 B.C.), farmers exploited loess soils left by glacial periods along river valleys of Europe, such as the Danube. The early inhabitants of the post-glacial forests farmed these alluvial fans and loess terraces of streams from the Mediterranean to the Baltic. Having thus implied a direct connection between life, agriculture and building, one of the earliest known examples of large-scale building comes, as an exception, from pre-agricultural levels. Jericho is located near the Dead Sea in Trans-Jordan (Figure 1.6). It is in the midst of a desert where archaeologists have found the layers of a permanent settlement reaching 20 meters the present ground level and representing over 7,000 years of occupation. Jericho is an oasis in a desert that lies 256 meters below the sea level.

Spring water from Elisha's Fountain provided a copious flow to the nearby Jordan River. 10,000 B.C., people raised a settlement boasting a massive stone wall. The base of the structure was a ditch 8.2 meters wide and 2.1 meters deep, carved out of solid rock. Today, after ten millennia, the wall is still almost four meters high.

The village within the wall contained huts made of sun-dried bricks. The people of this four-hectare village grew an early variety of wheat and raised goats. Jericho fits nicely into cultural evolutionary sequences that trace technology from hunter-gatherers of the late Pleistocene to the farming villages of the Neolithic.

Ceramics have never been a necessary condition for cooking or food storage-ground stone, wooden or leather containers and baskets, or leather work allowed many early cultures to manage nicely in this area. Pottery manufacturing may have been an antecedent for metallurgy – the next technological step taken by man. The early residents of Jericho excavated trench footer for the wall without metal tools. This following technological step was to wait another three millennia before man smelted copper ores.

The West of Jericho, in what today is Turkey, developed *Çatal Hüyük* which is perhaps one of the oldest cities in history (Figure 1.7). The process began around 10,300 B.P. and flourished to 7,600 B.P. on the Anatolian Plateau as a town of

Figure 1.7. Excavations at Çatal Hüyük

some 6,000 people trading obsidian and raising wheat, barley and domesticated animals. In a Europe of the early-mid Neolithic period, we see nothing to compare with Jericho or Çatal Hüyük.

After the glacial retreat, Europe experienced cyclical periods of wetter or drier climate over the hundreds of years. The retreat of ice caps was coupled with advance in the forests across central Europe where Palaeolithic peoples had followed the herds of reindeer across a taiga or steppe-like landscape. Those of the Neolithic had luxuriant, climax forests of mixed hardwoods to challenge settlers using stone tool inventory. Because of the forests, the Neolithic peoples of what is now Western France, Switzerland, Northern Italy, Germany and Poland, utilized wood as their principal building material. By 7000 B.C., they were building large dwellings or, "longhouses," that reached up to 30 meters in length with the widths of 5 to 8 meters. The interior surfaces of the walls were smoothed and whitewashed. The houses were divided into three sections – the central-living area, a southern-storage area and a northern stable for livestock.

Gradually, from 4,600–3,000 B.C., European cultures began the monumental use of large stones or "megaliths" some of which weighted up to 30 tons. The forms of construction included grave chambers, and upright stones (*menhirs*) often aligned in circles, avenues, barrows and tumuli. Information gained through the use of modern physical dating techniques suggests that the Europeans of the fourth to second millennia had developed a level of civilization, including metallurgy, quite independently of the Mediterranean cultures. Much debate exists over the function of Stonehenge (Figure 1.8) or other megalithic "*henges*". In recent years, it has been suggested to serve as an astronomical function, as a ritual or cult sanctuary or as an ancient administrative or political place such as that of Iceland's comparatively younger Althing.

Homo erectus and Neanderthal were hunters such as the world will never see again. We can only imagine the world they knew. For that springtime, 350,000 years ago, the camp in Nice required early humans or human-like beings to use cooperation in the construction of some of the world's first "planned" structures. The Neanderthals survived environmental conditions that would pull down our present-day specialized technology unless we displayed similar adaptability.

Figure 1.8. The construction of Stonehenge appears to have been created over a period of more than thousand years

A puzzle remains as to the origin of plant and animal domestication. We have the general picture with the early hillfolk of Iran and Iraq who tended the descendants of wild grasses and game. This allowed these groups to settle into small villages. Rather than following their game or being compelled to move from harvests of nut mast to seed crop, men could carry their grains in pouches. They acted in a way following the rivers across the Levant, the Balkans and into Europe. In each of these places, man adapted his technology to the local environment and vice versa. Engineers appeared as designers and builders of irrigation systems, fortifications, public buildings and great tombs.

Agriculture is not deterministic of cities or urbanization, but civilization is synonymous with cities. Populations grew rapidly in the fourth millennium, such that the children of the early farmers had to search for new lands to cultivate. Their Neolithic agricultural techniques accompanied them. Primitive societies are conservative owing, in great part, to their lack of economic surplus. Experiments, particularly the failed ones, in the area of subsistence technology could threaten these societies with death. The old adage, "if it works, don't change it," may have greater antiquity than we suppose.

Still, these primitive technologies could not support large populations. Humans lived in small bands, probably organized through extended family ties. The areas needed to support such groups were large and based on a prehistoric carrying capacity of a predator-prey balance. As good as these hunters might be, they were still inefficient as killers and wasteful of their resources to a great degree. Only when spread of forests cut off the migration routes of the great herds, man was forced to consider alternate strategies to survive. When man "considered" the edible equivalent of the field lily, only then did he begin the trek along the path to agriculture, economic surplus, population and civilization. This trek led early hill-farmers of the Zagros Mountains to chance the plains of the great Tigris-Euphrates valleys with their more fertile soils. This is the place innovation took a wing, as villages became towns and towns became cities. In Mesopotamia, as this area became known, development began in the middle of the fourth millennium with the Sumerians culture by 5500 B.C. This period saw the development of writing and metallurgy, together with building on a scale unseen heretofore. Evidence of this genesis is seen in the centralization of the population of Uruk. From 146 villages in 5000 B.C. to 76 villages 300 years later, and the final 24 in 4400 B.C. the villages were not depopulated by diseases or wars. The population moved to the cities such as Uruk, Ur and other early city-states or *poli*, as the later Greeks would call them. It is here where people started the conquest of materials by the engineers of these young civilizations.

CHAPTER 2 Dawn of the Conquest of Materials

This chapter will help in

- describing the dawn of the conquest of materials
  in Sumeria and Egypt,
  Assyria and Babylon;
- showing the earliest design of an ancient boat;
- explaining the appearance of a sail.

Humanity in the Palaeolithic Age confined their use of materials to the manufacture of tools. In the Neolithic Period, they continued the use of stone for such tools as hoes, axes, weapons and hunting gear. The *Natufian culture* existed from 13,000 to 9,800 years ago in the Levant, a region in the Eastern Mediterranean. The Natufian communities are possibly the ancestors of the builders of the first Neolithic settlements of the region that may have been the earliest in the world. There is some evidence for the deliberate cultivation of cereals, specifically rye. The Natufians used wild cereal grasses and *stone sickle blades*. The hunted animals included gazelles.

One reason for this lack of stone construction was the type of polity inherent to this level of human society. People lived in small villages of interrelated communal groups. This was reflected in their structures and common long-houses with living space for more than one family. In the forested regions, people utilized wood for construction, in Egypt – stone, in Mesopotamia – bricks and in China – wood. In five principal areas across the world, plant and animal domestication proceeded concurrently. These "nuclear" areas were located in

1. The Tigris-Euphrates Valley.
2. Egypt.
3. China and Southeast Asia.
4. Africa.
5. Mexico and Peru.

Sumeria and Egypt

It is no coincidence that these areas were the first to develop state-level societies and engineered works. An exception to early wood construction was in Mesopotamia where the people of the Tigris-Euphrates Valley-Plain, without such resources, used mud and reeds. It was in Sumeria that we can see the translation of one structural material into another form of the same material – from mud to brick. The early Sumerian societies of the Tigris-Euphrates delta built houses of reeds that grew to a height of 10 meters. Lashed together into tall thick bundles called *fascines,* they were placed as uprights equidistant along a straight line.

Thinner fascines laid horizontally were lashed to these vertical members, whereby they formed the framework for a wall. To complete the wall, the spaces between the framework were filled with reed matting. The final element was the roof – tall vertical *fascines* were bent inward and lashed together where they met. Reeds were tied to these rafters with matting thus completing the roof. Depending upon the extent of bending, the roof could assume either a semi-circular vault or a pointed arch.

It is still unclear how reed construction leads to the origin of brick. The Sumerian reed house gave adequate shelter until wind or rain came. The next step was to seal draughty walls and a leaky roof with mud plaster. This left only one serious defect in the design – a fire trap. Where the house was covered in plastered mud, the burnt shell often stood after the fire had consumed the reed frame. In all likelihood, the concept of masonry construction resulted from such chance experiences.

Early Sumerian society. The first significant change in human life was the rise of cities that occurred sometime before 3000 B.C. The principal difference between a village and a city was that most of the inhabitants in the village were directly engaged in the production of food, whereas very few city dwellers were so engaged. Large-scale cities and buildings did not arise overnight. For example, at Hassuna (Northern Mesopotamia), we can see the transition of building techniques.

It should be noted, however, that most early knowledge was purely empirical, gained from experience and handed down from generation to generation. Computational techniques and arithmetic gradually came to be used for commercial purposes and surveying in Sumeria in the third millennium. All of these changes had their influence on the engineering of the period. The growth of cities stimulated engineering in other ways. Before 3000 B.C., most buildings were reed or mud-clay homes. After that time, structural engineering was no longer merely functional, but architectural as well. The most important sources of information about the engineering of Mesopotamia are the ruins of the cities themselves. Archaeologists first found a ziggurat, or temple tower in Or (south of modern Baghdad, Iraq). High walls cannot be made with sun-dried brick. The masonry of this type cannot withstand much heavy pressure or compression before it fails. To relieve this vertical pressure or a static load, they started the second wall behind and above the first, each upon an interior mound of the earth. They often built a third or more walls. The outline of taller ziggurats must have been quite similar to our modern skyscrapers with set-back stories.

The ziggurat was a many-angled, solid pyramid, approximately 70 meters by 40 meters and 20 meters tall. It was made of sun-dried brick faced with burnt brick amid some stone. There were a number of stairways by which people could ascend to open-air shrines at the summit. The material commonly used for binding the masonry was asphalt or bitumen. The builders further relieved pressure on the brick facade and improved drainage by using reed mats mixed with bitumen.

Cities also produced problems that were gradually solved by the evolution of hydraulic engineering. Open drains were constructed to remove surface water. The need for increased efficiency in food production resulted in the construction of dams, reservoirs and canals for flood control and irrigation. City dwellers needed water, and tunnels were sometimes constructed to bring water from nearby springs to pools within the city or treadmill for raising water from an irrigation ditch. Water-raising wheels were 6 meters in diameter or more. Wherever feasible, there were spillways to carry excessively high water into great depressions in the desert. These depressions covered as much as 1700 square kilometres and could hold water to a depth of 7 meters. Irrigation requires a large amount of water. In distribution, about 50% of water gets lost and 25% flows away unused thus leaving 25% to be used on the crop.

The Tigris and Euphrates served as the source of water for Sumeria and all Mesopotamia respectively. The two rivers are quite different in their volume and flow. Rising earlier than the Euphrates, the Tigris carries about 2.5 times as much water, and it is deeper than its partner. The Tigris was used for watering lands on its left or west bank. The Euphrates bed is raised above the surrounding plain and therefore allowed gravity flow of waters into canals. Table 2.1 gives a comparison of Mesopotamian irrigation conditions with those of Egypt – its neighbour in antiquity.

Urban development, with its political, economic, religious and social structures, stimulated and interacted with the development of engineering. In turn, engineering influenced politics, economics, religion and social life making available the means applying which these activities could continue to evolve. Engineering came into being to solve the problems of these new societies and once established, interaction by which engineering, in turn, influenced the evolution of society arose.

Table 2.1. A comparison of conditions for irrigation in Ancient Egypt and Mesopotamia

	Egypt	Mesopotamia
Climate average summer temperature average winter temperature	Semi-tropical 43 °C 12 °C	Continental 48 °C 6 °C
Flood season	August to early October	April to early June
Season after floods	Winter	Hot summer
Relation between harvest and floods	In time for winter and summer crops	Too late for winter crops
Rise and fall of waters	Slow and clear rise and fall	Sudden rise and fall
Profile of the river valley	No stagnant water	Pools and swamps
Type and quantity of sediment	Salt-free sediment. Little silting-up of canals.	5x as much salty sediment, canals silting-up quickly
Effects of irrigation	As a result of irrigation and a type of soil tendency to extract salts present in soil. Very slow silting-up of canals.	Tendency of salts and alkaline compounds to accumulate in soil. Danger of silting-up many canals.

Mathematics and writing. Sumerian engineers understood right triangles and computed areas of land, the volumes of masonry and the cubic contents of a canal. *Cuneiform script* is one of the earliest known systems of writing, distinguished by its wedge-shaped marks on clay tablets. Emerging in Sumer in the late 4th millennium BC, cuneiform writing began as a system of pictographs.The original Sumerian script was adapted for the writing of the Akkadian, Eblaite, Elamite, Hittite, Luwian, Hattic, Hurrian, and Urartian languages. Writing tablets, like the bricks of their structures, made use of one of southern few abundant resources-clay in Mesopotamia. It was better to impress clay than to incise it. The Sumerians began to impress their signs using styli that were triangular in cross-section and giving a rise in their wedge-shaped (cuneiform) writing (Figure 2.1). Sumerian writing originated about 3200-3100 B.C. and is termed *ideo-*

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Figure 2.1. Sumerian wedge-shaped writing

graphic in that it renders concepts such as "walk", "take", etc. into symbols. It is a pictorial writing system based on common signs for these concepts. It required a large number of signs making about 2000.

The question why the first number systems were based on 60 rather than on 10, which is similar to the system we are using nowadays, still needs an answer. Further, the Sumerians and their successors constructed a nicely interrelated measure system of standard units with conversion factors such as 2, 3, 5, 6 and 10. Additional aspects of Mesopotamian civilization include the development of wheeled transport, legal codes and pyrotechnology (manufacture of glass, metals). The chronology of the Sumerian area is shown in Table 2.2 along with that of the Nile Valley.

Table 2.2. The chronology of Sumeria and the development of the Nile Valley

Sumerian Area	
534-330 B.C.	Persian Empire
612-534 B.C.	2nd Babylonian Empire
1350-612 B.C.	Assyrian Empire
1990-1790 B.C.	1st Babylonian Empire
2900-1990 B.C.	Sumerian Civilization
Nile Valley	
1085-30 B.C.	Late Period
1567-1085 B.C.	New Kingdom
1786-1567 B.C.	2nd Intermediate Period
1991-1786 B.C.	Middle Kingdom
2135-1991 B.C.	1st Intermediate Period
2658-2135 B.C.	Old Kingdom
3100-2658 B.C.	Archaic Period

Egypt. The rise of Egyptian civilization can be traced to before 4000 B.C. The state was unified following long periods of internal struggle. Egypt passed slowly from a kingdom to an empire. One significant early stage in that process was marked by the pyramids built in the Old Kingdom period from approximately 2658 to 2135 B.C. Lasting over 500 years, the Old Kingdom was a period of prosperity, stability and confidence. With the fall of the Eighth Dynasty, Egypt passed into the First Intermediate (2135–1991 B.C.) period, the one of political disunity. The Middle Kingdom transferred the capital to Memphis and the civilization of Egypt flourished anew. The southern frontier reached the second cataract of the Nile. The Pharoahs began to challenge an expanding Hittite Empire based in Anatolia (modern Turkey), in the area of Palestine. The Middle Kingdom ended in a collapse of unity between Upper and Lower (southern and northern) Egypt.

Almost simultaneously, Egypt suffered its first real foreign invasion by the Hykos, or "sea peoples". The Theban princes re-established a unified Egyptian rule and had founded the Eighteenth Dynasty by 1567 B.C. This is the period of Imperial Egypt with the rulers such as Seti I, Ramses II, Akhenaton and Tutankhamon. During the Late Period, Egypt slowly slipped from pre-eminence, being conquered by the Assyrians, Persians and, finally, the Romans.

Egypt, at its height, was one of the most populated countries in history. A population of eight million inhabited an area of roughly 29,000 square kilometres. Protected on the east by the Nile, on the west by the desert and on the north by the sea, the Valley of the Nile stretched southward into Africa, providing arable land to support this high population density (275 persons per kilometre).

Irrigation. Irrigation was fundamental to the survival of Egypt. As in Mesopotamia, nature could not be relied upon to always provide a steady and adequate supply of water. The Nile provided water for basin irrigation, which the Egyptians managed as a science. For the benefit of their people, the Theban kings (2000–1788 B.C) undertook to change the great oval basin in the Faiyum desert. The area of the West of the Lower Nile became a fertile and populated land. They threw dams across the ravines leading into the basin to impound the rains of the wet season against drought. One of these dams in a 250 m wide gorge had a base 143 m thick – four times its height. It was built in layers, and the bottom of rough stones was embedded in clay. The next of irregular limestone blocks and the top of cut stone were laid in steps so that the water pouring over the brim was checked in its fall and would not erode the structure.

In the third millennium B.C., the Egyptians were using engineering methods conditioned by their environment. They were still applying these techniques 30 or more centuries later when their country had ceased to exist as a separate entity. Three basic factors determined the nature of Egyptian engineering. One was the great supply of human labour. The horse was never used for work and was unknown in Egypt until about 1700 B.C. The Egyptian peasant fully expected to be drafted for public tasks in that portion of the year when the climate and river kept him from his own work in the fields. Another determining factor in Egyptian engineering was the concentration of these vast armies of workmen under the absolute control of a single man. The third factor was the great quantity of building stone in the ledges of the upper Nile Valley. From the quarries of limestone, sandstone and granite, pieces weighing from 2,5 to 30 tons were used for the largest and oldest stone structures in the world. Also, the obelisks weighing several hundred tons each where at least one huge block reached 1000 tons were cut.

For this quarry work, the Egyptians are supposed to have used only the simplest mechanical principles and appliances. Stone was marked off to the desired measurements; grooves were then cut employing a mallet and metal chisel or drilled. The pieces were finally split from bedrock by bronze plugs that were slid between thin metal feathers or wedges and driven in until rock split.

Surveyors and mathematics. Egyptian "rope stretchers," or surveyors, are said to have been applying knowledge that the angle between two sides of a triangle is a right angle if the sum of their squares is equal to the square of the hypotenuse. They understood also how to calculate the contents of solids and to determine the slope or amount of cutback necessary in terms of the height of a pyramid and the length of its side. They applied a system based upon 10, like ours, in finding the area of a circle; they used 3.16 for the value of pi (the ratio between the circumference and diameter of a circle). Egyptian signs for their numerals (Figure 2.2) were as follows.

1 was represented by a vertical line, 10 – by an upside down U, 100 – by a spiral, 1,000 – by a lotus plant, 10,000 – by an upraised finger, 100,000 – by a tadpole and a million – by a kneeling genie. Fractions were represented by the combination of the sign for "month" representing 1, with the denominator written below it (Figure 2.3).

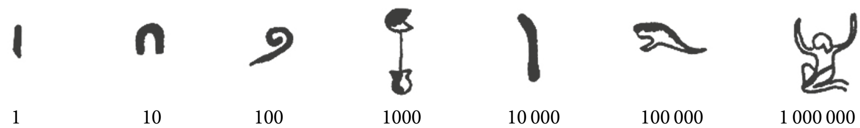

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Figure 2.2. Egyptian signs for their numerals

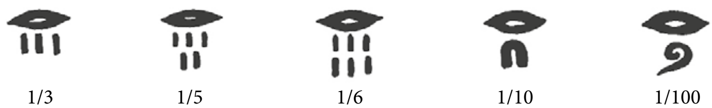

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Figure 2.3. Egyptian signs for fractions

Pyramids and temples. Around 3200 B.C., burials were in large mud-brick tombs recessed into the ground called *mastabas*, which, typically, were rectangular and not that large. The shape and size were unusual as well and made 70 meters square and 8 meters in height. The designers utilized stone rather than mud-brick.

These tombs underwent numerous changes. First, they were doubled in size, then again, and finally, a four or six-level step pyramid began to rise (almost 70 meters in height). A step pyramid is an architectural structure that uses flat platforms, or steps, receding from the ground up, to achieve a completed shape similar to a geometric pyramid. In profile, the structure resembles several *mastabas* stacked one on the top of the other (Figure 2.4).

During the Fourth Dynasty, in step with a new pharaoh Khufu (Cheops), planning a great pyramid that would be his tomb was initiated. Between roughly 2700–2550 B.C., approximately 11 million cubic meters of stone were quarried and transported for the purpose of forming pyramids, temples and causeways. The Khufu pyramid was constructed with more than two million stones weighing approximately

Figure 2.4. The Step Pyramid at Saqqara

two and one-half tons each. At completion, the pyramid was surfaced by white "casing stones" – slant-faced, but flat-topped, blocks of highly polished white limestone. With the ancient construction methods available, the pyramid was still completed at about 2600 B.C. The pyramid has an average length of the sides of 236 meters at the base with three or four minutes of error. The structure was carried upward at a uniform slope of 51° 51' thus reaching a height of 146 meters. It is a gigantic pile of 214 million rough looking, but carefully squared and placed, limestone blocks. To ensure precise and firm bedding, they used mortar made of gypsum with very little sand. Surveyors first marked the base to form a perfect square. The next step involved the construction of the multiple-step, the layered terraces on which formed the pyramid. The terraces had to be perfectly levelled to prevent the structure from being skewed. To achieve this, the builders erected a large system of water-filled trenches about its base. They used water for levelling the 5 hectare site to within a half-inch from one corner to the opposite diagonal corner. The ancient Egyptians never lifted stone blocks by means of pulleys or suspension by ropes. The stones were mobilized through jacking operations utilizing wedges, levers or rockers. It is now generally believed that the stones were transported mainly with the use of ramps. A two-ton block could not be dragged through loose sand. The ramp surface was hardened either with stone blocks or chips from quarries, or with a proportional mixture of wet sand, silt and clay. To date there is no way of knowing with certainty what methods were used by the ancient Egyptians for lifting the blocks of Giza.

Beyond the pyramids, the Egyptians continued their building in stone (Figure 2.5). The Middle Kingdom temple of Amon-Ra at Karnak was once, if no longer, the largest columnar structure in the world: its dimensions of 103 by 372 meters are sufficient to contain the combined ground arks of St. Peter's at Rome, the Cathedral of Milan and the Notre Dame at Paris. The great hall of the temple, known because of its rows of columns as the Hypostyle Hall, was 100 by 52 meters. The columns stood 21 meters in height in the centre of aisles and 13 meters along the sides (Figure 2.6). They were more than 3 meters in diameter supporting short architraves, or crossbeams, weighing from 60 to 70 tons each. These carried the flat roof at two levels, making room for clerestory windows, apparently the first in history. To place a column of several hundred tons upright is an engineering feat that requires the nicety of calculation and special equipment even in modern times.

Figure 2.5. The Pyramid of Khufu (Cheops)

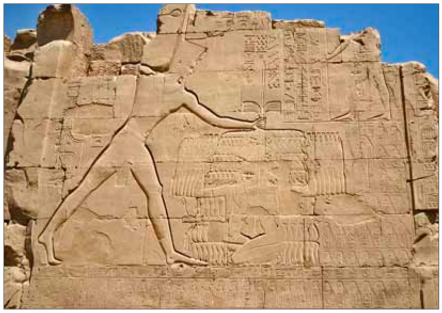

Figure 2.6. The temple of Amon-Ra at Karnak

Egyptian military engineering. Egyptian building in the Nile Valley below the First Cataract typically contained no militarism. Fortifications were present in the form of walls, but consisted of one rampart. Beyond the first Cataract in Nubia, the Egyptians built eight fortresses and beyond the second Cataract, they established six. The best preserved, until the construction of the Aswan High Dam, was Buhen, southern most of the first eight fortresses. Buhen was an ancient Egyptian settlement situated on the West bank of the Nile, 260 km south of Aswan, below the Second Cataract. It is well known for its fortress, probably constructed around the year 1860 BC. It covered 13,000 square metres. Its fortifications included a moat of three metres deep, drawbridges, bastions, buttresses, ramparts, battlements, loopholes and a catapult. The walls of the fort were about five metres thick and ten metres in height. At its peak, it probably had a population of around 3500 people. The masonry was of mud brick reinforced with timber on a base of rock and rubble fill. The fortress at Buhen today has been covered by Lake Nasser created by the Aswan Dam in 1964.

Mesopotamia: Assyria and Babylon

Assyria. The Assyrians came from a city-state on the upper Tigris River known as Ashur that was the capital of the Old Assyrian kingdom from the 3rd millennium B.C. King Sargon II (722-705 B.C.) moved the capital to Khorsabad. He built a new city with a palace flanked by huge winged bulls, temples, a ziggurat and a garden with exotic plants and trees. This city was one of the first to use city planning with crossing right angle streets and large squares. When he died, the capital was moved to Nineveh. This place was in a desert and the King wished to have greenery. The city was built in two years and completed with a palace and curtain wall. A park surrounded the palace and a broad greenbelt reserved for fruit orchards surrounded the city. To irrigate parks and orchards, the Khosr River 16 kilometres upstream of

Nineveh was straightened and canalized. Dikes raised the river level to permit irrigation around Nineveh. To expand this regime of irrigation and "greening" of the desert, by 700 B.C., the dams in the north of Nineveh were built, which joined three rivers 20 kilometres northeast of Nineveh to his canal system.

In 690 B.C., a 237 kilometre canal from the northern Tas mountains, which included the great 280-meter aqueduct of 20 meters wide, was started. Corbelled arches were used for crossing the river in the valley. The completed structure was 7 meters in height and contained about two million stone blocks dressed to 50 × 50 × 50 centimetres. The 80-kilometer long Bavian canal, including the aqueduct as a part of the structure, was completed in one year and three months.

The canal itself was built with a 40 cm bed of concrete poured on a 2,5 cm bed of mastic with a finely jointed stone pavement. The used concrete included one part of lime, two parts of sand and four parts of limestone aggregate. The "mastic" was bitumen mixed with fillers. Assyria had a special corps of men called *ummani* to level the ground for baggage carts and to build temporary bridges. The roads they constructed must have served for commerce and travel in times of peace and war. The walls were constructed of partially-baked brinks and faced either with stucco or enamelled brick. However, in some lower parts of the walls, great limestone monoliths weighing more than 20 tons are placed. The architectural and engineering purposes of the monoliths are not apparent. The courts of the palace were elaborately paved with stone set in asphalt. The perimeter walls totalled 80 kilometres. This curtain was 40 meters high, 10 meters wide and was equipped with 1500 towers. The wall was brick-faced and rubble-filled. Nineveh stood as a fortress city for less than a century, falling in 612 B.C.

Babylon was the capital of Babylonia, the alluvial plain between the Euphrates and Tigris. After the fall of the Assyrian empire (612 B.C.) and the fall of Nineveh in (612 B.C.), Babylon reached its greatest height and became the capital of the ancient Near East. The king adorned the city with several famous buildings. Even when the Babylonian Empire had been conquered by the Persian king Cyrus the Great (539 B.C.), Babylon remained a splendid city. Alexander the Great and the Seleucid kings respected the city, though after the mid-second century, the decline of the city started.

The new Babylon was laid out as a rectangle bisected into two parts by the Euphrates River. The famous walls around the perimeter estimated roughly 80 kilometres and were built to a height of 45 meters and a width of 10 meters. The battlement or outer parapet was crenellated in a square zigzag pattern (as in much later medieval fortresses).

Hard-surfaced streets were early achievements in engineering. Men walked on cobblestones in Assyria 4,000 years ago and there were brick and limestone pavements later in Babylon. Its famous Processional Street, leading to the bridge across the Euphrates, had a foundation of asphalt-covered brick and limestone flags with bevelled joints set in bitumen and mortar. The roadway was surfaced with large stones of one meter square. The sidewalks were of smaller slabs of red breccia, a sort of conglomerate. A large triangular salient formed by the outer wall on the east side of the river enclosed the residential area of the city. Over 20 meters high, the most fatuous gate of the city, the Ishtar Gate, was a vaulted passage finished in enamelled brick – blue on the towers and green and pink on the connecting walls. A relief showing bulls and dragons decorated the towers and walls leading to Processional Street (Figure 2.7). The ziggurat most remembered in the modern era was the so-called "Tower of Babel". Rebuilt several times, it reached nearly 100 meters in height and was covered in enamelled brick and reliefs. It had a base area of nearly 7000 square meters.

Figure 2.7. *The Ishtar Gate* in Babylon (Pergamon museum, Berlin)

Ancient Ships

The earliest boat design was that of a tree or log which probably evolved into a dugout. Reed and skin boats were suggested as early Mesopotamian vessels. The early boats in Egypt were made of reeds. Sail and rowing were utilized for propulsion using a square sail with a yard and boom. The sails of the early craft were fragile, and several support rings on the boom were necessary to relieve stress. The Bronze Age has been termed the "first great age of seafaring".

Starting from 2500 B.C., keelless seagoing vessels in Egypt traded with the Phoenician cities importing cedar wood and other merchandise, and were sent as the first Egyptian trade expedition to the Land of Punt.

The bipedal mast carried a vertical sail, was steered by six oars and had sixteen rowing oars. The bow was decorated with an eye. A rock served as anchor. Being rounded its action was solely based on its weight and the friction created when dragging over the bottom of the sea. When winds were strong, it was mostly useless, and the seamen were forced to seek shelter at the beach. Crew on an Egyptian ship was large, as their sailing capabilities were low and they had to be rowed often. Ancient ships could not sail into the wind so that tacking was impossible. If the wind was unfavourable, rowing was the only means to get anywhere.

Figure 2.8. The yard of fishing in Agadir (Marroco). The boats are about 15 m in length and all hand built from eucalyptus as they have been for centuries

New kingdom vessels (1500 B.C.) were about 22 metres long and 5 metres wide. They did not have a wooden keel. There were fifteen rowing oars on either side, two connected oars used as rudder, a single mast and a 15 metre wide horizontal sail. Bigger ships of seventy to eighty tons suited to long voyages and became quite common.

Cedar wood, imported from Lebanon, was worked with copper, later bronze, adzes, chisels, awls and saws. Ropes were made of the flax of halfa grass with the sails of linen. In the cross-section, the boat has no keel, but three broad, heavy bottom planks fastened edge-to-edge. The deck was supported by flat transverse beams that articulate with the sheer plank via holes drilled into the latter. On the centreline, directly under the deck beams, there was a long, continuous wood girder set on the edge and notched to all deck beams to cross over flush. This girder ends just short of the bow and stern. It is supported at regular intervals by column stanchions resting on transverse cross-framing. These cross-frames or floors were for providing lateral resistance to external stresses created by water pressure on the hull. A keel in Egyptian vessels appeared at about 1500 B.C. and was probably not Egyptian development.

The weakness in Egyptian truss construction was fasteners available to builders. These fasteners were rope fibre or heavy twine laced through the holes in adjacent wooden structural members. With stress and moisture, such fasteners became fatigued and loosened, resulting in short use-lives for the vessels.

CHAPTER 3 | The Greek Period

This chapter will help in

- explaining the mind and "wonders" of the Greek civilization;
- uncovering the reasons for a rise in mines;
- discussing Hellenistic engineering.

The Greek Mind and "Wonders"

In the history of engineering, science or technology, Greece presents an enigma along with the well-known "wonders" – the Colossus of Rhodes, the Pharos or Lighthouse of Alexandria and the Parthenon that is the only example which has remained. Technology, to the Greek mind, was mechanization.

The Mycenaean period of Greece is that which comes down to us in the epics of Homer. The cities of the Mycenaean period were built for both defence and habitation (Figure 3.1, Figure 3.2).

Figure 3.1. The Lion Gate of Mycenae (around 1250 B.C.). Two lionesses flank the central column

Figure 3.2. A storehouse for crops (excavations in Mycenae, 2013)

Late Bronze Age society used stone and wood in their constructions – arcaded facades, terraced staircases, windows and interiors the pillars of which opened space rather than enclosed it like in Karnak, Egypt. In Mycenae, the largest stones, including lintels and gate jambs weighed well over 20 tonnes; some may have been close to 100 tonnes.

The theatre of Epidaurus in north-eastern Peloponnese was designed in the 4th century B.C. (Figure 3.3). The theatre had seats up to 15,000 people and was marvelled for its exceptional acoustics that permitted almost perfect intelligibility of the unamplified spoken word from the proscenium or skênê to all 15,000 spectators, regardless of their seating.

The most famous temple of Asclepius was in Epidaurus, north-eastern Peloponnese. From about 300 B.C. onwards, the cult of Asclepius grew very popular and pilgrims flocked to his healing temples (Asclepieia) to be cured of their ills. The original Hippocratic Oath began with the invocation "I swear by Apollo the Physician and by Asclepius and by Hygieia and Panacea and by all the gods…".

While impressive, many of these structures call upon people to enter and enjoy them rather than being simply dim sanctuaries of god-kings (Figure 3.4). In the palaces of Crete, Minoan engineers moved water through terracotta pipes. Runoff from these pipes, after descending in parabolic-shaped gutters, was collected in cisterns. Settling basins collected sediment while sunlight helped with keeping the water pure in its passage to the storage cisterns. The road (in Knossos) itself represents an appreciation by those early engineers of a solid road base, along with drainage and paving. Their road foreshadowed Rome's stone paved roads build up 1,000 years later. The road was founded on rough stones embedded in subgrade. Smooth stone slabs then were set in clay cement, and meter-wide shoulders of

Figure 3.3. The theatre of Epidaurus (north-eastern Peloponnese)

Figure 3.4. The tholos tomb outside the citadel of Mycenae

a rammed gravel-pottery aggregated in clay made a total width of three and one-half meters. Unlike later Roman roads, it had only one drain and no camber to its surface to aid in runoff. No evidence of wheeled traffic has been found and these roadways have been termed "processional ways." Whatever their function, they represent one of the few examples of engineering we have from this period. By 1100 B.C., Mycenae, Knossos and many other centres had fallen. In the following centuries, the Greeks moved eastward to the Anatolian mainland coast while mainland centres remained ruined or abandoned.

After the "dark ages" (1000–700 B.C.) following the fall of Mycenae, the classical Greek civilization arose in the form of an individual *poli* or "city-states." The Greeks gave us a concept of the great works of art (sculpture, architecture). Persia invaded Greece. The Battle of Marathon in 479 B.C. dealt Persian troops a major defeat and Greece (from 467 B.C.) entered into its "golden age."

The philosophers of the Greek Illyrian coast were responsible for the beginning of scientific thought. These include Thales of Miletus (640–546 B.C.) who stressed empirical study, Pythagoras (582–500 B.C.) who developed abstract mathematics and Democritus (460–370 B.C.) who formulated atomic theory. Archimedes (287–212 B.C.) worked for clients (e.g. the defence of Syracuse against the Roman siege). In 600 B.C., Eupalinos of Megara constructed an aqueduct for Samos. Water was drawn from an inland lake on the island and passed through a tunnel 2.5 meters in diameter and a kilometre and one-half long. Within the tunnel, there was a meter-wide channel containing pipes for fresh water. Since being underground, it was not easily detected by an enemy who could otherwise cut off the supply. The Eupalinos tunnel is particularly notable because it is the second earliest tunnel in history to be dug from both ends in a methodical manner. Today, the Eupalinian subterranean aqueduct is regarded as one of the masterpieces of ancient engineering.

A slipway of 5 kilometres long was built in the Isthmus of Corinth for the movement of vessels or cargoes. This causeway for ships allowed the Greeks to avoid a journey of 720 kilometres around Peloponnesus (Figure 3.5). The vessels literally were portaged from one shore to the other using cradles on rollers.

Generally, the Greeks only built paved roads to connect a city with a sanctuary. These roads had built-in grooves or channels adapted to the span of a carriage for steadying and directing the course of the wheels, i.e. a "stone railway". The gauge was 138–144 cm (7–10 cm deep and 20–22 cm wide).

The engineer of the Classical Period used five simple machines: a lever, wheel, pulley, wedge and screw. The Greeks used iron in structures with great confidence. Iron was used as material for cantilevers in the Parthenon. The beams of wrought iron to transfer load to columns were used in column joints. Builders used limestone and marble, together with reinforcing iron members in their beams. The Greeks

a)

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b)

Photo by A. V. Valiulis

Figure 3.5. The Corinth canal: a – aerial picture; b – slipway

were familiar with mortar having developed hydraulic cement. Lead and wrought iron were materials used for joinery. The Greeks used post-and-beam construction in almost all of their buildings. They understood the arch but used it infrequently. The Greek builder commonly used headers (bricks the long axis of which lies through the wall with only the end showing on the wall face) in courses for binding face-stones to the body of the wall. The Greeks, using only basic structural forms such as post-and-beam, refined this method for building some of the most beautiful buildings of all time.

The Mines

The Greeks used bronze, a mix of tin and copper, for their weapons and tools. Later, iron was used because it produced harder tools and weapons. Athens found a large amount of silver in the mines at Laurium. Athenians learned to prospect, treat and refine the ore. At its peak, Athens had over 20,000 slaves mining in Laurium. All the miners were slaves. They were usually prisoners-of-war rather than criminals. Worse than torture or death was to find yourself a slave in the silver mines of Laurium, southeast of Athens, the source of much of Athens prosperity where the miners were routinely starved, savagely beaten, and, seldom seeing daylight, worked to death.

Athenians used money from these mines to pay for a large navy. There were very good mining areas around Greece. The northern mountains of Thrace had large gold and silver deposits. Cyprus was known for its copper, gold and iron. Laurium was known for the sliver and iron ore deposits. Typically, miners lived underground for a week and came out for a week. Abandoned places were utilized as kitchens and for sleeping quarters. Holes cut in the country rock were used for cupboards and lamp niches. The ore was lifted by small skips and hauled out by a rope often guided over a wheel on the rim of a shaft. Water was a limiting factor in the shaft depths. Pumping was not developed extensively. Deep shafts (110 m) necessitated air shafts (50 × 80 cm) for ventilation. Fires at the bottom of the shafts were probably utilized for draught. Shafts were sunk in pairs and parallel galleries driven from them. Frequent cross-cuts between galleries insured ventilation. A miner averaged 4.5 meter/month, timbering generally was not used. Mining and refining had environmental consequences. Mine tailings, the dross removed from the ore and litharge (lead oxide) cast off in smelting. In addition, we must reckon with deforestation, noxious fumes from smelting and smoke from combustion. The ancient mine drainage techniques did not allow for excavation below the level of subsoil waters. The passageways and steps of the Greek mines were dug out with the same concern for proportion and harmony found in their temples. Work was extremely difficult, and due to the depth of tunnels, they were sometimes more than 100 m deep. The miner, armed with his pick and iron hammer and hunched over in two, laboured to extract the ore (Figure 3.6).

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Figure 3.6. A painting on terra cotta depicting slaves in the mines of Ancient Greece (500 B.C.)

Hellenistic Engineering

By 352 B.C., new force – Macedonia – had arisen in the Greek political world. After Alexander the Great (356–323 B.C.) conquered the territories embracing Greek culture in varying degrees, Anatolia became Hellenistic, Egypt, superficially so, Mesopotamia, little or none. The capital of this pan-cultural expression was Alexandria. Greek learning, science and engineering thrived here. The Hellenistic world, with Alexandria at its centre, flourished perhaps as the most gifted of all Greek engineers. Aristotle (384–322 B.C.) and Straton published the first engineering treatise titled "Mechanics" or "Mechanical Problems" discussing the lever as well as gearing. Straton accepted Democritus's Atomic Theory and discovered compression. *Mechanics* discusses all simple machines except the screw. Archimedes of Syracuse (287–212 B.C.) discovered the principles of the screw and buoyancy. In Statics, he treats the theory of the lever, force parallelogram and the calculation of *pi*. Ctesibios of Alexandria (ca. 285–247 B.C.) invented the force pump, hydraulic pipe organ, metal spring, automated water clock and the keyboard. Hero of Alexandria (ca. 100 B.C.) understood the application of the vacuum and bent siphon. Euclid of Alexandria, was a Greek mathematician, often referred to as the "Father of Geometry". It was at Alexandria that the Pythagorean Theorem was proven, although we do not know if it was done by Euclid himself.

Important Hellenistic engineering accomplishments include:

- The Pharos of Alexandria – built in the 3rd century B.C was approximately 112 m in height with lighting on upper tier.
- The Colossus of Rhodes – built in 280 B.C. stood for 56 years. Constructed to between 27–37 meters, was shelled in bronze on an iron armature supported on columns.
- Pergamon Siphon was constructed at Pergamon (200–190 B.C.). In the pre-Greek eras, one had discovered that water in a U-shaped tube in both legs rises to the same level, which is the law of communicating vessels. This principle was applied in ancient siphons. Water in an open channel was poured into one or more closed pipes that descended to the bottom of a depression or valley and climbed up the other side to almost the original level. Water was then again transferred into an open conduit following its original course. Pergamon Siphon carried water 56 kilometres in 7.62–17.78 cm pipes. The pipes discharged into a reservoir 3–2 kilometres from the city and 30 meters above it. From there, a single 25 cm wooden or bronze pipe ran down across several valleys. The deepest point was 200 meters lower than the reservoir. At the low point in the line, water pressure approached 15 atm within the pipe. Lead pipes would have been unable to withstand this pressure.

Hero mentions the force pump constructed of bronze. The cylinders were oiled and valving allowed water to be raised to a height consistent with the ability of the pump to withstand internal hydrostatic pressure. The valves were held in place by tempered iron springs, which is the first known use of this metal in such a fashion. Very simple wooden pumps with a straight-through plunger were probably used even earlier to lift water from wells. Hero also invented a windwheel operating an organ, marking the first instance of wind powering a machine in history (Figure 3.7).

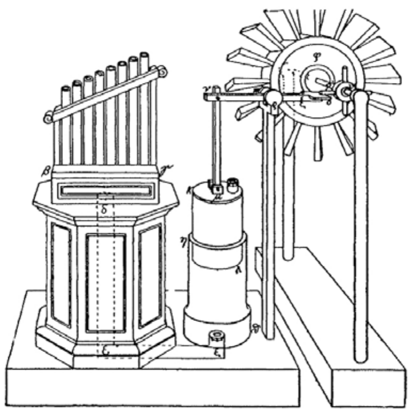

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Figure 3.7. Hero's wind-powered organ (reconstruction)

The other water-raising device from this period was the Archimedean screw (Figure 3.8). The rotor consisted of a wooden cylinder with a length/diameter ratio of 16:1. At either end, a circumference was

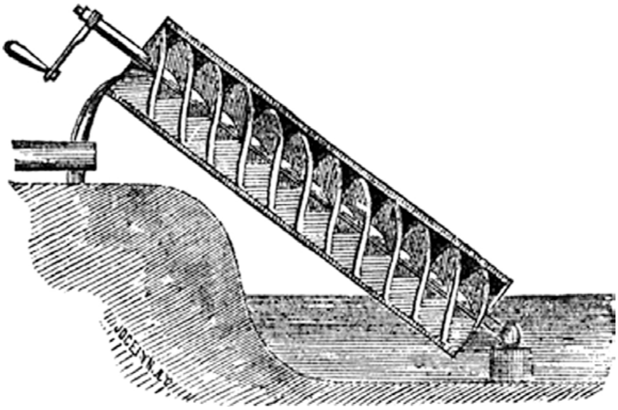

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Figure 3.8. The Archimedes screw

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Figure 3.9. Heron's Aeolipile

divided into eight equal arcs. Length was divided into sections, each equal to the one-eighth of the circumference. A blade was drawn obliquely around the rotor shaft one-eighth of the circumference along and at the same distance around the shaft with the resultant pitch of 45 degrees. Winding was continued in this way until the laminated blade was built up to twice the diameter of the shaft. Seven additional blades were formed in the same way. A wooden case fit around the blades. Barrel-like planks were painted with pitch and bound with iron hoops. The description of bearings is somewhat obscure, but the ends of the rotor were probably provided with iron caps.

An *aeolipile*, also known as a Hero engine, is a rocket style jet engine that spins when heated. In the 1st century A.D., Hero of Alexandria described the device, and many sources give him the credit for its invention. The aeolipile consists of a vessel, such as a sphere or cylinder, arranged to rotate on its axis and having oppositely bent or curved nozzles projecting from it (tip jets). When the vessel is pressurised with steam, it is expelled through the nozzles, which generates thrust due to the rocket principle. When the nozzles, pointing in different directions, produce forces along different lines of action perpendicular to the axis of bearings, the thrusts combine to result in a rotational moment (mechanical couple), or torque, causing the vessel to spin about its axis (Figure 3.9). It is not known whether the aeolipile was put to practical use in ancient times.

CHAPTER 4 Engineering in Ancient Rome, the Byzantine Empire, Muslim Countries and America

This chapter will help in

- introducing engineering in ancient Rome, the Byzantine Empire and Muslim Countries;
- presenting the civilizations of Ancient America (Olmec, Inca, Maya, Nazca);
- discussing achievements in the engineering of the Middle Ages;
- describing the design of Viking ships;
- giving the details of castles and domes;
- uprising and development of cathedrals.

Ancient Romans

The Greeks were great builders but, apart from a few exceptions such as Archimedes, were theoretical scientists rather than practical technologists. The Romans, being far more practical people, invented little of their own but did much to adapt the principles used by the Greeks. The Romans have been called the greatest engineers of Ancient Times. The key to success lay in their ability to organize and administer large-scale projects and their practical approach to engineering problems. The Roman Republic and the Roman Empire lasted for about 900 years, from 509 B.C. to about A.D. 400. Roman history consists mainly of two periods: the Republic (ca. 509–527 B.C.) and the Empire (27 B.C.–476 A.D.). The age of the Republic was the age of a military engineer; the age of the Empire was that of a civil engineer. It has been said in various forms that "Rome conquered Greece and, in turn, was conquered by Greece." It is obvious that Hellenistic design and principles did become the foundation of Roman engineering.

The military engineer was important to the Roman army. They were responsible for siege engines and artillery, built roads, bridges, baths and aqueducts. Mechanized arms were built of wood. An *onager* was a Roman siege engine that is a type of a catapult that uses torsional force, generally from a twisted rope, to store energy for the shot (Figure. 4.1). The onager used stone cannonballs of 1–5 kg attaining a range of roughly 0.5 km. The wooden arm was bent back, under pressure of catgut strings, to a catch. Catapults were mechanized bows.

In the time of Augustus, most Roman "architects" came from the ranks of military engineers. In the Roman army of the 2nd century A.D., the following specialists were attached to a legion: *architectus* – "master builder"; *mensor* – surveyor; *hydraularius* – water engineer; *ballistarius* – catapult artillery maker.

Rome built 300,000 kilometres of roads and 90,000 kilometres of "paved" viae. The armies could travel on highways from Britain to the Euphrates (4,000 km). They made possible trade, communications and postal service. The Roman engineer typically had very good materials at his disposal. He adopted local materials insofar as possible thus locating building quarries where roads were built. This resulted in the use of widely diverse materials as well as the types of construction. The Romans used lime mortar in combination with rubble, potsherds and crushed bricks to prepare grout that could penetrate between the stones of the foundation layers thus giving the solidity and resemblance of a wall. Grouting gravel surfaces resulted in a concrete road. The invention of true hydraulic concrete and its application in architecture and civil engineering was the only great discovery that can be ascribed to the Romans. About 150 B.C., they discovered the natural strata of *trass* (porous volcanic rock), a valuable substitute for lime that came to be known as *pozzolana*. Milestones were placed along the skies of the Roman roads with inscriptions indicating distances to one or more important places, road identity, the names of the road builder, the emperor in reign at the time and the name of the town or district that placed the milestones. The Romans classified their roads according to usage: *seinita* – a foot path of 0.3 m wide; *iter* – a path for horsemen and pedestrians of 1 m wide; *actus* – a single carriage of 1.2 m wide; *via* – a two-lane road of 2.5 m wide. Distances were

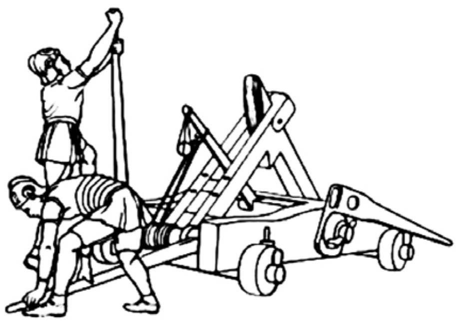

Figure 4.1. Sketch of an onager

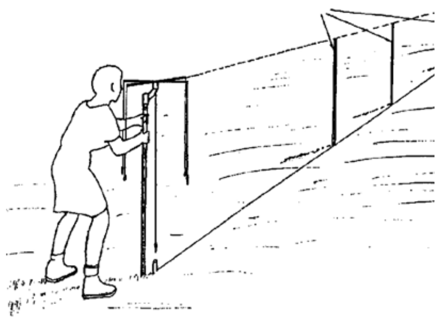

Figure 4.2. The use of the Roman groma

expressed in thousand-pace increments and denoted as M.P. in the inscriptions, The milestones were cylindrical columns three to nine feet (Roman) in length, commonly six feet (Roman) long and standing on a plain base with a cylindrical neckmould. The Roman surveyors laid out roads, but their principal task was dividing up land. For measurement purposes, the Romans used the groma or dioptra (Figure 4.2).

Water supply was of the greatest importance for the Romans in cities. Rome, in the 4th century A.D., had over 1350 fountains and 850 public baths while flowing water was also used for flushing plentiful sewers. The common use of lead pipes for water distribution producing lead poisoning and resulting in brain damage was held by some to have been one of the causes of the decline of the Roman Empire. In Rome alone, there were 11 aqueducts based on a low-pressure, continuous, gravitational flow theory of hydraulics. Technical divisions within Roman aqueduct engineering were *agrimensors* (land surveyors), *librators* (levellers), *mensors* (quantity measurers) and *aqualegus* (aqueduct inspector). In Rome, the purity of water in aqueducts was maintained by covered channels, reservoirs ("castellum") and settling basins. There were strict laws against pollution and water theft. Water from aqueducts was distributed to public fountains, industries (textiles), some private houses and public baths. Where hot springs were not available, furnaces raised water to the desired temperatures and heated the floors and walls of the baths (hypocausts). Excess water from aqueducts was used for removing sewage. Public latrines were maintained in Rome. The great sewer of Rome, the *Cloaca Maxima*, was five meters in height and, in some places, four meters wide. There was no sewage treatment, and waste discharged into the Tiber River and then to the Mediterranean Sea.

The Roman mill, described by Vitruvius in about A.D. 180, was the first machine in which gears were used for transmitting power. This mill had a vertical wheel driving a horizontal upper millstone through a lanthorn and trundle gears. Watermills did not increase greatly in number until the 4th and 5th centuries A.D., towards the end of the era of the Roman Empire. Watermills were used for corn milling, but possibly, a few – for ore crushing and driving forge hammers. In building, the Romans used cranes frequently fitted with a treadmill to turn the windlass and a rope running in pulleys. The extensive network of roads across the Roman Empire included many bridges. The construction of river bridges often involved building coffer dams of timber piles sealed with clay, and Archimedean screw to drain water from the completed coffer dam.

Hadrian's Wall was defensive fortification in Roman Britain, begun in 122 A.D. and was 117.5 km in length. The width and height of the wall were dependent on construction materials available nearby. The eastern part of the wall was made from squared stone while the western one was made of turf (Figure 4.3).

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Figure 4.3. Milecastle 39 on Hadrian's Wall

The earliest record of the name of Cappadocia (Asia Minor) dates from the late 6th century B.C. Cappadocia was known as *Hatti* in the late Bronze Age, and was the homeland of Hittite power. First built in soft volcanic rock possibly in the 8th–7th centuries B.C. The magnificent landscape has been formed from its solidified lava streams, ash and tuff stone, all dating from the Neocene period. Tuff is an ideal building material in the form of cut stones and has excellent insulating properties. The porous structure of rock is optimal for evening out swings in temperature between hot summers and cold winters. This relatively light stone, which can be cut with simple tools, provides the optimal basis for constructing rock-cut dwellings. Cappadocia contains several underground cities largely used by early Christians as hiding places before Christianity became an accepted religion. The underground cities have vast defence networks of the traps throughout their many levels. These traps are very creative and include such devices as large round stones to block doors and holes in the ceiling through which defenders may drop spears. This kind of a defence tunnel system also was made to have thin corridors that did not allow enemies moving in groups through making it easy to pick them off. Recently it has been shown mineral *andesite* was used as a pot to melt copper. Stone was hewn from an andesite layer within the complex. In order for it to be used in metallurgy, fifty-seven holes were carved into stone. The technique was to put copper ore into each of the holes (about 10 cm in diameter) and then to hammer the ore into place.

Engineering in the Byzantine Empire and Muslim Countries

After the fall of the Western Roman Empire, Europe, the Middle East and Africa became separate power blocks: the Frankish Kingdom (France-Germany, 481–814 A.D.), the Visigothic Kingdom (Spain, 507–711 A.D.), the Ostrogothic Kingdom (Italy, 493–526 A.D.), the Vandal Kingdom (Northern Africa, 429–533 A.D.) and the Eastern Roman or Byzantine Empire.

Constantinople, the Byzantine capital, was a civilized centre of the world. For several centuries, this Greek-speaking metropolis survived due the accumulated wisdom of antiquity. For a brief time, engineers, architects and scientists mounted scientific and technologic innovations. The Byzantine Empire survived for nearly a millennium after the fall of the West. At the heart of its military strength, lay the work of engineers. The walls of Constantinople rose to over 12 meters and were four meter thick with a core of concrete.

The most enduring construction was Haghia Sophia – the great domed church built between 532 and 537 using brick and concrete. The great height of Haghia Sophia comes from the use of 30 meter columns topped by 18 meter arches supporting the dome. Erected over an area of 1300 square meters, one can appreciate the impression of such a space. The use of chemical weapons particularly "Greek Fire" proved pivotal in the defence of fortification and at sea.

Parallel to the Byzantine Empire, fanatical Arab armies conquered Syria and Mesopotamia by 638, Egypt by 640, Persia by 641 and North Africa in 697 A.D. By 711, they conquered Spain, invading France. The Golden Age of Islam was from 900 to 1,100 A.D. During this period, the Arabs founded the centres of learning, copied Greek texts into Arabic, developed books from paper and introduced Arabic numerals, including zero and a decimal point. The Islamic world was a melting pot of numerous cultures. After the conquest period (ca. 750), learning the antiquity based civilization began to form. By 793, the first paper mill was operating in Baghdad.

Arab scientists and engineers developed an astrolabe for measuring angles and celestial positions, wrote more than 100 medical and over 200 works on meteorology and optics, promoted water raising machines and spherical trigonometry, invented sine, tangent, cotangent, improved the pendulum, studied optics and astronomy in physics, scientifically farmed making use of fertilizers, developed the horseshoe arch in architecture. Arab science invented chemistry that we can recognize with evaporation, filtration, sublimation, melting, distillation and crystallization. They developed sugar refining, alcohol, sulfuric and nitric acids, the production of gasoline as well as built windmills to pump water from wells for irrigation and ground corn. The Arabs also raised cooling towers where they place wetted sails of cloth.

These constructions, attached to larger buildings such as mosques, provided humidity and cool air to the interiors. The Arab windmills were wrecked by the Turks and Mongols in later 14^{th} and 15^{th} centuries. Arab engineers became the masters of hydraulics. They refitted ancient systems and built the new ones. To power these irrigation and water supplies, they typically used water wheels, built dams and tidal mills. To increase the rate of flow to the wheel, the engineer would set the mill between bridge piers to take advantage of higher current velocity through opening. Dams to create reservoirs for water supply were an Islamic innovation. Arabs repaired and used the dams built by the Romans and Greeks and later erected masonry dams set in concrete and reinforced with lead dowels. Cordoba became the capital of Moorish Spain. The Moors utilized the existing Roman systems such as Segovia, Toledo and Almunecar aqueducts. The engineers used the weight of the dam to resist uplift and water pressure.

Using the *Wootz* process for steel production, smiths in Damascus and Toledo made the finest swords of antiquity. Steel was produced by a crucible process using magnetite ore, bamboo charcoal and the leaves of certain carbonaceous plants. Smelting in the charcoal fire of hot blast using high temperatures produced a cake of metal. The crucible was heated to roughly 1200°C where wrought iron becomes austenite capable dissolve carbon atoms in its lattice structure. Wootz steel carbon content ranged from 1.5 to 2%, which after slow cooling, formed cementite that had to be forge-hammered to break up it into the final forms of Damascus steel. After hammering, the steel was hardened for forming martensite, a hard, iron-carbon phase.

Arabs equalled and surpassed the Romans in hydraulic engineering. Building and fortifications were borrowed without much innovation beyond adiabatic cooling, pointed arches and machicolated walls (cantilevered wall parapets). They cared little for roads never adapting the Roman system to their use. In fine technology and power, they surpassed their predecessors. Sciences such as chemistry, mathematics, astronomy and even that of agriculture are more derivative to Arabic knowledge than we admit.

Byzantium and the Caliphate fell before the Seljuk Turks. The Turks were barbarians and militarists. They were users rather than builders. Onto their realm, they grafted Arabic learning and Byzantine pomp. They were more successful in assimilating the latter. The new barbarians come from Asia under Jenghis Khan. The Mongol invasions resulted in the conquest of most of Eurasia. His horsemen destroyed Islamic kingdoms in Iran thus leaving Turkestan and Iraq only wastes where cities and irrigated fields once stood.

Civilizations of Ancient America

The *Olmec civilization* is the name given to a sophisticated Central American culture with its heyday between 1200 and 400 B.C. The Olmec heartland lies in the Mexican states of Veracruz and Tabasco. It is not known with any clarity what caused the eventual extinction of the Olmec culture. It has also been established that between 400 and 350 B.C., the population in the eastern half of the Olmec heartland dropped precipitously, and the area would remain sparsely inhabited until the 19^{th} century. The most recognized aspect of the Olmec civilization is enormous helmeted heads. It is now generally accepted that these heads are the portraits of rulers. The heads range in size from 3.4 m to 1.47 m in height. It has been calculated that the largest heads weigh between 25 and 50 tons (Figure 4.4). Olmec artworks are considered as most striking and beautiful among the masterpieces of Ancient America and the world as a whole.

The geographic extent of the *Mayan civilization* extended throughout southern Mexican states and Yucatan. The Maya area was initially inhabited around 2600 B.C. The first clearly "Maya" settlements were established in approximately 1800 B.C. The Classic period (c. 250–900 A.D.) witnessed the peak of large-scale construction and recording of monumental inscriptions (Figure 4.5).

The most notable monuments are *stepped pyramids* built in their religious centres (Figures 4.6–4.8) that went into a decline during the 8^{th} and 9^{th} centuries. The Mayan script was a *logosyllabic* system. Individual symbols ("glyphs") could represent either a word or a syllable; indeed, the same glyph could often be used for both. Mayan writing consisted of a relatively elaborate set of glyphs, which were laboriously painted on ceramics, walls or bark-paper codices, carved in wood or stone, or molded in stucco.

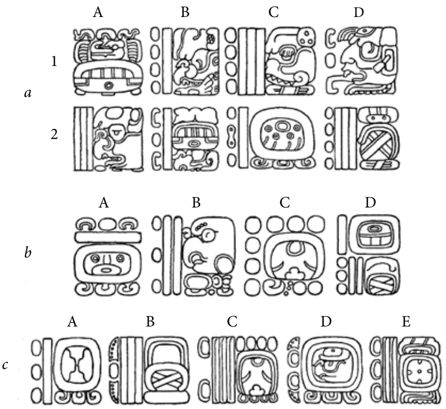
Figure 4.4. Olmec stone head

Figure 4.5. Maya maske

Figure 4.6. Maya stepped pyramids

Figure 4.7. Maya monumental inscriptions

a)



b)

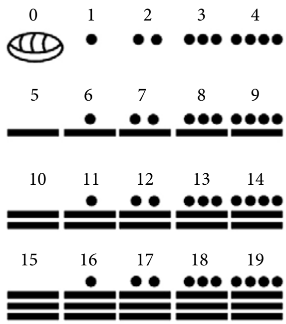


Figure 4.8 Maya numerals and scripts: a – Maya hieroglyphs; b – Maya numerals

The Mayan script combines about 550 logograms representing the whole words and 150 syllabograms (which represent syllables). There were also about 100 glyphs symbolising the names of places and gods. About 300 glyphs were commonly used. There is evidence that the Mayan population exceeded the carrying capacity of the environment including the exhaustion of agricultural potential and overhunting megafauna. An intense 200 year drought led to the collapse of the Mayan civilization.

The *Nazca civilization* was located in the Nazca region on the southern coast of Peru between about A.D. 1-750. The Nazca are known for their elaborate textile and ceramic art, including a mortuary ritual associated with warfare. The Nazca culture is characterized by its beautiful polychrome pottery painted with at least 15 distinct colours (Figure 4.9).

Figure 4.9. Nazca pottery

Figure 4.10. Nazca lines

The Nazca, like all other Pre-Columbian societies in South America, including the Inca, had no writing system, in contrast to the contemporary Maya. Thus, iconography or symbols painted on their ceramics served as a means of communication. The Nazca are also known for their technically complex textiles that were most likely made from spun cotton and wool. The Nazca people created a hydraulic system to sustain life in the exceedingly arid environment. This irrigation system was made up of underground channels, known as puquios, which tapped into subsurface water beneath the ground. The channels were dug into the mountainside until they reached the aquifers under the surface. The famous remnants of the Nazca civilization are the Nazca Lines – vast line drawings of animals and abstract figures that can be seen from Space (Figure 4.10). However, the Nazcans suddenly disappeared.

The Andean civilization probably began c. 9500 BP. Geographical conditions resulted in distinctive physical development characterized by a unique lung system with almost one third greater capacity than other humans. The Incas had slower heart rates, a blood volume of about 2 l more than other humans and a double amount of haemoglobin that transfers oxygen from lungs to the rest of the body. The Inca people began as a tribe in the Cuzco area around the 12th century A.D. The *Inca civilization* was the largest one in the Americas when Spanish conquistadors arrived in the early 16th century. Known for their unique writing system (called the *quipu*), a magnificent road system and lovely residence called Machu Picchu, the Inca also had some pretty interesting burial customs. Inca architecture was by far the most important of the Inca arts, with pottery and textiles reflecting motifs that were at their height in architecture. The Incas had no iron or steel, but they had developed an alloy of bronze. The Andean nations prior to the Incas used arsenical bronze at best. The Incas introduced to South America the tin / copper alloy today commonly associated with "Bronze Age" metallurgy. Spanish conquistadors led by Francisco Pizarro reached Inca territory by 1526. In 1572, the last Inca stronghold was discovered, and the last ruler was captured and executed bringing the Inca Empire to an end.

Engineering in the Middle Ages

The Middle Ages or Medieval period is the one in European history lasting from the 5th to 15th centuries. In the *Early Middle Ages*, depopulation, deurbanization and barbarian invasions, which had begun in Late Antiquity, continued. The barbarian invaders formed new kingdoms in what remained of the Western Roman Empire. During the *High Middle Ages*, which began after A.D. 1000, the population of Europe increased greatly, as technological and agricultural innovations allowed trade to flourish and crop yields to increase. The *Late Middle Ages* were marked by difficulties and calamities such as famine, plague and war, which much diminished the population of Western Europe. The Black Death (1347 and 1350) killed approximately a third of the European population.

Viking Ships

The genius of northern ship builders was in joining a strong hull to a frame rather than fastening individual planks to the frame as did the carvel hull designers of the Mediterranean (Figure 4.11). An example of a ship (a burial of an East Anglian king) is 24 meters in length, 4.27 meters in breadth and a draft of 6l centimetres. Copper nails were used as fasteners and lead rings were sewn to sail for taking brails (ropes that ran up from the deck through the rings, over the yard and down again). Stone was used as anchors and as ballast.

Their construction provided a relatively light hull with great longitudinal strength in the shell firmly held by a complete interior frame. With the continuous, single-piece external keel, scarfed to stern, and stern posts, the Norse boats provided the strongest craft of strength-to-weight ratio seldom exceeded today.

Figure 4.11. The Oseberg Ship in the Viking Ship Museum, Oslo

Castles

A medieval fortress (*castle*, Latin: *castellum*) is a type of fortified structure built in Europe and the Middle East during the Middle Ages. The medieval fortress was built about a central "donjon" or keep. The tower structure was first utilized by the Normans in the 11^{th} century. Simple motte-and-bailey castles from the 11^{th} century with their timber frames and palisades evolved into substantial concentric castles by the 14^{th} century. The "motte-and-bailey" castle became the standard of Medieval design. The motte was a man-made mound surrounded by a trench, on which a tower of timber was erected. The bailey was kidney-shaped enclosure at the bottom of the mound intended for horses and cattle. The works were surrounded by timber palisades. The motte measured 30-90 meters in diameter, and its height ranged from 3 to 30 meters. Subsequent changes in castle architecture tended to improve the existing ideas as opposed to developing completely new castle design. In many instances, the site itself would determine and even limit the castle plans. Concentric castles were built during the Crusades. Probably the earliest example would be the Land Wall in Constantinople built in the 5^{th} century. It had three distinct walls each one becoming progressively larger. The entrances to castles were, from a very early age, recognised as their weakest spot. This created the need for concentric castles with walls getting progressively smaller. The attackers would not only have to make more breaches but the design of the walls could force them into areas where they could be massacred. Defensive architecture covers drawbridges, spiral staircases, killing grounds, machicolations and loopholes. The influence of Medieval social changes reflected in the role and design of the castle. Along with weakening in central authority, the aristocracy

a) b)

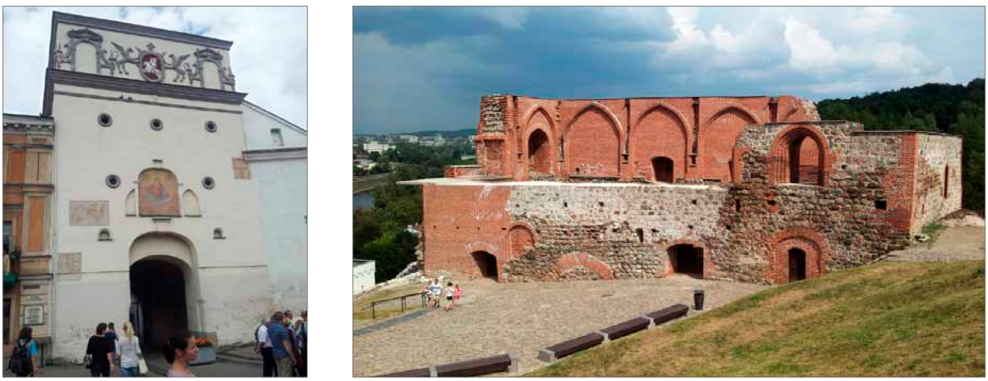


Figure 4.12. Vilnius fortifications: a – the gate of Dawn at the southern part of the city, b – remains of the Upper Castle in Vilnius built in 1422

became semi-autonomous of the king, holding large estates with wide powers of justice and administration. The castle was thus a caput of a lordship or chateaux of a seigneur instead of simply a military stronghold.

City gates were traditionally built to provide a point of controlled access to and departure from a walled city for people, animals, vehicles and goods. The Vilnius city wall was a defensive wall around Vilnius, capital city of the Grand Duchy of Lithuania. It was built between 1503 and 1522 as a part of defensive fortifications for the capital city. It contained nine gates. Medininkai Gate (Figure 4.12, a) guarded the entrance to the southern part of the city. One part of the Vilnius castle complex was built on a hilltop and is known as the Upper Castle. The hill on which it is built is known as Gediminas Hill, about 40 meters in height. The hill was strengthened with defensive wooden walls that were fortified with stone in the 9th century. Around the 10th century a wooden castle was built, and since about the 13th century the hilltop has been surrounded by stone walls with towers. In 1323 the castle was improved and expanded. After a major fire in 1419, was initiated a reconstruction of the Upper Castle, along with the fortification of other buildings in the complex. The present-day remains of the Upper Castle date from this era (Figure 4.12, b). Reconstruction of the castle ended in 1422. After the 16th century, the Upper Castle was not maintained, and it suffered from neglect.

Trakai Island Castle (Lithuania) was built in several phases (Figure 4.13). During the first phase, in the second half of the 14^{th} century, the castle was constructed on the largest of three lake islands. The expansion of the forecastle was finished in the early 15^{th} century. The walls of the forecastle were strengthened to a thickness of 2.5 metres and raised with additional firing galleries. Three major defensive towers were constructed on the corners. The top story of the towers was designed for soldiers and housed a large number of cannons. The main gatehouse was also constructed which, along with the Ducal Palace donjon, had movable gates. The gatehouse was reinforced with additional sections for firing galleries. The Ducal Palace and the forecastle were separated with a moat, just wide enough for small boats to sail through. They were connected by gates that could be raised in case of an enemy attack. Trakai Island Castle lost its military importance soon after the Battle of Grunwald in 1410.

Figure 4.13. View of Trakai Island Castle in winter

Figure 4.14. The inner yard of Medininkai castle (Lithuania)

Medininkai Castle (Lithuania) is a medieval castle built in the late 13th century or the first quarter of the 14th century (Figure 4.14). The defensive perimeter of the castle was 6.5 hectares; it is the largest enclosure type castle in Lithuania. The castle was built on plain ground and designed for flank defence. The rectangular yard of the castle was protected by walls 15 metres in height and 2 metres thick. The castle had 4 gates and towers. The main tower (donjon), about 30 metres in height, was used for residential quarters. Medininkai was first mentioned in 1392. Because of the increased use of firearms, this type of the castle was no longer suited for defensive purposes and therefore later used as a residence.

The construction of the earth bastion-type castle in Biržai (Lithuania) started in 1586. The major castle building works were finished in 1589 (Figure 4.15). Biržai Castle served as the main defensive structure during the wars with Sweden.

Klaipėda Castle, also known as Memel Castle, is a castle built by the Teutonic Knights in Klaipėda, Lithuania, near the Baltic Sea. The castle was first mentioned in written sources in 1252 and underwent numerous destructions and reconstructions in the centuries that followed. The castle of the year 1252 was wooden, protected by a tower and constructed in a marshy area. It soon became a prime outpost in the war between the Christian Orders and Lithuanian pagans allied with the Samogitians.

In 1379, the castle was destroyed in an attack by the Samogitians and Lithuanians. This destruction was followed by reconstruction; in 1393, the major defensive tower was erected, which was, however, destroyed by the Lithuanians in the same year. Continued expansions and renovations of the castle were systematically pursued until the 15th century. After the Teutonic Order lost the key Battle of Grunwald in 1410, the military importance of the castle was sustained, as Lithuanian rulers regarded these territories as a part of their patrimony. In the middle of the 15th century, the castle was again upgraded to withstand assaults using firearms. During the 16th century, it was further advanced into a bastion thus becoming one of the first such fortifications in the region (Figure 4.16). Between 1529 and 1559, the castle underwent a renovation by French engineers. After its reconstruction, the castle had five towers associated with the main building. The main tower probably had six floors and was about 15 meters in diameter. In 1629, the castle was devastated by Swedish attacks, and in 1757, the castle sustained severe damage during a war with Russia.

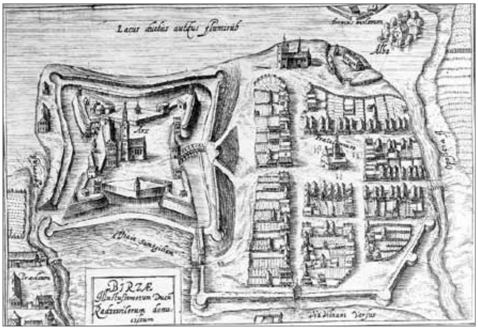


Figure 4.15. The earth bastion-type castle in Biržai

Figure 4.16. The earth bastion-type castle in Klaipėda

The Castle of the Teutonic Order in Marienburg (Malbork, Poland) is the largest one in the world by the surface area and the biggest brick building in Europe (Figure 4.17). The castle was built by the Teutonic Order after the conquest of Old Prussia. Its main purpose was to strengthen their own control of the area following the Order's 1274 suppression of the Great Prussian Uprising of the Baltic tribes. The work lasted until around 1300. The castle is located on the south-eastern bank of the river Nogat. It was named Marienburg after Mary, a patron saint of the religious Order. Soon, it became the largest fortified Gothic building in Europe. In the summer of 1410, the castle was besieged following the Order's defeat by the armies of Władysław II Jagiełło and Vytautas the Great in the Battle of Grunwald. During the Thirty Years' War, in 1626 and 1629, Swedish forces occupied the castle. They invaded and occupied it again for the period from 1656 to 1660 during the Deluge. After Prussia and the Russian Empire made the First Partition of Poland in 1772, the town became a part of the Kingdom of Prussia, a province of West Prussia. At that time, the officials used a rather neglected castle as barracks for the Prussian Army. The Teutonic Castle at Marienburg served as the blueprint for the Order Castles of the Third Reich built under Hitler's reign. In 1945, during World War II combat in the area, more than a half of the castle was destroyed.

Figure 4.17. Panorama of Malbork Castle. Teutonic Order named it Marienburg Castle

Domes

A dome is an element of architecture that resembles the hollow upper half of a sphere. The construction of the first technically advanced true domes began in the Roman Architectural Revolution, when they were frequently used by the Romans to shape large interior spaces of temples and public buildings such as the Pantheon in Rome. Domes in Western Europe became popular again during the Renaissance period and reached the zenith of popularity during the early 18th century in the Baroque period. Reminiscent of the Roman senate, during the 19th century, they became a feature of grand civic architecture. Many domes, particularly those from the Renaissance and Baroque periods of architecture, are crowned by a lantern or cupola, a medieval innovation that not only serves to admit light and vent air, but gives an extra dimension to the decorated interior of the dome (Figure 4.18).

a)

Public domain

b)

Photo by A. V. Valiulis

Figure 4.18. Domes: a – dome of St. Peter's Basilica in Rome crowned by a cupola; b – Saviours church crowned by a cupola (Vilnius)

Cathedrals

The word "cathedral" is derived from the Latin word *cathedra* ("seat" or "chair") and refers to the bishop or archbishop's chair or throne. The building itself, by its physical presence, symbolises both the glory of God and that of the church. Cathedral buildings, especially those dating from the medieval period, are frequently the grandest of churches in the diocese (and country). The ancient cathedrals of England, Northern France, Belgium, Spain, Portugal, Germany, South America and many individual cathedrals from Italy and other parts of Europe, are among the largest and finest

Figure 4.19. The Vilnius Cathedral

religious buildings. Many are renowned for their architecture or decorative features such as sculpture, stained glass and frescos. Cathedrals are very often oriented east/west, so that the worshippers look towards the rising sun symbolising the Risen Christ. The architectural form of the building most frequently has the ground plan of a cross. This form is both functional and symbolic, and such a symbolism refers to the cross on which Jesus was crucified. The main body of the building making a longer arm of the cross is called the *nave*, and is the place where worshippers congregate. The term comes from the Latin word for ship. The cathedral is symbolically a ship bearing the people of God through the storms of life. Cathedral buildings following Western European tradition symbolise the progression of the Christian soul towards Salvation. Many cathedrals of Eastern European tradition are centrally planned. These churches are almost always domed. Symbolism in these cathedral structures is of the hierarchy of Earth and Heaven, and therefore often reveals its meaning through the internal decoration of the building with frescos or mosaics. Because many cathedrals took centuries to build and decorate, they constitute a major artistic investment in the city they stand. Not only may the building itself be architecturally significant, but the church often houses treasures such as stained glass, stone and wood statues, historic tombs, richly carved furniture and objects of both artistic and religious significance, i.e. reliquaries. Moreover, the cathedral often plays a major role in telling the story of the town through its plaques, inscriptions, tombs, stained glass and paintings.

It is believed that in pre-Christian times, the Baltic pagan god Perkūnas was worshiped at the site of *Vilnius Cathedral* (Figure 4.19). It has also been postulated that the King of Lithuania Mindaugas ordered the construction of the original cathedral in 1251 after his conversion to Christianity and appointment of a bishop to Lithuania. The Cathedral was reconstructed to its present appearance according to the design of Laurynas Gucevičius in the Neoclassical style (Figure 4.19). Between 1786 and 1792, three sculptures by Kazimierz Jelski were placed on the roof of the Cathedral – Saint Casimir on the south side (originally symbolized Lithuania), Saint Stanislaus on the north (symbolized Poland) and Saint Helena in the centre (symbolized the dominance of Russia).

CHAPTER 5 Science and Technologies in the Renaissance

This chapter will help in

- discussing the rebirth of science and industry in the Late Middle Ages;
- explaining the importance of the invention of the printing press;
- describing the ideas of Leonardo da Vinci and Galileo Galilei;
- representing the achievements of Georgius Agricola, Isaac Newton, Leonhard Euler, Charles Auguste Coulomb, etc.;

The Renaissance and Scientific Revolution

The complete rebirth of science and industry in the Late Middle Ages suffered setbacks as a result of the Hundred Years War in Europe and the bubonic plague that killed up to one-third of the population of Western Europe. The Renaissance (*"re-birth"*) was a cultural movement that spanned the period roughly from the 14th to the 17th century, beginning in Italy in the Late Middle Ages and later spreading to the rest of Europe. The growth of a humanistic outlook fostered a new growth in sciences and had a profound effect on engineering. Though the availability of paper and the invention of metal movable type sped the dissemination of ideas from the later 15th century, changes in the Renaissance were not uniformly experienced across Europe. Technology in the Renaissance era is marked by profound technical advancements such as the printing press, linear perspective in drawing, patent law, double shell domes and bastion-type fortresses. The Renaissance science spawned the Scientific Revolution. Science and technology began a cycle of mutual advancement. Some important Renaissance technologies were mining and metallurgy; blast furnace enabled iron to be produced in significant quantities; finery forge enabled pig iron (from the blast furnace) into bar iron (wrought iron); slitting mill mechanized the production of iron rods for nail-making; smeltmill increased the output of lead over the previous methods.

The crank and connecting rod mechanism that converts circular into reciprocal motion is of utmost importance for the mechanization of work processes. Cranks and connecting rods become an integral part of machine design and are applied in ever more elaborate ways. The invention of the printing press by Johannes Gutenberg (1398–1468) is widely regarded as the single most important event of the second millennium and is one of the defining moments of the Renaissance. The mechanical device consists of a screw press modified for printing purposes and can produce 3.600 pages per workday. By the start of the 16th century, printing presses were operating in over 200 cities in a dozen of European countries. The earliest known parachute design appears in an anonymous manuscript from 1470s Italy. Around 1485, a more advanced parachute was sketched by polymath Leonardo da Vinci. In 1617, Veranzio successfully tests his parachute design by jumping from a tower in Venice. The earliest recorded uses of the astrolabe for navigational purposes are by the Portuguese explorers during their sea voyages around Africa (1481–1498). While dry docks were already known in Hellenistic shipbuilding, these facilities were reintroduced in Portsmouth navy base (1495/96). The earliest known description of a floating dock comes from a small Italian book printed in Venice in 1560.

The Renaissance engineers showed a strong proclivity for experimental study thus drawing a variety of technical devices, many of which appeared for the first time in history on paper. However, these designs were not always intended to be put into practice, and often practical limitations impeded the application of revolutionary designs. For example, Leonardo da Vinci's ideas of the conical parachute or winged flying machine were only applied much later.

Like the microscope, the telescope was invented in Holland and bent light to view a desired image. Galileo Galilei used the telescope to develop his theories and ideas about Earth and its relation to stars and the rest of the universe (Figure 5.1). By using convex and concave lenses, he was able to magnify stars and planets.

Another major development in the world of science was the microscope invented by Zacharias Janssen and his father in Holland in 1590. It was a compound microscope with two lenses. The microscope was used for viewing things too tiny to be

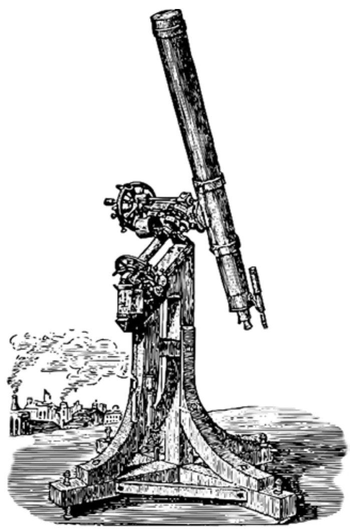


Figure 5.1. A telescope much like what Galileo would have used

seen by the naked eye. It used concave and convex lenses to bend light and greatly enlarge images. In 1660, Dutch scientist and microscope builder Anton van Leeuwenhoek was the first to study bacteria using the microscope. His microscopes had a magnifying power of up to 270 times larger than the actual size and used a single lens. This was considered to be the best available power for his time and was used for studying microorganisms and human blood cells.

The submarine was invented in 1624 by a man named Cornelius van Drebbel. However, Leonardo da Vinci drew out the basic concept of the submarine over one hundred years before. Drubbel's ship could stay underwater for a few hours, but it only went about fifteen feet under the surface.

Fire was difficult to create until Robert Boyle invented the match in 1680. Although fire could be made by rubbing sticks together or by striking flint to steel, this was a time consuming process. Boyle discovered that when phosphorus and sulfur were rubbed together, they would burst into flame. Boyle knew that this was not because of friction, but because of a chemical nature of these two substances.

In the beginning of the Renaissance, the first portable clocks developed in Florence, Italy appeared (1410). Before this time, mechanical clocks were large, fixed devices. The spring-driven clock made it possible to carry time around with you.

Fortress engineering, in addition to hydraulic engineering, continued to be that sphere in which theory and practice, architecture and mathematics were most interwoven. Military engineers had to demonstrate a more profound knowledge of mathematics being concerned with ballistics, which is the science or art of designing and accelerating projectiles so as to achieve a desired performance. Gun ballistics is the work of projectiles from the time of shooting to the time of impact with the target. The earliest known ballistic projectiles were stones and spears. Although gunpowder was invented around 1040, many destructive devices were invented thereafter. Rockets were launched as fireworks and weapons in China in the early 1230's. This led the way to William Congreve (1670–1729) developing rockets to use during wars. Launching tubes were developed by this engineer to improve their accuracy. Also, coming from the invention of gunpowder was the gun and other projectile-firing artillery. In the 16th century, a howitzer was developed by English and Dutch armies. The howitzer is a type of artillery piece characterized by a relatively short barrel and the use of comparatively small propellant charges to propel projectiles at relatively high trajectories with a steep angle of descent. The howitzer was used for firing explosive shells in a high arc and for intending to hit a distant target.

The use of prime movers, other than man, was rare in the work of cranes and hoisting gear. The treadmill was the principal motive device for cranes and systems of gearing.

In surveying up to the Renaissance, the number of instruments had hardly increased since antiquity. During the Renaissance, the problem of measuring height and distances by indirect methods was solved by elementary theorems of triangular. The German mineralogist and writer on mining Georgius Agricola (1494–1555) is the major figure in the history of technology. His main contribution was a book about mining and metallurgy "De re Metallica." Agricola pioneered in finding ores, the theory of forming mineral veins, mining law, mine surveying, tools, machines, pumps, hoists, water power, ore preparation and smelting, manufacturing salt, soda, sulfur, bitumen and glass. In his work on mineralogy published in 1546, the author set the initial standards for the future science. Knowing nothing of atomic theory, stoichiometry or crystallography, it was a colossal achievement to refute the ancient theory of the four elements: earth, water, fire and air. For classifying minerals, he used the criteria of outward appearance such as hardness, colour, consistence, solubility, smell or taste. To the then-recognized seven metals, including gold, silver, copper, tin, iron, lead and mercury, he added bismuth and antimony. The miners of his time wore a smock with belt, breeches, a short apron and a pointed cap. Agricola knew about personal protective clothing. He recommended elbow-high leather gloves for work with aggressive minerals and a veil worn before the face to protect from dust.

Renaissance Mathematicians and Scholars

For all revival of the Renaissance in Italy, its engineering provided little to the practical art of this type of activity after the 16th century. During the 18th century, the Enlightenment culminated in France and England. Philosophy and science increased in prominence. Philosophers dreamed of a brighter age. This dream turned into a reality with the French Revolution. France became the centre of engineering from the 17th century. In science, France must share the stage with England notably due to Sir Isaac Newton (1642–1727) who was the greatest scientist of the 17th–18th centuries. Apart from the theory of gravitation and optical theories, he was a co-inventor with Leibnitz of calculus. Newton stated his now famous laws that

- everybody continues at rest, or in uniform motion, unless compelled to change that state by forces impressed upon it;
- the change of motion is proportional to forces impressed upon it;
- to every action there is always an opposite and equal reaction.

Leonhard Euler (1707–1783), a Swiss mathematician, worked in bending and tried to solve buckling problems or determine elastic curves and associated centres of gravity. Euler proved that short columns fail by compression while the long ones by bending. He was the first to formulate laws governing the flow of fluids and explain the importance of pressure for the flow.

Robert Hooke (1635–1703) determined that elasticity was critical to a complete evaluation of displacement, i.e. force with the help of which a spring attempts to regain its natural position is proportional to the distance it has been displaced. Hooke analyzed forces acting on the arch and formulated a limit of elasticity to materials.

Charles Auguste Coulomb (1736–1806) formalized the universal treatment of the cantilever beam of a rectangular cross section. His treatment takes into consideration shearing strength as well as compressive and tensile strength, and admits, in principle, to the relationship between stress and strain. Coulomb summarizes three basic laws of statics and equilibrium:

- a sum of tensions must balance a sum of compressions;
- a sum of the vertical components of internal forces must equal the applied load;
- a sum of moments at internal forces must balance the bending moment produced by the applied load.

By 1779, Coulomb suggested the use of compressed air in caissons. The application of this technique by later engineers led to the construction of bridges across great rivers and bays together with tunnels in the areas never before attempted. Rene Descartes (1590–1650) invented analytical geometry.

Architecture is both the process and product of planning, designing and constructing space that reflects functional, social and aesthetic considerations. It requires the manipulation and coordination of material, technology, light and shadow. By the middle of the 18^{th} century, the principle of virtual displacement was known. The properties of many important building materials were tabulated. Further, the statical behaviour of buildings was being examined by scientists. This era was marked by the first attempt to theoretically survey the stability of structures. In 1742–1743, Pope Benedict XIV asked for a structural analysis of St. Peters Dome in order to determine the cause of cracks and damage. After evaluation, the design of the building was determined to be strengthened by the impost ring. The interesting part then began with an attempt to calculate the horizontal thrust and to prove the two iron tie rings, or bands, built into the dome were no longer able to withstand that thrust. London St. Paul's dome, built between 1675 and 1710, performed the role of St. Peter's in the Anglican church. Both are monumental structures topped by great domes. Chief divergence between the two lies in their structural difference which traces to the one harking backward to a classical, empirical tradition of building design vs. an experimental approach based on theoretical concepts of the structures newly proposed in the 17^{th} century.

Schools. Sebastien la Prestre de Vauban (1633–1707) was a great French military engineer who suggested the formation of a Corps of Engineers to be organized for the purpose of building roads, bridges and fortifications. Vauben developed polygonal and star-shaped fortifications, used canals as a military strategy to assist in defending the country, personally undertook to rebuild the harbour at Dunkirk into a fortified port and harbour. It was during his lifetime that the term "*ingenieur*" was used for the first time. Many higher education institutions were founded.

Canals, Locks and Hydraulics

In the post-medieval world, where communication by roads was problematic even in the best of times, navigable rivers presented avenues for travel and transport. The greatest obstacle to the navigation of rivers was their slope, rapids or bars. A lock has been considered the greatest single contribution to hydraulics ever made. The distinguishing feature of the lock is a fixed chamber in which the water level can be varied. Single-gate locks were difficult to use, and therefore they were replaced with multiple-gate locks subdividing stream elevation thus reducing a fall between barriers. A lock chamber is separate from the rest of the canal (river) by an upper and lower pair of mitre gates, which, in each pair, close against each other at an angle of 18° to approximate an arch against water pressure on the "upstream" side of the gates when the water level on the "downstream" side is lower (Figure 5.2). Leon Baltistą Alberti (1404–1472) took the design of the lock to the type that we associate with the device today.

Hydraulic engineering is concerned with the flow and conveyance of fluids, principally water and sewage. One feature of these systems is the extensive use of gravity as motive force to cause the movement of fluids. This area of civil engineering is intimately related to the design of bridges, dams, channels, canals, locks and levees, and to both sanitary and environmental engineering.

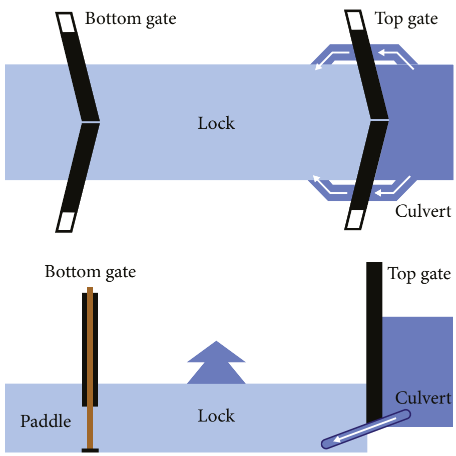


Figure 5.2. A plan and side view of a canal lock

Evangelista Toricelli (1608–1647) determined that the flow velocity varied in depth and posited that the velocity of water in a flowing stream increased with depth was wrong. DuBuat and Chezy reversed this theory correctly observing that the maximum flow velocity was the central channel at surface to mid-depth. Isaac Newton (1642–1727), by formulating the laws of motion and his law of viscosity, in addition to developing the calculus, paved the way for many great developments in fluid mechanics. Using

Newton's laws of motion, numerous 18th century mathematicians solved many problems of a frictionless (zero-viscosity) flow. However, most flows are dominated by viscous effects. Thus, the engineers of the 17th and 18th centuries found *inviscid flow* (the flow of an ideal fluid that is assumed to have no viscosity) solutions unsuitable, and developed, by experimentation, empirical equations thus establishing the science of hydraulics. Late in the 19th century, the importance of dimensionless numbers and their relationship to turbulence was recognized, and dimensional analysis was born. Ludwig Prandtl published a key paper proposing that the flow fields of low-viscosity fluids could be divided into two zones, namely a thin, viscosity-dominated boundary layer near solid surfaces and an effectively inviscid outer zone away from the boundaries. This concept explained many former paradoxes and enabled subsequent engineers to analyze far more complex flows. Assuming, the flow is bounded on one side only, and that a rectilinear flow passes over a stationary flat plate that lies parallel to the flow, the flow just upstream of the plate has a uniform velocity. As the flow comes into contact with the plate, the layer of the fluid actually "*adheres*'" to a solid surface. There is then a considerable shearing action between the layer of the fluid on the plate surface and the second layer of the fluid. The second layer is therefore forced to decelerate (though it is not quite brought to rest), creating a shearing action with the third layer of the fluid, etc. As the fluid passes further along the plate, the zone, in which the shearing action occurs, tends to spread further outwards. This zone is known as the "boundary layer". The flow outside the boundary layer is free of shear and viscous-related forces; thus, it is assumed to act like an ideal fluid. Intermolecular cohesive forces in the fluid are not great enough to hold it together. Hence, the fluid will flow under the action of the slightest street and the flow will continue as long as stress is present. The flow inside the layer can be either viscous or turbulent.

CHAPTER 6

Development of Metallic Materials

This chapter will help in

- presenting the birth of metallurgy;
- discussing the development of the metallurgy of non-ferrous metals;
- describing the development of powder metallurgy;
- dealing with the production of amorphous alloys;
- explaining the development of the production of ferrous metals;
- introducing technology for carbon capture and storage (CCS);
- getting acquainted with iron smelting in Lithuania.

Alchemy – a Protoscience of Modern Chemistry and Technologies

Alchemy is recognized as a protoscience that contributed to the development of modern chemistry, medicine and technologies. Alchemy covers several philosophical traditions spanning some four millennia. Alchemists developed a structure of basic laboratory techniques, theory, terminology, and experimental method, some of which are still in use today (see Table 6.1).

Practical applications of alchemy produced a wide range of contributions to physical sciences and medicine. Alchemists made contributions to the „chemical" industries of the day – ore testing and refining, metalworking, production of gunpowder, ink, dyes, paints, cosmetics, leather tanning, ceramics, glass manufacture, preparation of extracts, liquors, and so on. Alchemists contributed distillation to Europe. The attempts of alchemists to arrange information on substances, so as to clarify and anticipate the products of their chemical reactions, resulted in early conceptions of chemical elements and the first rudimentary periodic tables (Figure 6.1). They learned how to extract metals from ores, and how to compose many types of inorganic acids and bases.

Table 6.1. Basic alchemy symbols

Basic elements	Seven planetary metals	Mundane elements
Earth 🜃	Gold dominated by Sol ☉ ☼	Antimony ♁
Water 🜄	Silver dominated by Luna ☽	Arsenic 🜺
Air 🜁	Copper dominated by Venus ♀ (also: 🜠)	Bismuth ♉
Fire 🜂	Iron dominated by Mars ♂	Magnesium ⊛
	Tin dominated by Jupiter ♃	Platinum ☽☉
	Mercury (quicksilver) dominated by Mercury ☿	Potassium 🜪
		Stone 🝊
	Lead dominated by Saturn ♄	Sulfur 🜍 🜍

Alchemical compounds	Alchemical processes
Sal ammoniac ⚹	Calcination (Aries ♈)
Aqua Fortis A.F.	Congelation (Taurus ♉)
Aqua Regia A.R.	Fixation (Gemini ♊)
Spirit of Wine S.V.	Dissolution (Cancer ♋)
Amalgama 🝛	Digestion (Leo ♌)
Cinnabar (Mercury sulfide) 🜓	Distillation (Virgo ♍)
	Sublimation (Libra ♎)
	Separation (Scorpio ♏)
	Incineration (Sagittarius ♐)
	Fermentation (Capricorn ♑)
	Multiplication (Aquarius ♒)
	Projection (Pisces ♓)

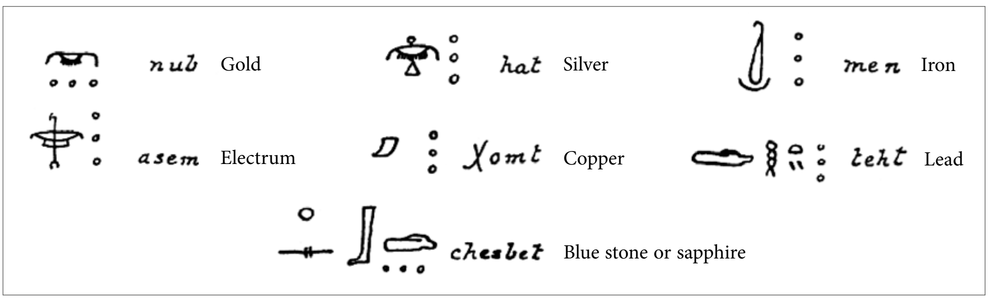


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Figure 6.1. Egyptian symbols for metals

The start of Western alchemy may generally be traced to Hellenistic Egypt, where the city of Alexandria was a center of alchemical knowledge, and retained its pre-eminence through most of the Greek and Roman periods. During the 17th century, practical alchemy started to disappear in favor of its younger offshoot chemistry. The decline of alchemy continued in the 18th century with the birth of modern chemistry, which provided a more precise and reliable framework within a new view of the universe based on rational materialism. The dawn of Western alchemy is sometimes associated with that of metallurgy, extending back to 3500 BCE. The original Egyptian documents on alchemy contain recipes for dyeing and making artificial gemstones, cleaning and fabricating pearls, and the manufacture of imitation gold and silver. The terms „chemia" and „alchemia" were used as synonyms, and the differences between alchemy, chemistry and small-scale assaying and metallurgy were not as neat as in the present day. The decline of European alchemy was brought about by the rise of modern science with its emphasis on rigorous quantitative experimentation.

Non-Ferrous Metals

The birth of metallurgy is shrouded in obscurity, although the weathered crystals of metals might well have attracted the attention of ancient man because of their remarkable colouration. Primitive man would have been most impressed by the malleability of relatively hard, heavy material copper nuggets, which allowed it, unlike wood and stone, to be hammered into a variety of useful shapes.

Cooper

Archaeological evidence indicates that, although lead was also known at a very early date, the first metal to be practically utilized was copper. Small pieces of hammered copper found in Western Iran and Anatolia date from the period between the 9th and 7th millennia B.C. and were made from native, unmelted copper. Apart from gold, silver and other noble metals, copper is the only metal found as a native crystal in metallic form (Figure 6.2). This is because its affinity for oxygen is lower than that of most other common metals, and, as a result, native crystals can be found in weathered out-crops of copper ore. During the later stages of the Neolithic

Figure 6.2. Natural copper nugget

period, more significant metallurgical developments appeared and were subsequently promoted. Native copper was hammered directly into small articles of jewellery or of ritual significance. The weapons from native copper were produced by cold forging with frequent annealing at temperatures below 800 °C. After an approximation to the final shape obtained in this way, the articles were finished by cold working. The craftsmen who produced these artefacts were obviously well aware that copper could be hardened by deformation and softened by annealing. At a later stage, the larger articles such as axe-heads were made by melting native crystals of copper in a crucible and casting molten metal. The beginning of the Copper Age is associated with the emergence of smelting processes that allowed copper to be extracted from its ores. Native copper must undoubtedly have been worked in sites close to the outcrops where metals had been found. It seems logical to assume, therefore, that copper ores were first reduced to metal, quite fortuitously, in the fires where native crystals were annealed at temperatures well below their melting point to soften them after cold forging, or in the furnaces where crystals were melted together.

Melted and cast native copper. Most of larger copper items produced in the Middle East between the 7^{th} and 4^{th} millennia B.C. have a micro-structure that consists of crystals that appear to have grown from melt by throwing out arms into the surrounding liquid. From these arms, secondary branches and spines have grown, thus producing a 'dendritic' structure characteristic of cast metal. During the long period when native copper was being worked and annealed, some of smaller crystals would, inevitably, have been accidently melted.

It seems most probable that when native copper was first intentionally melted, it would have been heated from above by a heaped charcoal fire and encouraged to run together and form a lens-like ingot in clay depression in the ground immediately beneath the fuel bed. Crucible furnaces must soon have been employed to produce items that were cast directly to size. The earliest crucible furnace remains so far identified were found in Israel in a site used between 3300 and 3000 B.C. The earliest vitrified copper-bearing slag was found in Catal Huyuk, Anatolia at a site dating from 7000 to 6000 B.C. The earliest Egyptian artefacts produced from cast and wrought native copper appear to date from the period between 5000 and 4000 B.C. The technique for melting and casting native copper originated in Anatolia and spread rapidly over much of the Middle East and Mediterranean area between 5000 and 4000 B.C.

Smelting oxide and carbonate copper ores. Towards the end of the 4^{th} millennium B.C., the supplies of native copper accessible to the ancient world were incapable of satisfying a rapidly increasing demand. Most of copper articles produced after 3500 B.C. contain substantial quantities of nickel, arsenic, iron or other impurities, which indicates that they had been produced from copper extracted from ore. The Copper Age began when improved copper extraction techniques meant that primitive copper workers were no longer dependent upon the supplies of relatively pure native metal.

The primary ores of copper are invariably complex sulphides of copper and iron, which rarely contains more than 2 per cent by weight of copper. Such deposits were too lean to be exploited by primitive producers. In some places, rich carbonate veins produced by leaching and weathering contained 32 per cent of copper and 26 per cent of iron. The earliest copper workers appear to have extracted their metal from oxide or carbonate ores that could be smelted in primitive furnaces. Methods were evolved, which allowed relatively pure copper to be separated in the molten state from iron and other unwanted materials in the ore. Quartz sand would have been added to such a smelting charge to ensure that most of iron separated into slag. The early smelting furnaces had no tap hole. After smelting, slag would have been broken up to remove the prills that were then remelted together in a crucible furnace.

Smelting techniques appeared to have reached their zenith around 1100 B.C. (Figure 6.3). After smelting operation, slag and metal appear to have been tapped simultaneously from the furnace into a bed of sand where they would have remained liquid for about fifteen minutes thus providing ample time for molten copper to sink beneath slag to form well-shaped ingots.

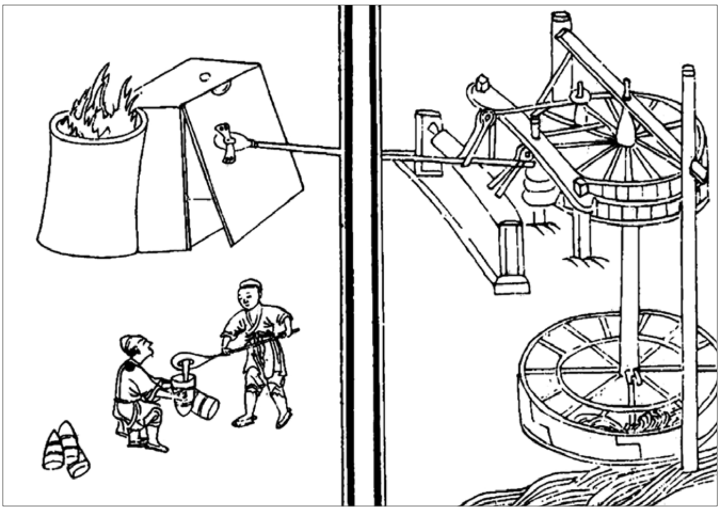

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Figure 6.3. Medievel smelting techniques

It would appear that when 'as smelted' copper was remelted in a crucible furnace in preparation for casting, any surplus iron it contained separated on the surface of melt to form a sponge-like mass permeated by molten copper. This layer would, in all probability, have been skimmed from the surface of melt before it was poured. Iron/copper residues could be consolidated by hot forging and worked to the shape required.

Smelting sulphide copper ores. From the presence of arsenic and other impurities in many artefacts, it must be concluded that much of the copper used was extracted from sulphide rather than oxide or carbonate ores. The sulphide copper ores exploited at the beginning of the Copper Age appear to have been thin, localized and very rich deposits. Due to atmospheric oxidation that occurs on the surface of chalcopyrite outcrops, sulphides are partially converted to more soluble compounds that are slowly leached away. The average copper content of thin secondary enrichment zones sometimes approaches 15–25%.

The beginning of the arsenical copper era is difficult to date with any certainty. Arsenical copper was extensively used by the ancient Egyptians (around 4000 B.C.) as a natural alloy obtained by smelting arsenical copper ore. In the cast condition, arsenical copper is only marginally harder than pure copper, although its hardness increases very rapidly as a result of cold working. This must have been a factor of great importance in the Copper Age when edge tools were invariably hardened and sharpened by hammering.

Tin and Bronze

For well over 1500 years, arsenical copper items were produced in the ancient world. After about 3000 B.C., however, the arsenical content of Middle Eastern copper items begins to decline. Artefacts containing small quantities of tin made their first appearance during the early stages of this decline although the quantities involved rarely exceeded about 2.5%. These low tin bronzes must be regarded as the precursors of the later true bronzes containing 8–10% of tin. They presumably emerged accidentally, either by smelting copper ores that contained tin minerals or by the use of tin-bearing fluxes. The sudden emergence of 8–10% tin bronzes in Sumer between 3000 and 2500 B.C. probably tells that that copper workers suddenly discovered the concept of alloying.

The most significant tin mineral is *cassiterite*, tin oxide SnO_2. Four cassiterite mines have been located in the eastern Egyptian desert. Copper artefacts containing significant quantities of tin were not produced in Egypt before the 2600 B.C., and the true Bronze Age in Egypt, when artefacts containing 8–10% of tin were produced, did not begin until 2000 B.C. The shortage of tin, whether it was caused by natural disasters or by political upheavals, was not prolonged; however, even after about 2500 B.C., the use of 8–10% tin bronzes expanded rapidly throughout the Middle Eastern world. Early Bronze Age developments have also been found in southern China, Thailand and Indonesia. Chinese bronze containing 8% of tin was being produced as early as 2800 BC, apparently independently of the emergence of high tin bronzes in Sumer. Bronze was first made in Italy between 1900 and 1800 B.C. Copper extraction started in Spain in Neolithic times, and, during the third millennium, its copper and precious metal deposits were extensively worked.

Although many arsenical copper artefacts of the Early Bronze Age appear to have been roughly cast to shape in open stone moulds and then forged to size, this practice died out soon after the 8% tin bronze alloy came into general use. Most artefacts were cast almost directly to the finished size moulds; many of those were made from clay into which internal cavities were precisely moulded before firing. Long thin sections, such as the blades of swords and daggers, were cast into thin shell moulds. By 2500 B.C., the Egyptians had developed considerable expertise in the production of hollow copper and bronze statuary. Many large Egyptian statues were cast with an internal sand core, which is still present in some of the found figures. Smaller components, like the spouts of copper water vessels, were undoubtedly made by ”lost wax“ technique obviously mastered by the Egyptian craftsmen before 2200 B.C.

In Europe, between prehistoric times and 500 B.C., bronze technology had developed in a slow and irregular manner. Around 1600 B.C., in late Minoan times, large copper ingots weighing more than 30 kg first appeared, and the usage of bronze throughout Europe and the Mediterranean region began to increase very rapidly. In the Royal Bronze Age tombs of Mycenae, opened 1876, weapons and most of the other metalwork which was found, are now believed to be of Cretan origin. Daggers with bronze blades were decorated by a black compound of sulphur with copper or other metals, which formed a background for lively and naturalistic pictures in gold and silver. The characteristic metal artefact of Crete was the double axe, an object of great ritual significance. Some compositions range from pure copper to ductile 4% tin bronzes, and finally to hard, brittle bronzes containing 18% tin. Arsenical coppers and low tin bronzes gradually were displaced by bronzes containing 8–10% tin. The emergence of the full Bronze Age, in the middle of the 2nd millennium B.C., is characterized by the almost universal employment of this type of alloy.

At a very early period, the Greeks attempted to standardize the composition and mechanical proportions of bronze. The alloy being made had twelve parts, eleven of copper to one of tin. It seems that the Greeks were aware of the weakening effect of lead on bronze and preferred the binary tin-copper alloy for structural items. From the weapons found in the Mycenaean tombs, it is evident that metal workers, even in the 16th century B.C., knew exactly how much tin could be put into bronze before the alloy became too brittle for hot or cold forging. They improved the hardness and elastic limit of their bronze sword blades by judicious cold working and knew that ductility could be restored, when required, by annealing.

The Roman world life stimulated metallurgical demand and encouraged the rapid diffusion of improved processes and techniques throughout the Empire. The primary metallurgical requirement for the army was iron. Bronze, in quite formidable quantities, was needed, however, for both military and non-warlike purposes, and coinage utilized vast quantities of gold, silver and copper. Roman civilization required large quantities of lead for plumbing, and the main metallurgical innovation of Roman times, however, was brass first used for coinage when tin became expensive. Mines appear to have reached their maximum output between the 2nd and 3rd centuries A.D. Most of the statuary of the Greco-Roman world was cast in bronze. In the 4th century, Greece leaded tin bronzes were generally used although zinc found its way into many of later Roman bronzes. The characteristics of these alloys covered their low melting points, and wide freezing ranges facilitated mould filling. Zinc would also have deoxidized the alloys and assisted greatly in the reduction of casting porosity.

The horses of San Marco (Venice) were cast by the indirect lost wax process. The wall thickness of the horses of San Marco varies between 7.5 and 10.5 mm, and originally they were gilded. Before the application of the gold leaf, the copper surface had been treated with an aqueous solution of gold mercury amalgam. When this had reacted chemically on the surface of copper to produce the required amalgamated layer, work would have been heated gently and the gold leaf applied. The castings from copper are easier to gild than bronze that might well have contained substantial quantities of lead.

Lead and Silver

Native lead is rarely found. Its sulphide (*galena*), upon which most lead ores are based, is fairly abundant and would have been instantly recognizable by primitive man. Lead is comparable to copper in its antiquity and was probably obtained in its metallic form by reducing the ore inside it. Galena usually contains between 0.03 and 0.1% by weight of silver, and the economics of lead extraction are generally determined by the value of the silver present in the ore. Metallic lead has been found in Iraq in about 6000 B.C. Lead artefacts have also been obtained from the 4th millennium sites in Iran and Egypt. The earliest evidence shows lead being used for aesthetic rather than practical purposes. Silver objects of the late 4th millennium have also been unearthed in Palestine, Ur and Mesopotamia as well as in various sites in Asia Minor. It seems probable that they emerged as a by-product of lead refining operations.

Lead artefacts of the 3rd and 2nd millennia B.C. include spindle whorls, weights for fishing and wire. Lead weights, based on a unit module of 61g, have also been found throughout the Aegean in Middle and Late Bronze Age contexts. Shapeless lumps of lead were found in the layer of the ruins of Hissarlik (Troy I), which corresponds to the period between 3000 and 2500 B.C.

Numerous deposits of silver-rich lead are known in the Aegean area. In the 5th century B.C., silver from the mines in Laurion provided wealth needed to support the Athenian Empire. Silver recovered from early Cycladic tombs in Greek islands located in the southern part of the Aegean Sea, appears never to have been melted down for recycling. Silver artefacts in the Cycladian tombs, which therefore appear to have been made from virgin metal and almost immediately interred, have isotope abundance ratios thus making it possible to very precisely identify the lead mines from which silver was obtained.

Rio Tinto is a part of the Iberian Pyrite Belt, a mineral deposit that stretches from Spain into Portugal. It is one of the largest known mining complexes in the ancient world. The Romans took over Rio Tinto in 206 B.C., after defeating and expelling the Carthaginians who had occupied the region since about 535 B.C. With a technical knowledge of Rome's military engineers and the availability of slave and convict labour, the Roman operations grew colossally, peaking from A.D. 70 to 180. Their magnitude far exceeded anything that came before. During that period, Rio Tinto was the largest silver and copper mining operation in the Roman Empire.

A chalcopyrite ore, which appears to have been worked for silver, have contained around 40 g of silver per tonne. Since Rio Tinto copper contained little lead, it is evident that lead oxide, or lead ore, must have been added to the smelting charge to take up silver. The chalcopyrite ore might have been fused directly in a small shaft furnace to form copper sulphide matte, from which silver was absorbed by molten lead, and the obtained alloy thus would then have been cupelled to extract silver.

During the long period, silver was extracted in Laurion (Greece). This region was the largest centre of mining in ancient times. The systematic exploitation of mineral wealth in the area began in 3200 B.C. and continued without a break until the end of the 6th century B.C. In the Bronze Age (2800–1100 B.C.), the mines of Laurion supplied all great cultures of the Aegean Sea (the Cycladic, Minoan and Mycenaean) with silver, lead and copper. The ore was particularly rich and is reputed to have yielded between 1200 and 1400 g of silver per tonne of the ore. During the 4th century, mines and mining rights in Laurion still belonged to the Athenian state. Industry declined very rapidly towards the end of the 4th century when the silver coinage standard of the Athenian state was superseded by the gold standard introduced by the Macedonian Kings.

a) b)

Figure 6.4. Silver and bronze in Lithuanian territory: a – the earliest Lithuanian money – silver bars (12^{th}–14^{th} century, Kernave excavation); b – hoard (Lithuanian long ingots, brooches, wristlets, etc.), (14th century, (Stakliškės, Lithuania, National Museum of Lithuania, The Old Arsenal)

In Lithuania, coins were usually melted down into a silver alloy, out of which the first Lithuanian money – the Lithuanian "long" – was fashioned (Figure 6.4). Notches in the alloy were designed to determine the quality of silver. 104 g equalled the weight of an average Lithuanian "long".

Some of produced lead was used to a limited extent for sealing and holding in place, and iron clamps were employed for preventing relative movement between large marble building blocks. It must be concluded that the value of lead at that time did not always justify the cost of recovery from waste.

Roman municipal life was characterized by the use of water. Lead piping was extensively used for water distribution and plumbing. Also, lead piping from the ruins of Pompeii varied in outside diameter from 30 to 40 mm and had a wall thickness of 5 mm. The method of production was very simple. Lead sheets 6 mm thick were cast on a flat stone surface, cut to size and then wrapped around an iron mandrel in such a way that a narrow longitudinal gap was left along the pipe. Into this gap, molten lead was poured at such a temperature that it was able to remelt the edges of the cast sheet before solidification. The pipes were then hammered on the mandrel over the whole of their cast surface to reduce their thickness to 5mm and to increase their diameter.

Lead from Britain contained very little silver, and lead mines were unable to compete with the deposits of Rio Tinto. Desilverized Roman lead rarely contains less than 0.007% of silver. Such levels, although considerably higher than those present in modern commercial lead, probably marked the feasible limits of cupellation in Roman times.

The origins of *brass* are almost as uncertain as those of bronze. Zinc, in its natural deposits, is rarely associated with copper. Certain ancient sources of copper, however, contained enough zinc to ensure that, when smelted, they produced low zinc brasses rather than copper. Thus, in Cyprus, many copper artefacts have been found containing zinc in quantities up to 9%.

Brass, containing only copper and zinc, dating from the period between 2200 and 2000 B.C. has been found in China where the deposits of copper ore containing high concentrations of zinc occur. Bronzes containing more than 10% of zinc were not made in China until about A.D. 1200, at which time zinc in metallic form must have become available and would have been deliberately incorporated into alloys.

Zinc is rarely found in significant quantities in early Hellenistic bronzes. Historical evidence indicates that the 6th century Greeks started using significant quantities of brass, which at that time, because of its colour and rarity was held in high esteem.

In the 4th century, brass was still being imported into Greece from Asia Minor. Irrefutable evidence that metallic zinc was known and used in the ancient world is also available in the form of a sheet of zinc, 6.5 cm long, 4 cm wide and 0.5 mm thick, dating from the 3rd or 2nd centuries B.C. and found in 1949 during an excavation of the Athenian agora. A statuette from the Agora was also made of 99% of pure zinc.

By the early years of the 1st century A.D., Rome was making brass on a large scale by the cementation process, even though metallic zinc was known and used in Athens three hundred years earlier.

The Etruscans were producing statuary from brass, rather than bronze, as early as in the 5th century B.C. (The Etruscans lived in what is modern Tuscany (Italy). Those dating from the 3rd or 2nd centuries contain around 12% of zinc. Egypt does not appear to have used brass before 30 B.C.

The Romans provided the following process. Calamine was ground and mixed in suitable proportions with charcoal and copper granules. This mixture was placed in a small crucible and carefully heated for some time to a temperature sufficient to reduce zinc in the ore to the metallic state, but not high enough to melt copper. The vapour of volatile zinc permeated copper fragments thus turning them to brass. Temperature was then raised; brass melted and was poured from the crucible into moulds. Such a process, operated at 1000°C, will produce brass containing about 28% of zinc.

The usage of brass in the 1st and 2nd centuries A.D. increased rapidly, particularly for manufacturing light decorative metalwork such as brooches, rings and horse trappings.

The original home of brass manufacture seems to have been Asia Minor where zinc in metallic form was being produced and used for brass making in the 4th century B.C. Technique for zinc manufacturing is described in Indian sources. *Sphalerite* (Zn, Fe)S was ground into fine powder and roasted to convert zinc sulphide to

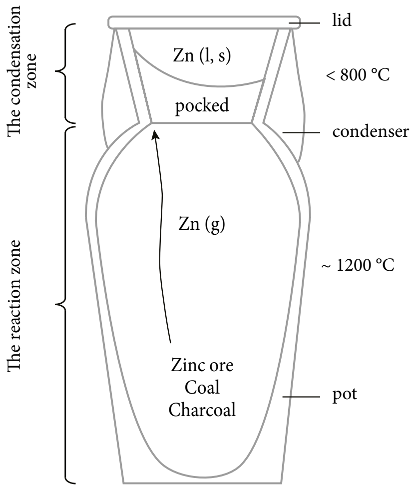

Figure 6.5. Zinc distillation

the oxide that was then rolled into balls with organic materials and fluxes. The charge have been heated to temperatures between 1050° and 1150 °C, which is considerably higher than the boiling point of zinc at atmospheric pressure (913 °C). Zinc vapour condensed on the outlet nozzles of the retort (distillation process) (Figure 6.5). From the size of waste production, it is clear that the vast quantities of zinc oxide must have been produced, and brass manufacturing appears to be the only technology that could absorb this output, whether zinc oxide was directly converted to metal.

Metallic zinc appears to have been a commodity well known to Chinese at an early date. Zinc coins were first struck during the early years of the Ming Dynasty (1368–1644). Some of later Ming bronzes also contain more than 30% of zinc, a level that could have been obtained only by the incorporation of metallic zinc in melt.

Zinc and its Applications

The currently available evidence indicates that certainly brass and probably metallic zinc were first made in Asia Minor before 800 B.C., and that by 700 B.C., golden-coloured brass was being imported by the Hellenes. The alloy may have been exported from Asia Minor to the Far East as early as 200 B.C. via Punjab (India). Brass manufacturing in this region of India started in the 1st century A.D. The manufacture of a natural brass alloy in China and the usage by Chinese of a complex copper-nickel-zinc ore for bronze manufacturing appeared around 2000 B.C.

Zinc ingots of Chinese manufacturers began to appear on European metal markets towards the end of the 16th century. Chinese zinc soon found extensive usage for producing golden coloured brass of the gilding metal type used for cheap jewellery. The availability of zinc in metallic form also made it possible to prepare the brass of a low melting point, containing 40–50% of zinc, which were then used as brazing alloys for joining copper and steel components.

European zinc manufacturing on a small scale seems to have started at Rammelsburg (Rammelsberg lies in the Harz Mountains (Lower Saxony), although the output was consumed locally and the initially employed techniques are unknown. In Britain, calamine deposits were discovered in the 16th century. *Calamine* is either a mixture of (ZnO) with about 0.5% of ferric oxide (Fe_2O_3) or a zinc carbonate compound. Zinc ore from this deposit contained less lead and was superior in quality. Britain brass industry declined very rapidly, since the Parliament refused either to reduce the duty on imported copper or to protect English manufacturers against the importation of cheap imported brassware. After 1650, hundreds of tonnes of high grade zinc ore were shipped to European brass-making centres to be turned into brassware, subsequently to be sold in English markets at prices well below those of the domestic product.

British non-ferrous metal industry began to revive in the 1670s when coal-fired reverberatory furnaces for refining lead and copper were first developed. Sulphide copper ores had been melted with lime in a reverberatory furnace to remove siliceous impurities. Iron and sulphur remaining in the purified matte thus attained was then gradually removed by reverberatory melting under oxidizing conditions until crude copper in metallic form began to separate from melt.

Metallic zinc was then being imported from the Far East at prices too costly for the manufacturers of sheet brass, although it was in great demand for producing low zinc content and attractive golden colour brass used for making cheap jewellery. It was also employed for making brazing alloys containing 40 per cent or more of zinc, which could not be obtained by cementation. English producers stared to produce metallic zinc from English calamine.

The essence of a new reduction process was to condense vapour rapidly into metal in the complete absence of air. For this purpose, a vertical retort was used. An iron tube led from the base of the reaction crucible to a cold chamber below the floor with its end being sealed by immersion in a bowl of water. This ensured that zinc vapour did not encounter significantly oxidizing conditions before it condensed and settled as granules below the water surface. 6 reaction crucibles, approximately 1m in height, were arranged in each furnace. The distillation process took about seventy hours, during which time around 400 kg of zinc were obtained from six retorts.

The horizontal zinc retort process, introduced in 1807, was far simpler and more economical to operate. Most of zinc produced in Europe and the United States between 1820 and 1940 was made in this way.

Attempts to obtain zinc directly from calamine were being made in Continental Europe, to be more precise, in Liège (Belgium) where the first commercially successful zinc reduction process was introduced. A rolling mill was producing zinc sheets 1.5 m long by 41 cm wide by 1811, and by 1813, the roof of St Paul's Cathedral in Liège had been sheathed in zinc.

Soon after the Belgian process emerged, Germany, which had large deposits of zinc ore in Silesia, developed a zinc furnace system that used retorts much larger and stronger than those used in Belgium. This approach was possible because

Silesian ore had coarse and porous texture. Regenerative furnaces became almost universal after 1900, and by 1920, fuel consumption had been reduced to 1 tonne of coal per tonne of zinc distilled instead of 24 tonnes of coal to produce a tonne of zinc in 1851. The first practical continuous zinc reduction process, introduced in the 1930s, utilized a large vertical retort built of carborundum bricks fed by the briquettes of roasted ore mixed with bituminous coal. The output of such a unit, around 9 tons of zinc per day, should be compared with 27 kg produced by the best horizontal retort of that period.

A modern zinc blast furnace is charged with a mixture of lead and zinc oxides reduced by hot coke as in a conventional iron blast furnace (Figure 6.6). The lead and slag produced by this reaction sink to the hearth of the furnace where they are tapped and separated. Zinc leaves the reaction zone as vapour and, mixed with carbon monoxide formed by the reduction process, is carried into the condensing chamber to be absorbed by the streams of molten lead. While dissolved in this way, it can safely be exposed to the atmosphere without a serious risk of oxidation. Lead is then slowly cooled and the layer of pure solid zinc, which separates out on its surface, is skimmed off. The remaining molten lead is then pumped back to the condenser. Each of these zinc blast furnaces can produce 60,000 tonnes of zinc per year. At least ten blast furnaces are in operation throughout the world.

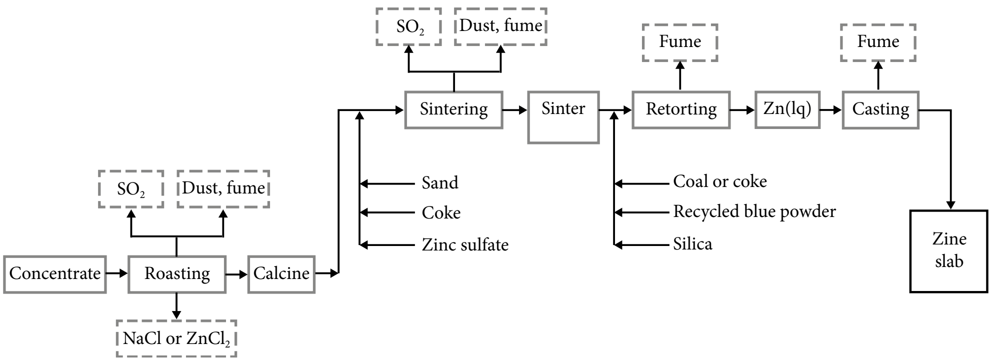

Figure 6.6. The pyrometallurgical smelting zinc process

As a new exotic metal that had only recently become an article of commerce, zinc was of great interest in the 18th century Europe. In 1819, French chemist Thenard began to appreciate that metallic corrosion was caused by electrochemical effects. Thus, an attempt to apply zinc to the surface of a rolled iron sheet for iron protection against corrosion was made. This new process was christened as „galvanizing". Many components such as iron sheets, chains, nails and wire were treated.

During the 19th century, a zinc sheet became a popular roofing material. Unlike corrugated iron, the zinc sheet had no structural stiffness, although when properly applied to a suitably supporting surface, it provided excellent protection. After about 1850, the zinc sheet was extensively employed for cladding the roofs of railway stations and other buildings.

Although the hot dip galvanizing process afforded excellent protection against corrosion, it had the disadvantage of increasing significantly the dimensions of the articles being treated and could not, therefore, be easily applied to the routine coating of small components such as nuts and bolts where close dimensional control was required. Galvanizing, normally carried out between 430 and 540 °C, also had a tendency to soften cold-rolled and heat-treated components. *Sherardizing*, a process introduced between 1900 and 1904, avoided many of the disadvantages of galvanizing. It was a cementation process made possible by the high vapour pressure of zinc. Steel or iron components to be treated were packed in sealed retorts in contact with zinc dust and heated for 4–5 hours at a temperature of about 375 °C, which is well below the melting point of zinc. At this temperature, zinc vapour soon saturated the interior surface of the retort and diffused well below the surface of the ferrous components being treated thus providing a corrosion-resistant surface without a serious change in external dimensions.

The Emergence of Nickel

The whitish alloy, *paktong,* found its way to Europe from the Far East since the 16th century. It was hard, whitish in colour and, relatively ductile in the cold condition. Being resistant to corrosion and free from the unpleasant taste of brass, it was much favoured for the manufacture of musical instruments. Although nickel metal was identified in 1751, the composition of paktong was unknown until 1822, when analysis showed that it was an alloy based on copper, nickel and zinc.

It is uncertain when Chinese first produced *paktong*, although they made the extensive use of cupronickel coinage as early as in the 1st century A.D. During the 18th century, cupronickel alloys from China contained approximately 30% of nickel and 20% of copper.

It was difficult to manufacture paktong in Europe since a suitable source of nickel had not yet been identified. Complex cobalt-nickel-iron-arsenic ores were, however, used for manufacturing cobalt blue and other pigments for ceramic industry. In 1824, a method for extracting reasonably pure nickel from the waste products of the cobalt-blue process was devised.

Nickel silver made in Europe was generally known as *Neusilber* or *Argentan*. In 1850, world nickel consumption was approximately 100 tonnes per annum.

The usage of the metal increased slowly until 1889 when the invention of nickel steel created a demand that was difficult to satisfy. Nickel usage worldwide increased slowly in the second half of the 19th century. Several European states used nickel silver as a coinage alloy. In 1879, the content of silver used for coins was omitted and the 75:25 copper-nickel alloy was adopted.

Pure unalloyed nickel was regarded as a type of metal that was difficult to melt, cast and fabricate. Although no serious problems were encountered in melting the metal, the obtained ingots were often too brittle to be worked either hot or cold. This problem was solved by adding an alloy of iron, manganese and carbon to the melt. By that time, it had been found that magnesium was far more effective than manganese in improving the ductility of nickel and that lower concentrations were required. The secret of success was to treat molten nickel with metal such as magnesium having higher affinity than nickel for sulphur.

In 1905, it was suggested that if the copper nickel iron converter matte then being received by the refinery was converted directly to metallic form instead of being laboriously separated into its constituents, it would yield a ductile, corrosion-resistant cupronickel alloy that could be of considerable value to chemical industry. Monel metal (was named after Ambrose Monell) contained approximately 67% of Ni, 28% of Cu and 5% of Mn. Monel metal is still widely used where a ductile corrosion-resistant alloy is required. In its colour, mechanical properties and ductility it resembles nickel and is resistant to anhydrous ammonia, sea water and a wide range of organic chemicals. It is used for marine propellers and applied to much marine hardware. Between the two world wars, it was also used extensively for manufacturing kitchen sinks and for producing the heads of golf clubs. From such applications, it has been largely displaced by austenitic stainless steel that is significantly cheaper.

Today, nickel oxide is turned into metallic form either by electrolytic refining or by the *carbonyl process* (a process of extracting and purifying nickel whereby nickel carbonyl is first formed by the reaction of the reduced metal with carbon monoxide, and then nickel carbonyl is decomposed thermally, resulting in the deposition of nickel). Monel metal, for the reasons of control and economy, is no longer a natural alloy, but is produced by alloying copper, nickel and iron together in the desired proportions.

Aluminium and Magnesium

In the middle of the 19th century, some known metals could not be obtained in metallic form because of their high chemical activity.

Particularly stable metallic oxide that could not be reduced was extracted from aluminum oxide. Danish chemist Hans Christian Oersted, in 1825, described the method of reducing aluminium chloride to metallic form with a mercury amalgam of potassium. Mercury from amalgam was subsequently removed by distillation, leaving behind a grey powder that was described as aluminium.

Later, the manufacturing process was switched to mineral bauxite. Since this material contained a good deal of iron, it required extensive purification after which it was mixed with sodium chloride and carbon and reacted with chlorine at a good red heat to sodium aluminium trichloride ($NaAlCl_3$). This, being vapour, distilled away from the reaction zone, and condensed as a crystalline deposit at temperatures below 200°C. Double chloride was then mixed with *cryolite* (Na_3AlF_6, sodium hexafluoroaluminate) and reacted with metallic sodium in a reverberatory furnace. The furnace charge consisted of double chloride, cryolite and sodium. The function of cryolite was to act as a flux and dissolve alumina on the surface of the produced aluminium globules, so that they were able to coalesce. It also produced slag that was fluid and light enough to let the reduced globules of aluminium sink to the base of the reaction bed and unite. After three hours, the reaction was completed and the products settled down into two layers at the bottom of the furnace. The upper layer is white fluid slag free from aluminium. Molten aluminium from the lower layer then run into a red hot cast-iron ladle from which it was cast into ingots.

The possibility of using cryolite not merely as a flux for the reaction process, but as the primary raw material for aluminium production, was first investigated around 1856.

The first serious attempts in the United States to obtain aluminium by the electrolysis of fused salts appeared about 1892.

In 1886, it was established that alumina could be dissolved in fused cryolite „like sugar in water“ and that the alumina/cryolite solution thus obtained was a good electrical conductor. By dissolving 15–20 per cent of alumina in cryolite, was obtained a bath the melting point of which was between 900 and 1000 °C, which was the temperature its electrical conductivity was high enough to permit electrolysis. The cast statue of Eros in Piccadilly Circus, London erected in 1893, appears to have been made from electrolytic aluminium.

Aluminium developed very rapidly in Europe where it was well appreciated as light and corrosion-resistant metal that had also very high thermal conductivity. This encouraged its use for cooking utensils. The first authenticated use of aluminium as a roofing material is provided by the dome of the Church of San Gioacchino in Rome, which was roofed with aluminium in 1897.

The Hall-Héroult process was invented independently and almost simultaneously in 1886 by the American chemist Charles Martin Hall and the Frenchman Paul Héro. The remarkable coincidence is that two young inventors, several thousand miles apart, should independently, at the same age arrived at the same technical solution. Even more remarkably, Hall and Heroult were both born in 1863 and both died in 1914. The electrolytic process of aluminium production introduced in 1886 (Hall process) used a bath of fused cryolite in which pure alumina was dissolved. This bath was maintained at temperature in the molten condition by the passage of a electrolysing current.

The Hall-Héroult process is the major industrial process for the production of aluminium (Figure 6.7) and involves dissolving alumina in molten cryolite and electrolysing the molten salt bath to obtain pure aluminium metal. In the Hall-Héroult process, alumina, Al_2O_3, is dissolved in an industrial carbon-lined vat of molten cryolite, Na_3AlF_6. Aluminium oxide has a melting point of over 2000 °C while pure cryolite has a melting point of 1012 °C. Some aluminium fluoride, AlF_3, is also added into the process to reduce the melting point of the cryolite-alumina mixture.

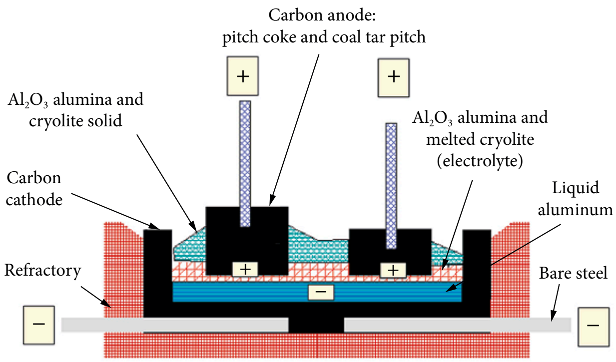

Figure 6.7. Principles of the Hall-Héroult Process

The molten mixture of cryolite, alumina and aluminium fluoride is then electrolyzed by passing a direct electric current through it. The electrochemical reaction causes liquid aluminium metal to be deposited at the cathode as a precipitate while oxygen from alumina combines with carbon from the anode to produce carbon dioxide, CO_2. An electric potential of 3–5 V is needed to drive the reaction. While solid cryolite is denser than solid aluminium at room temperature, the liquid aluminium product is denser than molten cryolite at temperatures around 1000 °C, and aluminium sinks to the bottom of the electrolytic cell where it is periodically taken out with the help of a siphon operating with a vacuum.

Bauxite, the most important ore of aluminium, contains only 30–54% of aluminium oxide, (alumina), Al_2O_3, whereas the rest covers a mixture of silica, various iron oxides and titanium dioxide. Aluminium oxide must be purified before it

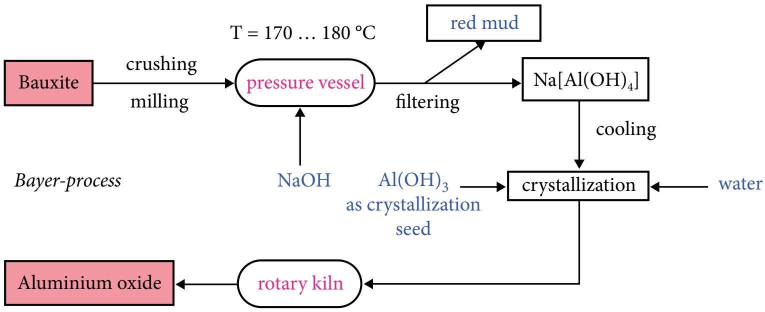

Figure 6.8. The Bayer process of producing alumina from bauxite

can be refined to aluminium metal. This is achieved through the use of the Bayer chemical process in alumina refineries (Figure 6.8). Aluminium oxide is released from other substances in bauxite in a caustic soda solution filtered to remove all insoluble particles. Aluminium hydroxide is then precipitated from the soda solution, washed and dried while the soda solution is recycled. After calcination, the end-product, aluminium oxide (Al_2O_3), is a fine grained white powder. Four tons of bauxite are required to produce two tons of alumina which, in turn, produces one tonne of aluminium at the primary smelter. Primary aluminium is produced in reduction plants (or "smelters") where pure aluminium is extracted from alumina by the Hall-Héroult process. Approximately 20,000 kWh of electrical power are needed for producing one ton of aluminium. Aluminium smelters are usually sited where inexpensive hydroelectric power is available.

Magnesium is the fourth most common element on Earth as a whole (behind iron, oxygen and silicon), making up 13% of the mass of our planet. The metal is not found naturally on Earth, as it is highly reactive (though once produced, it is coated in a thin layer of oxide, which partly masks this reactivity). The main sources of magnesium compounds include seawater (magnesium chloride, $MgCl_2$) and minerals dolomite ($CaCO_3{\cdot}MgCO_3$), magnesite ($MgCO_3$) and carnallite ($KCl{\cdot}MgCl_2{\cdot}6H_2O$).

The metal itself was first produced in England in 1808 using the electrolysis of a mixture of magnesium oxide (Mgo) and mercuric oxide (HgO). Bunsen (1852) from Heidelberg (Germany) devised a method of producing it continuously by the electrolysis of fused anhydrous magnesium chloride. In 1857, the metal was manufactured for the first time in quantities large enough to allow its properties to be evaluated. Magnesium metal was first obtained on a production scale by the electrolysis of fused *carnallite* ($KMgCl_3{\cdot}6(H_2O)$ in 1882. Magnesium is flammable, burning at a temperature of approximately 3,100 °C, and the autoignition temperature of magnesium ribbon is approximately 473 °C. By this time, it had been also known that magnesium was the metal burned readily to produce a very intense white light so that it could therefore be of value to the new science of photography and pyrotechnics. Magnesium, with the density of only 1.74 g/cm^3, could be of great value, particularly in the aeronautical field (aluminium density is 2.70 g/cm^3). The first hurdle to be overcome was to develop magnesium free from corrosion.

After 1905, attempts were made to evolve a magnesium bath using an analogy of the Hall/ Héroult method in which magnesium oxide was dissolved in fused salt which, like cryolite, did not participate in the electrolytic process that developed utilized magnesium oxide dissolved in a mixture of fluorides. A cathode of molten aluminium was used and the cell produced a magnesium-aluminium alloy. Magnesium oxide dissolved in a bath of fused sodium, barium and magnesium fluorides and electrolysed allows obtaining the magnesium of 99.99% purity. The fundamental obstacle in such processes is the low solubility of magnesia in fluoride baths; thus, it is difficult to ensure its constant replenishment.

Magnesium, unlike aluminium, is fairly volatile metal that, at high temperatures, behaves more like gas than liquid. It can, therefore, be obtained from some of its compounds by thermochemical reduction processes to which alumina, for example, would not respond. Because of its volatility, magnesium is able to vacate the vicinity of a high temperature reaction zone as soon as it is liberated. There are processes capable of producing high quality magnesium directly from magnesite without consuming the vast quantities of electric power, although the total energy requirements of such processes are usually greater than electrolytic technique. The ferro-silicon reduction process depended upon the reaction that occurs between dolomite and silicon:

$$2MgO \cdot CaO + Si \rightarrow 2Mg + (CaO)_2 \cdot SiO_2$$

In this reaction, the tendency of silicon towards reducing magnesia is assisted by the high affinity of lime for silica. Liberated magnesium that escapes from the reaction zone as a superheated vapour is rapidly condensed directly to the solid state. Resistor furnaces provided temperatures in the reaction zone of 1400°C and the furnace was run in a vacuum. Ferro-silicon was found to be the cheapest and most effective reduction agent, and magnesium liberated from the reduction zone condensed directly to the solid state in a cooled receiver at the end of the furnace. The furnaces of this type produced about a ton of magnesium a day. The ferrosilicon process was also very successfully used by the Allies during the World War II, more particularly in North America.

Magnesium can be obtained by a direct reduction in magnesite with carbon, although at atmospheric pressures, the reduction does not commence at temperatures below about 2000°C. Magnesium escapes from the reaction zone as a superheated vapour. To prevent from reoxidation, it must then be rapidly cooled, when a fine pyrophoric powder is obtained. Magnesium vapour leaving reaction furnaces can be condensed in a curtain of oil. The lethal potentialities of this type of the mixture were soon appreciated. Under the trade name of Napalm, it was soon in great demand.

Today, two principal magnesium extraction processes like *silicotermic process* and *electrolytic process* can be identified. The *Silicotermic process* involves reducing molten magnesium oxide by ferrosilicon under low gas pressure at a temperature of about 1400 °C. Metallic magnesium, formed in the process, evaporates and then condensates away from the hot region. Condensed magnesium having the purity of 99,95% is then remelt and cast.

The first stage of the *electrolytic process* is the preparation of magnesium chloride feed followed by the electrolytic dissociation of magnesium chloride. Industrial cell feeds consist of a mixture of dehydrated magnesium chloride, partly dehydrated magnesium chloride or dehydrated carnallite. Dehydrated magnesium chloride is prepared by one of the following two methods: the chlorination of magnesium oxide (IG Farben process) or dehydration of magnesium chloride brines (Figure 6.9). The electrolytic cell consists of a brick-lined vessel divided into anode and cathode compartments by a semi-wall. Air- or water-cooled graphite plate anode and steel cathode are submerged in an electrolyte composed of alkaline chlorides with the addition of magnesium chloride under the operating temperature of 680 °C to 750 °C. Magnesium chloride decomposes in the electrolytic cell according to the reaction

$$MgCl_2 \rightarrow Mg + Cl_2(gas)$$

Metallic magnesium is formed in the cathode and floats up (is lighter than an electrolyte) collecting in the cathode compartment. Chlorine, which is a by-product of the process, is collected in the anode compartment.

Magnesium is the lightest of all structural metals and offers a wide range of opportunities for transportation industries. While high-pressure diecast magnesium components have many applications, interest is now spreading to the implementation of gravity casting and advanced casting technology or to the use of wrought materials. These more advanced casting techniques offer the possibility of obtaining more homogeneous castings and better properties. Automotive applications around the globe utilizing magnesium today, such as cross car beams, seat frames, steering column brackets, vehicle front end structures, door closures, engine cradles and power train components are all examples of magnesium die casting.

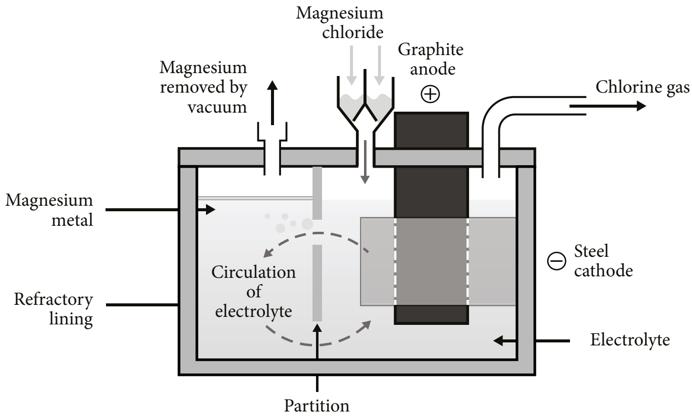

Figure 6.9. Electrolysis of magnesium chloride

The main applications of magnesium include a component of aluminium alloys in die casting (alloyed with zinc), removing sulfur in the production of iron and steel and producing titanium.

Magnesium, in the form of turnings or ribbons, is useful for preparing reagents in organic synthesis, and, as an additive agent is employed in conventional propellants and production of nodular graphite in cast iron. As a reducing agent, it can be applied to extracting uranium and other metals from their salts. As a galvanic anode, magnesium is used for protecting underground tanks, pipelines, buried structures and water heaters. Alloyed with zinc, magnesium is used for making a zinc sheet used in photoengraving plates in printing industry, dry-cell battery walls and roofing. As a type of metal, the principal use of this element is an alloying additive to aluminium with aluminium-magnesium alloys being employed mainly for beverage cans.

Age Hardening Alloys

Precipitation hardening, also called age hardening, is a heat treatment technique used for increasing the yield strength of malleable materials, including most structural alloys of aluminium, magnesium, nickel, titanium and some types of stainless steel. At the beginning of the 20th century, they were the only alloys intentionally strengthened by heat treatment. In 1909, however, it was noted that an aluminium alloy had the remarkable ability to harden slowly at room temperature after having previously quenched from a temperature just below its melting point.

Aluminium alloys. Between 1909 and 1911, this 'age-hardening' invention made in Durener-Metallwerke, Duren (Germany), was subsequently marketed under the trade name *Duralumin*. During World War I, large quantities of age-hardened aluminium alloys were used for Zeppelins and aircrafts. It was found that the strengthening reaction could be slowed down, very significantly, by refrigeration. This made it possible to quench aluminium alloy rivets and store them in the soft condition in a refrigerator. The age hardening process began only after the rivet head had been closed after insertion into the aircraft structure.

The outstanding result was the development of '*Y*' alloy (4% of Cu, 2% of Ni, and 1.5% of Mg) that retained its strength to moderately high temperatures and was extensively used for pistons and cylinder heads. "*Y*" alloy was originally used in the cast form and later in the forged form for aircraft engine use.

The first rational explanation of the age hardening process in light alloys such as Duralumin and Y alloy was provided in 1919. It was found that when an alloy, such as Duralumin, was heated to temperatures close to its melting point, most of the alloying constituents were taken into solution by the matrix. Quenching retained dissolved metals in this supersaturated solid solution which was, however, somewhat unstable at room temperature: small crystals of various intermetallic compounds were slowly precipitated and appeared below a certain critical size (invisible to the optical microscope). They strained and distorted the aluminium lattice and acted as mechanical obstacles, which inhibited plastic flow in the alloy and increased its strength. A wide range of other alloys were soon identified in which precipitation could be induced by ageing at elevated temperatures, and new precipitation hardening alloys of all types were rapidly developed.

Beryllium and beryllium alloys. Beryllium was first identified as an element in 1828. Attempts to produce beryllium by the electrolysis of a fused bath of beryllium chloride were first made in 1895. Relatively pure beryllium for the first time was yielded in 1921. The properties of the metal obtained confirmed theoretical predictions it would be a light metal with the density of 1.8 g/cm^3 and the modulus of elasticity would be significantly higher than that of steel. With a density equivalent to that of magnesium, stiffness higher than that of steel, and a melting point approaching 1300 °C, this seemed destined to become the light metal of the future. Unfortunately, solid beryllium produced by electrolysis was 99.95% pure but completely brittle when cold. If its ductility could be improved, it was likely to be light stiff metal which had long been sought by aircraft industry.

Up to 10% of beryllium could also be dissolved in copper to produce a pale yellow alloy. Copper containing around 2% of beryllium had a beautiful golden colour. Beryllium copper displayed a degree of age hardening far higher than that of any other copper-based alloy, and in spite of its cost beryllium copper, soon found extensive industrial applications. It was particularly valuable as a spring material: the alloys were soft and ductile in the treated conditions and could then be fabricated into complex shapes. After heat treatment at temperatures between 300 and 350 °C, the best alloys developed very high tensile strength (approaching 155 kg/mm^2), and as a result of precipitation, electrical conductivity improved to about 40 per cent of that of pure copper.

Cu–Ti alloys. The age hardening of Cu–Ti alloys containing approximately 1–5 wt.% of Ti (~1–6 at.% Ti) has been known since the 1930s. Mechanical and physical properties were found to be comparable to the widely used Cu–Be alloys with better high temperature strength and superior stress relaxation behaviour. The formation of fine-scale coherent/semi-coherent Cu_4Ti precipitates, under high supersaturation, imparts strength levels comparable to Cu–Be alloys to these age hardened alloys.

Titanium

Titanium is the fourth most abundant metal in the earth's crust after aluminium, iron and magnesium. The most valuable deposits are those based on the minerals rutile and ilmenite, first found in the Ilmen Mountains (Russia) and North Carolina (USA). Titanium was first produced in metallic form in France, in 1896, by electric furnace melting. This product, being heavily contaminated by oxygen, nitrogen and carbon, was very brittle. The metal containing less than 1% of impurities had a specific gravity of only 4.8 g/cm^3 and a melting point between 1795° and 1810 °C. The metal had a silvery-white colour and was hard and brittle when cold, although some material was forgeable at red heat. In the early years of the 20th century, ferro-titanium alloys were widely used for deoxidizing steel.

The titanium of very high purity, which was completely ductile even at room temperature, was first made in Holland in 1925. Also, for the first time, titanium was seen as a promising new airframe material free from many disadvantages inherent to beryllium.

Titanium metal was not used outside the laboratory until 1946 when William Justin Kroll (Luxembourg) proved it could be commercially produced by reducing titanium tetrachloride with magnesium. The Kroll process is a pyrometallurgical industrial process used for producing metallic titanium. Refined rutile (or ilmenite) from the ore is reduced with petroleum-derived coke in a fluidized bed reactor at 1000 °C. The mixture is then treated with chlorine gas affording titanium tetrachloride $TiCl_4$ (Figure 6.10). In a separate reactor, $TiCl_4$ is reduced by liquid magnesium or sodium at 800–850 °C in a stainless steel retort:

$$2Mg(l) + TiCl_4(g) \rightarrow 2MgCl_2(l) + Ti(s),$$

where T = 800–850 °C.

Titanium reduced by liquid magnesium was the first to show a high degree of ductility at room temperature.

Titanium has been commercially available, in pure and alloyed form, for over sixty years. Titanium alloys are beginning to find a place in the construction of a new supersonic aircraft, but as a constructional material, titanium suffers from one insuperable defect: an allotropic change in the structure occurs at 882 °C and even the best and strongest alloys start weakening catastrophically at temperatures above 800 °C. Titanium is therefore unlikely to become a high temperature material. However, the corrosion resistance of pure titanium is comparable to that of stainless steel, and this should eventually result in its wider usage in chemical industry. It is extremely resistant to prolonged exposure in sea water and, being resistant to both cavitation-erosion and corrosion-fatigue, has found many applications in shipbuilding and marine technology. Titanium is used for the construction of deep-water submarines the hulls of which, like all thin shells, fail by compressive buckling at extreme depths. By constructing the hull from titanium rather than from steel, the outer shell can be doubled in thickness without a serious increase in weight thus permitting safe descent to considerably greater depths without danger of collapse.

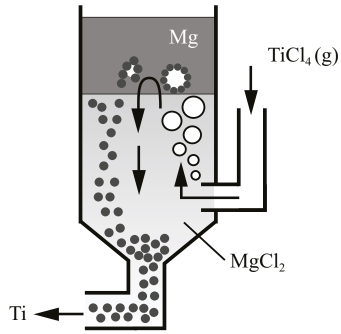

Figure 6.10. A pyrometallurgical industrial process of producing metallic titanium (Suzuki *et al.* 2000)

Because of its lightness, stiffness and abrasion resistance, a thin anodized titanium sheet is now the preferred material for low inertia camera shutters. Titanium electrodes are employed in the cathodic protection systems used for inhibiting the corrosion of ships and other marine structures immersed in sea waters.

The films developed on titanium are roughly proportional in thickness to the anodization voltage applied. Since they are transparent, interference colours are formed as film thickness increases, and this allows for the selective and intentional production of very attractive colour effects, which are widely used, particularly in titanium costume jewellery.

Niobium

Niobium was one of the last new metals to emerge into the industrial arena. Niobium is estimated to be the 33rd most common element in the Earth's crust. The free element is not found in nature, but it does occur in minerals that contain niobium and often include tantalum, such as columbite ($(Fe,Mn)(Nb,Ta)_2O_6$).

The two largest deposits of niobium ores were found in the 1950s in Brazil and Canada, and both countries are still the major producers of niobium mineral concentrates. After the separation from other minerals, the mixed oxides of tantalum Ta_2O_5 and niobium Nb_2O_5 are obtained. The first step in processing is the reaction of oxides with hydrofluoric acid $Nb_2O_5 + 10\ HF \rightarrow 2\ H_2[NbOF_5] + 3\ H_2O$ followed by

$$2\ H_2[NbOF_5] + 10\ NH_4OH \rightarrow Nb_2O_5\downarrow + 10\ NH_4F + 7\ H_2O;$$

$$3\ Nb_2O_5 + Fe_2O_3 + 12\ Al \rightarrow 6\ Nb + 2\ Fe + 6\ Al_2O_3$$

Niobium resembles in many ways its sister metal tantalum. In 1960, it was believed that niobium, being perfectly ductile refractory metal with density only a half that of tantalum, would provide a basis for the development of a new group of high temperature alloys capable of operating effectively at temperatures far in excess of those that nickel-base alloys could resist. Niobium also had unique superconducting characteristics. At atmospheric pressure, it has the highest critical temperature of elemental superconductors, which made 9.2 K. Hard superconducting alloys such as niobium-tin and niobium-titanium had transition temperatures very much higher than those of the alloys hitherto available; it was felt in the 1960s that such materials would soon be needed in large quantities for constructing magnets needed for the magneto-hydrodynamic generation of electrical power. The metal has a low capture cross-section for thermal neutrons; thus, it is used in nuclear industries. Unfortunately, however, no large-scale industrial applications for niobium emerged. Superconducting alloys were never required in significant quantities after the magneto-hydrodynamic approach to power generation was abandoned, and the role for niobium, as a canning material for fuel elements, disappeared when it was found that stainless steel was perfectly adequate. Interest in the prospects of niobium and its alloys as high temperature materials started fading rapidly after 1965 when it became very clear they had no inherent resistance to oxidation and could not be relied upon to function in the hotter regions of the gas turbine even if protective coatings could be developed.

Niobium is used mostly in alloys, and the largest part – in a special type of steel such as that used in gas pipelines. Although alloys contain only a maximum of 0.1% Nb, that small percentage of niobium improves the strength of steel. Other applications of niobium include its use in welding, nuclear industries, electronics, optics, numismatics and jewellery. In the last two applications, the low toxicity of niobium and its ability to be coloured by anodization are particular advantages. Since 2009, the production of niobium has been stable at around 63,000 tons per year.

Development of High Temperature Alloys

The rapid introduction of electricity into the domestic environment was an increasingly urgent requirement for improved alloys concerning electrical heating applications. The required alloy was the one that combined high electrical resistivity with extreme resistance to oxidation at high temperatures and mechanical properties high enough to ensure it did not fail by creep after prolonged use under a proper red heat. Prior to 1900, the only high resistivity alloys available were cupronickels such as *Constantan* and iron alloys containing up to 20% of nickel.

Nickel-chromium alloys. The first satisfactory electrical heating alloys, introduced in 1906, were based on nickel-chromium and nickel-chromium-iron systems. The wires of these alloys had electrical resistivity more than twice that of the best cupronickel alloys and were far more resistant to oxidation and stronger at high temperatures. The alloys containing around 80% of nickel and 20% of chromium showed a good balance between oxidation resistance and resistivity. They were also ductile enough to be drawn into wire. It is now known that the oxidation resistance of the alloys of such a composition depends on the formation of a protective oxide layer.

These alloys were initially induction melted in air and deoxidized with manganese. The air melted material was cheap to manufacture, although it was not always easy to draw into fine wire. Vacuum melted nickel-chromium alloys were easier to draw into wire and also had a longer high-temperature working life compensating in some degree for the added expense of vacuum melting.

Aluminium containing nickel-chromium alloys. It was found in 1929 that small quantities of aluminium improved the oxidation resistance and high temperature strength of nickel-chromium alloys and made them responsive to age-hardening. The strengthening effect of aluminium could be improved by small quantities of titanium.

When, around 1937, works began to develop a gas turbine for aircraft propulsion, the main technical problem encountered was that of producing turbine rotor blades that were strong enough at high temperatures to withstand high centrifugally imposed tensile stresses. The selected alloy was the 80/20 nickel-chromium solid solution alloy strengthened by small quantities of titanium and aluminium. The alloy was called *Nimonic 80*. The first Nimonic alloys were melted and cast in air and hot forged. As the alloys were strengthened and improved, however, very reactive alloying constituents were being added, and vacuum melting became mandatory. The alloys were progressively strengthened, initially by increasing the content of aluminium and subsequently by adding more refractory metals such as tungsten and molybdenum. The use of molten glass lubrication, introduced in the 1950s, made it possible to extrude the strongest Nimonic alloys down to rod as small as 20 mm in diameter in one hot working operation. Nimonic 115, the last of the alloys to be introduced, marked practical limit to workability. Stronger and more highly alloyed materials could not be worked and, by 1963, a new generation of nickel-based superalloys was being developed: they were not worked but cast directly to the shapes required.

Since the early years of the 20th century, it had been known that the grain boundaries of metals lost their strength far more rapidly than single crystals. This suggested that the strongest high temperature alloys would have large rather than small grains. The process, now known as *directional solidification*, resulted in an immediate improvement in high temperature blade performance even without change in alloy composition. Single crystal blades that contained no boundaries were soon produced as the logical development of a directional solidification concept. Because grain boundary strengthening additions such as hafnium were no longer required in single crystal turbine blades, they are now manufactured from the alloys having very much simpler compositions than conventional casting alloys.

Cobalt-base high temperature alloys. While the British were using the 80/20 nickel-chromium resistance wire alloy as a base for their first turbine blade, material American manufacturers were adopting a completely different approach to the problem of high temperature strength alloys. The blades and other components of gas turbines have been manufactured by casting them from cobalt chromium alloys.

The alloy „*Stellite*", based on the cobalt-chromium system, was patented in 1909. Further improvements in hardness were obtained by tungsten and/or molybdenum additions. The engine manufacturers understood that no long-term future existed for workable turbine blade alloys. High alloying content needed achieve high temperature strength, inevitably lowered the melting point of such alloys and simultaneously reduced their workability, which was illogical that an alloy specifically designed to withstand creep failure at high temperature should also be expected to undergo severe plastic deformation during manufacture. This group of alloys were named as *superalloys*. A superalloy, or a high-performance alloy, is an alloy that exhibits excellent mechanical strength and creep resistance at high temperatures, good surface stability, corrosion and oxidation resistance. From the very beginning of superalloy development, therefore, turbine blades were produced by vacuum melting and investment casting.

Superalloys develop high temperature strength through solid solution strengthening. In Ni-base superalloys, the present gamma prime phase [$Ni_3(Al,Ti)$] acts as a coherent barrier to dislocation motion and is a precipitate strengthener. The size of the gamma prime phase can be precisely controlled by careful heat treatments of precipitation hardening. Many superalloys have a two phase heat treatment that creates the dispersion of square gamma prime particles known as the primary phase with fine dispersion between these known as the secondary gamma prime. Cobalt base superalloys do not have a strengthening secondary phase like gamma prime but substantial quantities of stable carbides (e.g. $Cr_{23}C_6$). The historical development of superalloy processing has brought about considerable increases in operating temperatures of the superalloy. In the 1940s, investment casting of cobalt base alloys significantly raised operating temperatures. Within gas turbine engines, many forms of superalloys are present. They typically have a matrix with an austenitic face-centred cubic crystal structure. An alloying element of the superalloy base is usually nickel, cobalt or nickel-iron. Turbine blades can be polycrystalline, have a columnar grain structure or appear as a single crystal. Columnar grain structured blades are created using directional solidification techniques and have grains parallel to the major stress axes. Single-crystal superalloys (SC superalloys) are formed as a single crystal using a modified version of a directional solidification technique, so there are no grain boundaries in the material. The strongest wrought alloy to be developed was *Nimonic 115,* introduced in 1960. Stronger alloys could not be worked, and cast alloys were then introduced. These were subsequently improved, first, by directional solidification and finally, by turning them into single crystals. The upper curve on the diagram shows how the solidus of workable alloys began decreasing rapidly as the alloys were progressively strengthened by alloying. By 1960, the gap between the melting point of these alloys and their working temperature capability had decreased to about 100°C. It was found, however, that single crystal alloys could be based on very simple compositions, since no alloying additions were required to impart grain boundary strength. As these alloys developed, therefore, their melting points began to increase thus providing a much needed margin of high temperature safety.

Figure 6.11. Turbojet turbine disc blade made from superalloy

Superalloys are commonly used in gas turbine engines in the regions subject to high temperatures and require high strength, excellent creep resistance as well as corrosion and oxidation resistance (Figure 6.11).

Rare Earth Metals

Rare earth metals make a set of seventeen chemical elements in the Periodic Table. In the early 1930s, it was found that some of vacuum-melted alloys that had been processed from melts deoxidized with a rare-earth mixture were remarkably resistant to high temperature oxidation. The metal responsible for this effect was cerium. The mechanism that permits small quantities of reactive metals, such as cerium, yttrium, zirconium, lanthanum and hafnium to improve the protective nature of oxides that form on the surface of nickel chromium and other high temperature alloys is still imperfectly understood, although the effect is widely employed.

Today, China produces over 95% of the world's rare earth element supply, mostly in Inner Mongolia. Worldwide demand for rare earth elements is expected to exceed supply by 40,000 tons annually unless the major new sources are developed. Another recently developed source of rare earths is electronic and other kinds of waste that have significant rare earth components. New advances in recycling technology have made the extraction of rare earths from these materials more feasible, and recycling plants are currently operating in Japan where there is estimated 300,000 tons of rare earths stored in unused electronics. New efficient separation techniques, such as *ion exchange*, *fractional crystallization* and *liquid-liquid* extraction, are applied for the extraction of rare earth from ores.

Powder Metallurgy

Powder metallurgy is the process of blending fine powdered materials pressing them into a desired shape or form (compacting), and then heating the compressed material in the controlled atmosphere to bond the material (sintering). The powder metallurgy process generally consists of four basic steps: powder manufacture, powder blending, compacting and sintering. Compacting is generally performed at room temperature, and the elevated-temperature process of sintering is usually conducted at atmospheric pressure.

Powder metallurgical techniques were used in ancient Incas in pre-Columbian times to consolidate the fine grains of alluvial platinum they were unable to melt. These platinum grains were mixed with a little gold dust and heated with a blow pipe on a charcoal block. Platinum grains were, therefore, soldered together by a thin film of gold, and it was possible, applying this method, to build up solid blocks of metal that were malleable enough to withstand hot forging, and therefore were fabricated into the items such as rings and other articles. A crude form of iron powder metallurgy existed in Egypt as early as 3000 B.C. In these early manufacturing operations, iron was extracted by hand from metal sponge following reduction and then reintroduced as powder for the final melting or sintering. A mass manufacturing of powder metallurgy products did not begin until the mid-or-late 19th century.

A much wider range of products can be obtained from powder processes rather than from direct alloying of fused materials. In melting operations, the "phase rule" applies to all pure and combined elements and strictly dictates the distribution of liquid and solid phases that can exist for specific compositions. Other substances that are especially reactive with atmospheric oxygen, such as tin, are sinterable in special atmospheres or with temporary coatings. In powder metallurgy, it is possible to fabricate components that otherwise would decompose or disintegrate. All considerations of solid-liquid phase changes can be ignored, so powder processes are more flexible than casting, extrusion or forging techniques. Controllable characteristics of the products prepared using various powder technologies include mechanical, magnetic and other unconventional properties of such materials as porous solids, aggregates and intermetallic compounds. Competitive characteristics of manufacturing processing (e.g. tool wear, complexity or vendor options) also may be closely regulated.

Any fusible material can be atomized. Several techniques have been developed, which permits large production rates of powdered particles, often with considerable control over the size ranges of the final grain population. Powders may be prepared by atomizing, grinding, chemical reactions or electrolytic deposition. The powders of the elements such as titanium, vanadium, thorium, niobium, tantalum, calcium and uranium have been produced by high-temperature reduction of the corresponding nitrides and carbides. Iron, nickel, uranium and beryllium submicrometre powders are obtained by reducing metallic oxalates ($Be(COO)_2$; $UO_2C_2O_4(OH)_2{*}2H_2O$, etc.) and formats. Exceedingly fine particles also have been prepared by directing a stream of molten metal through a high-temperature plasma jet or flame simultaneously atomizing and comminuting the material.

The first critical metallurgical requirement for electrical engineers was a lamp filament that was more robust and could run at higher temperatures than a carbon filament. By the end of the 19th century, it was appreciated that the light emitted by an incandescent filament varied as the twelfth power of its temperature. The most efficient lamp, therefore, was that which could operate at the highest possible temperature, and attention was concentrated upon the melting points of the most refractory metals then known. These melting points reached a peak at tungsten. Rhenium and technetium were unknown at that time and hafnium and zirconium had never been made in pure metallic form. Osmium, however, was known to have a high melting point and could readily be produced as an exceedingly pure finely divided powder. The use of osmium as a light filament material was first proposed in 1898. The 'Osmi' lamp, first produced in Berlin, although more economical in operation than the carbon lamp was more expensive and rather delicate. It was soon overtaken by the tantalum lamp introduced in 1903. Although the melting point of osmium, 3050 °C, was marginally higher than that of tantalum at 2996 °C, the latter metal had the tremendous practical advantage of being ductile so that it could readily be drawn into fine wire. The tantalum lamp met with great and virtually instantaneous success. Between 1905 and 1911, over 103 million tantalum lamps were sold. Tungsten, which melted at 3422 °C, was certainly the ultimate material of the lamp filament, although tremendous practical difficulties were initially encountered in producing ductile tungsten wire. The first tungsten filaments, produced in 1904 by the process similar to that used for osmium filaments, produced a great deal more light than carbon filament lamps, but they were brittle and expensive. Ductile tungsten filaments were first produced by the US General Electric Company at Schenectady in 1909. Fine tungsten powder was pressed into bars about 14,5 mm in diameter. These bars were then sintered by heating them electrically in the pure hydrogen atmosphere to temperatures approaching their melting point. They were then reduced into wire by a process of hot fabrication within a well defined range of temperatures so that a fibrous structure was gradually developed within the rod or wire. Sintering was accomplished by passing a heavy alternating current through the bar so that its temperature was raised to about 3300°C. The bars thus sintered were then hot worked by rotary swaging and finally drawn to wire in diamond dies. Once compacted into the mold, the material is placed under a high heat for a long period of time. Powder metallurgy technique known as liquid phase sintering, in which completely dense, fully alloyed parts are formed from pressed metal powders at a temperature less than half the melting point of pure tungsten. Under heat, bonding takes place between porous aggregate particles, and, once cooled, the powder bonds to form a solid piece (Figure 6.16). Sintering can be considered to proceed at three stages. During the first, neck growth proceeds rapidly but powder particles remain discrete. During the second, most densification occurs, the structure recrystallizes and particles diffuse into each other. During the third, isolated pores tend to become spheroidal and densification continues at a much lower rate. The trade name *Osram*, first used by the Osram Lamp Works in Berlin, derives from the words osmium and wolfram.

Other products of powder metallurgy industry began developing very rapidly after 1910. In 1913, sintered porous self-lubricated bronze bearings were introduced, and this was followed by sintered metallic filters in 1923. Powder metallurgy was also used for producing a variety of magnetic alloys, since it was found that fine powders of metals such as iron, cobalt and nickel could be cheaply produced in a high state of purity employing chemical methods. Such powders could be fully or partly consolidated into any desired shape by powder metallurgical methods without contamination inevitably associated with conventional melting and alloying procedures. *Permalloy*, a very soft iron-nickel magnetic powder, was developed in 1928. In the 1940s, very fine powders were used for producing permanent magnets of very high coercive force.

Powder metallurgy also made it possible to produce a whole range of composite materials that could have been manufactured in no other way. Therefore, a whole generation of new electrical contact materials was developed. These included, for example, composite contacts which incorporated insoluble mixtures of silver and nickel, silver and graphite, silver-tungsten, copper tungsten and, probably the most important, silver-cadmium oxide. The last material is still widely employed because it combines the low contact resistance of silver with the ability of cadmium oxide to quench any arcs formed when the contacts open under load.

Powder metallurgy products are today used in a wide range of industries, from automotive and aerospace applications to power tools and household appliances. Many special products are possible with powder metallurgy technology. A list includes Al_2O_3 whiskers coated with very thin oxide layers for improved refractories, iron compacts with Al_2O_3 coatings for improved high-temperature creep strength, magnets, microwave ferrites, filters for gases, light bulb filaments made with powder technology, metal glasses for high-strength films and ribbons, heat shields for spacecraft re-entry into the Earth's atmosphere, linings for friction brakes, electrical contacts for handling large current flows and bearings that can be infiltrated with lubricants.

Sintered carbide cutting tools. Tungsten wire had to be hot drawn, and the deficiencies of steel dies that were first used for reducing the diameter of the hot swaged tungsten rod to the dimensions of wire that could be handled by diamond dies soon became apparent. Tungsten carbide seemed to fulfil most of the characteristics of the material required. Extremely hard carbide W_2C was first prepared in 1893 when tungsten was fused with carbon in an electric furnace. Sintered carbides were first produced in 1914. These compacts were rather brittle and produced by sintering the mixtures of WC and W_2C at temperatures close to their melting points. Cast tungsten carbides dies were produced by this method before 1914, although the cast product had a very coarse grain size and dies were again very brittle. In 1922, it was shown that tungsten carbide sintered in the presence of liquid cement could be very tough. The three metals, including iron, cobalt and nickel all provided satisfactory molten cement, although cobalt provided the best combination of hardness and toughness. The new sintered alloy was rapidly adopted for manufacturing wire drawing dies. In 1925, the first sintered cutting tool appeared. The tool material consisted of tungsten carbide powder sintered together with a 6% of cobalt. Nearly all well-known proprietary grades of carbide emerged from this operation. In 1932, tantalum carbide tools that were sintered with nickel rather than cobalt were introduced. Such materials were inferior in toughness to cobalt bonded composites and therefore never very successful.

The compaction of powdered metals. Powder compaction is the process of compacting metal powder in a die through the application of high pressure. The density of the compacted powder is directly proportional to the amount of the pressure applied. Typical pressure ranges from 15 to 80 kg/mm^2 and is commonly used for metal powder compaction. There are four major classes of tool styles: single-action compaction used for thin flat components, opposed double-action with two punch motions accommodating thicker components, double-action with a floating die and a double-action withdrawal die. Double action classes give much better density distribution than the single one.

Metal powders are still compacted in steel dies because this is generally the cheapest and most convenient method of producing large quantities of components on a routine basis. Technical limitations are imposed by die wall friction.

Considerable difficulties were encountered, however, in the early years of powder metallurgy industry when the production of larger and more complex components was attempted. It was then appreciated that such products would most effectively be compacted under pure isostatic pressure.

In the years immediately after the Second World War, shock, or dynamic, consolidation is an experimental technique for consolidating powders using high pressure shock waves generated by explosive detonation within a large bath of water. Despite being researched for a long time, the technique still has some problems in controllability and uniformity. However, it offers some valuable potential advantages. As an example, consolidation occurs so rapidly that metastable microstructures may be retained. This technique was among those investigated for consolidating large ingots of titanium from the sponge. The explosive approach, though versatile, is expensive and not well suited for routine manufacture.

In 1958, the isostatic pressing process began to modify and improve, so that it could be used for consolidating metal powders at high temperatures. Gas pressing equipment was developed, which allowed metals, alloys and ceramics to be consolidated in metal capsules at temperatures up to 2225 °C and pressure up to 39 kg/mm^2. The commercial units available in 1964 operated at similar temperatures, although routine operating pressures were limited to about 11 kg/mm^2. Pressing was accomplished in a cold wall autoclave with the furnace used for heating the compact enclosed within this chamber and thermally insulated from it so that the chamber walls remained cold. The unit was pressurized with inert gas such as argon. In HIP (*hot isostatic pressing*) pressing operations, compact formation and sintering occur simultaneously. This procedure is used extensively for producing high-temperature and high-strength parts such as turbine blades for jet engines. In most applications of powder metallurgy, the compact is hot-pressed heated to a temperature above which materials cannot remain work-hardened. Hot pressing lowers the pressure required to reduce porosity and speeds welding and grain deformation processes. Also, it permits better dimensional control over the product, lessens sensitivity to the physical characteristics of starting materials and allows powder to be driven to higher densities than with cold pressing thus resulting in higher strength.

Powders can also be rolled to produce sheets. The powdered metal is fed into a two-high rolling mill and is compacted into a strip at up to 3 m per minute. The strip is then sintered and subjected to another rolling and sintering. Rolling is commonly used for producing sheet metal for electrical and electronic components as well as coins.

Powders can also be extruded. There appears to be no limitation to the variety of metals and alloys that can be extruded, providing that the temperatures and pressures involved are within the capabilities of die materials. Extrusion lengths may range from 3 to 30 m and from 0.2 to 1 m in diameter.

Extrusion processes are of two general types. The first one includes the powder mixed with a binder or plasticizer at room temperature, whereas the second extrudes the powder at elevated temperatures without fortification. Extrusions with binders are used extensively for preparing tungsten-carbide composites. Tubes, complex sections and spiral drill shapes are manufactured in extended lengths and diameters varying from 0.5 to 300 mm. Hard metal wires of 0.1 mm in diameter have been drawn from powder stock.

Amorphous Alloys

An amorphous metal (also known metallic glass) is a solid metallic material, usually an alloy, with a disordered atomic-scale structure. Amorphous metals are non-crystalline and have a glass-like structure. Amorphous metals have good electrical conductivity. There are several ways in which amorphous metals can be produced, including extremely rapid cooling, physical vapour deposition, solid-state reaction, ion irradiation and mechanical alloying.

The first metallic glass was an alloy ($Au_{75}Si_{25}$) reported in 1960. The early glass-forming alloys had to be cooled extremely rapidly (on the order of 10^6 K/s) to avoid crystallization. Metallic glasses could be produced in a limited number of forms (typically ribbons, foils, or wires) in which one dimension is small so that heat could be extracted quickly enough to achieve the necessary cooling rate. As a result, metallic glass specimens (with a few exceptions) were limited to the thicknesses of less than one hundred micrometers. More recently, a number of alloys with critical cooling rates low enough to allow the formation of an amorphous structure in thick layers (over 1 millimetre) have been produced. These alloys are known as *bulk metallic glasses* (BMG).

In the 1990s, new alloys that form glasses at cooling rates as low as 1 K/s were developed. These cooling rates can be achieved by simple casting into metallic moulds. These "bulk" amorphous alloys can be cast into the parts of up to several centimetres thick (maximum thickness depending on the alloy) while retaining an amorphous structure. The best glass-forming alloys are based on zirconium and palladium, but the alloys based on iron, titanium, copper, magnesium and other metals are also known. In 2004, a product known as "glassy steel" was developed. The product is non-magnetic at room temperature and significantly stronger than conventional steel, though a long research and development process has remained before the introduction of the material into use.

Amorphous alloys tend to be stronger than the crystalline alloys of similar chemical composition, and therefore can sustain larger reversible ("elastic") deformations than crystalline alloys. Amorphous metals derive their strength directly from their non-crystalline structure, which does not have any defects (such as dislocations) that limit the strength of crystalline alloys. One modern amorphous metal, *Vitreloy*, has tensile strength that is almost twice than that of high-grade titanium. The zirconium and titanium based *Vitreloy* achieved the yield strength of over 1723 MPa, nearly twice the strength of conventional crystalline titanium alloys (Ti 6Al–4V is ~830 MPa) and about the strength of high-strength steel.

Ferrous Metals

Iron and steel possess an essential feature of the industrial civilization in which we live. They were largely responsible for it, and still remain an indispensable part of it.

It was man's ability to make and use tools that first distinguished him from other animals, and iron was crucial in this respect. Other metals were used before iron. However, the most important was bronze, but when iron came on the scene, it gradually took over, since it has been better, stronger and more abundant. Given weapons for the hunt man was assured of food, and the same weapons gave him some protection against his natural enemies. With tools he could more readily cultivate crops and prepare his food, clothes and shelter. Thus, man gained security, and starting with it, the human race was able to settle and develop. As the art of ironworking progressed, it became possible to harness natural forces more effectively. A windmill or waterwheel could be made of stone or brick and timber, but when ways were found to use the power of steam, only metal was strong enough for the machinery involved. And if iron made steam power practicable – and with it the industrial revolution – steam made possible the production of iron on an industrial scale and turned a domestic craft into an important industry.

Iron is the fourth most abundant element in the world and has been made for at least 4000 years. Throughout history, iron has been produced from naturally-occurring iron ores. Very small amounts of iron – more accurately, a natural alloy of iron and nickel – have been found as meteorites. They were hammered out into useful shapes, but quantities were so small that meteoric iron has never been more than a curiosity.

Iron ores are all, basically, a chemical mixture of iron and oxygen (iron oxide), including small quantities of other elements. Some impurities are easily removed; others are more difficult. Iron ore is an oxide because iron has a strong affinity for oxygen and there is always the supply of oxygen available in the air. If metallic iron is left exposed to the air, it will slowly become an oxide again and will rust. Fortunately for the iron-maker, carbon has an even greater affinity for oxygen than iron. If iron ore is heated strongly in contact with carbon, oxygen and carbon will unite to form gas that burns away leaving iron behind. That is the basis of iron ore conversion into iron-reduction or smelting.

For many centuries, the equipment used was very simple and the production of iron extremely small. But the trade of iron-making started, and villages began to get their iron-makers just as they had their millers, potters and weavers. In those parts of the world where there was no iron ore, traders began to take iron goods to exchange for other products and international trade in iron began to spread. The product made by the early workers from iron was wrought iron that can be shaped by hammering it while it is hot; if two pieces at the right temperature are hammered together, they weld into one piece. It is possible to melt wrought iron but of no practical value, thus it was never melted in practice. Iron was converted, or reduced, directly from the ore in what is therefore termed the *direct reduction process*.

Public domain

Figure 6.12. Bloomery smelting during the Middle Ages

The early equipment used for making wrought iron was simple and consisted of a small furnace heated by charcoal (called a bloomery), hand- or foot-operated bellows to blow a charcoal fire and some tongs to hold the hot metal while it was forged into shape. Bloomeries were made of clay that would resist the heat of fire (Figure 6.12). Charcoal was lighted inside the bloomery and then, while a continuous blast of air was kept up, more charcoal and some iron ore were fed in by hand through a small aperture at the top. As oxygen in the ore united with the carbon of charcoal, it became gas burned off at the top of the bloomery as a light blue flame. After a few hours, all oxygen left iron ore, and a small spongy ball of iron, the bloom from which the bloomery took its name, remained. Then, the front of the bloomery was broken open and the bloom was raked out and taken to an anvil for hammering to whatever shape was required. Ironworkers relied on their practical skills rather than on theoretical knowledge.

The *blast furnace* was introduced towards the end of the 15^{th} century and spread slowly throughout Europe. Technically it introduced a new product, cast iron, an alloy of iron and carbon, which, unlike wrought iron, is quite easily melted. Cast iron is very different from wrought iron. It is strong in compression but comparatively weak in tension. In addition, it is relatively brittle and cannot be forged or shaped by hammering; thus, its uses were limited in comparison with wrought iron. At first, a blast furnace, like a bloomery, needs a continuous blast of air to keep the fire burning, but its greater size demanded mechanical power to work the bellows. This meant, at the time, a waterwheel, and blast furnaces were built alongside streams where water was available. Second, blast furnaces needed timber for making charcoal. Economically, the introduction of the blast furnace meant that iron-making took the first real steps towards becoming an industry as distinct from a craft. Stonemasons and bricklayers were needed to build and maintain a blast furnace; millwrights were necessary to make the waterwheels and keep them in repair, and numbers of other specialized workers were also required. The blast furnace also brought about social change. It had to work continuously, twenty-four hours a day, seven days a week, and the workers had to be organized accordingly. Two teams of men, each working twelve hours, were needed to operate the furnace, so shift working became common. The blast furnace was in some respects like a bloomery (though much bigger) and it still used charcoal as its fuel. The major difference was that, because the furnace operated at a higher temperature and the ratio of charcoal to the ore was greater, iron absorbed a greater amount of carbon; therefore, it produced, instead of a spongy piece of wrought iron ready for forging, molten cast iron. This was allowed to accumulate at the bottom (or hearth) of the furnace and taken out, or tapped every twelve hours or so. Molten iron was allowed to run into channels moulded in a bed of sand where it solidified.

Earthy materials and other rubbish mixed with iron ore also melted and, being lighter than molten iron, floated on the top of it and were run off at intervals. Some limestone was also charged into the furnace, along with iron ore and charcoal, to act as a flux, that is, to combine with waste materials and help with forming molten waste called slag.

A few uses of cast iron were found as this came from the blast furnace – cast iron cannons were made in Sussex by 1543, and decorative fire-backs for domestic fireplaces are among the oldest existing forms of iron pastings. The conversion of cast iron from the blast furnace into wrought iron was done in a furnace called a *finery* that also used charcoal for fuel and was blown by waterwheel-driven bellows. Here, solid pieces of cast iron were remelted in the fire and carbon that had combined with iron in the blast furnace and was driven off thus leaving wrought iron.

Making iron in bigger pieces, although it was more economical, brought problems. Many iron users needed long thin bars, blacksmiths for horseshoes, for example, nail-makers. The power hammer could not forge a bar smaller than about 20 mm^2 simply because hot iron became too long and flexible to be handled. Furthermore, long thin lengths cooled down too quickly, and no furnace available at that time could reheat them.

A very effective way of solving the problem was slitting mills. This machine, driven by water powder, cut up long thin strips of iron into a number of small rods as the strip passed between rotating discs or cutters, and therefore could be adjusted to slit various sizes. To prepare a long thin strip for a slitting machine, another device was incorporated, a pair of smooth rolls of which were, in the long run, even more important than the slitting mill itself. They were the forerunner of all rolling mills vital in steel processing today. A piece of iron hammered out as long and thin as possible was passed between rolls while it was still red hot, and they squeezed and elongated it to make the required strip. But the expansion of output demanded increased supplies of raw materials, and charcoal was becoming scarce. By the early part of the 17th century, charcoal shortage was serious. Blast furnaces often had to be stopped for a time and wait while a stock of charcoal was built up.

Fuels for a Blast Furnace

For various reasons, raw coal, just as it was mined, could not be used in the blast furnace. The chief difficulty lay in the fact that coal, as found in the earth, contains impurities. Sulphur is one of them. Even a very small amount of sulphur gets into iron and makes the metal brittle. The year 1709 marks the second great step forward in the history of iron production because of successful making of iron in the blast furnace with artificial fuel – coke that was known and used making malt for brewing but not used in the blast furnace. Coke was made at the time by burning coal in large heaps until all unwanted impurities had gone off in smoke, and then cooling it quickly with large quantities of water. The modern coke-making process involves the carbonization of coal to high temperatures (1100 °C) in an oxygen deficient atmosphere in order to concentrate carbon. The process also includes processing coke oven gas to remove tar, ammonia (usually recovered as ammonium sulphate), phenol, naphthalene, light oil and sulfur before gas is used as fuel for heating ovens. Indirect heat is applied by means of gas firing and coal is baked for approximately 18 hours, during which time, 25% of its volume is released as gas. The average size of coke is 52 mm, ash – mean 8%. Coke is the most important raw material fed into the blast furnace in terms of its effect on blast furnace operation and hot metal quality. It took some time for the coke-smelting process to spread, and only in the 19th century the last of the charcoal furnaces stopped.

Coke smelting opened up a possibility of building blast furnaces near to new fuel production sites, but the blast furnace needed power. Windmills were of no use for iron-making. The wind varies in strength and sometimes does not blow at all, whereas the blast furnace needs a constant and continuous amount of power. Streams and rivers suitable for driving waterwheels were scarce and often in the wrong place. It was not practicable to locate a furnace in a place where there was a good stream but no coal or iron ore and then transport raw materials to it. There were no railways, and roads were very bad. A further difficulty was that, although the blast furnace could use coke fuel, charcoal was still needed at the finery to convert the blast furnace cast or pig iron into wrought iron. There was no point in making great quantities of cast iron if it could not be converted into the wrought product. Iron production and trade needed a new source of power which was the steam engine invented by James Watt and marketed in 1775. When Watt adapted his engine to produce rotation power in 1781 and improved it to become a better mechanical job in 1784, forges and rolling mills could use it as well as blast furnaces, and the problems of water power were over.

Because trade was no longer dependent on charcoal and water power, but did need iron ore and coal, it began, naturally, to move to the sites where these two minerals could be found. In Britain, these included South Wales and the West Midlands. The latter was to become, in the first half of the 19th century, the biggest iron-making area in the world. It was known long before the 18th century that there was plenty of coal and iron ore in the area, as well as limestone (for flux) and fireclay (for building furnaces), and they were all either at or near to the surface of the ground. The one thing it lacked was water power, but when the steam engine was perfected, the district had advantages no other area could equal for a time.

In 1784, Henry Cort suggested a process for making wrought iron from cast iron. It was pointed out that if sulphur in coal combined with iron, it spoilt it. In this process, coal was burned in a separate part of the furnace thus allowing only the flames to come into contact with iron. The flames were reflected or reverberated down on to iron to be treated by a specially shaped sloping roof. From this fact, the furnace was known as a reverberatory furnace.

A quantity of cast iron in the form of pigs, totalling about 250 kg, was put into the furnace, the coal fire was burning brightly and, in about two hours, the charge was decarburised and converted to wrought iron. Iron completely melts and carbon starts burning off as well. As carbon burns off, the melting temperature of the mixture rises from 1150 °C to 1540 °C. This, by the nature of the process, was not molten, but in the form of red hot spongy mass. The hook on the end of the bar (or large pair of tongs) is then used for pulling out large puddle balls of the material, about 35–40 kilograms each and 30–38 cm in diameter. Each ball was taken to a heavy power hammer and shaped into a rectangular lump called a *bloom*. This was wrought iron and could be rolled or forged into whatever shape was needed.

By the end of the 18th century, an average furnace made about 1000 tons of iron a year and would need 3000–4000 tons of iron ore (depending on how pure the ore was). It would also consume at least 2000 tons of coke. A little-known fact is that of all materials needed to operate a blast furnace, air for the blast is the largest in quantity. For every ton of iron made, the furnace would use about 5 or 6 tonnes of air.

By the beginning of the 19th century, iron trade was prepared to meet increased demands that the fast-developing industrial revolution was to make.

Steel

Cast and wrought iron were excellent materials for making many types of necessary machinery, but they could not be cast, forged or rolled to the precise shape and dimensions required. The parts of iron components, at least, had to be machined to shape. The man who found the first way of making better carbon steel was Benjamin Huntsman, a clockmaker. He was dissatisfied with the steel available for making clock springs. In 1740, he took some *shear steel* and melted it in a clay pot or crucible (shear steel is a type of steel produced by heating blister steel sheared into short lengths to a high heat, welded by hammering or rolling or both, and finally finished under the hammer at the same or a slightly greater heat). He poured out molten metal, let it solidify and tried working it: it was better than anything he had ever seen before. This steel soon became called crucible steel. In time it was found that the quality of steel could be varied according to need. Steel for making, say, a wood chisel, or a chisel for cutting metal, needed different grades. Crucible steel could be made to suit the application. Crucible steelmaking lasted all through the 19th century.

The Bessemer process. In August 1856, Bessemer read a paper called *The Manufacture of Malleable Iron and Steel without Fuel*. He took molten cast iron and blew a blast of cold air through it (Figure 6.13). It will be remembered that to convert cast iron into wrought iron, carbon has to be removed. In the puddling furnace, this was done by heating iron in contact with fluxes containing oxygen. Bessemer used

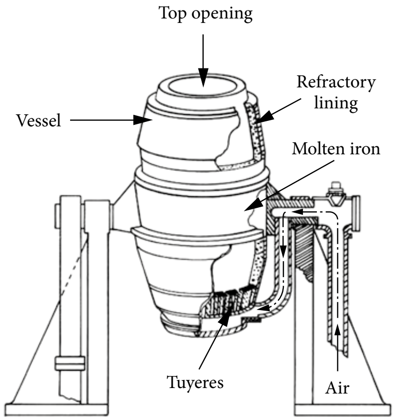

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Figure 6.13. A Bessemer converter

the cheapest form of oxygen, i.e. ordinary air, which contains about 21 per cent of oxygen. Because there was so much oxygen, carbon reacted very strongly with it in cast iron. The reaction was exothermic, which is, it actually generated heat; thus, iron became hotter instead of colder.

His experimental converter was big enough to deal with about 350 kg of iron at a time, but it did so in about thirty minutes, compared with the production of about 250 kg in two hours in the puddling furnace.

Thomas process. The most important problem affecting the Bessemer process derived from the presence of phosphorus that occurs naturally in small quantities in most iron ores found in Britain and on the Continent. As with sulphur, if even a minimal amount of phosphorus combined with wrought iron or steel, the product became weak and brittle. Processing in the puddling furnace removed phosphorus; the Bessemer process, as at first used, did not. The early experiments were carried out with cast iron made from the ore with negligible phosphorus. When cast iron contained phosphorus, the result was failure.

In 1879, a completely unknown man, P.G.Thomas, modified the process so that it could use phosphoric iron and the way was wide open for a great expansion of steelmaking. In his converter, Thomas used a special form of lining such as dolomite, which, being chemically basic, united with phosphorus and left the metal free of this troublesome element. Phosphorus went away with slag, and therefore he was able to sell this as an agricultural fertilizer.

There were then two bulk steelmaking processes: Bessemer's original that suited non-phosphoric iron, and Thomas' that would deal with phosphoric types. In the English-speaking world, the two processes were distinguished by names *acid Bessemer* (the original) and *basic Bessemer (Thomas's)* because of the chemistry involved.

The Siemens-Martin process. Meanwhile, the open-hearth steelmaking process was developed. The Siemens furnace was first applied to making steel in France in 1863 (Figure 6.14). Wrought iron now had some really serious competition. The Siemens and Bessemer processes were complementary, but both could use phosphoric or non-phosphoric iron. Bessemer was cheaper since it used no fuel but needed to be charged with molten iron. This was easy when the Bessemer plant was adjacent to blast furnaces. Molten iron could be used in the Siemens open hearth

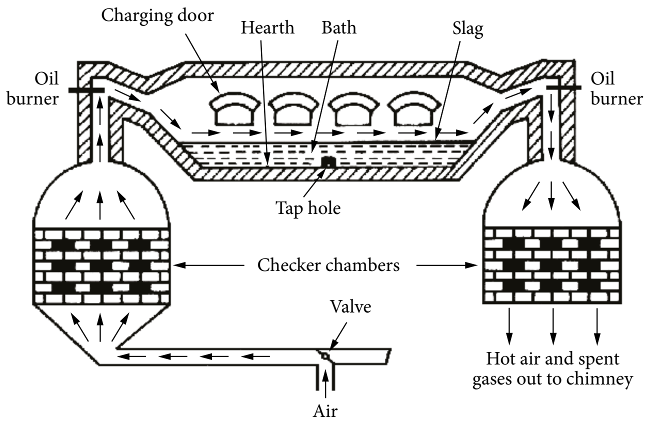

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Figure 6.14. The open hearth process

furnace, and it often was, but an advantage of this process was that it could melt scrap iron. With the spread of industry, scrap had become a useful raw material, as machinery of all kinds wore out or was replaced by new and better types, and it was cheaper than pig iron. The Siemens process was slower than the Bessemer's one the charge of whose of iron took about thirty minutes to convert to steel; in the Siemens furnace, it took from eight to twelve hours. This could be an advantage, because it enabled the furnace operator to make frequent checks on steel and to adjust its chemical composition as required.

Alloy Steel

Before 1856, there were only two important iron products – cast and wrought iron. The third, carbon steel, was essential for some purposes but its output was small. Bessemer added the fourth, mild steel. In 1868, the first of the fifth group, now called special, or alloy, steel appeared. The only means of cutting iron and steel was a tool made of carbon steel. This can be hardened by making it red hot and cooling it quickly. Such treatment makes steel very hard and brittle. It has to be heated again to a lower temperature and cooled once more to make it both hard and tough. Carbon steel, which has been hardened, can be softened again by making it red hot and letting cool naturally. When a carbon steel tool wore or was damaged, it was useful to be able to soften, file or machine it back to its original shape and reharden. However, softening was a distinct disadvantage when it occurred accidentally. Heat is always generated during cutting and when iron or steel are cut, temperature can easily reach a point where the tool is softened. Thus, there is a limit to the speed or depth of cut that can be made by a carbon-steel tool.

In 1900, tungsten steel, called high speed steel, was introduced. Tungsten steel could be forged to shape and left to cool naturally in the air when it became very hard and tough. It only needed grinding to a sharp cutting edge to be ready for use; when it became blunt from use, it was reground. The most useful feature, however, was that the new steel did not soften even at a dull red heat. Tungsten steel is still used today. In 1887, Hadfield made a type of steel that was particularly tough and wear-resisting. This steel was alloyed by manganese. It was used for railway points and crossings as well as in rock-crushing machinery.

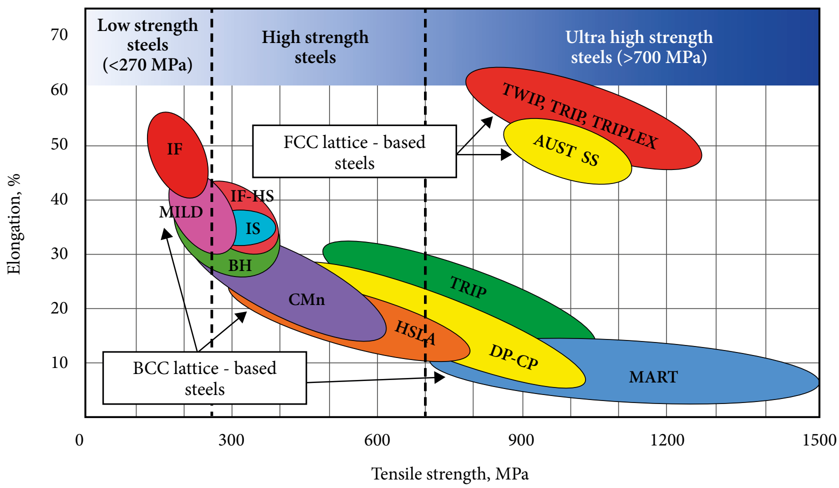


Figure 6.15. The current trends regarding steel development (Ozgowicz *et. al.* 2012)

Stainless steel was invented in 1913. While experimenting with better steel for rifle barrels, it was noticed that one of the steels was unaffected with acid. Attack by acid, or etching, is a form of corrosion. If steel did not etch, it would not corrode either, at least under many conditions where ordinary steel would. Experimental steel contained nearly 13% of chromium. Experiments were suspended during the World War I. One of the problems was steel hardening. Over time, several different kinds of stainless steel were developed for particular purposes. One of the commonest types of steel contains about 18% of chromium and 8% of nickel. The other types though of different composition but all correctly called stainless steel, are used for knives, razor blades and surgeon scalpels. Still, further varieties are used in industry to resist the heat of furnaces, or in other severe conditions. The current trends regarding steel development are visible from Figure 6.15.

Manufacturing Steel Products

While new types of steel were being invented and improved, developments were also taking place in the manufacture of the product. Rolling mills, in particular, changed drastically. Only a limited amount of shaping can be done in a single passage through rolls; metal must be passed through several times. With the Bessemer and Siemens processes, steel could be made not only quicker but in bigger individual pieces. Molten steel was cast into rectangular blocks, or ingots, rolled to the required shape. As ingots became bigger, men could not handle them and machinery had to take over.

In 1866, the first reversing mill was introduced. The continuous rolling mill was invented in 1862. The metal being rolled came out of one pair of rolls, passed immediately to the next pair, etc. down the line until the finished size and shape were achieved. Electricity is also used today in steel works as a direct source of heat, and its beginnings, for this purpose, go back a long way. In 1878, Siemens suggested that it could be used for steelmaking, but nothing came of it at the time. The electric furnace was ahead of its time and came only during the World War I for remelting quite large quantities of scrap. Neither the Bessemer converter nor the Siemens open-hearth furnace would do this adequately, but the electric furnace managed.

Modern Steelmaking

Steel makes up to 95 percent of all metal produced in the world. Today, iron and steel-making are technologically international (Figure 6.16). Steel is mainly made by the BOS (Basic Oxygen Steel) process that needs very large supplies of pure oxygen not available in Bessemer's time. Much of steel is made into large slabs by the process called *continuous casting*. The first commercial continuous casting machine was built at Bradford in 1946.

Economics, always important to iron and steel-making, is even more so today and this is the reason for changes in ore practicing and steelmaking in the blast furnace. In most cases, there are still large reserves of iron ore left. However, they are often of low grade. At present, in Europe and some other parts of the world, it is more economic

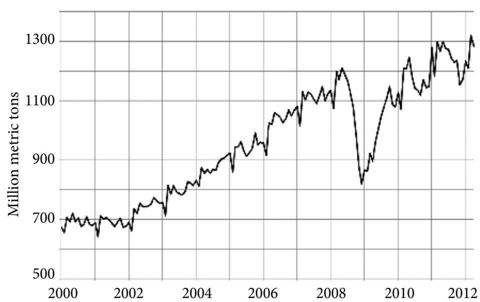


Figure 6.16. The dynamics of crude steel production in the world (1527 Mt in 2014)

Figure 6.17. Iron ore pellets

for ironworks to import high-grade foreign ores containing as much as 60% of iron. Vast new deposits have been opened up in Africa, Australia, Canada and South America. Consequently, steelmaking is being concentrated at the sites near to deep-water ports capable of taking super-carriers of 100,000 tons or more.

Today, pellets account for more than 97% of all agglomeration products. Pelletizing operations produce a "green" pellet or ball which is then hardened through heat treatment. These pellets are normally relatively large (10–12 mm) and usually contain at least 60% of iron (Figure 6.17). Pellets can include silica, alumina, magnesia, manganese, phosphorus, sulfur, limestone or dolomite. The fluxed pellets increase furnace efficiency.

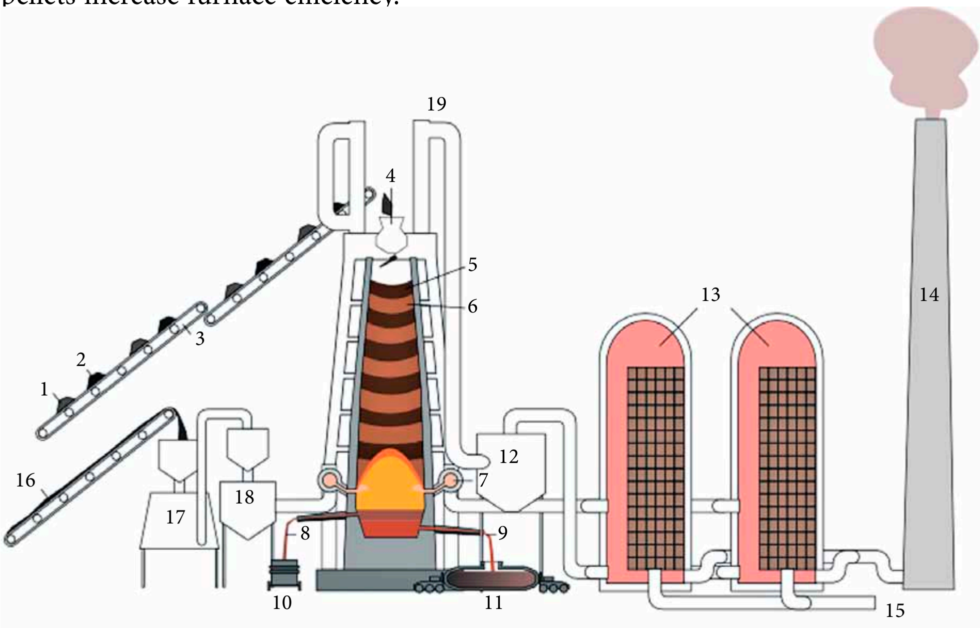


Figure 6.18. Blast furnace placed in an installation: 1 – Iron ore + limestone sinter; 2 – Coke; 3 – Elevator; 4 – Feedstock inlet; 5 – Layer of coke; 6 – Layer of sinter pellets of ore and limestone; 7 – Hot blast (around 1200 °C); 8 – Removal of slag; 9 – Tapping of molten pig iron; 10 – Slag pot; 11 – Torpedo car for pig iron; 12 – Dust cyclone for separation of solid particles; 13 – Cowper stoves for hot blast; 14 – Smoke outlet (can be redirected to carbon capture & storage (CCS) tank); 15 – Feed air for Cowper stoves (air pre-heaters); 16 – Powdered coal; 17 – Coke oven; 18 – Coke; 19 – Blast furnace gas downcomer

The blast furnace remains the same in principle, although very much bigger. A blast furnace of the 19th century made 100 tons or so of iron a week. A modern blast furnace can produce 10000 tons a day (Figure 6.18). All modern blast furnaces are mechanically charged and controlled according to the present programme.

The Bessemer process had been dying out for some time. The last British basic Bessemer converter was closed down in 1966. The BOS (Basic Oxygen Steel) process is the biggest single development of steelmaking in the present century. Oxygen steelmaking was first put to practical use in Austria in 1953. It looks rather like the Bessemer process, but uses pure oxygen instead of air and oxygen is blown on to the surface of molten iron at a very high speed instead of through it as air was blown in the Bessemer converter. BOS steelmaking is very fast. A charge of 300 tons of iron can be converted into steel in about thirty minutes. Enormous quantities of oxygen are used and BOS converters have to be equipped with their own oxygen plant in a nearby building. The process gives off a vast amount of gas and, under modern anti-pollution regulations, this has to be cleaned and disposed of harmlessly instead of being discharged to the atmosphere.

First, steel was cast into moulds to produce rectangular pieces (ingots) or heavy flat ones (slabs). These could then be processed into finished products by rolling. Ingots weigh from 45 kg to 20 tons. Continuous casting replaced ingots. Liquid steel is poured into an open copper mould of the required size for a billet, bloom or slab. The mould is water-cooled and the outer skin of the metal solidifies rapidly. At the bottom of the mould the metal is solid enough for it to be drawn out at the same rate as molten metal enters at the top. Water sprays below the mould to complete solidification. As long as molten steel is poured in, a solid product comes out in a continuous length. This is cut up automatically by a flame-cutting head into the lengths required for further processing.

Compared with steelmaking and casting, rolling mills have evolved into a rather less, but development has been impressive nevertheless. Continuous wide strip rolling now produces all steel sheets used for motor car bodies, cookers, refrigerators, washing machines and other industrial and domestic purposes.

A steel slab is heated to rolling temperature and passed through a series of rolls in line until it is reduced to a very long thin plate. This is automatically wound into a coil weighing several tons. Then, the coils are pickled in acid to clean the surface and passed through a further series of rolls, this time cold, to become finished coils of thin wide strip.

The coils of strips can be used in several ways. They can go to cut-up lines where they are unwound and continuous length is cut automatically, without stopping the line, into the sheets of a required length, or the coils can be passed through an automatic tinning line where they are given a microscopically thin coating of tin to become tinplate, which is used for making cans. Some coils go through a different automatic production line to be coated with zinc (galvanized).

Two other means of finishing strip steel are of growing importance: plastic coating and pre-painting. Sheets and coils can now be coated on either or both sides with a thin layer of a plastic film. It is bonded on to steel and adheres so tightly it cannot be pulled off without destroying it. Several different colours and surface finishes are available and plastic-coated sheets have many uses. In the pre-painted sheets and coils, there is also a wide range of colours and types of paint available. The thickness of sheet material is now usually stated in millimetres: it is technically a sheet if it is up to 3 mm; all above is a plate. Computers are now being used „on-line“ to control actual production processes.

Today, many alloys and special kinds of steel are made in electric furnaces that can be of two types, *arc* or *induction*. In the arc furnace, heat is generated by means of an electric arc, and metal is melted by this heat in a refractory-lined vessel of drum shape, which can be tilted mechanically to tap the finished steel (Figure 6.19, Figure 6.20).

Arc furnaces are now of many sizes and can hold from a few to 150 tons or more. The induction furnace was invented in Italy in 1877, but the time was not ripe for it to develop, and the first one in Britain was installed in 1927. There is no arc in this furnace; an electric current induces a secondary one inside the furnace itself and generates sufficient heat to melt steel. Induction furnaces are generally used for very special and expensive types of alloy steel and may range in capacity from a few kilograms to 5 tons or more.

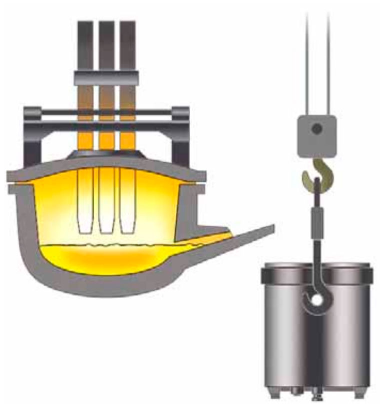


Figure 6.19. Overview of an electric arc furnace

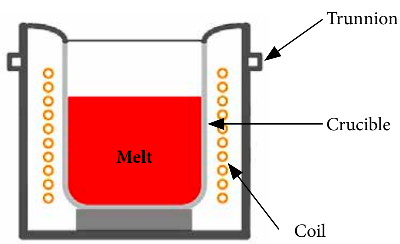


Figure 6.20. A coreless induction furnace

The electric furnace does not use any fuel, thus, the steel made in them cannot be contaminated from these sources. Unfortunately, some gases also act as contaminants in steel, and since these gases are present in the air, it is impossible to avoid getting them into steel. They must then be removed, and there are several ways of doing so. One is *vacuum melting*. Steel is made as carefully as possible in an ordinary electric furnace and then remelted in an induction furnace it is contained.

Future of Developing Iron Production

The blast furnace is becoming more and more a victim of its own success. With the recent boom in blast furnace capacity over the past decade in Asia (China in particular), raw material demand has escalated dramatically. With supply-constrained market conditions, prices have escalated to an alarming degree. The present steel production technology is based on coal, i.e. mostly on carbon, natural gas, a mix of carbon and hydrogen and on electric arc furnaces.

To identify CO_2-lean process routes, 3 major solution paths stand out and three routes only can be applied: either a shift away from coal called decarbonising, whereby carbon would be replaced by hydrogen or electricity in the processes such as hydrogen reduction or the electrolysis of iron ore, or the introduction of carbon capture and storage (CCS) technology, or the use of sustainable biomass.

Among all of these, 6 families of process routes have been selected for further investigation:

- a variant of a blast furnace where the top gas of the blast furnace goes through CO_2 capture, but the remaining reducing gas is re-injected at the base of the reactor, which is moreover operated with pure oxygen rather than hot blast (air). This has been called the *Top Gas Recycling Blast Furnace* (TGR-BF) or ULCOS-BF. The CO_2-rich stream is sent to storage;
- a smelting reduction process based on the combination of a hot cyclone and a bath smelter called *HIsarna*. The process also uses pure oxygen and generates off-gas that is almost ready for storage;
- a direct reduction process in a shaft furnace, either from natural gas or from coal gasification. Off-gas from the shaft is recycled into the process after CO_2 has been captured, which leaves the plant in a concentrated stream and goes to storage;
- two electrolysis variants. One operates slightly above 100 °C in a water alkaline solution populated by small grains of the ore (electro-winning process), the second – at steelmaking temperature with a molten salt electrolyte made of slag (pyroelectrolysis).

– two more options are available: one consists in using hydrogen for direct reduction without any carbon footprint; the other is based on the use of sustainable biomass, the first embodiment of which is charcoal produced from eucalyptus sustainable plantations grown in tropical countries.

HIsarna process. *HIsarna* is an emerging steel production technology the high-level attractiveness of which is based on the following points:

– ability to use thermal coals instead of metallurgical coals;
– ability to use low-quality iron ore feed materials;
– easy ability to capture a high proportion of CO_2 for possible geological storage;
– 20% of the primary energy and CO_2 saving (without geological storage).

The iron making process can be described as follows (Figure 6.21):

– Iron ore and oxygen are injected into the *Cyclone Converter Furnace* (CCF) where hot *Smelt Reduction Vessel* (SRV) off-gas is burned and the resulting heat is used for melting and partially reducing the ore. The resulting (partly molten) ore then runs downwards under gravity into the SRV below. The temperature of this material is expected to be around 1450 °C, and the degree of pre-reduction makes 10–20%.
– Coal is injected at high velocity (using carrier gas such as nitrogen) into the bath. The objective of the primary process for this component is to dissolve carbon into metal to replace dissolved carbon used in the smelting step. Coal injection conditions are critical, and the metal bath runs at 1400–1450 °C with dissolved carbon around 4.0%. There is essentially zero silicon present in the metal, and other minor impurities such as manganese are also present at very low levels (compared to hot metal in the blast furnace). Phosphorous and titanium partition are largely transferred to the slag phase as oxides.
– Molten ore dissolves directly into slag, and metal-slag mixing (generated by the coal injection plume) creates a large metal-slag interfacial area for smelting. Dissolved carbon in metal removes oxygen from the ore and a significant amount of CO gas is liberated. This reaction takes place in the reducing lower part of the vessel and is strongly endothermic.

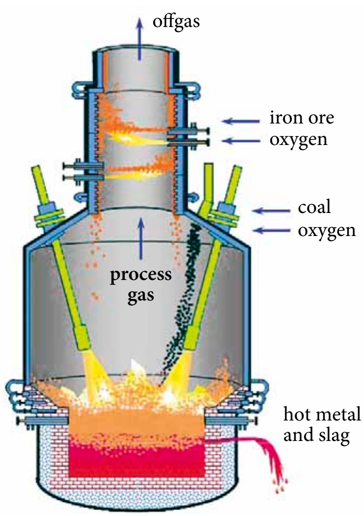


Figure 6.21. The concept of Hisarna furnace (Atkinson 2008)

– CO gas from smelting, together with conveying gas and devolatilisation products from coal, provide an upward-moving stream of hot fuel gas. This upward movement generates a large amount of droplets. Heat is carried from the upper to the lower region by these droplets. Oxygen is introduced into the upper section via lances, and heat is generated by combustion.
– Partly combusted gas leaving the Smelt Reduction Vessel then provides necessary hot fuel gas for the Cyclone Converter Furnace. This gas is typically around 1450–1500 °C and has a post-combustion degree of around 50%. The definition of post-combustion (PC) is as follows:

$$\%PC = 100(\%CO_2+\%H_2O)/(\%CO+\%CO_2+\%H_2+\%H_2O)$$

Steel pyroelectrolysis. Iron and steel industry generates roughly five percent of all greenhouse gases produced by human activity in addition to the major amounts of sulfur dioxide (gas that contributes to acid rain). Unlike the current technology for producing steel, the new process does not use the element at the root of steel problems, which is carbon. The principal by-product of the new system is oxygen. Iron oxide (derived from iron ore, the precursor to steel), is fed into a reactor called an electrolysis cell where it is made to dissolve in the solution of other molten oxides (Figure 6.22).

An electric current is passed through the cell from one end (anode) to the other (cathode). At the interface of the molten oxides and cathode, pure liquid iron- steel is formed. The principal by-product (oxygen) bubbles off the anode. Pyroelectrolysis could have a number of other advantages for steel production. For example, it could produce higher-purity alloys. Carbon used in the current steelmaking process contains sulfur, a contaminant that leads to lower-quality metal. The new technology would eliminate the steps currently in place to remove sulfur from steel thus saving energy and money. This technology is still at the early stages of research.

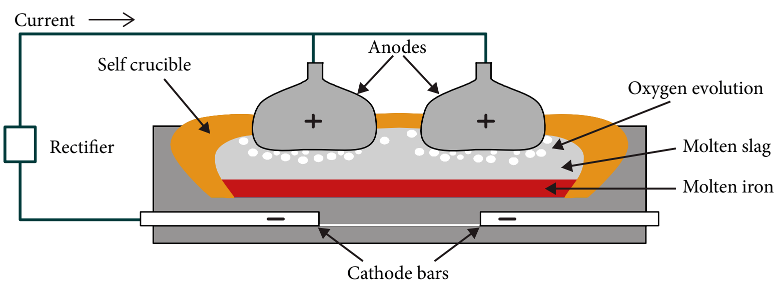


Figure 6.22. Electricity-based steelmaking: Molten Oxide Electrolysis (MOE) process (Atkinson 2008)

Iron Smelting in Lithuania

Iron smelting in Lithuanian territory must have occurred before our era and in the first half of the first millennium of our era. Up to now, archaeologists have collected very little direct information about ore deposits and their exploitation. The bog ore is often found and is widespread throughout Lithuania.

The bog iron deposits of North Eastern Europe were created after the Ice Age ended on postglacial plains. Bog iron refers to impure iron deposits that develop in bogs or swamps by the chemical or biochemical oxidation of iron carried in the solutions (FeO(OH)). Iron is oxidized to ferric hydroxide upon encountering the oxidizing environment of the surface. A large number of these springs and seeps on the flood plain provide iron for bog iron deposits (Figure 6.23). Also, it has been found that iron was smelted from sedimentary ore taken out from the bottom of lakes.

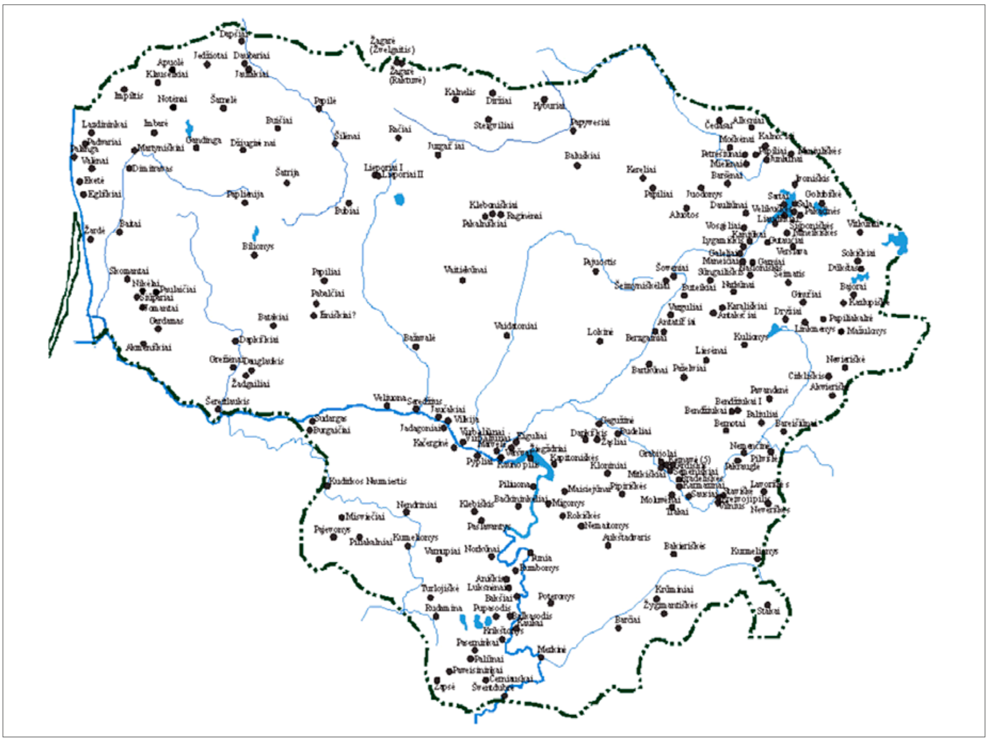
Hydrated iron oxides (unroasted ore) contain about 8.66–17.03% of Fe. After roasting, the ore contained 51–58% of iron. The ore was mined in summer while smelting occurred in the fall and winter. Charcoal is the only fuel that could have been used in Lithuania for smelting iron in a furnace (Figure 6.24).

At present more than 220 archaeological sites with iron metallurgy finds are known in Lithuania (Figure 6.25). In 180 of them only slag was discovered. In only 40 sites, besides slag, burnt and not burnt iron ore, ore excavation pits, ore burning hearths, ore sluice equipment were found.

Figure 6.23. Typical iron-bearing ground water emerging as a spring

a) b)

Figure 6.24. Smelting in Lithuania: a – slag bed (Kernave excavations), b – clay crucibles, I–XIII c. (National Museum of Lithuania)



With permission from Lithuanian University of Educational Sciences

Figure 6.25. The most important iron smelting findings in Lithuania (I millennium B.C. – I millennium A.D.) (Salatkienė 2008)

Figure 6.26. A preserved furnace and bellows

The hydrated ore was mined in an open fashion, washed with well water on wooden flooring and roasted in open fires in shallow pits. Flat rocks and ground stone were used for crushing and grinding it. Charcoal for iron smelting was made in round pits or stacks. The furnaces were roughly circular, about 80 cm tall and 30 cm in diameter (Figure 6.26).

From local metal, different items such as weapons, coat of mail, etc. were produced (Figure 6.27). In 1962, a modern cast iron plant was established in Kaunas to meet the demand of grey cast iron castings for agriculture, construction, forestry machinery, food and light industry, water, gas, heat and telephone companies. In 1966, for the first time in the Soviet Union, the production of synthetic cast iron was launched. The use of an electric arc furnace allows cast iron to be made from a 100% scrap metal feedstock. In arc furnaces, the charge material is directly exposed to an electric arc. The temperature of an industrial electric arc furnace can reach 1800 °C. Production capacity was 20000 tons of castings per year. Foundry exports its output to Sweden, Finland, Germany, Switzerland and Italy. Exports exceed 90% of the total production. In 2009, the company discontinued activities.

a)

b)

Figure 6.27. Iron items produced from local metal: a – iron axes from Kernave (Lithuania) excavations; b – two-edged swords, X–XI c. (National Museum of Lithuania, The Old Arsenal)

CHAPTER 7 The Development of Steam and Mechanical Engineering

This chapter will help in

- presenting the development of steam and mechanical engineering in the 18th and 19th centuries;
- introducing early water and steam machines;
- explaining the advent of steam and the Industrial Revolution (Boulton, Watt, Trevithick, Carnot and Otto works);
- describing the application of ferrous for railroads, bridges and towers;
- giving the details of building water supply systems, dams and canals (Suez, Panama);
- briefly discussing dredging techniques and different methods for canal excavation.

Early Water and Steam Machines

It is important to reflect on the meaning of the word "engineer" in relation to the late 18th and 19th centuries. Originally, it had only a military meaning – the man who erected and worked the intricate engines of war: a catapult, siege towers, fortifications or camps. In the 18th century Europe, the meaning was expanded and given a civil application – the builders of canals, bridges and wharves. It was now considered a highly skilled craft rather than a learned profession.

The ancient philosophers and men of science in Persia and Greece understood the importance of heat in the scheme of things as their designation of it and as one of the four basic elements of the universe – earth, air, fire and water. The ancient builders and engineers rarely utilized air and water for prime movers, but it happened in the Arabic and Byzantine periods where we can see their first serious utilization as power sources. With the exception of Hero's Pneumatics (The Pneumatics, 1851), the ancient world is silent on heat as a source of work.

Hero published a well recognized description of a steam-powered device called an *aeolipile*. The true utilization of heat as an energy source came along with the development of the steam pump and the engine. The progress of this new source of power preceded the theoretical description and analysis of that.

In the 17th and 18th centuries, Britain laid groundwork for the Industrial Revolution of the 19th century by maximizing natural prime movers, i.e. animal, wind and water power. Animal power was utilized in "horse mills" and "horse-ground gears" with the ratios of 100:1 for rotary mill power. Wind power came in the form of post mills, smock and tower mills ("cap mills"). Post mills were the earliest design going back to medieval times (Figure 7.1). Tower or "cap" mills appeared later, and only the upper part or a cap rotated to meet the wind. A fan tail was invented in 1750 and maintained the proper heading of the mill into the wind and prevented blow-over.

John Smeaton and Jean V. Poncelet (1778–1867) designed water wheels that more efficiently employed water power (Figure 7.2). The use of "buckets" instead of paddles utilized the force of gravity rather than incident current to make more efficient water-powered mills and blast furnaces. One in 1849 (Spain), five meters in diameter, nine meter width with only two meter head of falling water, generated 180 hp. The ultimate expression of the water wheel was (and is) the reaction turbine (Figure 7.3). Claude Burdin (1870–1873) carried the term "turbine" from the Latin *turbo* (to spin). Comparing to the efficiency of the water wheel (22%), Fourneyman turbines (80%) showed tremendous improvement these devices made.

a)

b)

Figure 7.1. Post (a) and tower (b) mills

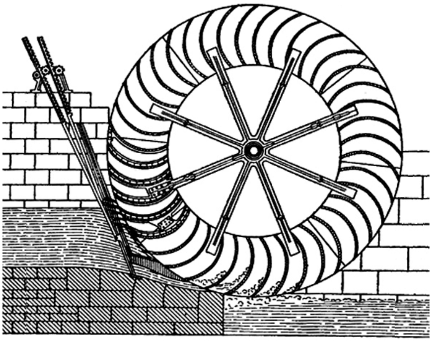


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Figure 7.2. The Poncelet wheel (1852)

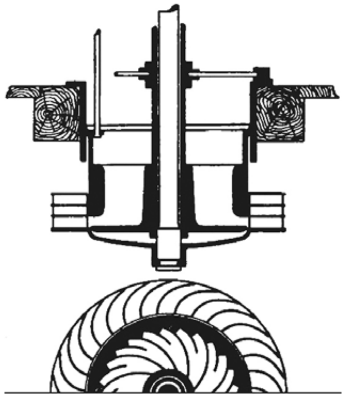

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Figure 7.3. Fourneyman's turbine

Advent of Steam and the Industrial Revolution

The effective use of steam was realized by Thomas Savery (1650–1715). The development of steam as a prime mover was linked to a fundamental problem encountered from classical antiquity to the Industrial Revolution – dewatering mines. The exploitation of wood for fuel and building materials created an "energy crisis" in the 16th century. Deforestation led to the use of coal in Britain thus changing economic history. Europe had major mining centres exploiting iron, copper and coal. As mines went deeper, damage caused by water increased and de-watering became critical. The use of horses and overshot waterwheels to run pumps were expensive. Horses ate expensive fodder and waterwheels required feedstock which in turn necessitated expenditures on dams for reservoirs. More importantly, it led to the development of steam power. In 1698, a mine drainage pump was patented. This device produced a steam-generated vacuum by spraying cool water on a steam-filled vessel. The resultant partial vacuum raised water in a suction pipe through a non-return valve. Steam pressure was then used for ejecting water up a delivery pipe. Unfortunately, this device consumed huge amounts of coal. In 1710, an engine with a piston within a cylinder was developed. Completed in 1712, it was employed in mine pumping.

In the Newcomen engine (Figure 7.4), the piston rod is connected to pumps. In operation, the engine was little more than a cylinder on the top of a boiler. Two jets – one steam, the other water – alternately bathed the cylinder creating a vacuum allowing atmospheric pressure to force the piston down. Circulation arcs at the beam end let the rods rise and fall in vertical lines. This engine practically changed

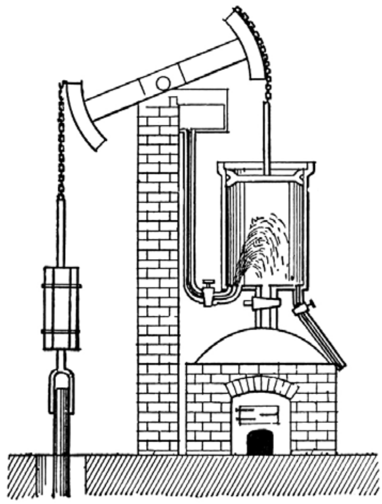

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Figure 7.4. Newcomen's atmosphere mine pump (1712)

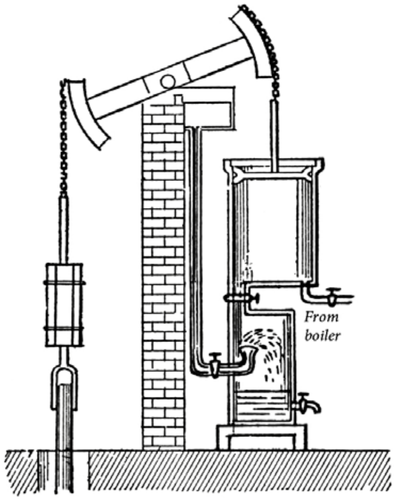

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Figure 7.5. Typical Watt steam engine (1860)

mining overnight. Automatic valves increased the frequency of power cycles to 15 strokes/min. The Newcomen engine with a 76 cm cylinder and 23 cm stroke (15 strokes/min) lifted water 27 meters. The brass cylinders of Newcomen engines were expensive and created frictional losses. They were not rotary but rather reciprocating.

James Watt (1736–1819), some say, reinvented the steam engine. Watt invented the steam condenser, parallel motion gearing, double-action engines and the governor (centrifuge) (Figure 7.5). Few technological innovations can truly be classified as singular in a revolutionary sense. If there was one device that earned this distinction, it is the condenser invented by James Watt. More than anyone thing, Watt's device for improving the thermal efficiency of steam engines allowed these machines to truly set the stage for today's industrial world. Matthew Boulton and Watt produced the first rotary design. Of 500 Boulton and Watt engines in 1800, 38% were pumped and 62% rotary. Trevithick laid the direct-acting engine horizontal and created the locomotive.

Early boilers were inefficient and low-pressure. First boilers were copper with cast iron heads-wrought iron with rivets was used by 1725. „Haystack" and „wagon" boilers were geometrically incorrect for steam; cylindrical cones were more stable. Increasing the heating surface led to the use of more flues or fire tubes.

Carnot cycle. The first person to analytically investigate the properties of heat energy was the French military engineer Nicolas Carnot (796–1832) who demonstrated that the most efficient thermodynamic attainable cycle was the one in which all heat was supplied at the fixed isothermal higher temperature, and all heat exhausted to the surroundings at the fixed lower temperature. It has been shown that the thermal efficiency of the engine operating upon the Carnot cycle is the ratio of the difference of higher and lower temperatures to the higher temperature or

$$\text{Thermal efficiency} = T_1 - T_2/T_1$$

All practical engine cycles, including Rankine (steam), Otto (gas), Diesel and Stirling (constant pressure), are compared to the Carnot cycle as the basis of comparison where the efficiency of the Carnot engine is

$$\text{Efficiency (Carnot)} = 1 - (T_c/T_H)$$

where T_H is intake temperature and T_c is exhaust temperature.

No engine operating between two given temperatures can be more efficient than the Carnot engine operating between the same two temperatures. All such engines have the same efficiency, irrespectively of the nature of working substance. It follows that the efficiency of the Carnot engine is independent of the nature of working substance and is a function of temperatures only.

In a heat engine, work is obtained by passing heat from a body at a high temperature – a boiler – to another body in which the temperature is lower than in the cylinder as in Newcomen's or Watt's condenser designs. The condition for the production of maximum work was that there should not occur any change of temperature in the bodies employed, which was not due to a change in volume, the main offender being direct conduction of heat through the work of the engine. The loss of heat to friction or convection was equivalent to wasting the difference in temperature that could have been utilized to produce useful power. Carnot defined this condition of maximum efficiency as reversibility, i.e. if the reverse of such a process was carried through, all effects would be of the same magnitude but reversed in direction. The maximum effect was obtained when the engine operated with reversibility. All reversible engines operating between the same two temperatures must have the same efficiency regardless of the medium utilized as the carrier of heat. The reversible operation of the engine alone determined its efficiency in producing mechanical work. Thus, the greater was relative difference in temperature between the boiler and condenser, the greater was work for the same difference in temperature, and the engine would operate more efficiently. The summarized principles regarding heat engines are as follows:

- maintaining difference in the greatest possible temperature between the boiler and condenser;
- if faced with a choice between equal temperature intervals, selecting the one lower on the temperature cycle;

- actual working medium is unimportant from the point of view of thermodynamic efficiency, except as its properties affect working temperatures;
- seeking to operate as closely to reversible conditions as practical.

This hypothetical cycle shows that, even under ideal conditions, the heat engine cannot convert all heat energy supplied to it into mechanical energy. Some heat energy must be rejected. The greater is the difference in temperature between the sources of a high and low temperature (sink), the greater is the efficiency of the engine. The Carnot cycle consists of (Figure 7.6):

- adiabatic (isentropic) compression (4–1);
- isothermal heat Q_{in} addition (from high temperature source);
- isothermal expansion (1–2);
- adiabatic heat rejection (expansion) (2–3);
- isothermal compression (3–4)

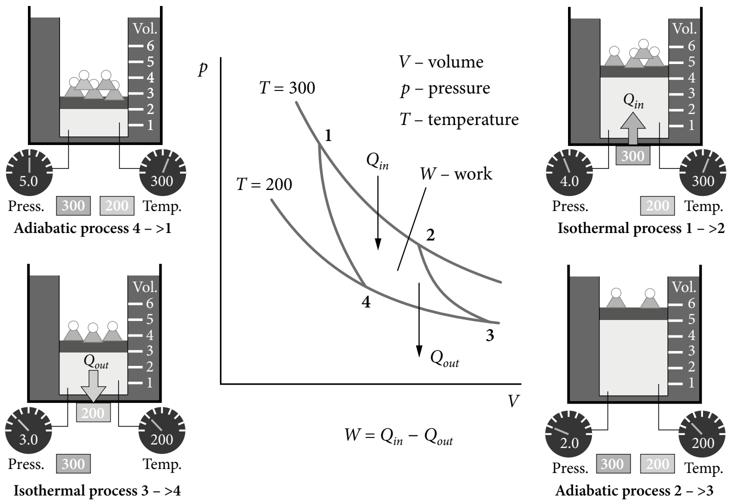

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Figure 7.6. The ideal gas-piston model of the Carnot cycle

To summarize the Carnot's idea (Figure 7.6), a piston, a frictionless cylinder, somewhere between the top and bottom of its stroke moves upward (mechanically) compressing working substance (air) and raising its temperature. The cylinder head is exposed to the high temperature source – heat flows to the cylinder with no real change in the high temperature body (isothermal), and gas (air) expands. The heat source is removed and the piston continues downward in adiabatic expansion; the last part of the cycle is the isothermal compression and rejection of heat to the cold body or sink (isothermal) returning the piston to its starting point. Carnot's work is made even more impressive as it was carried out before the formulation of the first and second laws of thermodynamics.

The early mechanical engineers designed and built the first steam cycle engines without reference to any elaborate theory on heat engines. As a result, their devices were not very efficient. Newcomen's engines were very costly to operate and not for the low cost of coal. Therefore, the use of these engines may never have been attempted. Watt's engines improved on the efficiency and economy of steam engines, but further refinements had to await thinkers like Carnot. Other engineers took different courses on higher efficiency and/or high economy. Two of these men – Otto and Diesel – are remembered in engine cycles that bear their names.

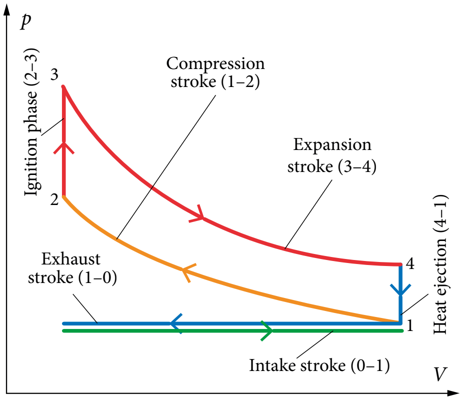

Figure 7.8. The Otto cycle. Pressure-Volume diagram

Nikolaus August Otto (1832–1891) was the German inventor of the first internal-combustion engine to efficiently burn fuel directly in a piston chamber. The Otto engine was designed as stationary, and in the action of the engine, the stroke is the upward or downward movement of a piston in a cylinder (Figure 7.8). Used later in an adapted form as an automobile engine, four strokes are involved:

- downward intake stroke: coal-gas and air enter the piston chamber;
- upward compression stroke: the piston compresses the mixture;
- downward power stroke: ignites the fuel mixture by electric spark;
- upward exhaust stroke: releases exhaust gas from the piston chamber.

Where the Otto engine took an air and fuel mixture at the suction stroke, Rudolf Diesel (1858–1913) developed one that took in only air with the fuel being injected later at the end of the compression stroke (Figure 7.9). Burning, theoretically, then proceeds at constant pressure. The remainder of the cycle is the same as in Otto's. The object of this device was to raise temperature by compression so that the fuel is self-ignited. This allows the diesel engine to be somewhat more thermally efficient than the internal combustion engine. Diesel engines tend to be slightly more economical in fuel costs as well and are typically used in the situations where size and power are not limiting factors. Many diesels today operate on both two-stroke and four-stroke cycles. The two-stroke cycle is popular in the motorcycle and small industrial designs, as it develops more power than a four-stroke design of the same dimensions.

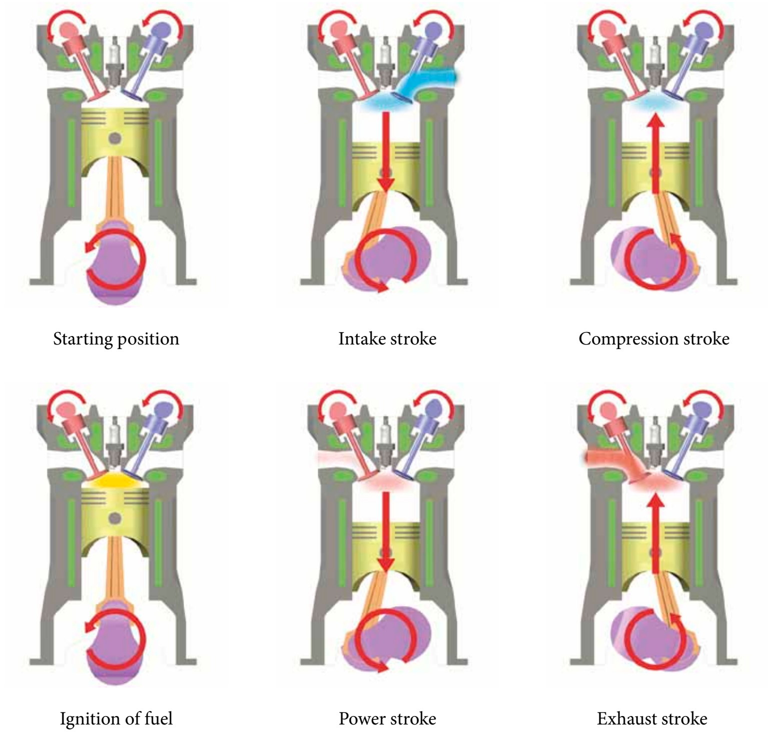

Figure 7.9. The diesel engine

Ferrous for Railroads and Bridges

Iron bridges began in Britain in 1779. In general, subsequent bridges were linked to railroads – bridges spanned streams and bays that blocked the most direct routes for expanding rail lines. Stephenson's *Rocket* (1829) carried a carriage with 30 passengers at a speed from 40 to 50 kph. The Great Western Railway (UK) had a wide gauge of 2.1 meters. Stephenson used a 1.3 meter standard gauge. The first rails to be used were cast iron; later wrought iron "T-rails" were employed. Designers understood that the stresses of compression and tension were greatest at the top and bottom, e.g. the neutral axis could be trimmed of excess metal. The new bridges of the late 18th and 19th centuries (Table 7.1) were first built of iron. The most unusual design combined in composite design covers three classical forms of bridges in one structure – arch, beam and suspension. The suspension cables of the bridge act like arches in reverse in what they pull rather than push on their supports. The first cable suspension bridge was built in the United States in 1816. The span of this bridge was 124 meters wide. Better design and mater-

Table 7.1. The evolution of bridge design

Antiquity to the 18th century	Masonry designs
Early 19th century	Iron and steel design; use of trusses; caissons for deep foundations
Mid-late 19th century	Suspension bridge design with steel cables
Late 19th to 20th century	Suspension, truss, pre-stressed and reinforced concrete design

ials could succeed in longer lasting bridges capable of spanning large distances compared to arch designs.

The early suspension bridge designers knew intuitively that the supporting members of a suspension bridge must be flexible and strong in tension. The tensile strength and ductility of wrought iron, unlike cast iron or steel, are greater in the longitudinal direction.

Eiffel Tower

The Eiffel Tower erected in 1889 as the entrance arch to the 1889 World's Fair has become both a global cultural icon of France and one of the most recognizable structures in the world. Eiffel built his construction out of iron, even though he was aware of the advantages of steel. The puddle iron structure of the Eiffel Tower weighs 7,300 tons while the entire structure, including non-metal components, makes approximately 10,000 tons. The structure took more than two years to complete. Each one of about 12,000 iron pieces were designed separately to give them exactly the shape needed. Three hundred workers joined together 18,038 pieces of puddled iron (very pure form of structural iron) using three and a half million rivets (Figure 7.10). The maintenance of the tower includes applying 50 to 60 tons of paint every seven years to protect it from rust. The height of the Eiffel Tower varies by 15 cm due to temperature.

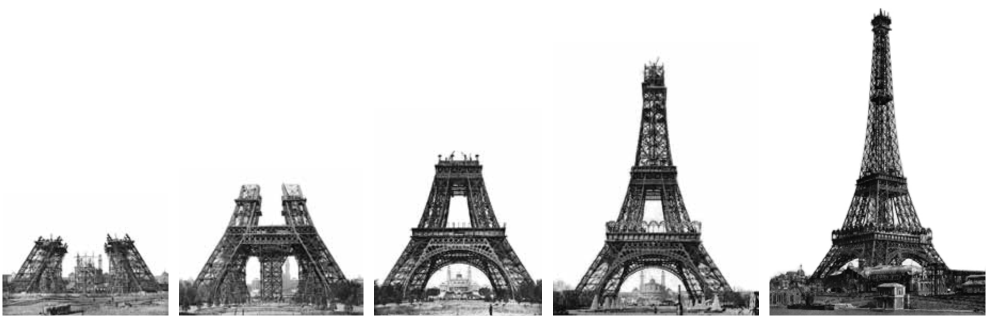

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Figure 7.10. The erection of the Eiffel tower

Table 7.2. The Eiffel Tower

Number of structural iron components in the tower	15,000
Number of rivets	2,5000,000
Weight of foundations, kg	277,602
Weight of iron, kg	7,341,214
Total weight, kg	8,564,816
Height of the first platform, m	57.63
Height of the second platform, m	115.73
Height of the third platform, m	276.13
Total height in 1889, m	300.51
Total height with the television antenna, m	320.755

Dams and Water Supply

Dams, generally earthen embankments, were built to hold reservoirs throughout England in the mid-to late-19th century. Within this period, over 90 major dams were built, and only two of the constructions of this era failed catastrophically. In the U.S., water supply projects produced similar successes and failures. St. Francis dam was completed on 1 March 1926 and collapsed on 12 March 1928 thus leaving in its wake a death toll of over 450. Southern California's need for an ever-larger water supply moved to create the 389 kilometres long Colorado River Aqueduct and its centrepiece – Hoover Dam, which, along with a power plant were completed in 1935. Upon completion, the dam's crest was 220 meters high, forming Lake Mead, the largest man-made lake in America (35 billion cubic meters). The Catskill water supply system, begun in 1907, was finished in 1937. Above it, a 76.8 meters high Olive Bridge dam of concrete with embedded large boulders and tanned cyclopean masonry construction were built. The Catskill Aqueduct, a part of the New York City water supply system, brings water from the Catskill Mountains. The aqueduct increased public water supply and facilitated the urbanization of the city area. Also, hydraulic engineers were called upon to create water courses for commerce. As water supply, these channels are *aqueducts* while as the routes of transport, they are *canals*. Four major classes of dams are based on the type of construction and materials used: *embankment, gravity, arch,* and *buttress*.

Embankment dams typically are constructed of compacted earth, rock, or both, making them less expensive than others that are constructed of concrete. Consequently, more than 80 percent of all large dams are of this type. Embankment dams have a triangular-shaped profile and typically are used to retain water across broad rivers.

Gravity dams consist of thick, vertical walls of concrete built across relatively narrow river valleys with firm bedrock. Their weight alone is great enough to resist overturning or sliding tendencies due to horizontal loads imposed by the upstream water.

Arch dams, also constructed of concrete, are designed to transfer these loads to adjacent rock formations (Figure 7.11, a). As a result, arch dams are limited .

Buttress dams are essentially hollow gravity dams constructed of steel-reinforced concrete or timber.

The world's two tallest dams are located in Tajikistan in the city of Vakhsh where they tower over 335 meters tall (Rogun) and 300 meters tall (Nurek). The Three Gorges Dam in China, a concrete gravity dam scheduled for completion in 2009, will be 175 meters tall, the equivalent of a 48-story building (Figure 7.11, b). *Three Gorges Dam* is the world's largest hydropower facility with a generation capacity of 18,200 megawatts. It will simultaneously supply flood storage and enhance navigation along the Yangtze River. The structure will create a reservoir more than 600 kilometers long and 1,100 meters wide, capable of storing 39.3 billion cubic meters of water.

Dams have long been acknowledged for providing electricity without the pollution of other methods, for flood protection, and for making water available for agriculture and human needs. Within recent decades, however, the *environmental impacts* of dams have been debated. While dams do perform important functions, their effects can be damaging to the environment. Prior to dam construction, most natural rivers have a flow rate that varies widely throughout the year in response to varying conditions. Once dam constructed, the *flow rate* of the river below a dam is restricted. Because water is held behind the dam and often released from some depth, the temperature of the water below the dam is usually lower than it would

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b)

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Figure 7.11. Arch-gravity dam in Hoover (USA) (a); view of the three Gorges Dam (China) (b)

be prior to dam emplacement. The *temperature of the water* flow is often constant, not reflecting the natural seasonal variations that would have been the case in the free-flowing river. Similarly, the chemistry of the water may be altered. Water exiting the lake may be higher in dissolved salts or have lower oxygen levels than would be the case for a free-flowing river. Impoundments increase the potential for *evaporation* from the river. Because the surface area of a lake is so great when compared to the river that supplies it, the loss of water to evaporation must be considered. In some desert areas, potential annual evaporation can be greater than 2.1 m, meaning that over the course of one year, if no water flowed into or out of the system, the reservoir would drop in elevation by 2.1 m.

The impoundment of water behind a dam causes the *velocity of the water to drop*. Sediment carried by the river is dropped in the still water at the head of the lake. Below the dam, the river water flows from the clear water directly behind the dam. Because the river no longer carries any sediment, the erosive potential of the river is increased. Erosion of the channel and banks of the river below the dam will ensue. Even further downstream, sediment deprivation affects shoreline processes and biological productivity of coastal regions. The environmental changes described above create a new environment in which native species may or may not be able to survive. The most commonly cited species affected by the presence of dams is the salmon. Salmon have been isolated from their spawning streams by impassable dams.

The main causes of *dam failure* include:

- spillway design error,
- geological instability caused by changes to water levels during filling or poor surveying,
- poor maintenance, especially of outlet pipes,
- extreme rainfall,
- human, computer or design error.

Canals

The Suez Canal was built from 1859-1869. With the extension of 163 kilometres, it had no locks and averaged 8 meters in depth. The Suez Canal connects the Mediterranean and Red Seas across 160 kilometres of the Isthmus of Suez (Figure 7.12). The construction of the Suez Canal reduced the travel distance from Liverpool to Bombay from 17,000 to 10,000 kilometres.

The canal is a relatively straight cut from Port Said on the Mediterranean to Ismailia from which it follows a series of shallow lakes to the Red Sea. The canal crosses no greater height than 15 meters at any point so it was just as it was styled – a long, shallow (8 meters) ditch (Figure 7.13). It has no elaborate lock or control structures. In 1855, it was deepened 3 meters and widened from 70 meters by the British to become the commercially viable waterway it is today. The project originator Ferdinand de Lasseps was not an engineer. Contractors used steam dredging equipment to complete work. Even with 8000 paid workers, the project was plagued by the debt and outbreaks of cholera killing hundreds.

The Panama Canal is a 77 km long international waterway allowing ships to pass between the Atlantic and Pacific Oceans and saving about 12,875 km from a journey around the southern tip of South America, Cape Horn (Figure 7.14).

It takes approximately fifteen hours to traverse the canal through its three sets of locks. The canal accommodates 14,000 ships a year on average, with over 200 million tons of cargo, representing five percent of the world's shipping. Vasco Nunez de Balboa first proposed the construction of a canal linking the Atlantic and Pacific Oceans in 1513. Centuries passed, and economic and technical difficulties prevented from such a massive undertaking. In March 1881, Ferdinand de Lessups, the builder of the Suez Canal, established the French Canal Company in an attempt to construct a sea-level canal across Panama. Work was halted in late 1888, with the construction of the one-fifth to be completed, due to a lack of money. The Panamanian government allowed the U.S. to build the Panama Canal and provided for perpetual control over the zone of five-mile wide on the either side of the canal. The construction of the Panama Canal was taken over by the United States in early 1904. By 1924, the canal was opened to traffic.

During the process of planning and building the project, the problems of sanitary and municipal engineering, construction management, excavation and dredging, lock,

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Figure 7.12. The route of the Suez Canal

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Figure 7.13. The Suez Canal

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Figure 7.14. Location of Panama Canal

dam and spillway construction were encountered. From the Atlantic Ocean to the Pacific Ocean, a vessel first passes through a man-made breakwater into Limon Bay from which it is lifted 26 meters above the sea level by a series of three locks at Gatun Dam. It then proceeds for 39 kilometres across the man-made lake (Gatun) following eight miles through Gailland Cut. Finally, the vessel descends Pedro Miguel and Miraflores locks down to the sea level and the Bay of Panama (Figure 7.15). To raise the ship, water is released from the lake (Gatún, Alajuela or Miraflores) or from a higher lock through the open valves on the upper end of the lock. To lower the ship, water drains to a lower lock or to the ocean through the valves open at the lower end of the lock.

Figure 7.15. The Gates of the Gatun locks

Sanitation measures were needed to combat diseases that stopped the French effort with 16,500 deaths. Yellow fever and malaria were responsible for deaths during construction by the French Canal Company. The United States realized control over these diseases was essential to the success of the project. The canal authorities led the fight on tropical diseases. Yellow fever was a problem at the end of 1904. An outbreak of the disease caused panic among workers. The control of the disease required the eradication of the *Aedes* mosquito by eliminating breeding places, destroying infected mosquitoes and preventing them from becoming infected by screening the patient at the infectious stages. Control was accomplished by:

- providing an abundant supply of piped water thus doing away with the need for open water containers used in the collection of drinking water;
- paving streets to eliminate puddles;
- fumigation;
- screening infected patients.

With these measures, yellow fever was virtually eliminated in the canal zone. Malaria was somewhat more difficult to control. The destruction of larvae by larvicidal poisons protecting man from mosquito bites by screening and immunization and the use of quinine were necessary. The steps that played an important role in the successful control of malaria by the engineers included the elimination of breeding places through filling lowlands, the construction of drainage ditches, cleaning stream and pond banks and cutting brush. To allow for the drainage of water, roads with hard surfaces were constructed, crowned and graded. Dust, the major problem of roads during the dry season, was controlled by the application of water or oil to the road surface.

House drainage and storm water runoff were carried in sewers. Both combined and separate sewage-water supply systems were built using gravity flow and pumping to move sewage to outfalls. Panama City used gravity flow for house and street drainage and discharged through pipes into the Gulf of Panama. Colon required a separate system for house and street drainage with sewage pumps, as elevations were insufficient for the gravity flow system.

Water purification plants were the most impressive feature of the water supply network. The treatment process, common today, consisted of aerating the incoming water and chemical treatment to promote settling, sedimentation, filtration and distribution. Water was pumped by pipes throughout cities and towns. The roads, sewers and water supply systems played a major role in making the project possible.

Excavation and Dredging Techniques

Excavation was accomplished by the use of explosives, steam shovels and trains. Earth slides and drainage proved to be other major problems during excavations. Explosives first loosened the earth and rock for removal by a steam shovel. An accidental explosion occurred involving 18,000 kilograms of dynamite and resulting in 23 deaths. Other disasters were prevented by requiring the loading and firing of explosives on the same day as well as by the careful handling and rotation of dynamite.

At a height of construction in 1909, 68 steam shovels were in use for loading loose earth and rock onto railroad cars. Excavation progressed through a series of stepwise cuts down to the final grade. The earth from the excavation was loaded onto rail cars for dumping at other construction sites. Rail lines were constructed along the length of the Cut thus following the paths of steam shovels. Work at the Gaillard Cut through the Continental Divide would have been impossible without such technology. Workers moved 73 million cubic meters of earth and rock and coped with huge dirt slides. The solution to drainage problems required keeping water out of surrounding areas and ridding the excavation of collected water. Channels were constructed on both sides of the Cut. Water from the area flowed into channels and was carried to a reservoir or river away from the Cut. Water in the excavation area was carried off by gravity flow. Later, depths used in the excavation required pumps. The excavation always moved from low to high elevation allowing gravity flow to clear the excavation.

Slides were always a danger due to the instability of geologic materials on over-steep slopes. Terracing slopes on the either side of the Cut aided in controlling slides. Dredging was used for clearing channels through the canal (Figure 7.16). Dredging operations were divided into three divisions, including Atlantic, Pacific and Central. The construction of the canal gave a rise in bucket-conveyer dredges with bucket volumes up to 1.5 cubic meters.

Atlantic and Pacific divisions were charged with the construction of locks, dams and spillways. The Gatun locks had three of those for a lift of 26 meters. The Gatun dam was an embankment dam. Its difference lay in its great length making 2500 meters, width – 610 meters and height – 30 meters. It remains one of the most strategic canals in the world, along with the Suez. It was the result of the innovative use of sanitary and hydraulic engineering that began earlier by the great engineers of the 19th century. In September 2007, work began to expand the Panama Canal. A Panamax cargo ships typically have a DWT of 65,000–80,000 tonnes. Expected to be complete in 2014, the Panama Canal expansion project will allow ships double the size of the current *Panamax* ships (length – 289.5 m, width – 32.3 m, draft – 12 m, air draft – 57.9 m) to pass through the canal thus dramatically increasing the amount of goods that can pass through the canal.

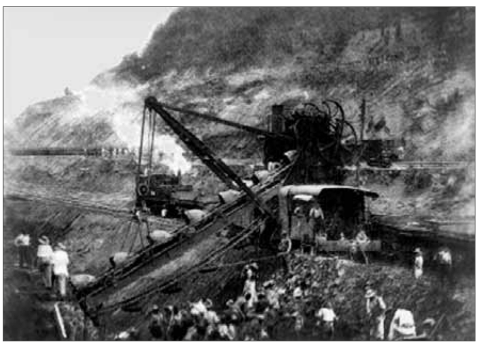

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Figure 7.16. Excavator at work, 1896

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Figure 7.17. A proposal for the Nicaragua Canal (1902)

The *Inter-Oceanic Nicaragua Canal* is a proposed waterway through Nicaragua to connect the Caribbean Sea and Atlantic Ocean with the Pacific Ocean (Figure 7.17). Such a canal would follow rivers up to Lake Nicaragua and then continue at least 10 kilometers through the isthmus of Rivas to reach the Pacific Ocean. Nicaragua's National Assembly (2013) approved a bill to grant a 50-year concession to the Hong Kong Nicaragua Canal Development Investment Company (HKND) to build the canal. The canal goes from the town of Bluefields on the Caribbean Sea (Isla del Venado) via Rio Escondido to Lake Nicaragua. Within 10–15 years, growth in global maritime trade is expected to cause congestion and delays in transit through the Panama Canal without a complementary route through the isthmus. Construction of a second canal, substantially larger than the expanded Panama Canal, across Central America, will help to solve this problem.

CHAPTER 8

Hydraulic Engineering and Sanitary

This chapter will help in

- discussing the development of water supply in the Middle Ages and 18^{th} –19^{th} centuries;
- explaining the relationship between water pollution and diseases (cholera);
- introducing measures for water treatment (filtration and chlorination).

Middle Ages

An adequate supply of drinking water and sewage disposal has been a concern of the civilization since its inception. Archaeological evidence from the 2^{nd} millennium B.C. indicates the ancient Indus Valley cities of Harappa and Mohenjodaro had water supply and sewage disposal systems. The extensive water supplies of Rome and its Imperial cities had the objective of providing adequate drinking water and sewage disposal. Certainly, the public authorities of those times were aware of the need for engineered facilities on a large scale to address water supply and sewage demands. In Asia Minor, the evidence of narrow streets and small rooms in houses tells us that crowding in the cities like Babylon and Troy was extreme. Garbage accumulated in the houses where dirt floors were continually being raised by debris. Human wastes were typically deposited in the nearest street. Water supply, generally open, was from wells, rivers and canals and was prone to pollution. Life expectancy was short with high infant mortality.

In Greece, aqueducts run underground much of its way. Further, water purity was aided by large settling basins where dirt and sediment would sink to the bottom. Water was not supplied to private homes but to public springs and fountains. The cities supervised water supply. The pollution of city water supply even could end in the death penalty. Laws in Athens and other cities required waste matter had to be carried outside the walls for a certain distance before it was dumped. Drainage sewers were often covered channels and any excess water in city supply might be used for flushing them. As for Athens, waste water was conducted through a series of canals to enrich the fields near the city.

In Rome, the purity of water in aqueducts was maintained by covered channels, reservoirs and settling basins much like those of the Greeks, but on a much greater scale. There were strict laws against pollution and water theft. Some aqueducts carried water unfit for drinking and were used for other purposes. The Romans recognized the temperature and taste of water from various aqueducts and had definite preferences. Water from aqueducts was distributed to public fountains, industries (textiles), some private houses and public baths. Furnaces, where hot springs were not available, raised water to the desired temperatures and heated the floors and walls of the baths. Excess water from aqueducts was used for removing sewage via the Cloaca Maxima. This large sewer system also doubled as a storm drain system. The Cloaca Maxima was 4.3 meter high and 3.2 meter wide in places. There was no sewage treatment, with waste discharged into the Tiber River and then to the Mediterranean Sea. Public latrines were maintained in Rome.

With the fall of the Western Roman Empire, a lack of central authority and large public funds led to a decline in public services such as road building and water supply. New cities in Western Europe depended on wells, springs and rivers. Many municipal laws forbade the contamination of rivers, but this did not work. Furthermore, wells were generally too close to cesspools and latrines. Contaminated water supply was the reason for epidemics killing thousands in the Middle Ages, mostly in the towns and cities of Western Europe. As a result of epidemics, England passed the Urban Sanitary Act of 1388 that forbade throwing filth and garbage in ditches, rivers, or waterways. In 1404, Charles III ordered all pollution of the Seine in Paris to cease.

The advent of the pump in the 15th and 16th centuries changed the picture of public water supply. Wooden pipes were preferred in many parts of Western Europe. In the 19th century, the water mains of 3,300 meters were still wooden in Augsburg (Germany). These pipes were 7 meters long and 10 to 14 centimetres in diameter. Pottery and lead pipes became less common. Lead pipes of 6.25 centimetres were soldered with tin solder. Cast-iron pipes were tried in Augsburg in 1412, but were changed to wooden pipes four years later.

18th–19th Centuries

The first clear proof that public water supplies could be a source of infection for humans was based on careful epidemiological studies into cholera in the city of London in 1854 and concerned with the spread of cholera through water supplied by the Southward and Vauxhall Company and the Lambeth Company. The former obtained its water from the Thames in the middle of London in an area almost certainly polluted with sewage, whereas the Lambeth Company obtained its water considerably upstream on the Thames, above the major sources of pollution. In one particular area served by these two companies and containing about 300,000 residents, the pipes of both companies were laid in the streets, and houses were connected to one or other source of supply. The examination of statistics on cholera deaths gave striking results. The houses served by the Lambeth Company had a low incidence of cholera, which was lower than the average population of London as a whole, whereas those served by water from the Thames in the middle of London had a very high incidence.

During the 17th to the early 19th centuries, a number of improvements in water supply were made, most of these related to improvements in filtration to remove the turbidity of waters. During this same period, the germ theory of disease became firmly established. In 1884, Koch isolated *Vibrio cholera,* the causal agent of cholera.

Importance of Water Filtration

In 1892, a study on cholera by Koch in German cities provided some of the best evidence of the importance of water filtration for protection against this disease. The cities of Hamburg and Altona both received their drinking water from the Elbe River, but Altona used filtration, since its water was taken from the Elbe below the city of Hamburg and hence was more grossly contaminated. Koch traced the incidence of cholera in the 1892 epidemic through these two cities. It was assumed that climate, soil and other factors were identical, with the principal variable being the source of water. The results of this study were clear-cut: Altona, even with an inferior water source, had a markedly lower incidence of cholera than Hamburg. Since by this time it was well established that cholera was caused by intestinal bacteria excreted in large numbers in the faeces, it was concluded that the role of filtration was to remove contaminating bacteria from water.

Experiments on water filtration were also carried out in the United States. The experiment station of the Massachusetts State Board of Health was established in 1887 in the city of Lawrence where the treatment of water as well as sewage was considered by an interdisciplinary group that included engineers, chemists and biologists. One important technological advance making water filtration adaptable to even turbid sources of water was the use of chemical/coagulation filtration processes. While the Lawrence experiment was going on, an epidemic of typhoid swept through the city thus hitting especially hard at those parts that were using the Merrimac River as its water supply. As a result, the city of Lawrence built a sand filter, and its use led to a marked reduction in the typhoid fever incidence. Filtration led to the elimination of turbidity and colour from water and to the removal of about 99% of the bacteria present.

Chlorination: Most Significant Advance in Water Treatment

Excellent water quality had been well established by filtration, but the most important technological advance in water treatment was yet to come and appeared as water chlorination the introduction of which, after 1908, provided a cheap, reproducible method of insuring the bacteriological quality of water. Chlorination has come down to us today as one of the major factors insuring the safety of our drinking water. Calcium hypochlorite $Ca(ClO)_2$ was manufactured industrially and used in paper mills and textile industries. It was a cheap chemical and, hence, readily adaptable to use on the large scale necessary for drinking water. The first practical demonstration of its use in water supply was introduced in Chicago 1908.

Chlorination has come as one of the major factors insuring the safety of drinking water. Moreover, it has been found a possibility of removing most of remaining bacteria so that the water supplied can be as easily and certainly held within the one-tenth of 1% of those in raw water. Soon after the introduction of chlorination, it became possible to obtain firm epidemiological evidence that the cities chlorinating water had lowered incidences of typhoid fever. Chlorination was introduced at about the time that the adequate methods of the bacteriological examination of water had been developed, thereby permitting an objective evaluation of the efficiency of treatment.

Present methods of water treatment (Figure 8.1) actually kill not just the bacteria, but the very life of the water. Chlorine eradicates all types of bacteria, beneficial and harmful alike. What science and manufacturers of traditional water filters completely overlook is that water is a life-carrier and is itself alive and cannot fulfill its original and intended natural function of life-giving, vitalizing, and even life-creating. All living organisms need to continue being nurtured by healthy living water in order to thrive.

Water supply system in Vilnius is the oldest one in Lithuania, dating back to the year 1501. Since the 16th century the city has had a gravitational water supply system from the natural springs. Water from these springs spurted out and was conducted by open canals and wooden pipes (Figure 8.2) down into the centre of Vilnius to the houses of the prosperous dwellers. Water system of ancient times relied on gravity for the supply of water and using pipes usually made of wood.

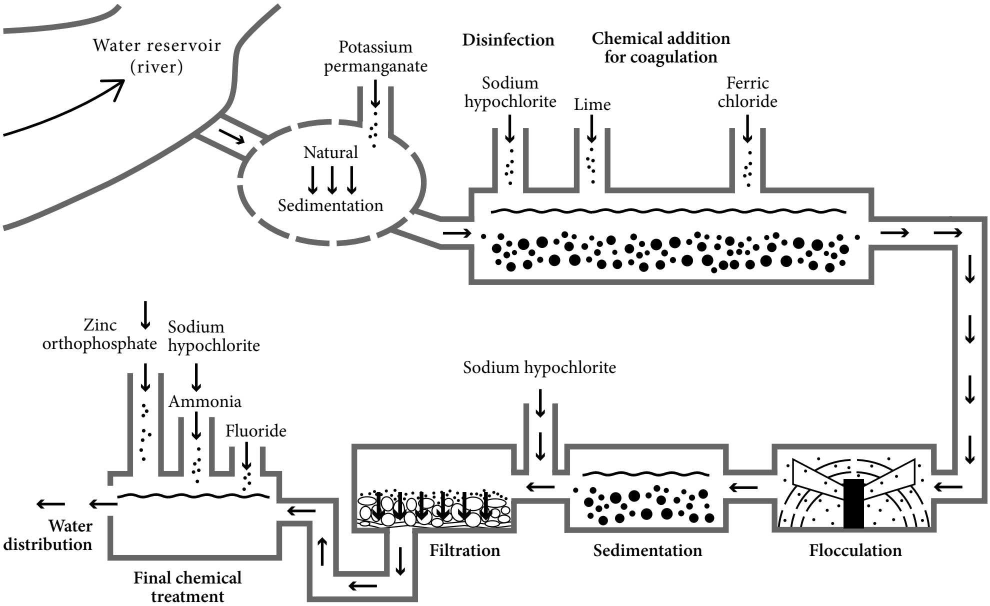

Figure 8.1. Typical drinking water treatment process

Figure 8.2. Hollowed wooden log pipe (National Museum of Lithuania, The Old Arsenal)

Hollowed wooden logs wrapped in steel banding were used for plumbing pipes. Logs were used for water distribution close to 500 years ago. The capacity of the water supply system in Vilnius was approximately 1,500 m^3 per day. Now – 90 thousand m^3 water per day. There are many natural springs in the city and its surroundings. Such water has a good taste and low concentrations of iron and manganese. On the other hand, bacteriological pollution is substantial because of urbanisation and pollution. Vilnius consumers use only underground water from deep wells. The depth of the wells is 40–180 m. The water treatment facilities are installed, where iron and manganese are removed. Concentration of iron before the treatment amount to 1.1 mg/l and 0.00 mg/l after the treatment.

CHAPTER 9 Manufacture and Production

This chapter will help in

- discussing the development of early machines (pole lathes, cross-slide lathes, screw cutting lathes, fusees, etc.);
- explaining *cubit* – the first "standard" of measurement;
- talking of the birth of the machines used in the 18^{th}–19^{th} centuries (boring, steam hammers, turret lathe, grinding and milling machines);
- introducing the tools produced of high speed steel, chromium, cobalt, tungsten alloys, carbides and ceramics;
- describing joining by welding, weldability, arc, resistance, gas, laser, electron beam as well as ultrasonic, electroslag, explosion welding;
- presenting joining by brazing, soldering (capillary action, wetting) and braze welding.

Manufacturing processes feature the grounds for satisfying the needs of contemporary societies. Manufacturing consists of making products from raw materials in various processes, using different machines and operations organized according to the well-prepared plan. Therefore, the manufacturing process is composed of a proper use of such resources as materials, energy, capital and people. Nowadays, manufacturing is a complex activity merging people working in various professions and carrying out miscellaneous jobs using diverse machines, equipment and tools and automated to a various extent, including computers and robots. Traditional mechanical products are consumer goods, including those made from leather, wood, textiles and polymer materials as well as products related to food, medicines, electronic and information technology.

Early Machines

By the end of the Roman Empire, virtually all the forms of modern hand tools had been devised. The second step in the development of manufacturing was the introduction of the mechanical means of cutting and forming to take advantage of high cutting speed possible with iron tools. Drilling was the first mechanically assisted cutting operation. The earliest illustration of this type of a lathe, on a wall of the Egyptian tomb (3rd century B.C.), shows an assistant holding each end of the cord to give the rotational movement to a spindle the driving methods of which advanced slowly. The difficulty of manipulating the bow while guiding the tool with intermittent cutting calls for a very high degree of manual skill and is only possible to make light cuts.

As the size and complexity of work for the lathe increased so did the need for growing rigidity and more power, which was met by a heavier construction of the wooden frames of lathes and the development of the pole lathe. This type of drive probably existed in the 12th century (Figure 9.1).

The pole lathe, with its intermittent cut, was not adequate for turning metal and, as the need for machined metal products increased, the continuous method of driving was developed, first of all through the use of a large wheel in separate bearings carrying round its periphery a cord that also passed round the work spindle. The large wheel was turned by an assistant using a cranked arm or by using a treadle and crankshaft. Such machines had to enable the use of other power sources: horse gins, water wheel, steam engine and electric motor. The development of continuous drive also made possible the control of the cutting tool relative to work through the systems of gears, screws and guides (Figure 9.2).

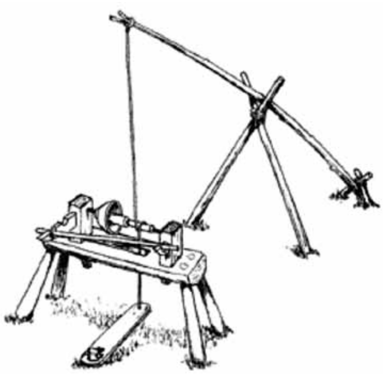

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Figure 9.1. Scheme for a pole lathe

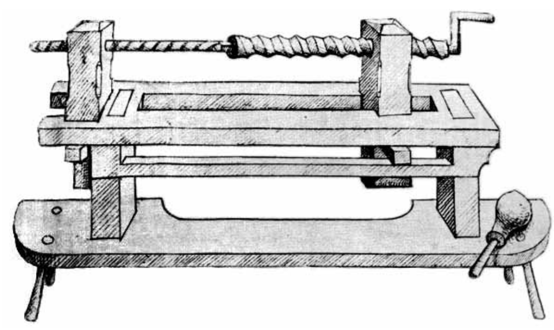

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Figure 9.2. Thread-cutting machine (Battison 2010)

Measurements

From the earliest days of the man's use of tools, measuring the size and shape of things produced has been of prime importance to satisfy the performance required. The first "standard" was the *Egyptian Royal Cubit* equivalent to the length of the Pharaoh's forearm and a palm made of black granite. This master standard was subdivided into finger widths, palm, hand, large and small spans, one remen (20 finger widths) and one small cubit equivalent to six palms. The small cubit was used for general purposes and made in granite or wood for working standards. These principles and the use of cubits and their subdivisions became the basis of Roman, Greek, Middle East and later European measures. Once the standards of length were established, these were used for checking parts for accuracy, measuring tools and applying for transferring sizes in comparison with the standard. Calipers, dividers and proportional dividers were evolved by the Greeks and Romans some 3000 years ago. These instruments, with the cubit measuring stick, made possible accurate calculations by measuring the shadow cast by the sun, determining the heights of buildings, agreeing the time of the day, establishing the calendar and navigating with reference to the stars.

Time measurement based on the observation of the sun's passage has always exercised the mind of man, and therefore many methods have been explored. The first all-mechanical clocks were made in Europe in 1364, which incorporated elliptical gear wheels and sun and planet gears, all cut by hand.

These construction methods, while adequate for large public clockwork, were not suitable for smaller timekeepers the people sought. The development of design led to improvements in machining, including accurately cut gears, turned spindles and screws that were in advance of general manufacturing by 300 years. During this period, the spring-driven pocket watch made its appearance. An important component of the watch – a *fusee* was invented (1450). The fusee is a cone-shaped pulley with a helical groove around it, wound with a cord or chain attached to the mainspring barrel. Fusees were used from the 15^{th} to the early 20^{th} century to improve timekeeping by equalizing the uneven pull of the mainspring as it ran down. Simple fusee engines, employing a screw-controlled linear motion but hand-controlled infeed to the profile, were designed from 1740.

General Machine Tools

Clockmakers also required accurate screw threads and these were cut by machine since c. 1480. Many machines were constructed on the master screw principle; however, greater flexibility in screw making was achieved by the use of a sliding spindle controlled by a set of master threads brought into mesh as required. Instrument makers had special problems involving the need for accurate screws to be used in obtaining precise linear and circular divisions. Work mounting required heavy cutting loads in machining metal, and were also the basis for the design of metal cutting machine tools.

The greatest users of machine tools of the day were the engine builders. In 1797, the true industrial lathe came into being. The machine was the synthesis of all previously recognized desirable features: rigid construction, prismatic guide bars for the carriage, changes in gear drive from the spindle to lead screw, graduated infeed for the tool holder and substantial centres for work holding readily adjustable for length by moving the tailstock.

While this general development of the lathe was in progress, special needs for manufacturing in different fields was being met in a variety of ways. The need for accurate gear wheels in clock and instrument making resulted in the dividing engine. A formed rotary file cuts the teeth of the gear in turn indexed using a plate containing a number of hole circles. Very little difference is apparent between this machine and those produced 100 years later.

Engineering grew due to demand for producing the engines of war. The water-powered cannon boring mill of the 16th century was used for finishing the cored hole in the casting. This method continued until the 18th century. In 1795, a horizontal boring mill that employed a stationary boring bar and rotated the cannon in bearings was invented. In 1776, John Wilkinson produced a cylinder boring machine producing the cylinder sufficiently accurate to satisfy the design of James Watt's steam engine. The machine consisted of a carriage on which the cylinder to be bored was mounted with the boring bar, running through its centre and fixed in bearings at each end one of which was driven through gearing from a water wheel. The boring bar was hollow, with a rod running through its centre, which advanced the cutter head along the revolving bar by means of a longitudinal slot and a slider attached to the cutter head and traversed by a rack and gear with a weighted lever to provide moving force. In 1800, a cylinder of 162.5 cm in diameter was completely machined on the boring mill in 27.5 working days.

A vertical boring mill was also built in 1854 to bore four cylinders for Brunel's Great Eastern steamship to 2.13 m in diameter. This engine produced 2000 hp (1491.4 kW) and was the most powerful in the world at that time.

In 1820, a planning machine in which work was clamped on a reciprocating table passing beneath a cutting tool that could be moved horizontally and vertically was introduced. This sort of machine was the key to satisfying the need for accurate plane surfaces in building lathes as well as machines and engines applied for special purpose.

The steam hammer appeared around 1839 to forge the paddle wheel shafts of 75 cm in diameter for a steamship.

To provide for the required accurate measurement of producing standard gauges used in manufacturing in the middle of the 19th century, two different measuring machines were constructed in the UK: the first was designed for reading up to 0.00254 mm and the second, for reading up to 0.0000254 mm. Both were comparators, set first to the standard and then on the gauge to be measured. In 1841, the standardization of screw threads appeared. A table was produced giving the pitches of the screws of different diameters and a constant proportion between depth and pitch by adopting an angle of 55° for the 'V' profile. By 1858, the standardization of screw threads had been implemented. In some time in continental Europe, the decimal scale or the metric thread system was implemented and an angle of 60° for the 'V' profile was adopted.

The need for many small parts in mass-produced goods encouraged the use of specialized machines designed for the purpose in small workshops. Ball bearings were gradually introduced into bicycle manufacture, and cycle bearing companies were formed producing balls either by rolling hot metal or turning from bar on special machines. All balls were finished by grinding, and cups and cones were turned, case-hardened and ground. A vast number of screws required for the percussion locks of pistols caused demand for designing and building the first *turret lathe* in 1845. This had a horizontal axis and carried eight tools that could be fed into work in turn to perform eight successive operations without stopping to change tools. This idea was so time-saving that it was rapidly taken up by other manufacturers. Many of the mass-produced goods also required gears and, more importantly, new manufacturing machinery also depended on the ability to transfer motion and change speeds between shafts accurately and without vibration through gears made in large quantity.

The cast gears used extensively in early machines and continuing in textile industry were no longer satisfactory for high-speed accurate machining by new machine tools. In 1839, J. Bodmer patented a gear cutting machine capable of cutting internal gears, spur and bevel in addition to external spur, worm wheels and racks.

The described generating method of producing the gear tooth profile was first applied practically in 1877 and used a single point reciprocating tool. All motions to generate the gear tooth were being given to the blank by various attachments to make spur or bevel gears with *epicycloidal* or *involute teeth*. In 1835, a worm cutter as a 'hob' to make spiral gears was created.

By the end of the 19th century, general machine tools were established in their operating principles. The features of interchangeable manufacture were well known and mass production machinery ready for explosion in manufacturing in the 20th century with the proliferation of the motor car and aero-plane.

The early 20th century also saw the introduction of an individual electric motor drive for machines, which was eventually to eliminate the 'forest of belts' typical of all early production factories. The two significant machine tools of the first half of the 20th century were a heavy production grinding machine by Norton in 1900 and the Bridgeport turret milling machine. The former was designed with the size and power to plunge grind cylindrical parts using wide wheels to eliminate finish turning in the lathe. The Bridgeport machine consists of a multi-purpose head for drilling and milling with its built-in motor drive mounted on an arm that can be swung over the work table.

In 1921, the jig borer (Figure 9.3) was produced in Switzerland. Scaled up to engineering standards, it gave 0.00254 mm standard in locating holes and was invaluable in the manufacture of gauges and jigs. The jig borer machine, with the hydraulic work table and optical scale, was introduced in 1934. The general purpose centre and turret lathes continued to be improved to give greater accuracy and flexibility of operation, and by the late 1920s had hardened and ground beds, thrust bearings using balls and rollers and special turret tools. The major improvements were made in spindle speed capacity and rigidity to match a continued increase in the capability of cutting tool material.

In 1900, cutting mild steel at 36.5 m/min with the cutting tool red hot (*red-hot property* – is being hard enough to cut metals even when heated to a dull-red color) was demonstrated, and this revelation was so fundamental that all types of machine tools were needed redesign so that to increase rigidity, speed range and power in order to take advantage of the new tool material. The introduction of Stellite, chromium, cobalt, tungsten alloys (1917), carbides (1926) and ceramic tools were to have similar effects later. A comparison of cutting speeds possible with these tool materials for grey cast iron shows *high speed steel* containing tungsten (HSS) 22.8 m/min., stellite – 45.7 m/min., carbide – 122 m/min. and ceramic – 183 m/min. Many tool types were produced from tungsten carbide and titanium carbide developed in 1938. They were all very brittle, and cutting tools were made with steel shanks and tips of carbide brazed to them. Modern tools use specially designed tips to be carried in a holder fitted to the tool shank. In 1914–15, the systems of selecting grinding wheels in terms of grit,

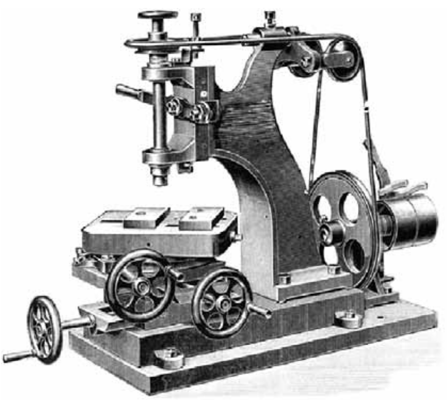

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Figure 9.3. An early jig borer

grade and bond were devised in England and USA; however, carbide tools were too hard to be ground by ordinary abrasive wheels and, only after production, a small diamond bonded wheel (1934) was possible to deal with them.

Lithuanian Input

In 1990, Lithuania accounted for only 0.3% of the territory of the Soviet Union and 1.3% of its population, but it turned out a significant amount of the industrial and agricultural products of the U.S.S.R., including 22% of electric welding apparatus, 11,1% of metal-cutting lathes, 2.3% of mineral fertilizers, 4,8% of alternating current electric motors, 2% of paper, 2.4% of furniture, 5.2% of socks, 3.5% of underwear and knitwear, 1.4% of leather footwear, 5.3% of household refrigerators, 6.5% of television sets, 3.7% of meat, 4.7% of butter, 1.8% of canned products and 1.9% of sugar.

Figure 9.4. Desktop drilling machines produced in Vilnius factory “Praktika” (1979)

Fastest-growing industry, in comparison to the total industrial production, was automotive, tool and metal processing industry. Only Vilnius counted were four big metal-cutting lathe factories. The first one producing metal-cutting lathes was established 1947. The same year, it manufactured only 13 desktop drilling machines (Figure 9.4).

a) b)

Figure 9.5. Milling machines produced in Vilnius factory “Žalgiris”: a – for horizontal milling (1978), b – for vertical milling (1979)

Figure 9.6. Coordinate drilling machines produced in Kaunas factory (1964)

For the period 1960–1991, Lithuania operated 6 metal lathe factories (4 in Vilnius, 1 in Šiauliai and 1 in Kaunas). In 1980, the factories produced 24800 lathes per year for grinding, milling, drilling and special applications (Figure 9.5).

Coordinates drilling machines are high precision machines whose application lies in high precision tool manufacturing of precision engineering and machine tools manufacturing. They are mainly used for precision drilling and light milling operations. The single-column-machine is equipped with a cross table. The cross table guides the movement in the horizontal X and Y axes, the drilling unit performs vertical movement (Z-axis) in addition with the vertical feed of the drilling spindle (Figure 9.6).

Vilnius State Grinding Machines factory produced universal high-precission cylindrical grinding machines (Figure 9.7). Vilnius branch of the Experimental Scientific Research Institute of Machine Tools was established in 1961. Over the period of 50 years, the team of the institute was engaged in the creation, production and implementation of precision machine tools, optoelectronic measuring systems and components, dividing machines and CMMs. For the period 1970–1990, the company produced more than a half of encoders and CMMs produced in the former U.S.S.R.

Photoelectric linear and rotary position encoders were widely used in the former Soviet Union, and some of them have been operating until now in machine tool and metalworking industries. After joining the European Union,

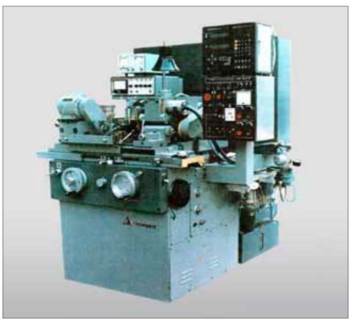


Figure 9.7. High-precission cylindrical grinding machines produced in Vilnius (1990)

Figure 9.8. Electro-discharge jig-copying-broching machine produced by Joint stock company "Kauno stakles" (Lithuania)

Lithuanian machine tool manufacturing has been specializing in manufacturing high precision jig boring, jig grinding, EDM, boring, drilling, milling machine tools and accessories, electro – discharge jig-copying-broaching machines, etc. (Figure 9.8).

Joining by Welding

Welding is a fabrication process that joins materials by causing *coalescence*. This is often done by melting workpieces and adding a filler material to form a pool of a molten material (weld pool) that cools to become a strong joint with pressure sometimes used in conjunction with heat, or by itself, to produce the weld. This is in contrast with soldering and brazing, which involves melting a lower-melting-point material between workpieces to form a bond between them without melting workpieces. Many different energy sources can be used for welding, including a gas flame, an electric arc, a laser, an electron beam, friction and ultrasound. Welding may be performed in many different environments, including open air, under water and in outer space. Welding is the most efficient method for joining metals and alloys. Over 50% of the gross national product of the developed countries is related to welding in one way or another. The welded products include heat exchangers, tanks, pressure vessels, sheet metal, prefabricated metal buildings, architectural and ornamental works, shipbuilding, production of aircrafts and spacecrafts, railroads, automobiles, trucks, buses, etc. All metals can be welded, but some are easier to weld than others. In other words, they possess *weldability* – the capacity of materials to be welded under the imposed fabrication conditions. All metals and alloys cannot be joined by each welding process. Certain metals require specific precautions and procedures.

Joining metals by heat and filler was practiced in bronze statuary c. 3000 B.C. and the first welded iron joint appears in a headrest from the tomb of Tutankhamun, c. 1350 B.C. During the Iron Age, the Egyptians and people in the eastern Mediterranean area learned to weld pieces of iron together by forge welding (Figure 9.9, a). Many tools made approximately 1000 B.C. have been found. Pattern welding is the practice in sword and knife making of forming a blade of several metal pieces of differing composition that are forge-welded together and twisted and manipulated to form a pattern (Figure 9.9, b).

The heat required achieving local melting of iron and steel had to await the production of oxygen in 1774, acetylene in 1800, and the Linde's method of extracting oxygen from liquid air in 1893. The oxy-acetylene welding torch appeared in France in 1903 and gave an adjustable flame of 3250 °C. A patent for

a)

b)

Figure 9.9. A typical smithy (a) and pattern welded knife blade (b)

welding by electric arc, filed in 1849, devised a method using carbon electrodes, but it is the use of the consumable bare steel rod by the Russian N.G. Slavianoff in 1888, which makes this device recognized as the beginning of metal arc welding. Considerable skill was required to strike and maintain the arc with a bare metal electrode, and coatings were developed to improve arc stability and provide a protective layer for the weld. Later in 1890, (C. L. Coffin) arc welding method that utilized a metal electrode was invented. The process, like *Manual metal arc*, deposited melted electrode metal into the weld as filler. Around 1907, Oscar Kjellberg released the first coated electrodes and dipped iron wire into the mixtures of carbonates and silicates to coat the electrode. In 1909, an asbestos yarn-covered electrode was used. Many different materials, including cellulose, mineral silicates and carbonates, have been employed as coatings to suit welding under different conditions. In 1927, the development of the extrusion process reduced the cost of coating electrodes while allowing manufacturers to produce more complex coating mixtures designed for specific applications. In the 1950s, manufacturers introduced iron powder into the flux coating thus making it possible to increase welding speed.

Manual metal arc is the most used welding process, but Elihu Thompson's invention (1886) of using electric power to achieve welded joints by resistance heating is very significant, as it took welding from the maintenance function into production. This butt-welding machine was the forerunner of the resistance spot welder between 1900 and 1905 and the seam welder. Welding and joining processes are grouped as follows:

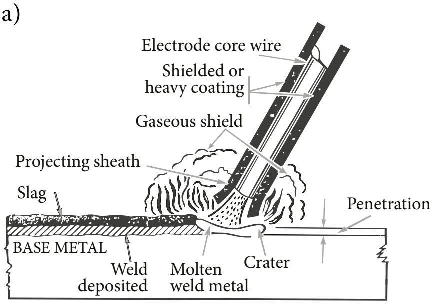
- arc welding;
- resistance welding;
- gas welding;
- forge welding;

– other welding processes (laser, electron beam, ultrasonic, explosion);
– brazing, soldering and braze welding.

Forge welding, or fire welding, uses the blacksmith's hearth. Before making a fire-welded joint, it is essential to have a clean clinker-free fire, and the parts to be joined must be upset, scarfed and are heated to creamy white heat and removed from fire and tapped on the edge of the anvil to shake off dirt. Care must be taken to ensure that the scarves are in a proper relationship to each other before the first blow is struck in the centre of the work. It must now be repeatedly hammered, working from the centre, to drive out the molten scale. It may be necessary to reheat for the weld to be completed.

Manual metal arc (MMA) welding, also known as shielded metal arc welding (SMAW) or informally as stick welding, is a manual arc welding process that uses a consumable electrode coated in flux (Figure 9.10). An electric current (alternating current or direct current) from a welding power supply is used for forming an electric arc between the electrode and metals to be joined. As the weld is laid, the flux

a)



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b)

Photo by A. V. Valiulis

Figure 9.10. Manual metal arc welding: a – sketch of the process; b – welding transformer for 300 A welding current

Figure 9.11. Welding equipment produced in Vilnius Electrical Welding Equipment factory: electric motor and engine driven welding machine

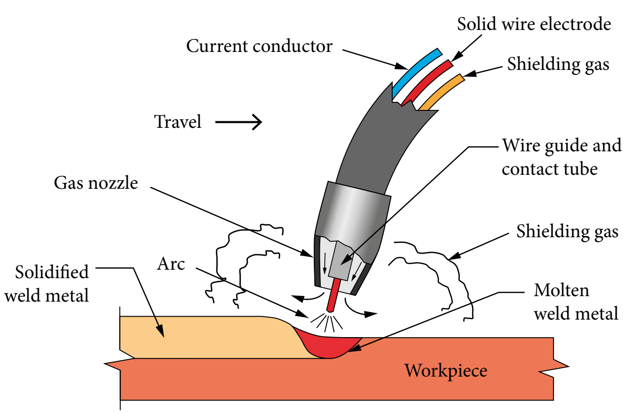
coating of the electrode disintegrates giving off vapours that serve as shielding gas and providing a layer of slag, both of which protect the weld area from atmospheric contamination. The producer of welding equipment for manual arc welding during 1950–1990 was Vilnius Electrical Welding Equipment factory (Figure 9.11).

Gas metal arc welding (GMAW), also known as metal inert gas or MIG welding, is a semi-automatic or automatic process that uses continuous wire feed as an electrode and an inert or semi-inert gas mixture to protect the weld from contamination (Figure 9.12). Gas shielding of the arc and weld metal from atmospheric contamination were considered from 1919, and in the 1930s, an interest cantered on inert gases. With these developments, electric welding became the key process in the rapid fabrication of locomotives, ships and automobiles and played a huge part in wartime production effort required from 1939.

The inert gas metal arc process was introduced in 1948 and used an electrode of metal consumed in the process as filler. Many other gases have been used as shields, notably argon and CO_2, and it is particularly useful and economic in welding aluminum using small-diameter wire and direct current with the electrode as positive. Since the electrode is continuous, welding speeds are greater for GMAW than for SMAW.

Flux-cored arc welding (FCAW) was first developed in the early 1950s as an alternative for shielded metal arc welding (SMAW). *Flux-cored arc welding* uses similar equipment but applies a type of wire consisting of a steel electrode surrounding a powder fill material. This cored wire is more expensive than the standard solid one and can generate fumes and/or slag but permits even higher welding speed and greater metal penetration. One type of FCAW requires no shielding gas. This is made

a)



b)

Figure 9.12. Gas metal arc welding: a – scheme for the welding process; b – modern GMAW equipment

possible by the flux core in the tubular consumable electrode. However, this core contains more than just flux – it also contains various ingredients that, when exposed to the high temperatures of welding, generate shielding gas for protecting the arc.

Gas tungsten arc welding (GTAW), or tungsten inert gas (TIG) welding, is a manual welding process that uses a non-consumable tungsten electrode, an inert or semi-inert gas mixture and a separate filler material (Figure 9.13). During and after World War II, many developments took place in standard electric welding processes such as inert gas tungsten arc welding. GTAW had the addition of an insulated water cooled nozzle to form a chamber around the electrode to produce arc plasma in the form of a flame when the arc is struck from the electrode to the nozzle.

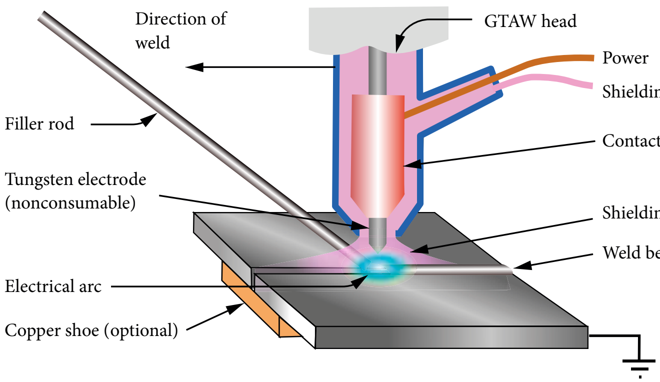


Figure 9.13. Gas tungsten arc welding

GTAW can be used on nearly all weldable metals, though it is most often applied to stainless steel and light metals. It was not until 1940 that the Northrop Aircraft Company experimented with tungsten electrodes and helium gas, which was used successfully on thin gauge stainless steel. Further advances using other gases followed to give techniques able to cope with welding aluminum alloys and other difficult metals.

Plasma arc welding (PAW), as a welding process, was introduced to welding industry in 1964 as a method of bringing better control to the arc welding process in a lower current range. PAW also uses a tungsten electrode but applies plasma gas to make the arc. The latter has had the addition of an insulated water cooled nozzle to form a chamber around the electrode to produce arc plasma in the form of a flame when the arc is struck from the electrode to the nozzle. The arc is more concentrated than the GTAW arc and is much faster thus making transverse control more critical and generally restricting the technique for a mechanized process. It can be applied to all the same materials as GTAW, except magnesium, and the automated welding of stainless steel is one important application of the process. This is a non-transferred arc used for metal spraying with the addition of powdered metal to plasma. If the arc is struck between the electrode and work, it is called a *transferred arc*. The arc is constricted in passing through the nozzle orifice and used at low currents for sheet metal at about 400 A for welding thick metal employing the keyhole technique. It requires an additional inert gas shield when used for welding. The plasma arc cutting process, introduced in 1954, can be applied to cutting metals that form refractory oxides such as stainless steel, cast iron, aluminum and other non-ferrous alloys.

The *submerged arc welding* (SAW) process, developed for shipbuilding industry in the USA and USSR in the mid-1930s, employs powder flux to completely cover the weld pool and the end of the electrode wire (Figure 9.14). It allows a high welding current and low electrode usage in automatic processes. SAW is a high-productivity welding method in which the arc is struck beneath a covering layer of flux.

This increases arc quality, since contaminants in the atmosphere are blocked by flux. The slag that forms on the weld generally comes off by itself, and combined with the use of continuous wire feed, the weld deposition rate is high. Working conditions are much improved over other arc welding processes, since flux hides the arc and almost no smoke is produced. The process is used for large products and in the manufacture of welded pressure vessels.

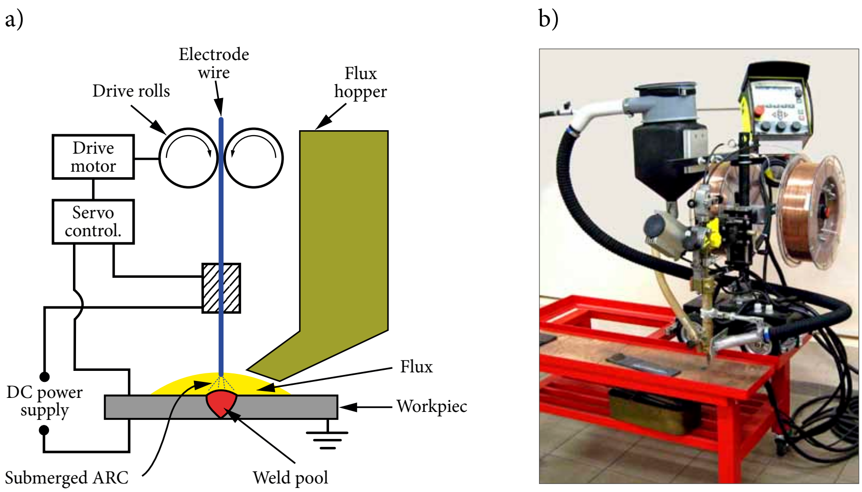


Photo by A. V. Valiulis

Figure 9.14. Submerged arc welding: a – scheme for the welding process; b – modern SAW equipment

Electron beam welding (EBW) is a fusion welding process in which a beam of high-velocity electrons is applied to the materials being joined (Figure 9.15). Workpieces melt as the kinetic energy of the electrons is transformed into heat upon impact, and filler metal, if used, also melts to form a part of the weld. Welding is often done under conditions of a vacuum to prevent the dispersion of the electron beam. The first practical electron beam welding machine began operation in 1958.

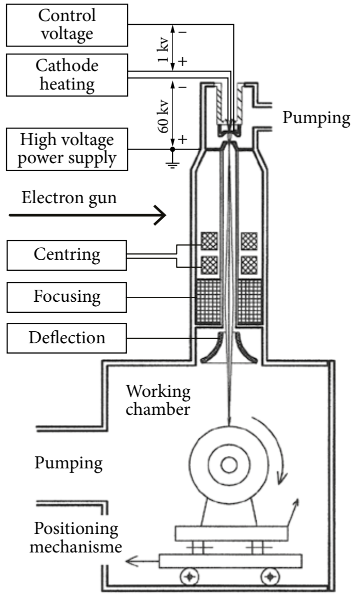


Figure 9.15. Electron beam welder

Laser beam welding (LBW) is a welding technique used for joining the multiple pieces of metal through the use of a laser. The beam provides a concentrated heat source allowing for narrow, deep welds and high welding rates. LBW has high power density (on the order of 1 MW/cm²) resulting in small heat-affected zones and high heating and cooling rates. The spot size of the laser can vary between 0.2 mm and 13 mm, though only smaller sizes are used for welding. Some of the advantages of LBW, in comparison to EBW, are as follows: the laser beam can be transmitted through air rather than requiring a vacuum, the process is easily automated with robotic machinery, x-rays are not generated and LBW results in higher quality welds. There are many types of laser beam welding but the most popular types in industry are Nd:YAG (neodymium-yttrium aluminum garnet) laser using a man-made crystal as its active medium and producing light on a 1.06-micron wavelength, Carbon Dioxide Lasers using a mixture of gases, including CO_2 as the active medium and producing light on a 10.6-micron wavelength; the Diode Laser using a semi-conductor diode material as its active medium can be manufactured to produce one of several wavelengths. In 1970, laser technologies were first implemented in industry: in 1975 – the first applications of laser beam cutting in sheet fabrication industry, in 1983 – introduction into the market of 1kW CO_2 lasers and in 1984 – the first applications of laser beam welding in industrial serial production.

Ultrasonic joining requires a transducer assembly operating at about 20 kHz (source of the ultrasound) coupled with a sonotrode the tip of which is placed in contact, usually under a clamping load of 1–10 Nmm^{-2}, with the workpiece. The generated heat is localized at the interface, creating a temperature of up to 600 °C when using aluminum inter-layers. The bonding mechanism relies on the vibratory shear

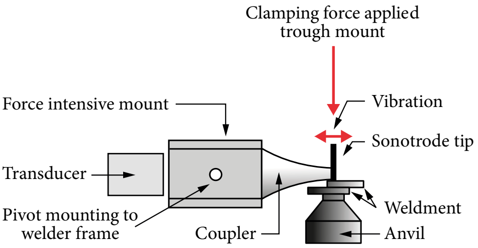


Figure 9.16. Ultrasonic joining

stress of metal exceeding its elastic limit, coupled with the breakdown of surface oxide films exposing atomically clean metal (Figure 9.16). Clamping force exerts plastic deformation on metal, which increases the interfacial contact between metal and ceramics. Then, mechanical keying occurs across the interface and the joint is formed, perhaps along with some chemical interactions. One limitation of the ultrasonic joining of ceramics is that only films or thin sheets of metal can be joined. In 1960, the first metal ultrasonic welding machine (US) was patented. The patent for the ultrasonic method of welding rigid thermoplastic parts was issued in 1965. The first car made entirely out of plastic was assembled using ultrasonic welding in 1969. Automotive industry has used it regularly since the 1980s. Ultrasonic welding processes can lead to highly durable bonds between light metals and *Carbon-fiber-reinforced polymer* (CFRP) sheets.

Electroslag welding is a process in which heat is generated by an electric current passing between the consumable electrode (filler metal) and the workpiece through molten slag covering the weld surface (Figure 9.17). Prior to welding, the gap between two workpieces is filled with welding flux. Electroslag welding is initiated by an arc between the electrode and the workpiece (or starting plate). Heat, generated by the arc, melts fluxing powder and forms molten slag, which, having low electric conductivity, is maintained in liquid state due to heat produced by the electric current. Slag reaches a temperature of about 1930°C, which is sufficient for melting the consumable electrode and work piece edges. Metal droplets fall to the weld pool and join the work pieces. The process was patented by Robert K. Hopkins (US) in 1940 and developed and refined at the Paton Institute, Kiev, USSR during the 1940s.

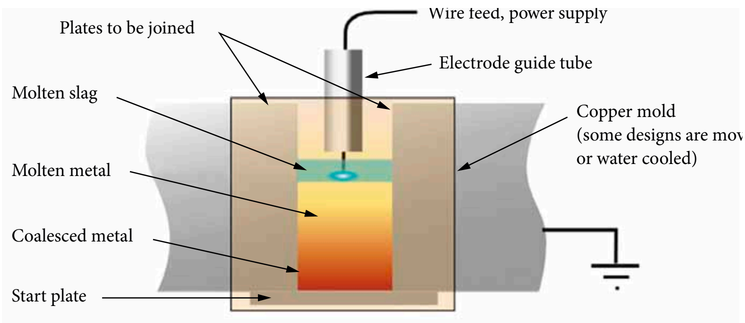


Figure 9.17. Electroslag welding

Figure 9.18. Resistance seam welding machine

Resistance welding is a process in which workpieces are welded due to a combination of pressure applied to them and localized heat generated by a high electric current flowing through the contact area of the weld. To create heat, copper electrodes pass the electric current through workpieces. The generated heat depends on the electrical resistance and thermal conductivity of metal and the time that the current is applied. The generated heat is expressed by the equation $E = I^2 \cdot R \cdot t$, where E is heat energy, I is the current, R is electrical resistance and t is time the current has been applied. Heat produced by the current is sufficient for the local melting of the workpiece at the contact point and formation of a small weld pool ('nugget'). Molten metal is then solidifies under pressure and joins the pieces. AC electric current (up to 100000 A) is supplied through copper electrodes connected to the secondary coil of a welding transformer. The most popular methods of resistance welding include Spot Welding (RSW), Flash Welding (FW), Resistance Butt Welding (UW) and Seam Welding (RSEW) (Figure 9.18).

Friction welding, introduced in the Soviet Union in 1953, uses heat developed by contra-rotating parts held in contact at high speed until temperature is high enough for welding when rotation is stopped and the ends forced together. *Friction-stir welding* (FSW) is a solid-state joining process (metal is not melted) that uses a third body tool to join 2 faying surfaces (Figure 9.19). Heat is generated between the tool and material which leads to a very soft region near the FSW tool. It then mechanically intermixes the two pieces of metal at the place of the join, and the softened metal (due to the elevated temperature) can be joined using mechanical pressure (applied by the tool), much like joining clay or dough. It is primarily used on aluminum, and most frequently, on extruded aluminum (non-heat treatable alloys), as well as on structures that want superior weld strength without a post weld heat treatment. It was invented in the UK in 1991.

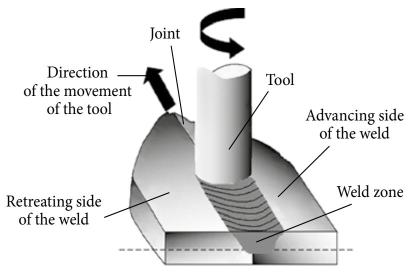


Figure 9.19. Scheme for the friction stir welding process

Oxyfuel gas welding methods use a gas flame as a source of heat (Figure 9.20). In the oxy-fuel gas welding process, heat is produced

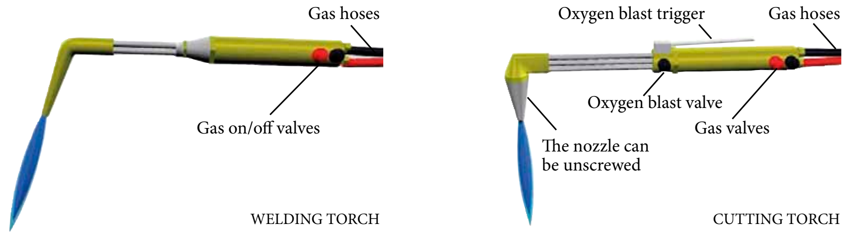


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Figure 9.20. Oxyfuel gas welding

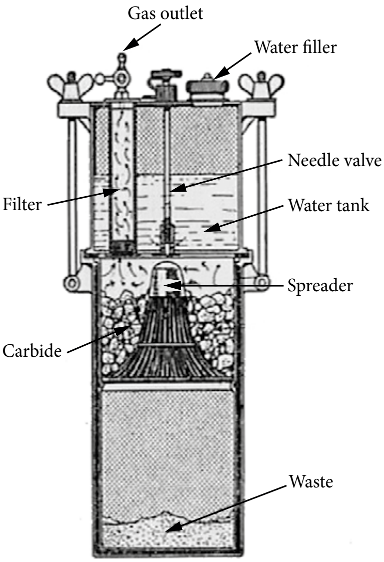

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Figure 9.21. A drip-type carbide acetylene generator

by burning combustible gas, such as acetylene (C_2H_2), mixed with oxygen. Gas welding is widely used in maintenance and repair work because of the ease in transporting oxygen and combustible gas cylinders. French engineers Edmond Fouché and Charles Picard became the first to develop oxygen-acetylene welding in 1903. A common propane/air flame burns at about 2,000 °C, a propane/oxygen flame burns at about 2,500 °C and an acetylene/oxygen flame burns at about 3,500 °C.

An acetylene generator is an apparatus for producing acetylene, C_2H_2, by decomposing calcium carbide, CaC_2, with water by reaction

$$CaC_2 + 2H_2O = C_2H_2 + Ca(OH)_2.$$

The acetylene generator can produce 0.8–150 m^3 of acetylene per hour (Figure 9.21).

Joining by Brazing and Soldering

Soldering and brazing are thought to have arisen very early in the history of metal working, probably before 4000 B.C. Sumerian swords from ~3000 B.C. were assembled using hard soldering.

Brazing is a metal-joining process whereby *filler* metal is heated above the melting point and distributed between two or more close-fitting parts by *capillary action*. Liquid filler metal is protected by a suitable atmosphere, usually flux, flows over the base metal (known as wetting, Figure 9.22) and is then cooled to join workpieces together. In most cases, joint clearances of 0.03 to 0.08 mm are recommended for the best capillary action and joint strength. The best temperature is usually selected so as to be the lowest possible braze temperature, minimize any heat effects on the assembly, keep filler /base metal interactions to a minimum and maximize the life of any fixtures or jigs used.

The process known as *soldering* is generally similar to brazing except that the used filler metals melt at temperatures below 427 ˚C. In actual practice, most brazing alloys melt at temperatures well above 427 °C and most solders – at temperatures well below 427 °C . Many of the brazing alloys based on silver (all of which melt above 600 °C) were formerly termed '*silver solders*'. Soldering is used in plumbing, electronics and metalwork from flashing to jewellery.

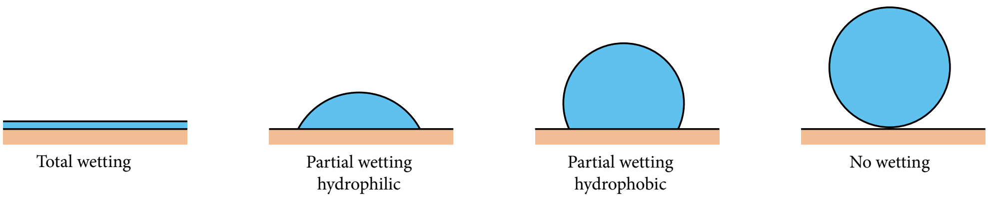

Figure 9.22. Wetting the surface

One of the obstacles to a successful solder joint are an impurity at the site of the joint, for example, dirt, oil or oxidation. Impurities can be removed by mechanical cleaning or chemical means by flux. For many years, the most common type of flux used in electronics (soft soldering) was rosin-based, using rosin from selected pine trees. It was ideal due to the fact it was non-corrosive and non-conductive under normal temperatures but became mildly reactive (corrosive) at the elevated soldering temperatures. Plumbing and automotive applications, among others, typically use the acid-based (muriatic acid) flux that provides cleaning the joint.

CHAPTER 10

Development of Electrical Engineering

This chapter will help in

- observing the beginning and discussing the development of electrical engineering;
- introducing lighting equipment (arc lamp, incandescent lamp, sodium vapour lamp, light-emitting diodes used in fluorescent lamps);
- presenting measurement instruments (current and voltage measuring devices, electrodynamometer, galvanometer for following variations in an alternating current waveform);
- describing electromagnetic engines (reciprocating engines, rotary engines, direct current generators);
- characterizing electric motors (linear low and high-acceleration, induction, pole amplitude modulated motors);
- discussing the growth of electrification in the 19^{th} – 20^{th} centuries;
- explaining the working principle of a combined cycle power plant;
- talking about electrification in Lithuania;
- showing reasons and consequences for the Sayano–Shushenskaya Dam disaster.

Beginning of Electrical Engineering

Electrical engineering is a field of engineering that generally deals with the study and application of electricity, electronics and electromagnetism. Electrical engineering is closely related to *power engineering* that discusses the generation, transmission and distribution of electricity as well as considers the process of designing a range of related devices, including transformers, electric generators, electric motors, high voltage engineering and power electronics. Power engineers may work on the design and maintenance of a power grid as well as on power systems connecting to it. Such systems are called on-grid power systems and may supply the grid with additional power, draw power from the grid or do both.

Electricity has been a subject of scientific interest since at least from the early 17th century. The first electrical engineer was probably William Gilbert who designed the *versorium* – a device that detected the presence of statically charged objects (Figure 10.1).

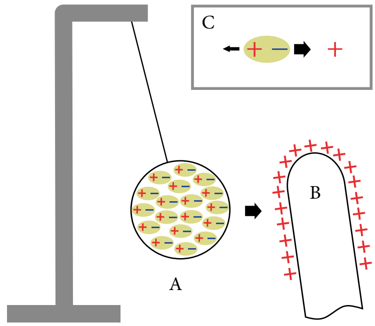

Figure 10.1. The versorium: a device that detects the presence of statically charged objects

He was also the first to draw a clear distinction between magnetism and static electricity and is credited with establishing the term of electricity. An electric current, as opposed to static charges, was first made readily available in 1800 as a result of work done by the Italian Alessandro Volta. In 1775, his scientific experimentations devised an *electrophorus*, a device that produced a static electric charge. By 1800, Volta developed the *voltaic pile*, a forerunner of the electric battery. The first mass-produced battery was designed in 1800 by William Cruickshank. He soldered together the pairs of copper and zinc plates and set them in wax in grooves across a wooden trough that later was filled with acid. By far the best known of all primary cells was the one developed by G. Leclanché in 1868. It is a carbon-zinc cell with ammonium sulphate as an electrolyte. The depolarizing agent is manganese dioxide packed around carbon. Most modern "*dry cells*" are those containing the electrolyte made into a paste. *Accumulators,* lead-acid battery, familiar as the car starter battery, can be re-charged from an electricity supply. Lead-acid batteries are heavy, which is their main disadvantage. The principal alternative is cells based on nickel and iron or nickel and cadmium, with sodium hydroxide as the electrolyte. Such cells are more robust than the lead-acid ones and not that heavy and so efficient. Despite their expense, chemical cells were the main source of electricity until the development of practical generators in the 1860s. Oersted's experiment disclosed that an electric current had produced magnetism. The question was if magnetism could produce electricity? M. Faraday discovered electromagnetic induction in 1831. He rotated a copper disc between the poles of this magnet and showed with his galvanometer that a current was produced between the axis and the edge of the disc.

Notable developments in electricity include the works by Georg Ohm. In 1827, he quantified the relationship between the electric current and potential difference in a conductor. In 1873, James C. Maxwell published a unified theory of electricity and magnetism. By the end of the 19th century, the rapid growth of communication was possible thanks to the engineering development of land-lines, submarine cables and wireless telegraphy. Practical applications and advances in such fields created an increasing need for the standardized units of measure, which led to the international standardization of such units as *volt, ampere, ohm, coulomb, henry* and *farad*. Thomas Edison built the world's first large-scale of 110 volts (1882), an electrical supply network of a direct current (DC). Nikola Tesla developed transformers and induction motors for use in AC. In 1884, Sir Charles Parsons invented the steam turbine. By 1890, power industry had flourished and power companies had built literally thousands of power systems (both direct and alternating current) in the United States and Europe, and these networks were effectively dedicated to providing electric lighting. Steam turbines are now generating mechanical power for about 80 percent of electric power in the world using a variety of heat sources. In 1880s, electric power distribution by the alternating current (AC) system was introduced. The method of AC won over DC for generation and power distribution because of its superior technology, especially the use of transformers to increase and decrease voltages. The use of high-voltage AC vastly extended the range of electric power distribution, and the use of transformers improved both the efficiency and safety of electric power distribution. The output voltage of early generators varied considerably with a load. As the load increased, voltage would fall, and, if the load was shed suddenly, voltage would surge. This did not matter when the load was arc lighting, but with filament lamps it was critical. During the 1880s, generator design gradually became science rather than art.

The machines described above were applied for a direct current. The early alternating current generators had armature coils arranged around the edge of a fairly thin disc and moving between the poles of a multi-polar field system. It was essential to use multi-polar machines if the generator was to be coupled directly to the steam engine. Even the fastest reciprocating engines ran at only about 500 rpm. A twelve-pole generator running at that speed would give a 50 Hz output. The disadvantage of the disc generator was that it was impossible to make such a machine for three phase operation. However, before three-phase supplies came into general use, the turbine had replaced the reciprocating steam engine, and generators were being designed for a higher running speed of the turbine.

Lighting

Although the possibility of electric arc lighting was demonstrated very early in the 19th century, it could not be a practical proposition until a supply of electricity was readily available. In an electric *arc lamp,* two carbon rods are connected to the opposite poles of the supply. The rods are briefly touched together and then drawn a few millimetres apart. This draws a spark, or "arc". The current in the arc produces considerable heat, and the contact points on carbon quickly become white hot, which is the source of light. White hot carbon burns in the air, and therefore some arrangement is necessary to feed carbon closer together so that the gap is kept constant. Without any adjustment, the gap widens and within a minute or two the supply of electricity will be unable to maintain the arc across a wider gap and the lamp will be extinguished. Two improvements in arc lighting were made during the 1890s. One was an enclosed arc having the arc contained within a small glass tube that restricted air flow. The effect was to reduce the rate of burning carbon. The second improvement was the addition of the cores of flame producing salts, mainly fluorides of magnesium, calcium, barium and strontium, to carbon rods. They increased light output and gave some control over the colour of light.

An *incandescent lamp* is electric light produced employing filament wire heated to a high temperature by an electric current passing through it, until it glows. Hot filament is protected from oxidation with a glass bulb filled with inert gas or evacuated. Incandescent bulbs are much less efficient than most other types of lighting. Most incandescent bulbs convert less than 5% of the energy they use into visible light, with the remaining energy being converted into heat. In 1802, Humphry Davy created the first incandescent light by passing the current through a thin strip of platinum chosen because the metal had an extremely high melting point. Over the first three-quarters of the 19th century, many experimenters worked with various combinations of platinum, iridium or osmium wire, carbon rods and evacuated or semi-evacuated enclosures. Tungsten filament lamps were first marketed by the Hungarian company Tungsram in 1904. A tungsten filament lamp lasted longer and gave brighter light than carbon filament. The conducted experiments also showed that the luminosity of bulbs filled with inert gas was higher than in a vacuum. Tungsten filament outlasted all other types. In 1906, the General Electric Company patented a method of making filaments from sintered tungsten and in 1911, used ductile tungsten wire for incandescent light bulbs. In 1913, Irving Langmuir found that filling a lamp with inert gas instead of a vacuum resulted in twice luminous efficacy and reduction in bulb blackening.

In 1920, Compton made the first successful sodium vapour lamp that consisted of a section of borate glass tubing which had been blown to an approximately spherical shape at the centre, and a coiled tungsten electrode was sealed into either end of the discharged vessel. Electrodes were brought to incandescent temperatures whereupon thermionic emission took place; they formed good cathodes, and discharge could then be

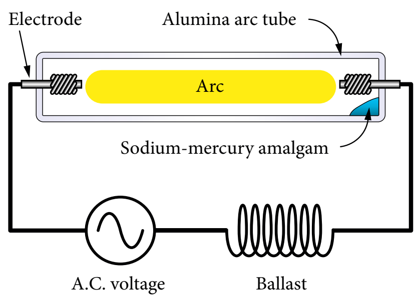

Figure 10.2. High pressure sodium lamp

struck across the gap. The distance between electrodes was relatively short and the lamps therefore had low operating voltage; a high current was necessary to deliver a useful amount of light. High-pressure sodium lamps are smaller and contain additional elements such as mercury, and produce a dark pink glow when first struck, and an intense pinkish orange light when warmed (Figure 10.2).

Since incandescent light bulbs use more energy than alternatives such as *compact fluorescent lamps* (CFL) and *light-emitting diodes* (LED lamps), today, many governments have introduced measures to ban their use (Figure 10.3). The LED consists of a chip of semiconducting material doped with impurities to create a *p-n* junction. As in other diodes, current flows easily from the *p*-side, or anode, to the *n*-side, or cathode, but not in the reverse direction. Charge carriers – electrons and holes – flow into the junction from electrodes with different voltages. When an electron meets a hole, it falls into a lower energy level and releases energy in the form of a photon (Figure 10.4). In 2011, a typical 6-watt LED lamp emitted 450 to 650 lumens, which is equivalent to a standard 40-watt incandescent bulb. A standard 40-watt incandescent bulb has an expected lifespan of 1,000 hours, whereas an LED can continue to operate with reduced efficiency for more than 50,000 hours, 50 times longer than the incandescent bulb. In the developed countries, 1 kWh of electricity will cause 610 g of CO_2 emission. Assuming the average light bulb is on for 10 hours a day, one 40-watt incandescent bulb will cause 89 kg of CO_2 emission per year. The 6-watt LED equivalent will only cause 14 kg of CO_2 over the same time span. A building's carbon footprint from lighting can be reduced by 85% by exchanging all incandescent bulbs for new LEDs.

a) b)

Figure 10.3. Tubular and helical-type integrated compact fluorescent lamps (a), LED spotlight using 60 individual diodes (b)

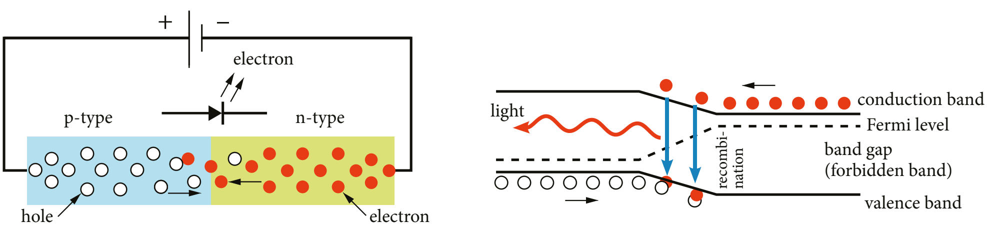

Figure 10.4. The inner workings of an LED showing circuit (left) and band diagram (right)

All public electricity supply systems are now entirely AC, but for transmission over long distances, DC systems are sometimes preferred. Very high power mercury arc valves, developed in Sweden in the 1930s, permit very heavy currents to be rectified, transmitted over a DC line and then inverted back to AC. DC transmission is also used for the purpose of linking two AC systems operating at different frequencies or at the same frequency but cannot be kept in synchronism. The cables under the English Channel linking the British and French grids also operate on DC.

Measuring Instruments

The rapid progress of electric lighting in the 1880s created a demand for practical and reliable measuring instruments. The most important devices were those for measuring voltage and consumption. Accurate voltage measurements were essential because the life of filament lamps was critically dependent on supply voltage. Devices for measuring electricity consumption were required so that customers could be charged according to the electricity they had used. Until the relatively recent advent of electronic measuring instruments, most current measurements have depended on measuring magnetic force created by a current in wire. This magnetic force has been balanced against controlling force that may be produced by a spring, gravity or another magnet. Simple galvanometers, in which a large diameter coil surrounded a pivoted magnetic needle, with the earth's magnetic field providing the controlling force, were made in 1837. A current measuring instrument that did not depend on magnetic materials was an *electrodynamometer*. A moveable coil, usually supported by torsion suspension, hangs within a fixed coil. When the current flows in both coils, magnetic forces tend to twist the moveable coil which is usually returned to its original position by twisting the top of the suspension through an angle that is a measure of the current.

The described Edison, Thomson and Ferranti meters were essentially DC instruments. After the invention of the induction motor, the instruments of an induction type were adopted for all AC systems and are now used virtually in all electricity supplies.

With alternating current systems, there was interest in the nature of the involved wave-forms. A *mechanical oscillograph* giving a visual display of a complete waveform was suggested in 1892, but only in 1897 succeeded in making a galvanometer the movement of which was light enough to follow the variations of an alternating current waveform. The discovery of the electron in 1897 led to the cathode ray tube. The voltage being studied was then applied to deflect a beam in the other axis, and a complete waveform was drawn out on the screen.

Electromagnetic Engines

The early *electromagnetic engines* can be broadly divided into two groups. The first to appear were *reciprocating engines*. One problem that faced the experimenters was converting the linear motion of an electromagnet into rotary motion. Contemporary steam engines solved this problem by the use of connecting rods and cranks, and initially it seemed logical to follow suit. These reciprocating electromagnetic engines are a fascinating example of one of the dead ends of technology.

The second group involved *rotary engines* that did away with the connecting rods and linkages of the reciprocating engine and evolved into a modern electric motor. Although the electric motor was important development as a new power source, alternatives at that time covered water, wind, or steam, and therefore it would be many years before it became widely used. The problem was that the only practical source of electricity supply in the 1840s was batteries. Reliable electrical power distribution appeared in the second half of the 19th century. In the 1840s, it was estimated that the electric motor powered by zinc/carbon batteries cost seventy times more to run than a coal fired steam engine of equivalent power.

In 1844, Paul Froment (France) developed a motor that utilized an electromagnet to attract iron bars mounted on a flywheel. This design did away with the need for mechanical linkages, and therefore reduced the number of moving parts. Froment's engine was more efficient and was the next step that would eventually lead to the development of the modern electric motor. The Froment-type motor was probably made in the second half of the 19th century (Figure 10.5). Four iron bars in the brass wheel are attracted to two electromagnets set below the middle of the wheel, and thus allowing them to attract the bars when energized alternately giving eight impulses per revolution.

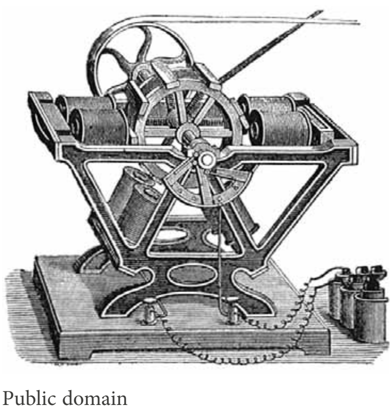

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Figure 10.5. The Froment-type motor

Zénobe Gramme (Belgium) invented the machine that was a DC (Direct Current) generator capable of producing much higher voltage than dynamos known up to that point. The Gramme machine used a ring armature, with a series of armature coils, wound around a revolving ring of soft iron. The coils are connected in series, and the junction between each pair is connected to a commutator on which two brushes run. Permanent magnets magnetize the soft iron ring producing a magnetic field that rotates around through the coils in order as the armature turns. This induces voltage in two of the coils on the opposite sides of the armature picked off by the brushes. With enough coils, the resulting voltage waveform is practically constant, thus producing a near direct current supply (Figure 10.6). This type of machine needs only electromagnets producing the magnetic field to become a modern generator. The Gramme machine was the first powerful electric motor useful more than a laboratory curiosity. Today, such design forms the basis for nearly all DC electric motors.

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Figure 10.6. The Gramme machine or Gramme magneto

The fact that the machines used as generators could also be used as motors was recognized quite early. In 1872, it was noted that a small rotating machine run just as well as a motor as it did as a generator.

Electric motors were more efficient than human or animal power. Conversion efficiency for animal feed to work is between 4 and 5% compared to over 30% for electricity generated using coal. The first company to exploit electric motors on a large scale was Siemens, and their first major customer was the state mines. Electricity in mining, however, developed fairly slowly. Only in the 1920s, mine electrification became widespread. The major industrial use of electric power was iron and steel industry. Electric motors proved to be very good for driving machine tools and rolling mills where the combination of power and precise control was valuable. The first permanent public electric railway was opened in Germany in 1881. It ran for about 3 km. Each carriage had a motor under the floor connected to the wheels through a belt drive. The first really practical electric tramway system was built in U.S. in 1888. Forty cars powered from overhead conductors ran over 20 km of streets.

Most early electric traction systems used a direct current because the DC series wound motor has good operating characteristics for the purpose. The advantages of alternating current transmission encouraged engineers to develop AC motors the first practical which were produced by Nicola Tesla in 1888. In his machines, a piece of a magnetic material free to turn followed a rotating magnetic field. He created a rotating magnetic field by using two coils energized from supplies that were in synchronism but not in phase. His first machine had two coils with axes at 90° to each other supplied with alternating currents and 90° out of phase. The resultant of the two oscillating magnetic fields was a rotating field. This was a synchronous motor. The rotating member either was or became a permanent magnet, and its poles followed the rotating field round, keeping in synchronism.

In the 20th century, most of the world's electric motor power comes from induction motors. Their disadvantage is that they are essentially constant speed machines, but for many applications that is perfectly satisfactory. The inherent simplicity and robustness of induction motors, which usually have no brush gear, make them the first choice for many applications.

By the mid-20th century, it seemed that electric motor development was complete; however, in 1957 the *pole amplitude modulated*, or PAM, motor was developed. This is a synchronous or induction motor the field coils of which are so arranged that by interchanging a few connections the number of poles can be changed. Since speed is determined by the number of poles, this gave the motor the speed of which could be switched between two distinct values. The PAM induction motor therefore retains the reliability and robustness of the conventional induction motor but can work at two different speeds.

The other approach to variable speed control is to change supply frequency, which is possible employing power semiconductors. By 1960, the available semi-conductor devices were capable of controlling a few tens of amperes, but progress in the following decade was so rapid that by the end of it semi-conductor frequency convertors became capable of supplying the largest motors and controlling their speed.

Linear motors represent another area of motor research that still remains active. Linear motors have also been a subject of research at least since 1841. The idea was revived in 1901 when a linear motor as a silent gun was tried to be used. The linear motor is an electric motor that has had its stator and rotor "unrolled" so that instead of producing torque (rotation), it produces linear force along its length. The view is perpendicular to the channel axis. Two coils at the centre are mechanically connected and energized in "quadrature" (with a phase difference of 90° (π/2 radians)). If the bottom coil (as shown) leads in phase, then, the motor will move downward (in the drawing), and vice versa (Figure 10.7). Many designs have been put forward for linear motors falling into two major categories: low-acceleration and high-acceleration motors. Low-acceleration linear motors are suitable for *maglev trains* and other ground-based transportation applications. High-acceleration linear motors are normally rather short and designed for accelerating an object to a very high speed, for example *railguns*.

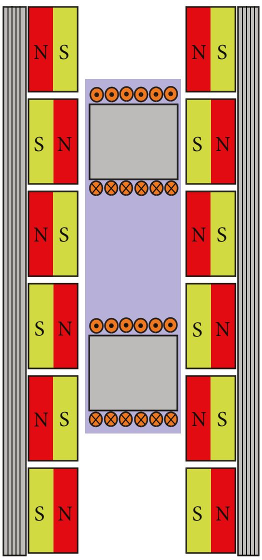

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Figure 10.7. Diagram of a U-channel synchronous linear motor

Research on the linear motor continues, and the widespread application of these machines will probably use semiconductor controls.

Electrification

Electrification is the process of powering with the help of electricity. Electrification was the build-out of electrical generating and distribution systems that occurred in the United States, Britain, Germany and other countries from the mid-1880s until around 1940 and is in progress in rural areas in some developing countries. Electrification was called the greatest engineering achievement of the 20th century.

The earliest commercial uses of electricity were telegraph and electroplating. The *electrification* of U.S. and Europe began with the demand, in urban areas, for public lighting. The *electrification* of the city went hand-in-hand with the demand for electrically powered public transport in the form of trolleys. The first central station providing public power is believed to be one in Surrey, U.K. in 1881.

The reason for the U.S. lack of rural electrification in the early part of the 20th century was no real electrical power demand for agriculture. By the 1920s, only 600,000 of 6.5 million U.S. farms had electricity. The Midwest and South of USA were simply kerosene-lit. In 1935, only 10 percent of national farms had electricity. During World War II, a shortage of materials forced a halt to construction of rural electric lines, but after the war, construction boomed as poles and wires became available. By the early 1970s, nearly all farms in the United States had electricity.

The first-ever Soviet plan for national economic recovery and development (State Commission for Electrification of Russia – GOELRO) was adopted in 1920. The Plan represented a major restructuring of the Soviet economy based on the total electrification of the country. The goal was the organization of industry on the basis of modern advanced technology and electrification which would provide a link between town and country, put an end to the division between town and country, make it possible to raise the level of culture in the countryside and overcome, even in the most remote corners of land, backwardness, ignorance, poverty, disease and barbarism. 112 regional power stations were built according to the state plan (1920–1940). The objective of 48 billion kWh was reached in 1940.

Nowadays, most electricity is generated by thermal power stations or steam plants, the majority of which are fossil fuel power stations that burn coal, natural gas, fuel oil or bio-fuels such as wood waste and black liquor from chemical pulping. The most efficient thermal system is a *combined cycle* in which a combustion turbine powers a generator using high temperature combustion gases and then exhausts cooler combustion gases to generate low pressure steam for conventional steam cycle generation.

Emerging Renewable Energy

Fossil fuels are nonrenewable and draw on finite resources that will eventually dwindle thus becoming too expensive or too environmentally damaging to retrieve. Renewable energy is generally defined as energy that comes from resources naturally replenished on a human timescale such as sunlight, wind, tides, waves and geothermal heat. Renewable energy resources exist over wide geographical areas, in contrast to other energy sources, that are concentrated in a limited number of countries. The rapid deployment of renewable energy and energy efficiency are resulting in significant energy security, climate change mitigation and economic benefits. Globally, the long-term technical potential of wind energy is believed to be five times total current global energy production or 40 times current electricity demand, assuming all practical barriers needed were overcome.

Wind Energy

This would require wind turbines to be installed over large areas, particularly in those of higher wind resources such as offshore. As offshore wind speed average is ~90% greater than that of land, offshore resources can contribute substantially more energy than land stationed turbines. A wind farm or wind park is a group of wind turbines in the same location used for producing energy (Figure 10.8 a). A large wind farm may consist of several hundred individual wind turbines and cover an extended area of hundreds of square miles; however, the land between the turbines may be used for agricultural or other purposes. A wind farm can also be located offshore.

Airflows can be used for running wind turbines. Modern utility-scale wind turbines range from around 600 kW to 5 MW of rated power, although turbines with the rated output of 1.5–3 MW have become the most common for commercial use; the power available from the wind is a function of the cube of wind speed, so as wind speed increases, power output increases dramatically up to the maximum output for the particular turbine.

The wind at higher altitudes becomes steadier, more persistent and of higher velocity. Because power available in the wind increases as the cube of velocity (the velocity-cubed law), doubling the velocity of the wind gives $2 \times 2 \times 2 = 8$ times the power; tripling the velocity gives $3 \times 3 \times 3 = 27$ times the available power. With steadier and more predictable winds, high-altitude wind has an advantage over the wind near the ground. Various mechanisms are proposed for capturing the kinetic energy of winds by kites, kytoons, tethered gliders, tethered sailplanes, aerostats (spherical and shaped kytoons), bladed turbines, airfoils, airfoil matrices, balloons, parachutes, variable drogues, multiple-rotor complexes, fabric parafoil kites, uni-blade turbines,

a)

b)

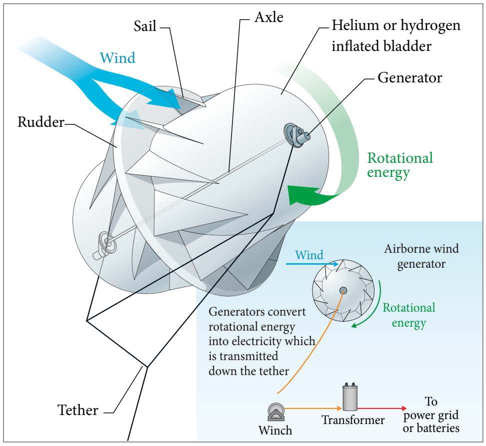

Photo by A. V. Valiulis

Figure 10.8. A wind farm in Western Lithuania (a) and airborne wind generator of flip-wing style (b)

flipwings, tethers, bridles, string loops, wafting blades, undulating forms, piezoelectric materials and more (Figure 10.8 b). Energy generated by a high-altitude system may be used aloft or sent to the ground surface by conducting cables, mechanical force through a tether, the rotation of an endless line loop, the movement of changed chemicals, the flow of high-pressure gases, the flow of low-pressure gases and laser or microwave power beams.

Solar energy

Solar power is the conversion of sunlight into electricity, either directly using *photovoltaics* (PV), or indirectly using *concentrated solar power* (CSP). Concentrated solar power systems use lenses or mirrors and tracking systems to focus a large area of sunlight into a small beam. Solar technologies are broadly characterized as either *passive solar* or *active solar* depending on the way they capture, convert and distribute solar energy.

Active solar techniques include the use of photovoltaic panels and solar thermal collectors to harness energy. The first solar cell was introduced in 1893. In 1956, Bell Laboratories accidentally discovered the use of silicon as a semi-conductor, which led to the construction of a solar panel with an efficiency rate of 6%. In 1956, the first commercial solar cell was made available to the public at the cost of \$300 per watt. In 1958, the first satellite that used solar energy to generate electricity was launched. During the energy crisis in 1970, solar energy became important to find an alternative form of energy. The price of solar cells dropped dramatically to about \$20

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Figure 10.9. The Nellis Solar Power generating complex in the United States

per watt. Photovoltaics converts light into electric current using the photoelectric effect. Silicon cell conversion efficiency is about 25%, whereas module efficiency makes about 20.4%. Photovoltaics is an important and relatively inexpensive source of electrical energy where grid power is inconvenient, unreasonably expensive to connect, or simply unavailable.

Passive solar techniques include orienting a building to the Sun, selecting materials with favorable thermal mass or light dispersing properties and designing spaces that naturally circulate air. In 1860–1881, A. Mouchout was the first man to patent the design of a 0.5 horse-powered steam engine running on solar energy. In 1876–1878, W.Adams experimented with mirrors and was able to power a 2.5 horsepower steam engine. In 1885–1889, Charles Tellier (a Frenchman) installed the first solar energy system for heating household water on the top of the roof. F. Shuman, an American inventor (1912–1913), used parabolic troughs to power a 60–70 horsepower engine that pumped 22.7 m^3 of water per minute from the Nile River to adjacent cotton fields. Commercial concentrated solar power plants were first developed in the 1980s.

The past few years have seen enormous investment in utility-scale solar plants (Figure 10.9). As of 2012, the largest solar energy plant in history is the Golmud Solar Park in China, with an installed capacity of 200 megawatts. This is surpassed by India's Gujarat Solar Park boasting a combined installed capacity of 605 megawatts. It is supposed to have been operating a concentrated solar power plant of 2,000 MW in Tunisia by 2016.

Biomass Energy

Biomass has been an important source of energy ever since people first began burning wood to cook food and warm themselves against the winter chill. Wood is still the most common source of biomass energy, but other sources of biomass energy include food crops, grasses and other plants, agricultural and forestry waste and residue, organic components from municipal and industrial wastes and even methane gas harvested from community landfills. Biomass, as a renewable energy source, refers to living and recently dead biological material that can be used as fuel or for industrial production. As an energy source, biomass can either be used directly via combustion to produce heat, or indirectly after converting it to various forms of biofuel. In this context, biomass refers to plant matter grown to generate electricity or produce, for example, trash such as dead trees and branches, yard clippings and biofuel from wood chips. It also includes plant or animal matter used for producing fibers, chemicals or heat. Biomass may also include biodegradable waste that can be burnt as fuel. Industrial biomass can be grown from numerous types of plants, including miscanthus, switchgrass, hemp, corn, poplar, willow, sorghum, sugarcane and a variety of tree species ranging from eucalyptus to oil palm (palm oil). The particular plant used is usually not important to the end products, but it does affect processing the raw material.

The processes of biomass *thermal conversion* (torrefaction, pyrolysis and gasification) use heat as the dominant mechanism to convert biomass into another chemical form. During *chemical convertion,* a range of chemical processes may be applied for converting biomass into other forms such as to produce fuel that is more conveniently employed, transported or stored, or to exploit some property of the process itself. *Biochemical conversion* makes the use of the enzymes of bacteria and other microorganisms to break down biomass. In most cases, microorganisms are used for performing the conversion process: anaerobic digestion, fermentation and composting. Biomass manufacturing is a growing branch of industry, as interest in sustainable fuel sources is increasing. Biomass can be used for producing electricity, or for manufacturing products that would otherwise require the application of non-renewable fossil fuels, or conversion into other usable forms of energy like methane gas or transportation fuels such as ethanol and biodiesel.

Hydrogen

Hydrogen has tremendous potential as a fuel and energy source, but the technology needed to realize that potential is still at the early stages. Hydrogen is the most common element on Earth; for example, though water is two-thirds of hydrogen, in nature, it is always found in combination with other elements. Once separated from other elements, hydrogen can be used for powering vehicles, replacing natural gas for heating and cooking and for generating electricity. The hydrogen fuel cell operates similar to a battery. It has two electrodes, an anode and a cathode separated by a membrane. Oxygen passes over one electrode and hydrogen over the other. Hydrogen reacts to a catalyst on the electrode, anode that converts hydrogen gas into negatively charged electrons (e-) and positively charged ions (H+). Electrons flow out of the cell to be used as electrical energy. Hydrogen ions move through the electrolyte membrane to the cathode, electrode where they combine with oxygen and electrons to produce water. Unlike batteries, fuel cells never run out.

The history of hydrogen fuel cells started in 1839. The first fuel cell was conceived by William Grove, a Welsh inventor and physicist. He mixed hydrogen and oxygen in the presence of an electrolyte and produced electricity and water. The invention did not produce enough electricity to be useful. In the 1920s, research on the fuel cell in Germany paved the way to the development of the carbonate cycle and solid oxide fuel cells of today. Early cell designers used porous platinum electrodes and sulfuric acid as an electrolyte bath. Using platinum was expansive and employing sulfuric acid was corrosive. Engineer Francis T. Bacon improved expensive platinum catalysts with a hydrogen and oxygen cell using a less corrosive alkaline electrolyte and inexpensive nickel electrodes. In 1959, he demonstrated a five-kilowatt fuel cell that could power a welding machine. During the early 1960s, General Electric produced a fuel-cell-based electrical power system for NASA's Gemini and Apollo space capsules. Space Shuttle electricity was provided by fuel cells, and the same fuel cells provided drinking water for the crew. The first bus powered by a fuel cell was completed in 1993, and fuel-cell cars are now being built in Europe, Japan and the United States.

Ocean Energy

The ocean provides several forms of renewable energy, and each one is driven by different forces. Energy from ocean waves and tides can be harnessed to generate electricity, and ocean thermal energy – from the heat stored in sea water. While using current technologies, most ocean energy is not cost-effective compared to other renewable energy sources, but the ocean remains an important potential energy source for the future.

Wave power is the transport of energy by ocean surface waves, and the capture of such energy does useful work, for example, electricity generation, water desalination or pumping (into reservoirs). Machinery able to exploit wave power is generally known as a *wave energy converter* (WEC). The first known patent to use energy from ocean waves dates back to 1799 and was filed in Paris. Modern scientific pursuit of wave energy was pioneered in the 1940s when various concepts of wave-energy devices at sea, with several hundred units used for powering navigation lights, were tested. Among these was the concept of extracting power from angular motion at the joints of an articulated raft. A renewed interest in wave energy was motivated by the oil crisis in 1973. Stephen Salter's 1974 invention became known as Salter's duck or *nodding duck*. In small scale controlled tests, the Duck's curved cam-like body can stop 90% of wave motion and can convert 90% of that to electricity giving 81% efficiency. In the 1980s, as the oil price went down, wave-energy funding was drastically reduced. Nevertheless, a few first-generation prototypes were tested at sea.

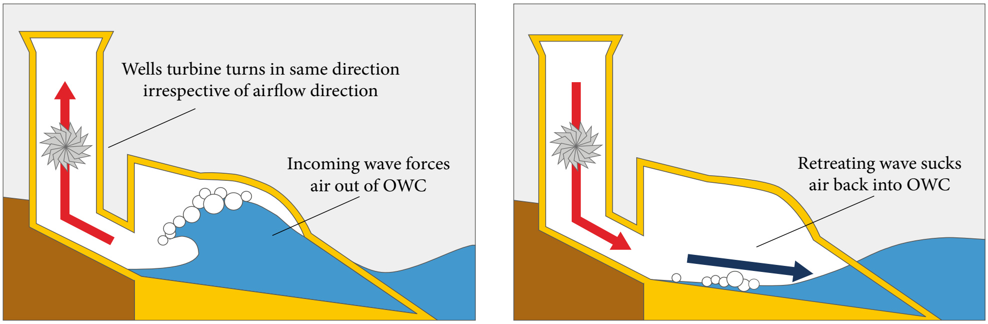

Figure 10.10. The oscillating water column process

Wave power devices are generally categorized by the method used for capturing the energy of waves, by location and by the power take-off system. The types of the method are *point absorber or buoy*, *surfacing following or attenuator* oriented parallel to the direction of wave propagation, *terminator* oriented perpendicular to the direction of wave propagation, *oscillating water column* (Figure 10.10) and *overtopping*.

The *Oscillating Water Column* generates electricity in a two step process. As a wave enters the column, it forces air in the column up the closed column past a turbine, and increases pressure within the column. As the wave retreats, air is drawn back past the turbine due to the reduced air pressure on the ocean side of the turbine.

Locations are shoreline, nearshore and offshore. Some of these designs incorporate parabolic reflectors as a means of increasing wave energy at the point of capture. These capturing systems use the rise and fall motion of waves to capture energy. Once wave energy is captured at a wave source, power must be carried to the point of use or to a connection to the electrical grid by transmission power cables. More recently, following the issue of climate change, there is again a growing interest worldwide in renewable energy, including wave energy.

Harvesting energy from *tides* is a surprisingly old method. In the Roman Era, tide mills were used. A tide mill is a dam constructed across a tidal inlet turning the estuary into a reservoir. At high tide, sea water was flowered into the reservoir through a one way gate and then it closed automatically when the tide started to fall. When the tide was low enough, the stored water was released to turn a water wheel which would then turn the millstone. At one time, there were 750 tide mills operating along the shores of the Atlantic Ocean: approximately 300 in North America, 200 in the British Isles and 100 in France. By the mid 20^{th} century, the use of water mills had declined. Modern tidal energy dates to 1966. The tidal power station on the Rance River in France became the world's first tidal power station. Various types of construction and concepts are currently being employed through these power stations and only the long-term will prove which concept, perhaps all, is the most sustainable form of renewable energy from tides.

Geothermal Energy

Geothermal electricity is generated from geothermal energy. Technologies in use include dry steam power plants, flash steam power plants and binary cycle power plants. *Dry steam plants* are the simplest and oldest design. They directly use the geothermal steam of 150 °C or greater to turn turbines. *Flash steam plants* pull deep, high-pressure hot water into lower-pressure tanks and use the resulting flashed steam to drive turbines. They require fluid temperatures of at least 180 °C, usually more. This is the most common type of a plant in operation today. *Binary cycle power plants* are the most recent development and can accept fluid temperatures as low as 57 °C. Moderately hot geothermal water is passed by secondary fluid with a much lower boiling point than water. This causes secondary fluid to flash vaporize, which then drives turbines. The thermal efficiency of this type of the plant is typically about 10–13%.

The first industrial use of geothermal energy began near Pisa, Italy in the late 18th century. Steam coming from natural vents (and from drilled holes) was used for extracting boric acid from hot pools. In 1904, Italian scientist Piero G. Conti invented the first geothermal electric power plant in which steam was used for generating power (Figure 10.11).

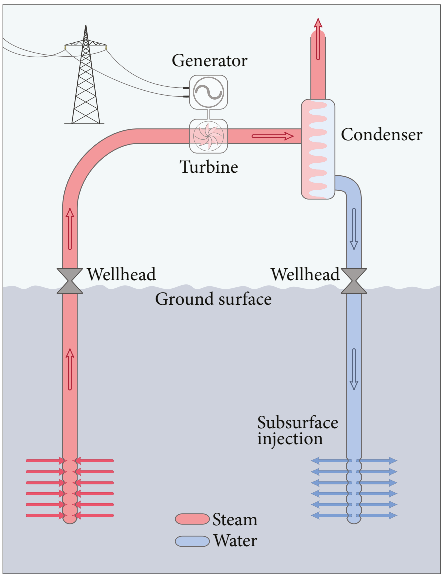

Figure 10.11. A geothermal electricity generation plant

The thermal efficiency of geothermal electric plants is low, around 7–10%, because geothermal fluids are at a low temperature compared with steam from boilers. By the laws of thermodynamics, this low temperature limits the efficiency of heat engines in extracting useful energy during the generation of electricity. Exhaust heat is wasted, unless it can be used directly and locally, for example, in greenhouses, timber mills and district heating. In order to produce more energy than pumps consume, electricity generation requires high temperature geothermal fields and specialized heat cycles. Because geothermal power does not rely on variable sources of energy, unlike, for example, wind or solar, its capacity factor can be quite large – up to 96% has been demonstrated.

Geothermal electricity generation is currently used in 24 countries (USA, Philippines, Indonesia, India, Italy, Iceland, etc.). Geothermal power today supplies less than 1% of the world's energy in 2009 needs but is expected to supply 10-20% of the world's energy requirement by 2050.

Hydropower

Hydro-power or water power is derived from the energy of falling or running water. Humans have been harnessing water to perform work for thousands of years. The power of water was used for flour grinding, saw wood and power textile mills. For more than a century, technology for using falling water to create hydroelectricity has existed. The evolution of a modern hydropower turbine began in the mid-1700s in France. In 1880, a brush arc light dynamo driven by a water turbine was used for providing lighting. Between 1880 and 1895, hydropower was started to be used for generating electricity; these first hydroelectric plants produced direct current (DC) used mostly to power nearby arc and incandescent lighting. Alternating current is used today. That breakthrough came when the electric generator was coupled to the turbine.

Dams and canals were necessary for the installation of successive waterwheels when a drop was greater than 5 m. Large storage-dam construction, however, was not feasible, and low water flows, during summer and autumn, coupled with icing during winter and led to the replacement of nearly all waterwheels by steam when coal became readily available. The basic principle of the operation of most major installations has remained the same since then. Plants depend on a large water-storage reservoir upstream of a dam where water flow can be controlled and a nearly constant water level can be assured. Water flows through conduits, called penstocks, are controlled by valves or turbine gates to adjust the flow rate in line with demand for power. Water then enters turbines and leaves them through the so-called tailrace. Power generators are mounted directly above turbines on vertical shafts (Figure 10.12). The design of a hydroelectric plant became fairly well standardized after World War I with most development in the 1920s and 1930s being related to thermal plants, transmission and distribution.

Hydropower is a very efficient energy source, because some turbines can achieve the efficiency of 95% and more. Hydropower today provides about 20% of the world's electricity. The advantage of hydropower over other renewable energy sources is the fact that the average rainfall is highly predictable, and therefore output

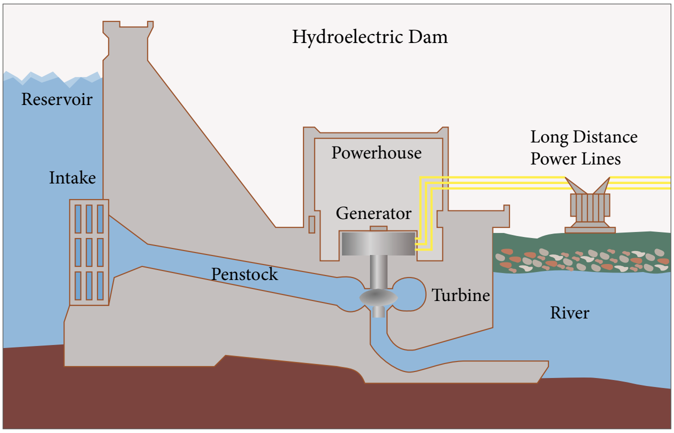

Figure 10.12. A conventional dammed-hydro facility

is reliable and river flow does not fluctuate from minute to minute like is the case with wind energy. Hydropower can have negative ecological impacts, especially on fisheries and water eco-systems, which especially applies to large-scale hydropower.

The largest hydroelectric power station in the world is the Three Gorges Dam in China (installed capacity 22,500 MW). The main generators weigh about 6,000 tons each and are designed to produce more than 700 MW of power. The flow rate varies between 600–950 cubic meters per second. The turbine is 10 m in diameter, and rotation speed is 75 revolutions per minute. The outer diameter of the generator stator is 21 m. The inner diameter is 18.5 m. The dam project was completed and became fully functional as of July 4, 2012.

Electrification of Lithuania

In 1892, the first power plant in Lithuania was launched in Rietavas, and the power line (110 V) to the park and palace was built. In 1911, the plant had 69.5 m^2 surface steam boiler, steam engine and Gram electrical generator. Incandescent lighting used 16 and 25 candle bulbs. It was the first electric lamp in Lithuania. The plant operated until 1915. During the First World War, the generator was damaged. In 1919, Rietavas municipality built a new, more modern generator in the local mill and transmitted electricity to the town by the power line. The power station was in operation until World War II: in the first day of the war, the German shell hit the plant and completely destroyed it. In 1898, the first industrial power plant in Lithuania was launched in R. Tilman's metal factory in Kaunas. In 1900, Kaunas and Klaipėda initiated first public power plants in Lithuania. The public power plant in Vilnius started working in 1903, whereas in 1903, the first hydropower generator on the Virvytė river (Kairiškiai village, Mažeikiai district) was installed. In 1921, Tauragė owned 3000 V, and in 1922–1923, Kaunas and Šiauliai operated 6000 V alternating current supply lines. In 1930, a thermal power plant in Petrašiūnai and the first 15 kV voltage transmission line Petrašiūnai-Jonava in Lithuania were built. In 1941, the first 30 kV voltage transmission line in Lithuania connecting Šiauliai, Radviliškis and Panevėžys came into use. In 1955, the construction of Kaunas Hydroelectric Power Plant began. The first unit (22,5 kW) started working in 1959. The same year, Antalieptė Hydroelectric Power Plant on the river Šventoji was developed. In 1960, 110 kV power transmission line Kaunas–Marijampolė–Sovetskas (Tilžė)–Šilutė–Klaipėda and the first transmission line in the area between Lithuania and Kaliningrad (Russia) area were opened. Upon completion of the first phase of energy system development in Lithuania, all major power plants in Vilnius (two), Petrašiūnai, Rėkyva, Klaipėda and Kaunas were incorporated into the common system. In 1960, the construction of Lithuanian power plant in Elektrėnai started, the first 150 MW turbine of which ran two years later. In 1962, the first 330 kV energy transmission line Šiauliai–Jelgava (Latvia) was introduced. In 1964, in general, the electrification of the rural areas of the country was finished. In 1972, the eighth (last) 300 MW turbine in Lithuanian Power Plant started running, the plant reached the power capacity of 1.8 million kW, and the construction was completed.

In 1974, the works of the Ignalina Nuclear Power Plant started, and in 1983, the first 1500 MW power block was launched. The Ignalina Nuclear Power Plant contained two RBMK-1500 water-cooled graphite-moderated channel-type power reactors that were originally the most powerful in the world. As a condition for joining the European Union, in 1999, Lithuania agreed on closing up the existing units of the station. , The Ignalina plant has not a containment building and such structure bring on a high risk. Unit 1 was closed in 2004. Unit 2 came online in 1987 and was closed in 2009.

In 1977, the construction of Kruonis Pumped Storage Power Plant (PSPP) began, and in 1992, the first 200 MW unit (Figure 10.13) was launched. During the periods of low demand, usually at night, Kruonis PSPP operates in the pump mode, and, using cheap surplus energy, raises water from a lower Kaunas reservoir to the upper one. The station is designed to have an installed capacity of 1600 MW; however, only four 225 MW generators currently operate. With the fully filled upper reservoir, the plant can generate 900 MW for about 12 hours. Lithuanian Green Movement, in 1989, blocked the construction of the fifth turbine of Kruonis PSPP (Figure 10.14). Lithuania is planning construction works for the fifth hydro unit of 225 MW to expand the generation capacity of Kruonis Pumped Storage Power Plant (PSPP). The unit will be able to operate at 110–225 MW

Figure 10.13. Kruonis Pumped Storage Power Plant

Figure 10.14. A rotor of the derelict turbine

capacity in the pumping mode and at 55–225 MW capacity in the generating mode, which will increase the overall capabilities of the power plant. The new unit is expected to achieve the coefficient of performance reaching 78%.

A new Combined Cycle Unit (CCU) was opened at Elektrėnai Power Plant in 2012. The CCU has an efficiency of more than 58%. Meanwhile, using an ordinary technology only, 25–40% of fuel used for generation is converted into electricity. In the combined cycle power plant, the heat of gas turbine exhaust is used for generating steam by passing it through a *heat recovery steam generator* (HRSG) with a live steam temperature between 420 and 580 °C (Figure 10.15). It is planned that the CCU will produce 1.374 TWh of electricity in 2013, which is 14% of the country's demand.

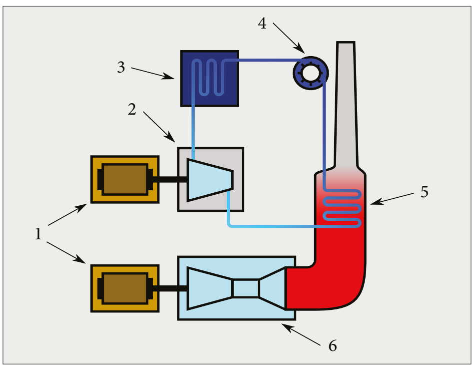

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Figure 10.15. The working principle of a combined-cycle power plant: 1 – electric generators, 2 – steam turbine, 3 – condenser, 4 – pump, 5 – boiler/heat exchanger, 6 – gas turbine

Starting in the 1960-ies, the 330 kV transmission network of the Baltic States was forming as a part of the global electricity generation and transmission network of the Soviet Union. In 1992, the Heads of the Baltic States (Estonia, Latvia and Lithuania) decided on integrating three Baltic power systems, with the united coordinating dispatcher centre aimed at maintaining the parallel synchronous operation of the Estonian, Latvian and Lithuanian power systems in the Electrical Ring that also included Russian areas of Leningrad, Novgorod, Pskov, Tver, Moscow and Smolensk as well as Belarus . Currently Lithuania is synchronized with Russia. Elimination of the Baltic States' electric energy system isolation and its integration into the common European electricity market (Continental Europe grid) is the ultimate goal of the Baltic states. In case of merging Baltic, Russian, Ukrainian and Byelorussian power systems with the United power system of the European Union, one huge power system with ~ 800 000 MW consumption would form. Such a global power system would allow reducing considerably the expenses of different specific global services, e.g. those involving frequency regulation with the use of primary power reserve.

Sayano–Shushenskaya Dam Disaster

The Sayano–Shushenskaya Dam is located on the Yenisei River in Khakassia, Russia and is the largest power plant in Russia and the sixth-largest hydroelectric plant in the world by average power generation. The plant operated ten hydro turbines each with a capacity of 640 MW. The total installed capacity of the plant is 6,400 MW. The arch-gravity dam is 245.5 metre high. It has a crest length of 1,066 metres and base width of 105.7 metres. Water pressure for the dam is approximately 30 million tons, 60% of which is neutralized by the own weight of the dam and 40% is carried to rock on the bank. Construction started in 1968, and the plant was opened in 1978 (Figure 10.16).

Figure 10.16. The generator hall of the Sayano-Shushenskaya hydroelectric power station seen before the accident

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Figure 10.17. Damage to the Sayano-Shushinskaya power station

On 17 August 2009 at 8:13 AM, the hydro-electric plant suffered a catastrophic accident that caused flooding of the engine and turbine rooms and the explosion of two 711 MVA electric generators under water as a result of a short circuit (all other machinery was damaged in different measures, only four hydro-aggregates were possible to repair and all other six demanded for replacement with new machines).

The official report states that the accident was primarily caused by vibrations in turbine No. 2, which led to fatigue damage of the mountings of the turbine, including its cover. At the moment of the accident, the nuts on at least 6 bolts keeping the turbine cover in place were absent. Shortly afterwards, the bolts keeping the turbine cover in place were broken. Under water pressure (about 20 atmospheres) spinning the turbine with its cover, rotor and upper parts jumped out of the casing thus destroying machinery hall equipment and building (Figure 10.17). Pressurised water immediately flooded the rooms and continued damage to the plant. At the same time, the power output fell to zero resulting in a local blackout. It took 25 minutes to manually close water gates to other turbines; during that time, they continued spinning without a load. After the accident, 49 found bolts were investigated: 41 had fatigue cracks. On 8 bolts, the fatigue-damaged area exceeded 90% of the total cross-sectional area. The accident caused an oil spill with at least 40 tons of transformer oil released, spreading over 80 km downstream of the Yenisei. 75 people were confirmed dead.

CHAPTER 11

Information Transfer and Audiovisual Technologies

This chapter will help in

- presenting information transfer and audiovisual technologies;
- explaining magnetic sound recording (Edison, Marconi, CD recording, wire recording) and reproduction;
- introducing a compact cassette, play-only cassette and multi-channel sound machines;
- discussing the production of sound recording devices in Lithuania.

Sound Recording and Reproduction

Sound recording and reproduction is electrical or mechanical recording and re-creation of sound waves such as a spoken voice, singing, instrumental music, etc.

The two main classes of sound recording technology include *analog recording* and *digital recording*. Acoustic analog recording is achieved by a small microphone diaphragm that can detect acoustic sound waves and record them on a medium such as a phonograph. In magnetic tape recording, sound waves vibrate the microphone diaphragm and are converted into a varying electric current, which is then converted to a varying magnetic field by an electromagnet, which makes a representation of the sound on a plastic tape with a magnetic coating. Analog *sound reproduction* is the reverse process with a bigger loudspeaker diaphragm causing changes to atmospheric pressure to form acoustic sound waves.

Digital recording and reproduction converts the analog sound signal picked up by the microphone to a digital form by a process of digitization allowing it to be stored and transmitted. Digital recordings are considered higher quality than analog recordings because the digital format can prevent much loss of quality found in analog recording due to noise and electromagnetic interference in playback and mechanical deterioration or damage to the storage medium. A digital audio signal must be reconverted to the analog form during playback before it is applied to a loudspeaker or earphones.

Figure 11.1. Edison tinfoil phonograph

The automatic reproduction of music can be traced back as far as the 9th century when "the earliest known mechanical musical instrument" was invented. In this case, a hydro-powered organ played interchangeable cylinders automatically. Such a cylinder with raised pins on the surface remained the basic device for producing and reproducing music mechanically until the second half of the 19th century. Later, an automatic flute player was invented, which appears to have been the first programmable machine, piano rolls, the use of which began declining in the 1920s although one type is still being made today. The first practical sound recording and reproduction device was a mechanical phonograph cylinder invented by Thomas Edison in 1877 (Figure 11.1). In 1888, Edison patented a wax-cylinder machine.

The next major technical development was the invention of a gramophone in the United States in 1889. Discs were easier to manufacture, transport and store. They were single-sided. The double-sided 78 rpm shellac disc was the standard consumer music format from the early 1910s to the late 1950s (Figure 11.2). Discs were made of *shellac* or similar brittle plastic-like materials. Shellac is a natural bio-adhesive polymer. This resin is secreted by the lac bug on trees in the forests of India and Thailand. Around 1895, a shellac-based compound was introduced and became standard.

a)

b)

Figure 11.2. View of the gramophone (a) and portable pathepfone (1950) (b)

The exact formulas for this compound varied by manufacturer and over the course of time, but it was typically composed of about one-third of shellac and about two-thirds of "mineral filler", which meant finely pulverized rock, usually slate and limestone, with an admixture of cotton fibres to add tensile strength and carbon black for colour (without this, it tended to be a "dirty" gray or brown colour that most record companies considered unattractive). Discs were played with needles made from a variety of materials, including mild steel, thorn and even sapphire. Discs had a distinctly limited playing life heavily dependent on how they were reproduced. The acoustic methods of recording had limited sensitivity and frequency range. Mid-frequency range notes could be recorded but very low and very high frequencies could not. Long-playing (331/3 rpm) *vinyl* discs (polyvinyl chloride plastic) were introduced in US in 1948. Although vinyl was much more expensive than *shellac*, but with a long-playing disc the added cost was acceptable. Vinyl offered improved performance both in stamping and playback. Vinyl records were much less fragile than shellac.

Between the invention of the phonograph in 1877 and the advent of digital media, the most important milestone in the history of sound recording was the introduction of «electrical recording», in which a microphone was used for converting the sound into an electrical signal that was later amplified and used. During World War I, experiments were undertaken in the United States and Great Britain to record and reproduce, among other things, the sound of a German U-boat (submarine) for training purposes.

In 1920, motion picture sound systems employed optical recording technology, in which the audio signal was graphically recorded on a photographic film. Sound signals were used for modulating a light source imaged onto the moving film ("sound track"). Optical sound became the standard motion picture audio system throughout the world.

In 1878, Edison proposed magnetic recording, but only in 1898 a system for recording on steel wire was invented. One of the biggest advantages of a magnetic-recording medium was its erasability so that it could be used again and again. In 1920, a plastic magnetic tape was developed, but only in 1934 a magnetic recording tape was first manufactured in Germany. In the 1930s, radio pioneer Guglielmo Marconi promoted a system of magnetic sound recording using a steel tape that was of the same material used for making razor blades, and not surprisingly that these recorders were considered so dangerous. The first practical tape recorder was produced in Germany in 1935.

Other important inventions of this period were magnetic tape and a tape recorder. The paper-based tape was first used but was soon superseded by polyester and acetate backing. This technology was employed for recording from the 1950s to

1980s. Around 1958, reel-to-reel recorders using 1/4in-wide magnetic tape became popular, because the tape could play for hours without interruption. In 1963, a compact cassette as the worldwide standard for audio recording was made. Any cassette could work in any recorder and it led to the explosive growth of cassette industry. The compact cassette is a two-hub system, a miniaturized version of reel-to reel; the tape never leaves the enclosure but is exposed only as it passes by recording or playing heads. In 1979, Sony introduced its Walkman, play-only cassette machines that were half the weight and a fifth the size of older cassette recorders. This technology is now being challenged by another, even more revolutionary audio system, the Compact Disc or CD.

A typical *Compact Cassette Magnetic* tape recorder was launched in 1948 and brought sweeping changes in both radio and recording industry. Sound could be recorded, erased and re-recorded on the same tape many times; sounds could be duplicated from tape to tape with only minor loss of quality. The magnetic tape transformed recording industry, and by the late-1950s, the vast majority of commercial recordings were being mastered on tape. The compact cassette also benefited enormously from developments in the tape material itself as materials with wider frequency responses and lower inherent noise were manufactured and often based on cobalt and/or chrome oxides as the magnetic material instead of more usual iron oxide (Figure 11.3).

There had been experiments with *multi-channel sound* for many years, but the first commercial application of the concept came in the early 1970s with the introduction of Quadraphonic sound. This multi-track recording used four tracks (instead of two used in stereo) and four speakers to create a 360-degree audio field around the listener.

Figure 11.3. Some products of Walkman line (2006)

Figure 11.4. Wire recorder from 1945

CD recording initiated another massive wave of change in consumer music industry with vinyl records effectively relegated to a small niche market by the mid-1990s. Digital audio technology is used in all areas of audio, from casual use of music files of moderate quality to the most demanding professional applications. Today, the process of making recordings is separated into tracking, mixing and mastering. Multi-track recording makes it possible to capture signals from several microphones or from other sources to the tape or disc thus allowing mixing and mastering stages for editing, level balancing or adding different effects.

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Figure 11.5. Cockpit voice recorder and flight data recorder, each with an underwater locator beacon on the front

By 1930, advances in electronics allowed the first commercially successful wire recorders to be introduced as dictating machines and telephone recorders (Figure 11.4).

Data recorders on modern commercial jetliners may track as many as 3,000 data points, including the status of every system on the aircraft, the positions of cockpit controls as well as pressure and temperature readings from fuel tanks and hydraulic systems. The black box casing is made of stainless steel and/or titanium, and can withstand impacts at up to 3,400 g's. Thermally insulated solid-state flash drives store up to several gigabytes of data. The box distress signal can be detected through 4200 m of water (Figure 11.5).

Production of Sound Recording Devices in Lithuania

National factories VILMA and ELFA (both in Vilnius) produced different recorders and/or players (Figure 11.6) and other sound recording (tape, wire, cassette) devices (Figures 11.7). The factory VENTA produced different electronic musical instruments (Figure 11.8).

Figure 11.6. Products by ELFA Electrotechnical Works: tape recorder Gintaras (1960)

Up to 1995, VILMA National devices factory, founded in 1960 in Vilnius, produced top secret mini cassette (metal wire) recorders (black boxes) for USSR planes and soviet secret service. Recording time was 5,5 hours and wire speed – 145 mm/s. Also, for the period from 1970 to 1995, the company manufactured some cassette recorders *Vilma* for domestic use.

a) b)

Figure 11.7. Tape recorder *Elfa 201* Stereo manufactured by Vilnius Electrotechnical Works (1986). Museum of Energy and Technology, Vilnius (a). Portable cassette recorder Tonika-310 Stereo manufactured by VILMA National devices factory, Vilnius (1977–1982) (b)

Figure 11.8. Electronic musical instrument "Vilnius" produced by "Venta" plant in Vilnius (1970–1990). Museum of Energy and Technology, Vilnius

CHAPTER 12 Telegraph, Telephone and Radio

This chapter will help in

- explaining information transfer by visual (or aerial) telegraphy and semaphore;
- revising data-communication by electric telegraph;
- describing the Morse Code Alphabet;
- presenting telex;
- clarifying transmission by the telephone, payphone, smartphone;
- defining wireless transmission (radio);
- specifying Lithuanian radio products.

Visual Telegraphy

The roots of telegraph reach 1600s when a suggestion that two men at a distance might communicate by the use of magnetic needles pointing towards letters on a dial was put forward. Nevertheless, electricity was still poorly understood to provide a practical means of communication. In 1790, France put into use the first practical system of visual (or aerial) telegraphy, a *semaphore*. This invention was favourably accepted as a war aid. However, some vital questions such as Would it work in fog? or Could secret codes be used? were raised. In 1794, the line from Paris to Lille (14 km) was completed. With the help of the semaphore, the capture of Quesnoy from Austrians was made known in Paris only one hour after troops entered the town.

To operate the semaphore, the station agent manipulated levers from the ground level, and movements were faithfully followed by wooden arms placed three metres above, on the top of a stone tower (Figure 12.1). The central member, regulator, was 4 m long, and the arms of 1.8 m were pivoted from both ends. The regulator and arms could be placed in 196 recognizable positions. For communication at night, a lantern was put at each end of the arms and at each of the pivots.

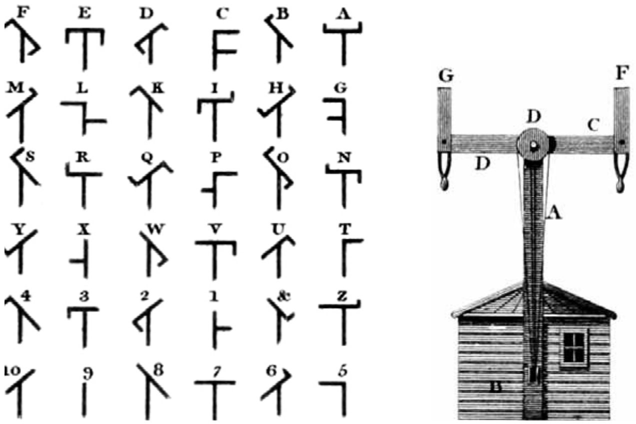

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Figure 12.1. Chappe telegraph

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Figure 12.2. Optical telegraph erected on the top of the western tower of Gediminas Castle in Vilnius (1838)

A code of 9999 words was published with each word represented by a number. Aerial telegraphers could send a dispatch at a speed of 3000–7000 km/h. It was three hundred times better than the available fastest system – a relay of horses and riders.

In 1839, the optical telegraph line from St Petersburg to Warsaw, the longest one in the world, was laid through the territory of Lithuania (Figure 12.2).

Electric Telegraph

In 1809, a telegraph system using the principle where 26 parallel wires were used for transmitting the letters of the alphabet at a distance of up to two miles was invented in Germany. This was really the first relay, although not electromechanical. In 1820, the use of a galvanometer needle for telegraphy purposes was suggested. By 1839, a 13-mile long telegraph line for British railways was set. Thus, needle telegraph became the first serial data-communication system with the codes defined by the sequences of needle deflections to make up each character. In Samuel Morse telegraph built in 1837, the sender used notched metal strips to encode the alphabet, and the receiver was an electromagnetically-driven pendulum with a pencil attached to which wrote coded signals on a moving roll of paper (Figure 12.3). It replaced the type-bar sender with what we now call a *telegraph key* and simplified the receiver to a pen that put marks (dots and dashes) on

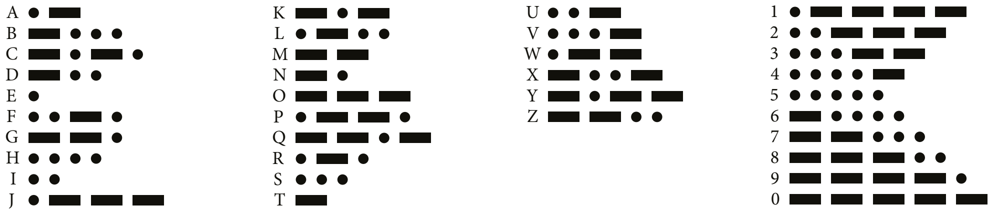

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Figure 12.3. Morse code alphabet. The length of a dot is one unit. The dash is three units. The space between letters is three units and the space between words is seven units

a paper tape (Figure 12.4). With some government support, Morse started a telegraph service between Washington and Baltimore in 1844. Later, the visual receiver was replaced by a sounder.

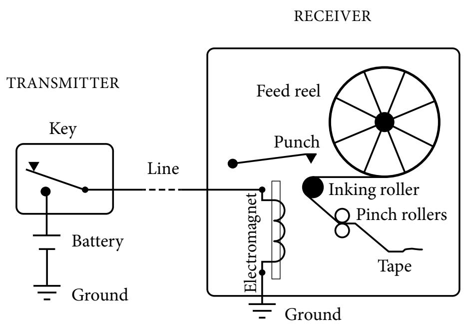

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Figure 12.4. Morse telegraph

The most radical departure, which led the way to the modern era of telegraphy, was the invention of a teletypewriter (or teleprinter) in U.S. in 1928. This allowed operators to compose messages using a typewriter-like keyboard to punch the paper tape that was then torn off and fed into a tape reader for transmission. At the receiving end, the message was printed out on paper. Morse is still used in radiotelegraphy where it can get a message through static and difficult transmission conditions when electromechanical and electronic alternatives are unworkable.

Telex

Sending and delivering a telegram was slow and labour-intensive. The sender had to go to a telegraph office and print out his message in capital letters on a special form. The clerk had to count the words and compute the amount due from a tariff schedule; then the telegraph operator had to key in the message. At the other end, the reverse took place; the message would be printed out on a paper strip cut into segments, pasted on a form and then delivered by hand to the recipient. A unique feature of the telegram among public telecommunication services has always been the delivery of a written message providing legal proof to the sender and receiver.

Most customers would insist on physical delivery even though the message had been read to them over the telephone. For business users, the whole system seemed archaic. In U.S. and Europe, they turned increasingly to the telephone whenever written proof was not essential. The first telex network was put into service in Germany in 1933. By the early 1980s, the number of telex subscribers world-wide exceeded 1.5 million, which made more than a half of them in Europe. Today, telex is a world-wide switched public teleprinter service. Therefore, it is like the telephone where subscribers are loaned terminal equipment, can dial up other subscribers themselves, receive messages and are charged on the basis of time and distance. However, unlike the telephone, messages must be keyed in and received on teleprinters.

One of the greatest advantages of telex is that teleprinter equipment and service available to subscribers is uniform throughout most of the world. The speed of transmission is very slow and makes 50 bits per second, equivalent to 66 words per minute (a word is defined to be six characters, including space). In the early 1980s, a completely new public switched message service called *teletext* was introduced. It operates 40 times faster than telex and allows upper-and-lower-case letters as well as formatting codes.

Telephone

A telephone, or phone, is a telecommunication device that converts sound into electronic signals suitable for transmission via cables or other transmission media over long distances.

Acoustic transmission. The speaking voice is audible only up to a few hundred metres. Mechanically aided sound-generation systems, such as drums, can carry messages for many kilometres. Today, police and ambulance sirens can be heard over a radius of several kilometres even in noisy cities. Sounds also can be sent through hollow tubes, and through denser matter (liquids or solids) they can be transmitted much further and at higher speeds than in air. Modern hydrophones listen in on the cries of dolphins and whales and detect other underwater sounds over a range of frequencies far beyond human hearing. In 1837, the first commercial telegraph system was build. In 1861, a telephone line was constructed in Germany; however, it was able to transmit only tones rather than speech. The breakthrough came in 1876 when A. G. Bell and Elisha Gray filed patents for successful speaking telephones. The first transmission of a human voice over wire happened only three days after the first telephone patent had been issued to Bell. In 1877, Bell patented an electromagnetic telephone using permanent magnets, iron diaphragms, carbon (graphite) transmitter and a call bell. In 1883, Edison discovered signal amplification in both wire and wireless communication. The early history of the telephone still remains a confusing, but the patents obtained by Bell and Edison, however, were forensically victorious and commercially decisive.

Early telephones were technically diverse. Some used a liquid transmitter, some had a metal diaphragm that induced current in an electromagnet wound around a permanent magnet and some were "dynamic" – their diaphragm vibrated a coil of wire in the field of a permanent magnet or the coil vibrated the diaphragm (Figure 12.5). Early telephones were locally powered, using either a dynamic transmitter or by powering a transmitter with a local battery. Rural and other telephones that were not on a common battery exchange had a magneto or hand-cranked generator to produce a high voltage alternating signal to ring the bells of other telephones on the line and to alert the operator.

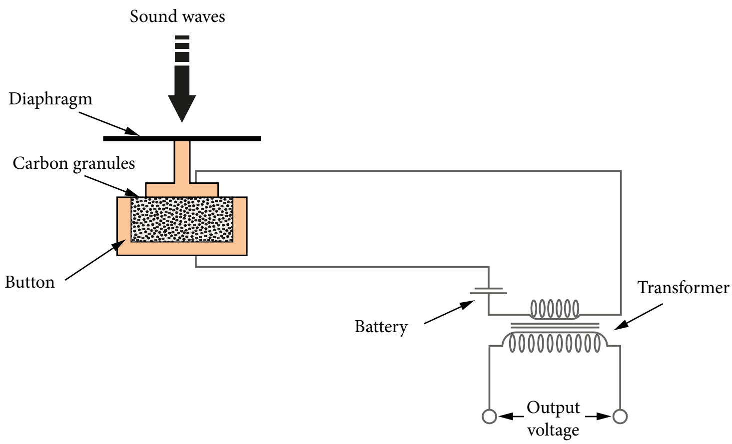

Figure 12.5. Scheme for a simple carbon "button" phone

What turned out to be the most popular and longest lasting physical style of telephone was introduced in the early 20^{th} century. A carbon granule transmitter and electromagnetic receiver were united in a single molded plastic handle, which when not in use sat in a cradle in the base unit. The circuit diagram of the phone shows the direct connection of the receiver to the line, while the transmitter was induction coupled, with energy supplied by a local battery. The coupling transformer, battery, and ringer were in a separate enclosure. The dial switch in the base interrupted the line current by repeatedly but very briefly disconnecting the line 1–10 times for each digit, and the hook switch (in the centre of the circuit diagram) disconnected the line and transmitter battery while the handset was on the cradle.

Figure 12.6. A wooden wall telephone with a hand-cranked magneto generator

Figure 12.7. An old rotary dial phone

Figure 12.8. A touch-tone dialling phone

After the 1930s, the base also enclosed the bell and induction coil obviating the old separate ringer box. Power was supplied to each line by central office batteries instead of a local battery, which required periodic service. For the next half of the century, the network behind the telephone became progressively larger and much more efficient, but after the telephone dial (Figures 12.6 and 12.7) was added, the instrument itself changed little until the introduction of touch-tone dialling in the 1960s (Figure 12.8).

Each telephone line has an identifying number called its telephone number. To initiate a telephone call, the user enters the number of the other telephone into a numeric keypad on the phone. A public telephone (payphone) on the streets of different world cities has been used until today (Figure 12.9).

Payphones (public phones) were preceded by pay stations manned by telephone company attendants who would collect payment for calls placed. William Gray patented his coin-operated telephone in 1891. In 1889, a public telephone with a coin-pay mechanism was installed in the bank (USA). It was a "post-pay" machine; coins were inserted at the end of a conversation. In the last years, *customer-owned coin-operated telephones* (COCOT) have also appeared in the market, but their numbers are smaller due to the emergence of cellular phones.

A *mobile phone* (also known as a cellular phone) is a device that can make and receive telephone calls over a radio link while moving around a wide geographic area. It does

Photo by A. V. Valiulis

Figure 12.9. A public telephone (payphone) on a street in Vilnius (2013)

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Figure 12.10. The evolution of the mobile phone

so by connecting to a cellular network provided by a mobile phone operator thus allowing access to the public telephone network. The first hand-held mobile phone was demonstrated by Motorola in 1973, using a handset weighing around 1 kg. In 1983, the first mobile phone was commercially available.

In 1983, U.S. launched the first 1G mobile phone network. The phone had a talk time of just half an hour and took ten hours to charge. In 1991, the second generation (2G) cellular technology was launched in Finland on the GSM Standard. Ten years later, in 2001, the third generation (3G) was launched in Japan on the WCDMA standard. By 2009, it had become clear that, at some point, 3G networks would be overwhelmed by the growth of bandwidth-intensive applications like streaming media. Consequently, industry began looking to data-optimized 4th-generation technologies, with the promise of speed improvements up to 10-fold over the existing 3G technologies. The first two commercially available technologies billed as 4G were offered in USA (WiMAX Standard) and Scandinavia by TeliaSonera (LTE standard) (Figure 12.10).

Although originally designed for simple voice communications, most modern telephones have many additional capabilities. They may be able to record spoken messages, send and receive text messages, take and display photographs or video,

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Figure 12.11. Smartphone Nokia N8 (2010)

play music and surf the Internet. A current trend is phones that integrate all mobile communication and computing needs; these are called *smartphones*.

A smartphone is a mobile phone built on a mobile operating system, with more advanced computing capability and connectivity than a feature phone. Smartphones combine the functions of a personal digital assistant, portable media player, low-end compact digital camera, pocket video camera and GPS navigation units to form one multi-use device. Many modern smartphones also include high-resolution touchscreens and web browsers that display standard web pages as well as mobile-optimized sites (Figure 12.11). High-speed data access is provided by Wi-Fi and mobile broadband.

Wireless Transmission

Radio is the wireless transmission of signals through free space by the electromagnetic radiation of a frequency significantly below that of visible light, in the radio frequency range, from about 30 kHz to 300 GHz. Broadcasting has a unique advantage over other communication media – its messages are carried without any physical link between the sender and receivers. Unlike newspapers, magazines and the post, no rails, roads or vehicles are needed to carry messages, which results in large savings of time and expense. Unlike telegraph or telephone, no wires need be laid, and therefore radio messages can reach the remotest and most inaccessible locations, including a ship at sea or an aircraft.

The meaning and usage of the word "radio" has developed in parallel with developments within the field of communications and can be seen to have three distinct phases: electromagnetic waves and experimentation, wireless communication and technical development as well as radio broadcasting and commercialization. James Clerk Maxwell predicted the propagation of electromagnetic waves (radio waves) in 1873, and Heinrich Rudolf Hertz made the first demonstration of transmitting radio waves through free space (1887). Development from laboratory demonstration to a commercial entity spanned several decades and required the efforts of many practitioners.

The Scottish physicist and mathematician Maxwell, whilst experimenting with magnetic fields and light, predicted the existence of radio waves. In 1864, before the Royal Society of London, he said: "We have strong reason to conclude that light itself – including radiant heat and other radiation, if any – is an electromagnetic disturbance in the form of waves propagated through the electro-magnetic field according to electro-magnetic laws". Heinrich Hertz was a German professor of physics and the one who clarified and expanded the electromagnetic theory of light that had been put forward by James Clerk Maxwell. Hertz proved that an electric current swinging rapidly, back and forth in conducting wire would radiate electromagnetic waves into a surrounding space (an antenna) and travel at the speed of light. The carried out experiments on electromagnetic waves led to the development of wireless telegraph and radio. In recognition of his work, the unit of frequency of a radio wave – one cycle a second – is named Hertz (Hz). In 1878, David E. Hughes noticed that sparks could be heard in a telephone receiver when experimenting with his carbon microphone. He developed this carbon-based detector further and eventually could detect signals over a few hundred yards. In 1884, Temistocle Calzecchi-Onesti from Fermo, Italy performed experiments with tubes containing powder and nickel silver with traces of mercury metal filings and their reactions when conducting electricity. This would lead to the development of the iron filings filled coherer, a radio detecting device usually credited to Edouard Branly in 1890.

A detector is an electronic component in a radio receiver that recovers information contained in a modulated radio wave (Figure 12.12). The diode detector consists of a diode (D1) connected between the input and output of the circuit, with a resistor and capacitor in parallel from the output of the circuit to the ground to form a low pass filter. Early radio receiving equipment did not have to extract a sound wave from the incoming radio signal, but only "detect" the presence or absence of the radio signal and to produce clicks in the earphones (E1) of the receiver representing the symbols of Morse code (Figure 12.13).

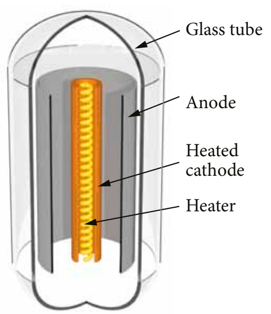


Figure 12.12. Thermionic diodes

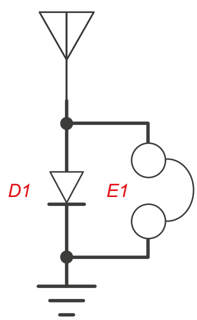


Figure 12.13. Detector radio set

In 1904, as a result of experiments conducted on Edison effect bulbs, John Ambrose Fleming developed a device he called an "oscillation valve" (passed a current in one direction only). The heated filament, or cathode, was capable of the thermionic emission of electrons that would flow to the plate (or anode) when it was at higher voltage. Electrons, however, could not pass in the reverse direction, because the plate was not heated and thus not capable of the thermionic emission of electrons (Figure 12.14). The result is a net flow of electrons from the filament to plate.

Around 1891, a circuit invented by Nikola Tesla could produce high voltage, a low current, high frequency and an alternating current that commonly became known as the Tesla Coil that helped him with making a discovery of transmitting and receiving powerful radio signals when tuning to resonate at the same frequency. By early 1895, Tesla was ready to transmit these signals over 50 miles to New York. In 1895, Alexander Stepanovich Popov built his first radio receiver that contained a coherer. Popov's receiver was created on the improved basis of Lodge's receiver and originally intended for the reproduction of its experiments.

In 1895, Italian inventor Guglielmo Marconi built equipment and transmitted electrical signals through the air from one end of his house to the other, which really was the birth of practical wireless telegraphy or radio. Shortly after the 1900s, Marconi held the patent rights for radio.

In 1901, signals were transmitted across the Atlantic ocean. The world's first wireless transmission of speech and music was performed at the Graz University of Technology in 1904. American Inventor Lee De Forest invented a three electrode vacuum tube that boosted radio waves, which allowed any broadcast signal to be heard loud and clear. On 12 January 1910, using his 'wireless telephony', De Forest sent a signal from the Metropolitan Opera House in New York to listen-

Figure 12.14. Modern vacuum tubes

ers up to 50 miles away. On April 15 at 12.30 am, in the middle of the night, the Titanic struck an iceberg in the North Atlantic and sank. Thanks to wireless transmission, 745 passengers were saved, but 1595 people lost their lives. Two wireless operators were employed by Marconi Telegraph Company onboard and sent these signals by Morse Code using a spark gap transmitter. In 1905, the naval battle of Port Arthur in the Russo-Japanese war was reported by wireless. In 1909, Robert E. Peary, Arctic explorer, radiotelegraphed: "I found the Pole". Inventor De Forest's invented *amplitude-modulated* (AM) radio that allowed for a multitude of radio stations. The earlier spark-gap transmitters did not allow for this. The Triode amplifier had now made transcontinental telephony (wire & wireless) more practical and led to the foundation of radio industry. Radio was used for passing on orders and communications between armies and navies on both sides in World War I. Germany used radio communications for diplomatic messages once it discovered that its submarine cables had been tapped by the British. Another use of radio in the pre-war years was the development of detecting and locating aircrafts and ships by the use of radar (RAdio Detection And Ranging).

After World War I, the Marconi Company in Great Britain began producing vacuum-tube continuous-wave transmitters (Figure 12.15). By early 1920, it was operating with a 15,000 watt transmitter. In 1913, a cascade-tuning radio receiver and a heterodyne receiver were introduced, and in 1919, shortwave radio was developed (Figure 12.16).

Public domain

Figure 12.15. A tube audio amplifier

Public domain

Figure 12.16. Back view of the vacuum tube radio chassis

A long wave occupies a part of the electromagnetic spectrum stretching from a frequency of 148.5 to 283.5 kilohertz (kHz), a medium wave occupies a part of the electromagnetic spectrum stretching from a frequency of 526.5 to 1606.5 kilohertz and VHF (Very High Frequency) broadcast band stretches from 87.5 megahertz (MHz) to 108.0 MHz. Long wave: 148.5 kHz to 283.5 kHz or expressed in wavelengths as 2020 meters to 1058 meters. Medium wave: frequencies from 526.5 kHz to 1606.5 kHz or expressed in wavelengths as 570 meters to 187 meters. VHF: frequencies from 87.5 MHz to 108.0 MHz could be expressed as 3.428 meters to 2.778 meters. As the frequency of the wave is increased the length of the wave decreases. The mathematical relationship between wavelength and frequency is expressed by the formula:

Wavelength (in meters) = 300,000 / Frequency (in kHz)

For Medium and Long Wave bands, it was decided from the outset of radio broadcasting in the 1920s that the method of modulating the radio wave would be *Amplitude modulation* (AM).

Frequency-modulated or FM radio was invented in 1933. FM improved the audio signal of radio by controlling the static caused by electrical equipment and the earth's atmosphere. Frequency modulation (FM) is a method of impressing data onto an alternating-current (AC) wave by varying the instantaneous frequency of the wave. This scheme can be used with analog or digital data. In analog FM, the frequency of the AC signal wave, also called the carrier, varies in a continuous manner. Thus, there are infinitely many possible carrier frequencies. In digital FM, carrier frequency shifts abruptly, rather than varying continuously. The number of possible carrier frequency states can be two, four, eight or even more different frequency states. Each specific carrier frequency represents a specific digital input data state.

The first Germanium transistor was built by John Bardeen, William Bradford Shockley and Walter Houser Brattain at Bell Telephone Laboratories USA in 1947. The transistor replaced vacuum tubes and mechanical relays and changed the future of electronics. The Regency TR-1 used Texas Instruments' NPN transistors and was the world's first commercially produced transistor radio. In 1954, the Regency company introduced a pocket transistor radio, the TR-1, powered by a "standard 22.5 V Battery." In 1955, the newly formed Sony company introduced its first transistorized radio. It was small enough to fit in a vest pocket and was powered by a small battery. It was durable, because it had no vacuum tubes to burn out. Over the next 20 years, transistors replaced tubes almost completely except for high-power transmitters. In Europe, FM radio stations in 1986 began using the subcarrier signal of FM radio to transmit digital data. Today, radio takes many forms, including wireless networks and mobile communications of all types as well as radio broadcasting.

By 1963, the first (radio) communication satellite, Telstar, was launched. In the late 1960s, long-distance telephone networks began converting to digital networks thus employing digital radios for many of its links.

Lithuanian Radio Products

The first radio sets were produced in Šiauliai (1925–1931), Kaunas (1932–1937) (Figure 12.17) and Vilnius 1922–1939 annexed by Poland (Figure 12.18).

On 8 February 1952, J. Stalin signed a precept ordering to establish a radio factory in Kaunas Resurrection Church that was not even furnished yet. In 1956, first tube television switches were manufactured. The production of Kaunas radio factory included radio *Mayak* (1956–1958), radio/record player Minija-4 (1966) (Figure 12.19), car radio *Neringa* (1959–1962), radio/record players *Minija* and *Nida* (1963–1973). Since 1978, the factory started manufacturing satellite TV tuners. In 1995, the company went bankrupt.

In the early 1990s, amateur radio experimenters started using personal computers with audio cards to process radio signals. Digital transmissions began to be applied to broadcasting in the late 1990s.

Figure 12. 17. Radio set „KARADI B37“ made in Kaunas, 1937

Figure 12.18. Radio set Elektrit Allegro made in Vilnius factory of electrical equipment

Figure 12.19. Radio/tape record Minija-4 made in Kaunas, 1966 (Museum of Energy and Technology, Vilnius)

CHAPTER 13 The Internet, Television and Radar

This chapter will help in

- revising the development of television;
- explaining scanning and rasterization processes, cathode ray tube oscilloscope;
- describing the transmission of colour television;
- discussing the production of TV sets in Lithuania;
- introducing the Internet – the global information system;
- presenting the Lithuanian Academic and Research Network LITNET;
- defining the space-based satellite navigation system GPS;
- dealing with radio detection and ranging (radar) development.

Development of Television

Television (TV) is a telecommunication medium for transmitting and receiving moving images that can be black-and-white or coloured and with or without accompanying sound. Television was not invented by a single inventor, instead, many people working together and alone over the years, contributed to the evolution of television.

In its early stages of development, television employed a combination of optical, mechanical and electronic technologies to capture, transmit and display a visual image. By the late 1920s, however, those employing only optical and electronic technologies were being explored.

The first images transmitted electrically were sent by early mechanical fax machines developed in the late 19th century. The concept of electrically powered transmission of television images in motion was first sketched in 1878 as the telephonoscope, shortly after the invention of the telephone. At the time, it was imagined that light could be transmitted over copper wires, as sounds were.

The idea of using scanning to transmit images was put to actual practical use in 1881 through the application of a pendulum-based scanning mechanism. The concept of "rasterization" explains the process of converting a visual image into a stream of electrical pulses.

In 1884, the first electromechanical television system that employed a scanning disk, a spinning disk with a series of holes spiralling toward the centre, for rasterization purposes was patented in Germany. The holes were spaced at equal angular intervals such that in a single rotation the disk would allow light to pass through each hole and onto a light-sensitive selenium sensor producing electrical pulses. This design would not be practical until advances in amplifier tube technology became available. Later design would use a rotating mirror-drum scanner to capture the image and a *cathode ray tube* (CRT) as a display device; however, moving images were still not possible due to the poor sensitivity of selenium sensors.

In 1926, a television system utilizing fully electronic scanning and display elements and employing the principle of "charge storage" within the scanning tube was designed in Hungary. In some time in Japan, a television system with a 40-line resolution was demonstrated. This was the first working example of a fully electronic television receiver. Picture quality was very poor, and the screens were only an inch or so wide and usually made up of only 30 to 60 lines. Most of these sets used a motor to rotate a metal disk to produce the picture, with a neon tube behind the disk to provide light (Figure 13.1). Later mechanical systems used the lens disk, mirror screw and mirror drum. In UK, regularly scheduled 30 line television programming was first broadcast by the BBC in 1929. At first, only the picture was transmitted, but from 1930, sound and pictures were transmitted together. Due to the poor quality of the picture, mechanical television was not a success. The transmission of mechanical television by the BBC continued until 1935 and in the Soviet Union – until 1937.

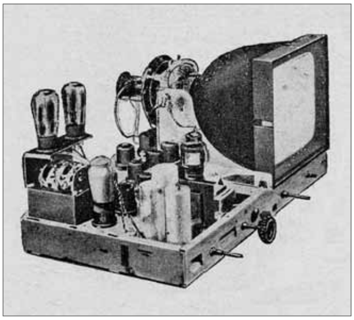
a)

b)

Early Television Museum

Figure 13.1. Mechanical television. A General Electric TV set *Octagon* made in 1928: a – front view of *Octagon*, b – rear view of *Octagon*

By the early 1930s, it was clear that mechanical television systems could never produce the picture quality required for commercial success. Electronic



Early Television Museum

Figure 13.2. Chassis of the 1938 German set *Volkfernseher*

television requires a cathode ray tube (picture tube) to display the picture and some sort of an electronic camera tube to capture the image (Figure 13.2). The cathode ray tube was the easier of these to develop, but the emergence of electronic television was delayed for years until a suitable camera tube could be developed. Electronic television is based on the development of the cathode ray tube, which is the picture tube found in modern TV sets. German scientist Karl Braun invented the cathode ray tube oscilloscope in 1897. Russian inventor Vladimir Zworykin created an improved cathode-ray tube called the *kinescope* in 1929.

In 1936, the Olympic Games in Berlin were carried by cable to television stations in Berlin and Leipzig where the public could view the games live. In 1936, the BBC began transmitting the world's first public regular high-definition service. It therefore claims to be the birthplace of television broadcasting as we know it today (Figure 13.3).

The outbreak of World War II prevented it from TV being manufactured on a large scale until the end of the war. Since the 1950s, television has been the main medium for molding public opinion (Figure 13.4).

Since the 1970s, the availability of video cassettes, laserdiscs and DVDs have resulted in the television set frequently being used for viewing recorded as well as broadcast material. In recent years, Internet television has seen a rise in television available via the Internet.

Early Television Museum

Figure 13.3. A TV set made in 1939 by the Andrea Radio Corp

Early Television Museum

Figure 13.4. French TV set *Grammont* apparently made around 1951

The broadcast television system is typically disseminated via radio transmissions in the 54–890 MHz frequency band. Signals are often transmitted with stereo sound in many countries. Until the 2000s, broadcast TV programs were generally transmitted as an analog television signal, but during the decade, several countries went almost exclusively as a digital signal (in Lithuania from 2012).

A standard television set comprises multiple internal electronic circuits, including those for receiving and decoding broadcast signals. A visual display device is called a video monitor. A television system may use different technical standards such as digital television (DTV) and high-definition television (HDTV). Television systems are also used for surveillance and industrial process control in the places where direct observation is difficult or dangerous.

Colour TV was by no means a new idea. The earliest proposal was a German patent in 1904, while in 1925 V. Zworykin (USSR) filed a patent disclosure for an all-electronic colour television system. A successful colour television system began commercial broadcasting in 1953.

Formerly known cable television was born in Pennsylvania in the late 1940's. The first successful colour television system began commercial broadcasting in 1953. The first television remote control was created in 1950. It could turn a television on and off as well as change channels. However, it was not a wireless remote control and was attached to the television by a bulky cable.

In 1961, Moscow welcomes home astronaut Yuri Gagarin and these meetings was the first western viewing of live television from the USSR. In 1962, the first satellite *Telstar* to send television signals was launched. In 1964, the first prototype for a plasma display monitor was invented at the University of Illinois. In 1969, Neil Armstrong walked on the moon and a worldwide audience of 720 million watched the event live. In 1975, the distribution of a TV program via satellite began. In 1986, scrambling satellite-fed cable television programming started. The sale of decoders and program subscriptions to home dish owners began. In 1995, the first television program was delivered via the Internet. In 1996, one billion television sets worldwide were counted.

Production of TV Sets in Lithuania

In 1930, a radio laboratory in Šiauliai produced the first mechanical TV set in Lithuania. This receiver was tested by the reception of television broadcasts from Berlin. The TV sets in the Lithuanian house appeared in 1955 and received signals from the TV Centre launched in Riga. Since 1957, Lithuania started broadcasting its own TV programmes. In 1961, a television factory was founded in Šiauliai, and two years later, the first TV set (Temp-6 model) appeared in the market. In 1968, the factory moved to a new building and started producing black and white TV sets with the diagonal tube of 59 cm and from 1973 with the diagonal of 61 cm. In 1978, the first colour TV Tauras-714 and, in 1984, a TV set with semiconducting-integral schemes were assembled. Television became lighter as it used much less electricity and was of good quality. In 1970, the TV factory in Šiauliai started producing stationary and mobile broadcasting centres equipped with coloured television. The factory became the most reliable hardware developer and manufacturer in the former Soviet Union. After regained independence, the factory had to compete in the global market. From 2004, LCD TVs and the TVs with built-in digital tuners have been produced. The TVs have been designed to accept all formats of digital television signals: besides a traditional cathode tube, TV sets and those with liquid crystal displays (LCD) as well as exceptional products – TV sets with integral digital receivers are proposed to markets. The TV sets are able to accept all formats of digital TV signals, including ethereal, cable, multi-record MPEG –2 and MPEG-4 standards, high definition and with a port for coded signals. Cathode ray tubes (diagonal of 14", 15", 20" and 21") were produced in Panavėžys and cathode ray tube deflection coils – in Vilnius. About 98% of Lithuanian TV production is sold in European countries.

The most famous products of the Kaunas radio factory were TV sets *Šilelis* (Shilyalis). Particularly saleable was the model *Šilelis* (Figures 13.5 and 13.6). From 1972, its design was widely accepted on the world scale.

Figure 13.5. Portable TV set *Šilelis S-445D*, 1979 (diagonal of 30 cm)

Photo by A. V. Valiulis

Figure 13.6. Portable TV set *Šilelis 403D*, 1977

The Internet

The Internet is the worldwide publicly accessible network of interconnected computer networks that transmit data by packet switching using the standard Internet Protocol (IP). It is the "network of networks" that consists of millions of interconnected smaller domestic, academic, business and government networks that together carry various information and services such as electronic mail, online chat, file transfer, interlinked Web pages and other documents. The World Wide Web ("www" or simply the "web") is global information medium the users of which can read and write via computers connected to the Internet. The term is often mistakenly used as a synonym for the Internet itself, but the Web is a service that operates over the Internet, just as the e-mail also does.

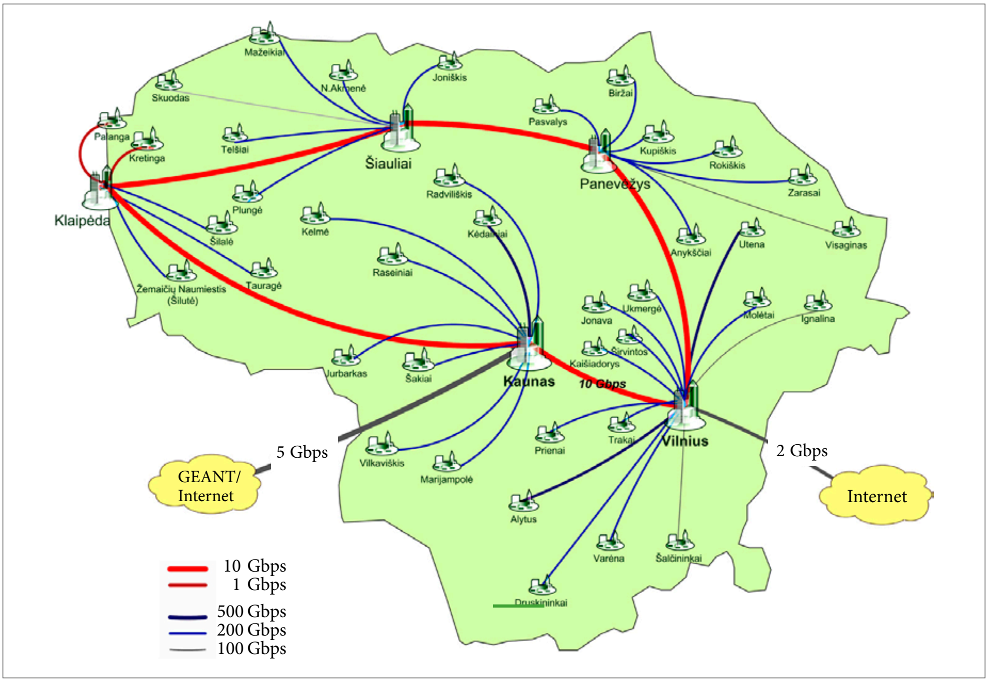
The history of the Internet began with the development of electronic computers in the 1950s. The concept of data communication – transmitting data between two different places connected via some kind of electromagnetic medium such as radio an electrical wire predates the introduction of the first computers. Such communication systems were typically limited to point to point communication between two end devices. Early computers used the technology available at the time to allow communication between the central processing unit and remote terminals. As the technology evolved, new systems were devised to allow communication over longer distances (for terminals) or with higher speed (interconnecting local devices) necessary for the mainframe computer model. Using these technologies made it possible to exchange data (such as files) between remote computers. However, the point to point communication model was limited, as it did not allow for direct communication between any two arbitrary systems; a physical link was necessary. Fundamental theoretical work concerning the data transmission and information theory was developed during the early 20th century.

The public was first introduced to the concepts that would lead to the Internet when a message was sent from computer science laboratory at the University of California to the University of Utah (1969). ARPANET (Advanced Research Projects Agency Network) became the technical core of what would become the Internet, and a primary tool for developing the technologies used. Many different networks were developed in the late 1960s and early 1970s using a variety of protocols. In 1982, the Internet protocol suite (TCP/IP – Transmission Control Protocol/ Internet Protocol) was standardized, and consequently, the concept of a world-wide network of interconnected TCP/IP networks, called the Internet, was introduced. Commercial Internet service providers (ISPs) began to emerge in the late 1980s and early 1990s. Since the mid-1990s, the Internet has had a huge impact on culture and commerce, including communication by electronic mail, two-way interactive video calls, World Wide Web forums, blogs, social networking and online shopping sites. The takeover of the Internet over the global communication was almost instant in historical terms: it only communicated 1% of the information flowing through two-way telecommunications networks in the year 1993, already 51% by 2000 and more than 97% of the telecommunicated information by 2007.

In 1980, ENQUIRE (Enquire Within Upon Everything) – a personal database of people and software models was built in the European Organization for Nuclear Research (CERN), Switzerland, which also was a way to play with hypertext. In 1990, Berners-Lee built all tools necessary for a working Web: the HyperText Transfer Protocol (HTTP) 0.9, the HyperText Markup Language (HTML), the first Web browser (named WorldWideWeb that was also a Web editor), the first HTTP server software (later known as CERN httpd), the first web server (http://info.cern.ch) and the first Web pages that described the project itself. The WWW project was started to allow high energy physicists to share data, news and documentation. There was still no graphical browser available for computers. This gap was filled in 1992 by creating embedded graphics, scripting and animation. The turning point for the World Wide Web was the introduction of the Mosaic web browser – a graphical browser developed in 1993. The first Microsoft Windows browser (Cello) was released in 1993. The period 1996–1998 was the years of WWW commercialization. Many companies that began as online retailers blossomed and became highly profitable. Traditional media outlets (newspaper publishers, broadcasters and cable casters in particular) also found the Web to be useful and profitable. During 2002–present time, a handful of companies found success developing business models that helped with making the World Wide Web more compelling experience. These include airline booking sites, Google's search engine, Facebook, Wikipedia, YouTube video viewing website, etc. The continued extension of the World Wide Web has focused on connecting devices to the Internet.

Lithuanian Academic and Research Network LITNET was established in 1991 and had X.25 satellite connectivity to the University of Oslo. Invaluable assistance was provided by the University of Oslo, (Norway), Denmark, Iceland, Finland, Sweden and NORDUnet academic computer networks. The goal of developing LITNET infrastructure and services has been to guarantee data transfer to the institutions and inter-institutional IT projects, to incorporate Lithuanian educational institutions and to transfer the most advanced technologies and services of European academic networks to Lithuania. LITNET was the first Internet service provider in Lithuania. Schools started connecting to this network in 1997 with their own initiative and financial resources. In 2011, 880 educational institutions using LITNET Internet, including 668 secondary schools, were counted.

LITNET is integrated into the unified European academic network GEANT (Gigabit European Academic Network Technology). The optical fiber ring was con-



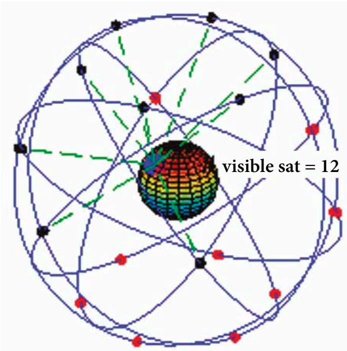
With permission from Lithuanian Research and Education Network (LITNET)

Figure 13.7. Scheme for the LITNET data transfer channel (2014)

necting 5 biggest Lithuanian cities (for the period 2005–2008). At the end of 2007, the RAIN project of the rural broadband network opened new possibilities of providing the broadband Internet. The goal of the RAIN project is to create equal opportunities in rural as well as urban regions for installing as similar as possible technological solutions based on contemporary and enduring technologies. There is a constant growing trend towards IT services the examples of which include library systems, campus information systems, distance learning systems, wireless networking, virtual private networks, etc. (Figure 13.7).

Global Positioning System

The Global Positioning System (GPS) is a space-based satellite navigation system that provides location and time information in all weather conditions, anywhere on or near the Earth where there is an unobstructed line of sight to four or more GPS satellites. The inspiration for GPS came when the Soviet Union launched the first man-made satellite Sputnik in 1957. Two American physicists made their own decision on monitoring radio transmissions from the Sputnik. Within hours, they realized that, because of the



Public domain

Figure 13.8. The orbits of GPS satellites

Doppler effect, they could pinpoint where the satellite was along its orbit from the Doppler shift. The first satellite navigation system (US) *Transit* was first successfully tested in 1960. It used a constellation of five satellites and could provide a navigational fix approximately once per hour. In the 1970s, the ground-based Omega Navigation System based on the phase comparison of signal transmission from the pairs of stations became the first worldwide radio navigation system. This system drove the need for a more universal system the navigation solutions to which were of low accuracy. In 1960, the US Air Force proposed a radio-navigation system called MOSAIC (MObile System for Accurate ICBM Control), and in 1963, the GPS concept was born. In 1973, the program *Navstarn* was launched. The constellation of Navstar satellites, *Navstar-GPS*, later was shortened simply to GPS. When the Korean Air Lines aircraft carrying 269 people was shot down in 1983 after straying into the USSR's prohibited airspace, US President R. Reagan issued a directive making GPS freely available for civilian use. The first satellite was launched in 1989, and the 24th satellite was launched in 1994 (Figure 13.8).

The basic concept of GPS is that a GPS receiver calculates its position by precisely timing the signals sent by GPS satellites high above the Earth. Each satellite continually transmits messages that include the time the message was transmitted and the satellite position at the time of message transmission. The receiver computes the distance to each satellite. Each of these distances and satellite locations define a sphere. The receiver is on the surface of each of these spheres when the distances and satellite locations are correct. This location is then displayed, perhaps with a moving map or latitude and longitude; elevation information may be included. Many GPS units show derived information such as direction and speed calculated from a changing position. As regards typical GPS operation, four or more satellites must be visible to obtain an accurate result (Figure 13.10). The current GPS consists of three major segments, including a *space segment* (SS), a *control segment* (CS) and a *user segment*. GPS satellites broadcast signals from space, and each GPS receiver uses these signals to calculate its three-dimensional location (latitude, longitude and altitude) and the current time.

While originally a military project, GPS is considered a dual-use technology, meaning it has significant military and civilian applications. GPS has become a widely deployed and useful tool for commerce, scientific uses, tracking and surveillance. GPS accurate time facilitates everyday activities such as banking, mobile phone operations and even the control of power grids by allowing well synchronized hand-off switching. For example, a mobile phone tracking refers to attaining the current position of a mobile phone, stationary or moving. Localization may occur via GPS. To locate the phone using the multilateration of radio signals, it must emit at least a roaming signal to contact the next nearby antenna tower; however, the process does not require an active call.

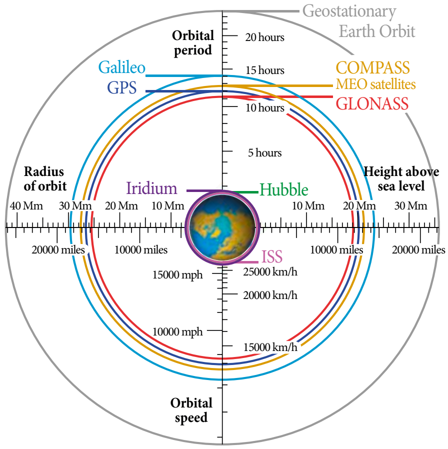


Figure 13.9. A comparison of GPS (US), GLONASS (Russia), Galileo (EU) and Compass (China) satellite navigation system orbits with the International Space Station, Hubble Space Telescope, Iridium constellation orbits and Geostationary Earth Orbit

GSM is based on signal strength to nearby antenna masts. Mobile positioning, which includes location based service that discloses the actual coordinates of a mobile phone bearer, is a technology used by telecommunication companies to approximate where a mobile phone, and thereby also its user (bearer), temporarily resides (Figure 13.9).

Other satellite navigation systems are in use or various states of development include (Figure 13.9):

- GLONASS – Russia's global navigation system; fully operational worldwide;
- Galileo – a global system being developed by the European Union and other partner countries and planned to be operational by 2014 and fully deployed by 2019;
- Beidou – People's Republic of China's regional system currently limited to Asia and the West Pacific;
- COMPASS – People's Republic of China's global system planned to be operational by 2020;
- IRNSS – India's regional navigation system planned to be operational by 2014 and covering India and Northern Indian Ocean;
- QZSS – Japanese regional system covering Asia and Oceania.

The Gulf War from 1990 to 1991 was the first conflict where GPS was widely used. In 1993, GPS achieved the initial operational capability indicating a full constellation (24 satellites) was available. In 2000, the precision of civilian GPS from 100 to 20 meters was improved. GPS is owned and operated by the United States Government as a national resource. In 2004, U.S. signed an agreement with the

European Community establishing cooperation related to GPS and Europe's planned *Galileo* system, and in some time, information on the successful tests on the assisted GPS for mobile phones was announced.

The current motion of tectonic plates is nowadays revealed from remote sensing satellite data sets calibrated with ground station measurements. Depending on how they are defined, there are usually seven or eight "major" plates: African, Antarctic, Eurasian, North American, South American, Pacific and Indo-Australian (Figure 13.10). The latter is sometimes subdivided into Indian and Australian plates. There are dozens of smaller plates, the seven largest of which include the Arabian, Caribbean, Juan de Fuca, Cocos, Nazca, Philippine Sea and Scotia.

The movement of the plates has caused the formation and break-up of continents over time, including the occasional formation of a supercontinent that contains most or all of the continents. Plate motions range from up to a typical of 10–40 mm/yr (Mid-Atlantic Ridge; about as fast as fingernails grow), to about 160 mm/yr (Nazca Plate; about as fast as hair grows). The supercontinent of Columbia or Nuna formed during a period of 2,000 to 1,800 million years ago and broke up about 1,500 to 1,300 million years ago. The supercontinent of Rodinia is thought to have formed about 1 billion years ago and to have embodied most or all of the Earth's continents, which therefore were broken up into eight continents around 600 million years ago. The eight continents later re-assembled into another supercontinent called Pangaea that broke up into Laurasia (became North America and Eurasia) and Gondwana (became the remaining continents).

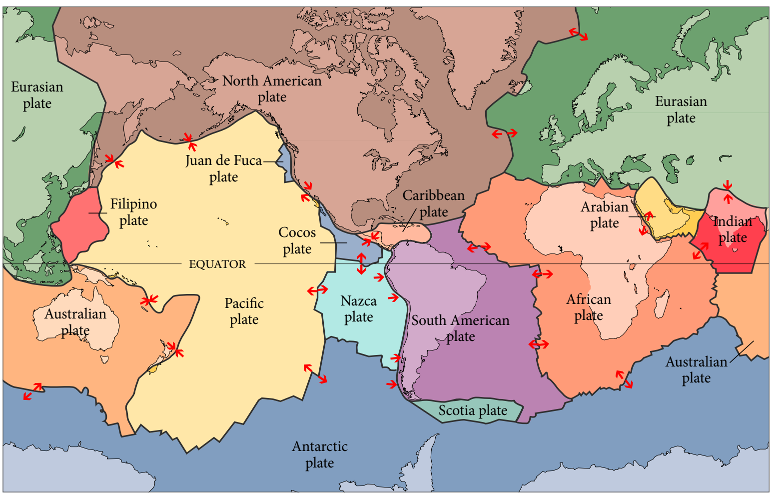


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Figure 13.10. Tectonic plates of the world

Radar

Radar (Radio Detection and Ranging) is an object detection system that uses radio waves to determine the range, altitude, direction or speed of objects. It can be used for detecting aircrafts, ships, spacecrafts, guided missiles, motor vehicles, weather formations and terrain. The radar dish or antenna transmits the pulses of radio waves or microwaves that bounce off any object in their path. The object returns a tiny part of the wave's energy to the dish or antenna usually located at the same site as the transmitter. The delay caused by the echo measures the distance. The direction of the beam determines the direction of the reflection. The polarization and frequency of the return can sense the type of the surface. *Navigational radars* scan a wide area two to four times per minute. They use very short waves that reflect from earth and stone. They are common on commercial ships and long-distance commercial aircrafts. *Search radars* scan a wide area with the pulses of short radio waves. They usually scan the area two to four times a minute. Sometimes search radars use the Doppler effect to separate moving vehicles from a clutter. *Targeting radars* use the same principle as search radars but scan a much smaller area far more often, usually several times a second or more. *Weather radars* resemble search radars but use radio waves with circular polarization and a wavelength to reflect from water droplets. Some weather radars use the Doppler effect to measure cloud speed (Figure 13.11).

The radio detection system firstly was patented in U.S. in 1933. However, for the next decade, research became shrouded in wartime secrecy. In 1935, the first

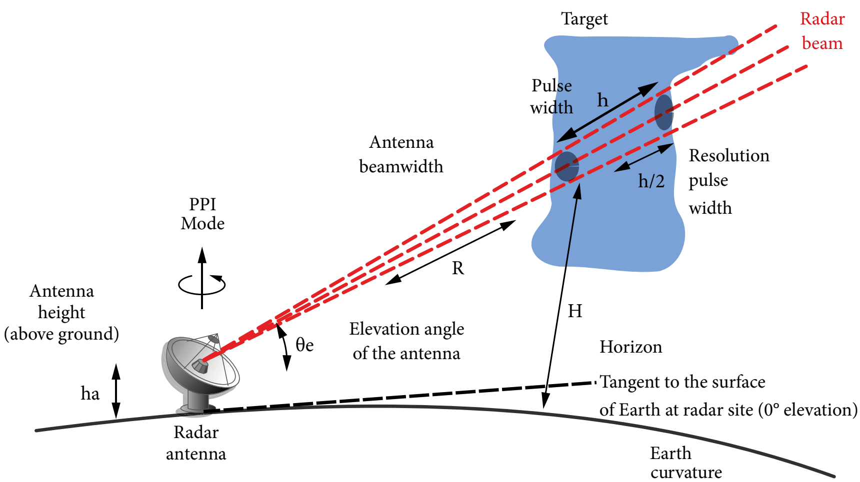

Figure 13.11. The path of the radar beam considering height

practical radar (Radio Detection and Ranging) equipment for detecting an aircraft was developed in the UK. Both the Allies and Axis powers had radar during the Second World War. Since the war, it has become essential not only for a host of military purposes, but also for civilian air traffic control. Radar operates by sending out pulses of radio energy, a tiny portion of which is reflected by the objects in its path and picked up by highly directional antennas. The range of such objects can be calculated automatically from the difference in time between transmitted and received pulses. Early-warning radars use 10 m wavelengths, and high-resolution system wavelengths are down to 10 cm.

In 1933, the Bell Telephone Laboratories discovered that radio waves were being generated by stars. This began a new era in astronomy – *radio-astronomy* that has greatly expanded our knowledge of star creation and life processes and has detected many stars invisible to light telescopes. Immense antennas are used for this purpose. The largest one in Puerto Rico, with 3 km in diameter, is capable of detecting phenomena out to 15,000 million light-years away.

CHAPTER 14 | Printing

This chapter will help in

- introducing the Gutenberg's early printing process;
- explaining hot-metal typesetting in printing;
- discussing a rise in colour printing (chromolithography, zincography);
- describing the development of writing machines for office printing;
- defining the duplication of information by copying (hektograph process, xerography).

Gutenberg's Early Printing Process

Printing is a process for reproducing text and images, typically with ink on paper using a printing press. The earliest form of printing was woodblock printing, with is dating to before 220 A.D. Printing as a means of making many identical copies arose in the late Middle Ages. Johannes Gutenberg was a German printer and publisher who introduced printing to Europe. His invention of mechanical movable type printing (around 1439) started the printing revolution. It played a key role in the development of the Renaissance, the Reformation, the Age of Enlightenment and the Scientific Revolution. In 1455, Gutenberg completed his 42-line Bible known as the Gutenberg Bible. About 180 copies were printed. Gutenberg's early printing process, and what tests he may have made with movable type, are not known in great detail. In the standard process of making type, a hard metal *punch* (made by punchcutting with the letter carved back to front) is hammered into a softer copper bar creating a matrix. This is then placed into a hand-held mould and a piece of type is cast by filling the mould with molten type-metal that cools almost at once, and the resulting piece of type can be removed from the mould. The matrix can be reused for creating hundreds, or thousands, of identical types so that the same character appearing anywhere within the book will look very uniform.

Figure 14.1. Metal movable type

After casting, types (sorts) are arranged into type-cases and used for making up pages that are inked and printed, which is a procedure that can be repeated hundreds, or thousands, of times (Figure 14.1). The sorts can be reused in any combination earning the process having the name of "movable type".

As regards printing development, J. Gutenberg's contribution in respect of the composition of type metal should not be overlooked. He spent years searching for an alloy with certain properties such as a low melting point for the convenience of casting yet strength enough to resist the wear of thousands of impressions. Such alloy consisted mainly of lead and tin with some antimony. The Gutenberg's process survived virtually unchanged for three and a half centuries.

Printing spread rapidly. Italy gave us two faces on which nearly all western types have been based to this day: roman and *italic*. Early in the 15th century, technological books on mining, the assay of metals, glassmaking and distilling began to appear. The books were joined by periodicals. The earliest periodical *Philosophical Transactions* of the Royal Society is dated 1665. In 1814, a newspaper using a machine press was printed for the first time. Production shot up to 1100 sheets an hour. In 1827, printing rate grew up to 4000 sheets an hour, and twenty years later, the type-revolving press reached up to 8000 sheets an hour. The next stage was the replacement of the reciprocating flat-bed by a continuously rotating cylinder press carrying the type on a curved form. The first press of this kind was constructed in the U.S.A. in 1846. An essential element in the quest for speed was the rotary principle, by which the type form was curved to follow the shape of a cylinder replacing the flat form on a flat-bed press. At first, type was cast in wedge-shaped moulds and clamped to cylinders, but this led to problems with dislodged type, so curved stereotypes were introduced. During the 18th century, attempts were made to produce a mould of the type form, and cast a printing plate from it to enable copies to be run off while the original type was distributed to the use of a fresh piece of work (Figure 14.2). From 1830, lithography virtually triumphed over printed illustrations. The great mass of produced printed illustrations required services of the armies of artists. The invention and development of photography from the late 1820s provided a method of factual representation, but the large-scale mechanical reproduction of photographs was not easy. Only in the 1880s, a photo-process engraved plate that could be printed in a press like a hand-engraved one became a possibility. The

Figure 14.2. Heidelberg sheet-fed offset printing machine (Museum of Energy and Technology, Vilnius)

original was copied by a process camera in which a cross-line screen was interposed in front of the negative producing a photograph consisting of thousands of dots which, viewed from a distance, gave the illusion of tone variation. From the 'dotted' negative, the 'dotted' sensitized printing plate was prepared, which could be printed off with the rest of the text. In 1887, the Monotype system was developed. Its keyboard unit produced a strip of paper tape with holes punched in patterns corresponding to the characters required, and this controlled the assembly of matrices in the caster, from which types cast singly were ejected and fed to an assembly point until a complete line was formed.

Computers were introduced in order to decide about line length thus justifying the right-hand margin, spacing between words, etc. The next stage used a system of electronics to build up the images of the characters in a cathode-ray tube. Beyond that, the character image could be stored in the form of magnetic impulses in a computer (digital storage) cutting out the need for a photographic matrix where the stage when hot-metal technology became the past technology was reached.

Colour Printing

Colour printing is the reproduction of an image or text in colour. The earliest way of adding colour to items printed on paper was hand-colouring. Early European printed books often left spaces for initials, rubrics and other elements to be added by hand, just as they had been in manuscripts, and a few early printed books had elaborate borders and miniatures added. The 19th century also saw the flowering of colour printing. Until that time, nearly all the colour that appeared in books was applied by hand, but then, several processes were developed for colour printing. Technical progress in this field depended on a proper understanding of the theoretical background of the nature of colours and how they could be produced by mixing the three primary colours – red, blue and yellow. Some progress was made in applying the three-colour theory to photography. The problem was how to convince the eye that it sees the range of colours in the original when what it really sees is a collection of dots of yellow, magenta (red) and cyan (blue) ink of varying intensities. The fourth colour added was black giving a greater depth of tone.

Chromolithography was another process, which by the end of the 19th century had become dominant, although this used multiple prints with a stone for each colour the mechanical separation of which, initially using the photographs of the image taken with three different colour filters, reduced the number of prints needed to three. Zincography, with zinc plates, later replaced lithographic stones and remained the most common method of colour printing until the 1930s.

Before digital imaging was developed, the traditional method was photographing the image three times thus using a filter for each colour. However, this can be achieved, and the desired result is three grayscale images representing the red, green and blue (RGB) components of the original image. The process of colour separation starts by dividing the original artwork into red, green and blue components (for example by a digital scanner). Cyan, magenta and yellow are the three basic colours used for colour reproduction. When these are variously used in printing, the result should be a reasonable reproduction of the original, but in practice this is not the case. Due to limitations to inks, darker colours are dirty and muddied. To resolve this, a black separation is also created, which improves the shadow and contrast of the image.

Office Printing

Ideas for 'writing machines' go back to the 18th century when many designs were conceived. The first commercially successful typewriter was evolved by the American inventors Christopher Sholes and Carlos Glidden to type letters on paper. The first practical typewriter appeared in 1873. During the 1880s, many different designs appeared. Most used type bars for carrying type striking down, up or sideways on paper. Others carried types on a wheel or cylinder. It was the Underwood No. 1 of 1897 with side-strike type bars enabling the typist to see the things she was typing, which proved to be the standard arrangement for many years to come (Figure 14.3).

In 1873, Sholes and Glidden first arranged the keys alphabetically (Figure 14.4). In 1881, classes for young ladies were delivered in New York. Five years later, 60,000 female typists were estimated in the USA. The typewriter revolutionized office work and brought about a social revolution providing millions of women with the possibility of finding employment in the places where none existed before.

Figure 14.3. Underwood No 5 typewriter

Although the Sholes & Glidden is generally regarded as the first production typewriter in history, the Hansen writing ball, in fact, beat the S&G by no less than four years. The most striking feature of the writing ball is the semi-sphere on the top of the machine with 52 keys sticking out like a giant pin cushion (Figure 14.5). At the lower end of each stem is a character, cast in exactly the right angle to create a perfectly even print on the central printing point under the ball. The escapement mechanism moved the paper frame that held paper one space until the end of the line was reached. By pushing the button on the left in front of the ball all the way down, the carriage was turned concentrically back to the beginning of the line and moved one line to the left. The first writing ball model was built in 1865.

The first electric typewriter was made in 1902. Other designs followed, but it was not generally in use in offices until the late 1950s (Figure 14.6). In 1961, IBM abandoned the type bars and moving carriage and substituted a moving 'golf ball' typeahead. Another improvement was the introduction of automatic control, first mechanically with punched paper tape and then electronically by the computer.

A significant innovation was the *shift key* that physically "shifted" either the basket of type bars, in which case the typewriter is described as "basket shift", or the paper-holding carriage, in which case the typewriter is described as "carriage shift".

Either mechanism caused a different portion of the type bar to come in contact with the ribbon/platen. The result is that each type bar could type two different characters, cutting the number of keys and type bars in half.

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Figure 14.4. Early Sholes and Glidden typewriter (1874)

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Figure 14.5. Hansen writing ball (1870)

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Figure 14.6. IBM Electric typewriter (1950)

The obvious use for this was to allow letter keys to type both upper and lower case; however, normally the number keys were also duplexed thus allowing access to special symbols such as percent (%).

Electrical typewriter designs removed a direct mechanical connection between the keys and the element that struck the paper. Not to be confused with later electronic typewriters, electric typewriters contained only the motor – a single electrical component. Where the keystroke had previously moved a type bar directly, now, it engaged mechanical linkages that directed mechanical power from the motor into the type bar.

At the beginning of the third millennium, the increasing dominance of personal computers, the introduction of low-cost, truly high-quality laser and inkjet printer technologies and other electronic communication techniques largely replaced typewriters.

Copying

Copying is the duplication of information not using the process that originally generated it. Persons or offices often need more than one copy of a document. They usually require a copy of outgoing correspondence or a letter. Sometimes they want to circulate the copies of documents (projects, reports, etc.) to several interested parties. Hand copying was invented between the mid-17th century and late 18th century. In 1780, the created letter copying press had a significant impact in offices.

In 1876, the hektograph process was introduced. A text was made with special aniline ink. The sheet was then placed face down on a tray containing gelatine and pressed gently for a minute or two with the result that most of the ink transferred to the surface of the gelatine. Copies were made by using a roller to press blank papers onto the gelatine. Each time, a copy was made, and some ink was removed from the gelatine. Consequently, successive copies were progressively lighter. The cylinder duplicator appeared in 1905 and contained a composition to receive a negative of pen or typewritten matter made with duplicating ink. Duplicate copies were made by running the roller over blank papers. The maker claimed that the device would make 50 to 75 copies of letters written with a typewriter and 100 to 125 copies of letters written with a pen.

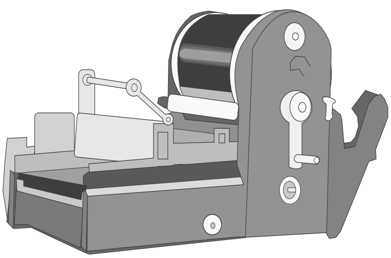
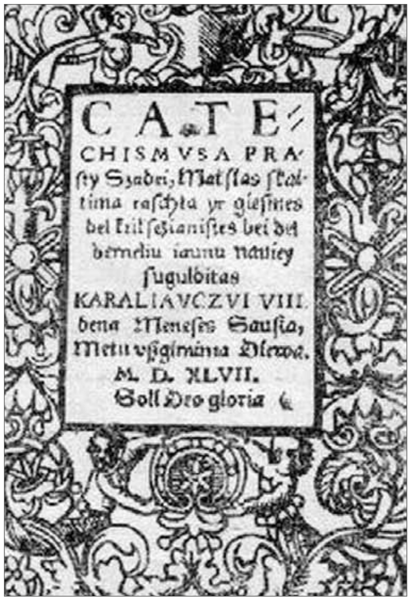


Figure 14.7. Illustration of a typical mimeograph machine

The Chester Carlson (USA) applied for photocopier patent in 1937. The first photocopiers were introduced in the market in 1949. A single copy has rarely sufficed in offices, and various processes were developed to make additional copies. In 1806, "*carbonic*" or "carbonated" paper – the first recorded appearance of carbon paper was patented. The stencil and duplicator were a form of screen printing. In this process, ink is forced through the fine mesh of a silk screen partly blanked off by a stencil plate. The stencil and duplicator gave good service for many years and are still in use, but they have been in a large measure replaced by two processes (Figure 14.7). One was a small offset litho machine introduced in 1927. The other was an electrostatic copying machine that emerged in the 1950s under the name of *xerography* (Greek words: *xeros* for "dry" and *graphos* for "writing").

Printing in Lithuania

The first printed books reached Lithuania before the beginning of the Reformation (16th century). The first dated printed book in Russia (*Apostolos*) appeared in 1564. Right after the issue of "Apostol", the printers became subject to persecutions from books' copyist fearing competition. After the arson that destroyed printing, workshop printers were forced to run away to the Grand Duchy of Lithuania. The first Russian printers were received by the Lithuanian great hetman H. Khodkevich at his estate in Zabludov near Grodno; they were used for publishing *Ievanheliie uchytel'noie* (Didactic Gospel, 1569) and *Psaltyr* (Psalter, 1570). The first typography in Vilnius was opened in approximately 1575. During the 16th–17the centuries, there were 36 typographies in the territory of the Grand Duchy of Lithuania. However, both multiplied religious discussions and an increase in printing in Western Europe activated interest in printing books in Lithuania. The first book in the Lithuanian language was printed in 1547 in Königsberg (it was a Protestant Catechism by Martynas Mažvydas) and was a small format publication (18×11 cm) that consisted of 79 pages printed in Gothic lettering (except for some


CATE-
CHISMVSA PRA-
sty Szadei, Makslas skai-
tima raschta yr giesmes
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KARALIAVCZVI VIII.
dena Meneses Sausia,
Metu vzgimima Diewa.
M. D. XLVII.
Soli Deo gloria

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Figure 14.8. Catechism of Martynas Mažvydas (1547)

titles and the text in Latin). The Latin text is printed very evenly, but the Gothic letters of the Lithuanian text are often uneven and not clear; technically, the printing shop must have been of very poor quality. Insufficient attention was paid to typefaces: when running out of one typeface, printers simply replaced it by another. The number of copies printed was probably somewhere between 200 and 300. The book also includes a Lithuanian primer with an alphabet adapted for the first time to Lithuanian sounds, which clearly follows the Latin pattern.

The presence of printed books became the signal factor in changing the ancient Lithuanian cultural attitude against literacy (Figure 14.8). The necessity of literacy became evident. Unfortunately, at the same time, a growth in literacy coincided with the refusal of Lithuanian culture among the nobility in the Grand Duchy of Lithuania.

CHAPTER 15 | Photography

This chapter will help in

- discussing the creation of durable images by recording light;
- describing the camera obscura, Niépce, Daguerre and Talbot inventions;
- presenting photographic emulsions on metal plates, paper and glass;
- introducing the invention of roll film and a charge-coupled device (CCD) for digital photography;
- explaining the rise of electrophotography.

Optical Photography

Photography is the science and practice of creating durable images by recording light or other electromagnetic radiation, either chemically by means of a light-sensitive material such as photographic film, or electronically by means of an image sensor.

The world's first true photograph was made in 1826 in France by the inventor Joseph Nicéphore Niépce. The photograph was produced on a polished pewter plate (Figure 15.1). The light-sensitive material was a thin coating of bitumen, a naturally occurring petroleum tar, dissolved in white petroleum, applied to the surface of the plate and allowed to set before use. After a very long exposure in the *camera obscura* (traditionally said to be eight hours, but possibly several days), bitumen was sufficiently hardened in proportion to its exposure to light that the unhardened part could be removed with a solvent thus leaving a positive image with the light regions represented by hardened bitumen and the dark regions – by bare pewter. To see

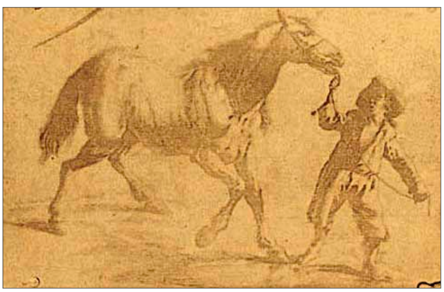

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Figure 15.1. The earliest known surviving heliographic engraving made by Nicéphore Niépce in 1825

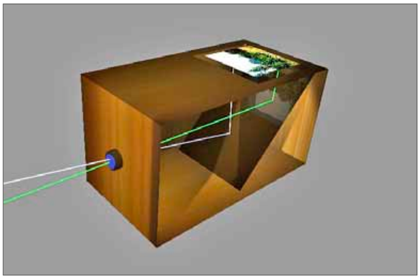

Figure 15.2. The Camera obscura

the image plainly, the plate had to be lit and viewed in such a way that bare metal appeared dark and bitumen – relatively light. The bitumen process was substituted by more sensitive resin and a very different post-exposure treatment that yielded higher-quality and more easily viewed images. Exposure times in the camera, although somewhat reduced, were still measured in hours. The *camera obscura* is an optical device that projects an image of its surroundings on a screen (Figure 15.2). The device consists of a box or room with a hole in one side. Light from an external scene passes through the hole and strikes a surface inside where it is reproduced, upside-down, but with the colour and perspective preserved.

In 1833, Daguerre experimented with photographing camera images directly onto a silver-surfaced plate fumed with iodine vapour that reacted with silver to form a coating of silver iodide. Exposure times were still impractically long. Then, by accident, Daguerre made the discovery that an invisibly latent image could be "developed" to full visibility by mercury fumes. This brought the required exposure time down to a few minutes. A hot solution of common salt served to stabilize or fix the image by removing the remaining silver iodide (Figure 15.3). William Talbot succeeded in creating stabilized photographic negatives on paper in 1835. In 1839, he acquired an effective fixer – the hyposulfite of soda that dissolved silver salts. This method replaced less effective hot salt water treatment.

Talbot's early experiments on silver chloride "sensitive paper" required camera exposures of an hour or more. In 1840, Talbot invented the *calotype* process that reduced exposure time to a few minutes. Paper with a coating of silver iodide was exposed in the camera and developed into a translucent negative image. A calotype negative could be used for making a large number of positive prints by simple contact printing. Later, George Eastman refined Talbot's process, which is the basic technology used by chemical film cameras today. Janez Puhar invented a process of making photographs on glass in 1841. In 1851, the *collodion* process was discovered and displaced calotype and daguerreotype processes as the most popu-

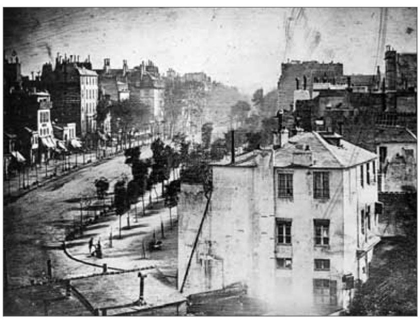

Figure 15.3. A daguerreotype made in 1838 is generally accepted as the earliest photograph of people

lar photographic medium until the dry-plate process was introduced in 1870. Collodion was made by dissolving guncotton, ether and alcohol thus blending it with a solution of silver and iron iodides. This was applied to a glass plate that was then immersed in silver nitrate solution and exposed wet in the camera. The plate had to be developed while collodion remained moist, giving the photographer only ten minutes after the exposure to develop the plate. This was physical development: silver was deposited to make the latent image visible and then fixed in sodium cyanide. The great advantages of the wet collodion process were that exposure required only two to three seconds. However, a photographer in the field had to carry more than 50 kg of equipment with him, including cameras, distilled water, chemicals, plates, trays and even a darkroom (Figure 15.4).

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Figure 15.4. A photographic van (1855)

In 1839, the first microphotograph was made and the remarkable reduction of 160:1 was attained. Microphotography is copying objects on a very reduced scale, in contrast to photomicrography, which is recording the enlarged image of an object under a microscope. During the Franco-Prussian War, when Paris was under siege, carrier pigeons were sent out in balloons and provided the only means of communication over enemy lines. However, a pigeon could only carry a tiny piece of paper. The system was devised and all letters were printed on large sheets of paper, each of which could hold about 100 typewritten pages today. These sheets were photographically reduced on to thin collodion films, each 50×70mm. 20 of those films could be rolled into a quill. Flown by pigeon to Paris, they were projected on to a large screen and transcribed by a small army of copyists. 100,000 dispatches were sent in this way, forty copies of each being made to ensure that at least one got through.

The inexpensive material became available only in the early 20th century – the 35 mm motion-picture film appeared. Special cameras and emulsions were developed for this purpose, and the microfilms of everything, from rare books to fragile newspapers and bank cheques, are now archived using this strong, long-lived medium. Photographic emulsions were supported by metal plates, later – by paper and glass.

In 1889, roll film and simple cameras preloaded with 100 exposures were invented. Amateurs would take ‘snapshots’ and send the camera and film to a developing company for developing and printing. The company *Kodak* helped photography with becoming a branch of industry with the mass market.

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Figure 15.5. Photograph of the first flight made by Wright brothers, 1903

Today, non-digital photographs are produced with a two-step chemical process where a light-sensitive film captures a negative image (colours and lights/darks are inverted). To produce a positive image, the negative is most commonly transferred onto photographic paper. Alternatively, the film is processed to invert the negative image yielding positive transparencies. Such positive images are usually mounted in frames, called slides.

Originally, all photographs were monochromatic or hand-painted in colour. Although methods for developing colour photos were available as early as 1861, they did not become widely available until the 1940s or 1950s, and even so, until the 1960s, most photographs were taken in black and white. Since then, colour photography has dominated popular photography, although black and white is still used, being easier to develop than colour. Panoramic format images can be taken with special cameras on a standard film.

Photography has become ubiquitous in recording events and data in science, engineering or accident scenes (Figure 15.5). The method has been much extended by using other wavelengths such as infrared photography and ultraviolet photography as well as spectroscopy.

The first color photo was made in 1861 by the Scottish physicist James Clerk Maxwell. Practical methods for sensitizing a silver halide film to green and then orange light were discovered in 1873 and 1884, but full sensitivity to red light was not achieved until the early years of the 20^{th} century. The first fully practical color plate *Autochrome* reached the market 1907 and was developed to the negative and reversed to the positive, which when viewed through the screen, restored the colours approximating the original. Other systems of color photography involved three separate monochrome exposures of a still scene through red, green and blue filters. These required a special machine to display, but the results are impressive even by modern standards.

The charge-coupled device (CCD) is the most important invention for digital photography. It was created in 1969 at AT&T Bell Labs. The essence of design was the ability to transfer charge along the surface of a semiconductor. In 1973, the first large image forming CCD chip consisting of 100 rows and 100 columns was released. In 1986, Kodak scientists developed the world's first megapixel sensor. The advent of a microcomputer and digital photography has led to the rise of digital prints created from stored graphic formats such as JPEG, etc. The types of the printers used include inkjet printers, dye-sublimation printer, laser printers and thermal printers.

Electrophotography

The photographic process had always involved the use of chemicals, which usually had to be in solution or vapour form; other imaging processes, such as television, were based on electronics and could not provide a fixed image. In 1938, the process named *xerography* was patented. It used dry powder that clung to the electrified parts of a semi-conductive plate. The first automatic system for making dry images on ordinary paper Xerox 914 was marketed.

The *xerography* process involves the following steps (Figure 15.6). *Charging*. An electrostatic charge of 600 volts is uniformly distributed over the surface. A corona discharge is generated by a narrow wire of 6,35 to 12,7 mm apart from a photoconductor. A negative charge is placed on the wire that ionizes the space between the wire and conductor so that electrons will be repelled and pushed away onto the conductor. *Exposure*. The document to be copied is illuminated by flash lamps such that its image is projected onto and synchronized with a moving drum surface. Where there is text or image on the document, the corresponding area of the drum will remain unlit. Where there is no image, the drum will be illuminated and the charge will be dissipated. The charge that remains on the drum after this exposure is a 'latent' image and is the negative of the original document. In a laser or LED printer, modulated light is projected onto the drum surface to create the latent image. *Development*. In copiers, the drum is presented with a slowly turbulent mixture of toner particles and larger iron reusable carrier particles. Toner is carbon powder later melt-mixed with a polymer. Carrier particles have a coating which, during agitation, generates a triboelectric charge (form of static electricity) attracting a coating of toner particles. By contact with the carrier, each neutral toner particle has an electric charge of polarity opposite to the charge of the latent image on the drum. The charge attracts toner to form a visible image on the drum. *Transfer*. Paper is passed between the drum and a transfer corona having the polarity opposite of the charge on toner. The toner image is transferred from the drum to the paper by a

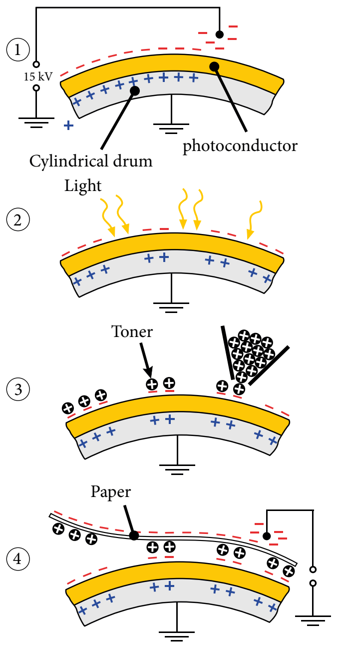

Figure 15.6. Schematic overview of the xerographic photocopying process: 1 – charging, 2 – exposure, 3 – development, 4 – transfer

combination of pressure and electrostatic attraction. *Separation*. Electric charges on paper are partially neutralized by AC from the second corona usually constructed in tandem with the transfer corona and immediately after it. As a result, the paper complete with most (but not all) of the toner image, is separated from the drum or belt surface. *Fixing or fusing*. The toner image is permanently fixed to paper using either a heat and pressure mechanism (Hot Roll Fuser) or a radiant fusing technology (Oven Fuser) to melt and bond toner particles into the medium (usually paper) being printed on. *Cleaning*. The drum, having already been partially discharged during detack, is further discharged by light. Any remaining toner is removed from the drum surface by a rotating brush or a cleaning blade.

For the period 1957–2003, the Scientific Research Institute for Electrography was operating in Vilnius. The Institute was engaged in the development of information systems, specific materials, medical software and electronic control and measurement devices. In 1956, engineer Jonas Žilevičius invented equipment that assisted in creating the first copiers which used plates coated with amorphous selenium and began systematic electrographic layers and imaging studies. In 1957, the Research Institute for Electrography was founded. The institution designed the first Soviet copier ERA produced by the Experimental Factory of Automation Tools in Kaunas and by Vilnius Factory of Computing Machinery.

CHAPTER 16

Counting, Calculating and Computing

This chapter will help in

- introducing early calculators (abacus, slide-rule);
- discussing the invention of logarithms;
- describing mechanical calculating machines (pin-wheel calculator, perforating machines);
- explaining the rise and development of electronic calculating machines (colossus computers using vacuum tubes, mainframes, microchips);
- presenting the main components of the computer: an arithmetic logic unit (ALU), control unit, memory, input and output devices;
- adding principal computer varieties: random-access memory (RAM) and read-only memory (ROM);
- looking at computer development in Lithuania.

Early Calculators

The first mechanical calculator (indeed, the world's first digital computer) was the *abacus*, the exact origins of which are lost in prehistory. An abacus is a small wooden frame having fastened wires each of which represents a numerical place value (units, tens, hundreds, etc.). The frame is strung with a number of beads representing a digit corresponding to the value of that place. By manipulating the beads, addition and subtraction may be performed. The abacus is a calculation tool used primarily in parts of Asia for performing arithmetic processes (Figure 16.1).

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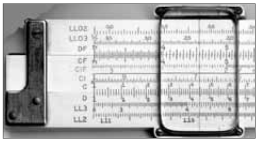
Figure 16.1. Abacus

a)



b)

c)

Figure 16.2. Logarithmic rulers: a – the binary liner slide rule; b – circular; c – cylindrical

About 1600, the Scottish mathematician John Napier invented logarithms. The beauty of this calculation system was that the problems of multiplication were reduced to those of addition, and division reduced to subtraction. In 1624, the first tables of logarithms to the base ten (decimal logarithms) were published thus becoming the basis of all surveying, astronomy and navigation up to these days.

In 1622, the English mathematician William Oughtred invented a slide-rule, an analogue calculator based on logarithmic scales. The slide-rule was much quicker than using tables, but the accuracy of practical-sized instruments was limited to three or four figures. The slide rule is a mechanical analog computer used primarily for multiplication and division as well as for performing functions such as roots, logarithms and trigonometry, but not normally applied for addition or subtraction. Slide rules generally appear in linear or circular form with a standardized set of markings (scales) essential to performing mathematical computations. The slide rule was developed in the 17th century and is based on the mathematical principle of logarithms developed by John Napier in 1614. Before the advent of the pocket calculator, it was the most commonly employed calculation tool in science and engineering. It was in constant use for a period of almost 300 years, from the time of Newton, through the industrial revolution, and on into the space age. The application of slide rules continued to grow through the 1950s and 1960s even when digital computing devices were being gradually introduced. Around 1974, the electronic scientific calculator made it largely obsolete and most suppliers left the business.

There are the three most common types of slide rules (Figure 16.2):
- traditional engineer's slide rule with straight-line scales;
- circular slide rule;
- cylindrical slide rule having one or more scales arranged in a helix around a sliding cylinder so as to provide a much longer scale with greater resolution and accuracy.

Mechanical Calculating Machines

The *pin-wheel calculator* was a simple and reliable machine capable of all four arithmetic functions. Addition and subtraction were performed directly while multiplication and division were accomplished by simple procedures involving repeated additions (or subtractions) and carriage shifts.

W. Odhner's machine was developed in St. Petersburg from about 1874. Odhner's design had greater commercial success, and it was his name that became more generally associated with the pinwheel mechanism. By about 1890, Odhner's calculator developed into the form that was to endure more or less unchanged for the next eighty years. Pin-wheel machines were in continuous production for almost one hundred years, from their beginning in the 1870s until overtaken by four-function electronic calculators in the 1970s (Figure 16.3).

Figure 16.3. Odhner mechanical calculator

The first calculating machine appeared in 1642. There was a wheel for units, one for tens, one for hundreds, etc.; gear wheels were used for carrying. In 1671, a calculating machine that could multiply numbers was created. In 1805, perforated cards which, linked together, were used for controlling weaving patterns were introduced in France. This system was the direct ancestor of punched cards, the key element of tabulating machines that were the principal product of IBM during the first half of the 20th century and the main form of external storage for digital computers until the 1960s.

Electronic Calculating Machines

In 1847, the English mathematician George Boole published *The Mathematical Analysis of Logic* showing that logic could be expressed using binary symbols. Boolean algebra has only two symbols, 0 and 1; however, instead of using them as binary digits for arithmetic computation, they are interpreted as the outcomes of logical operations, 0 denoting 'false' and 1 standing for 'true'. In 1938, the Bell Telephone Laboratories realized that the Boolean 'true' and 'false' could be represented by the '*on*' and '*off*' states of bistable electronic components, and this became the basis of computer logic design.

During the first half of the 20th century, many scientific computing needs were met by increasingly sophisticated analog computers that used a direct mechanical or electrical model of the problem as a basis for computation. However, these were not programmable and generally lacked the versatility and accuracy of modern digital computers. The non-programmable Atanasoff-Berry Computer (completed in 1941) used vacuum tube based computation.

The first programmable electronic computer of limited programmability was *Colossus* built in 1943 (UK). It used 1500 valves or vacuum tubes and was employed for breaking German wartime codes. John von Neumann, a Hungarian immigrant, became a consultant on the atomic bomb project at Los Alamos (USA). A report written by von Neumann with Goldstein and Arthur W.Burks of the University of Michigan first described the design of an electronic stored-program digital computer. In 1948, John Bardeen, Walter Brattain and William Shockley of the Bell Telephone Laboratories published the results of their work on solid-state electronic devices. This paper described the invention of the point-contact transistor that almost completely displaced the valve as the active component in electronic circuits. In 1949, magnetic-core storage that became the standard internal memory for large-scale digital computers until semiconductor memory was introduced in the mid-1960s was presented.

During the 1950s, computers, known as *mainframes*, were introduced. They cost from hundreds of thousands to millions of dollars, required special housing and air conditioning and demanded sizeable staff. The users never got near the actual hardware. Punched cards or paper tape were the primary input media. Users entered their programs and data by keypunching. Hours, sometimes days, later users would get their output, usually in the form of paper printouts. Frequently, finding some slight mistake, they would have to resubmit the whole job again. The earliest computers had to be programmed in the machine language, which meant that both instructions and data had to be input in binary form. Many users wanted to be able to specify their instructions to the computer in a language closer to algebra, and therefore, in 1954, IBM published the first version of the FORTRAN language. By the mid-1960s, 90 per cent of all scientific and engineering programs were written in FORTRAN and most business data-processing – in COBOL. From its beginning in 1964, BASIC was designed for interactive use via 'time-sharing' on minicomputers. BASIC has become the standard higher level language of microcomputers.

The development of the integrated circuit (IC), popularly known as a *microchip*, enables digital-circuit designers to put essential mainframe functions into one or more tiny chips of silicon. In 1962, the first MOS (metal-oxide semiconductor) chip was made. The first electronic calculator with MOS chips was introduced in 1963 by the Bell Punch Company in the U.K. CMOS (*Complementary Metal-Oxide Semiconductor*) chips appeared in 1968. They were more expensive but faster and draw less current than MOS chips. Later, CMOS chips have made portable battery-operated microcomputers possible.

The idea of creating a simple, low-cost computer from MOS chips, a microcomputer-on-a-chip, seems to have occurred first to enthusiastic amateurs. In 1973, Scelbi Computer Consulting company offered the minicomputer as a kit. It used the Intel 8008, the first 8-bit microprocessor and had 1 kilobyte of memory. Microcomputers remained largely in an amateur market until the Apple II and Commodore PET came on to the market in. For a couple of thousand dollars, customers were offered an alphanumeric keyboard for data entry and control, a colour display, a beeper that could be programmed to play simple tunes, a pair of paddles to play video games, a high-level language (BASIC) in Read-only memory (ROM) and an inexpensive means of storing and retrieving data and programs (their audiocassette recorder). In 1981, Osborne 1, the first complete, portable microcomputer was introduced. It included a full-sized keyboard, cathode ray tube (CRT) display, dual floppy drives and was 'bundled' with some of the best software available at the time. Only after few hour-instruction, a non-specialist could learn a user-friendly microcomputer software package directly relevant to his work. As a result of their convenience, versatility and low cost, micros have been already far more numerous than mainframes accounting more than half the revenue of the entire computer market in the last half of the 1980s. The rapid development of micro technology led to smaller and smaller system, and eventually battery-operated laptop portables (Figure 16.4). Modern laptop micros can communicate with remote systems by means of a modem that does not have to be physically connected to the public switched telephone network or use Wi-Fi technologies (Wi-Fi is a popular technology that allows an electronic device to exchange data wirelessly using radio waves) over a computer network, including high-speed Internet connections.)

A number of projects to develop computers based on the *stored-program architecture* commenced around that time, the first of which was completed in 1948. Computers using vacuum tubes as their electronic elements were in use throughout

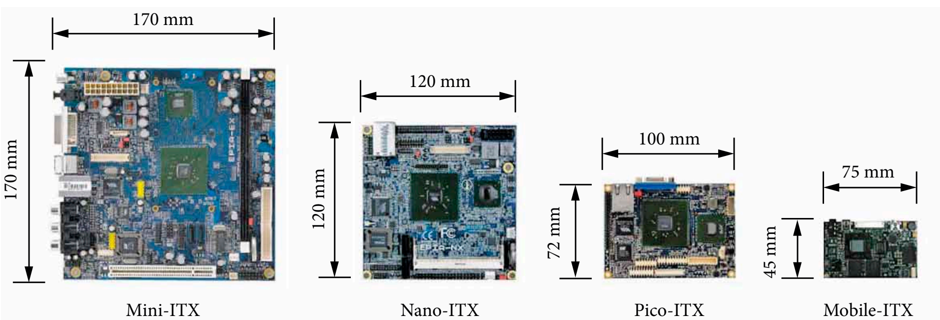


Figure 16.4. Motherboard form factor comparison

the 1950s; however, by the 1960s, they had been largely replaced by transistor-based machines that were smaller, faster, cheaper to produce, required less power and were more reliable. The first transistorised computer was demonstrated in 1953. In the 1970s, integrated circuit technology and the subsequent creation of microprocessors, such as the Intel 4004, further decreased the size and cost as well as increased the speed and reliability of computers. By the late 1970s, many products such as video recorders contained dedicated computers called microcontrollers. They started to appear as a replacement of mechanical controls in domestic appliances such as washing machines. The 1980s witnessed home computers and the now ubiquitous personal computer.

A general purpose computer has four main components: the *arithmetic logic unit* (ALU), *control unit*, *memory* and *input and output devices*. These parts are interconnected and each of them inside contain thousands to trillions of small electrical circuits that can be turned off or on by means of an electronic switch. Each circuit represents a bit (binary digit) of information so that when the circuit is on it represents a "1" and when off it represents a "0" (in positive logic representation). The circuits are arranged in logic gates so that one or more of those may control the state of one or more of other circuits.

The control unit, ALU, registers and basic input and output devices are collectively known as the *central processing unit* (CPU). The key component common to all CPUs is the program counter – a special memory cell (register) that keeps the track of which location in memory the next instruction is to be read from. The ALU is capable of performing two classes of operations: arithmetic and logic.

The computer's memory can be viewed as a list of cells into which numbers can be placed or read. Each cell has a numbered "address" and can store a single number. The computer can be instructed to "put the number 23 into the cell numbered 135" or to "add the number that is in cell 135 to the number that is in cell 246 and put the answer into cell 159". The information stored in memory may represent practically anything. Letters, numbers and even computer instructions can be placed into memory with equal ease. Since the CPU does not differentiate between different types of information, it is the responsibility of software to give significance to what the memory sees as nothing but a series of numbers.

In almost all modern computers, each memory cell is set up to store binary numbers in groups of eight bits (called a byte). Each byte is able to represent 256 different numbers ($2^8 = 256$): either from 0 to 255 or from −128 to +127. To store larger numbers, several consecutive bytes (typically, two, four or eight) may be used. A computer can store any kind of information in memory if it can be represented numerically. Modern computers have billions or even trillions of bytes of memory.

The main memory of the computer comes in two principal varieties: *random-access memory* or RAM and *read-only memory* or ROM. RAM can be read and written to anytime the CPU commands it, but ROM is pre-loaded with data and software that never changes; therefore, the CPU can only read from it. ROM is typically used for storing the initial start-up instructions of the computer. In general, the contents of RAM are erased when the power to the computer is turned off, but ROM retains its data indefinitely. In a PC, ROM contains a specialized program called BIOS that orchestrates loading the operating system of the computer from the hard disk drive into RAM whenever the computer is turned on or reset. Software stored in ROM is often called firmware, because it is notionally more like hardware than software.

The devices that provide input or output to the computer are called *peripherals*. On a typical personal computer, peripherals include input devices like the keyboard and mouse and output devices such as the display and printer. Hard disk drives, floppy disk drives and optical disc drives serve as both input and output devices.

Historically, computers evolved from mechanical computers and eventually from vacuum tubes to transistors. There active research on making computers out of many promising new types of technology such as optical computers, DNA computers, neural computers and quantum computers is taking place. Most computers are universal and able to calculate any computable function as well as are limited only by their memory capacity and operating speed.

Computer Development in Lithuania

The idea of computer applications first rose in Lithuania in 1939. However, the outbreak of World War II and the subsequent Soviet occupation postponed the realization of this idea for two decades. Cybernetics, until the middle of 1950s, was claimed as "pseudoscience." In 1956, computers became legal, and research into theoretical physics became a high priority. Computer industry began in Lithuania in the late 1950s. The Vilnius Computer Factory (VCF) and the Special Computer Design Bureau (SCDB) started functioning in Vilnius in 1959. The first computer used for scientific research purposes in Lithuania was the BESM-2M installed in the Lithuanian Academy of Sciences in 1962. It was the most powerful Soviet computer at that time. The computer had the following characteristics: 2,047 words of 39 bits of RAM, two magnetic drums of 12,290 words and four drives of magnetic tapes (each of 131,000 words). The computer also consisted of 3,000 electronic tubes and 10,000 diodes and consumed 43 kW of electric power. Only 50% of computer time was productive; the other half was spent performing preventive maintenance or repairs. The programmers used the Algol language. The BESM-6 computer system, installed in 1979, operated 24 hours per day, had a RAM of 192,000 words of

50 bits, 24 drives of exchangeable drums of 512,000 words and 20 magnetic tape drives. Apart from usual I/O devices, the system had digital-to-analog and analog-to-digital converters. Magnetic tapes were used for data exchange among computers. When IBM PCs appeared in 1987, they very soon dominated all applications. In comparison with PCs, old-fashioned mainframes were slow, expensive and not user-friendly.

Computer industry. Lithuanian computer industry passed all typical stages of design, technology and production. In 1959, manufacturing 80-column punched-card peripheral devices, including card punches, card readers and reproducers started (Figure 16.5). Card sorters were constructed starting in 1961.

The production of the first electronic calculator began in Vilnius in 1960 (Figure 16.6). Designers created the first Lithuanian original computer in 1962. The process of building it continued from 1964 to 1967. This specialized computer was designed for a statistical analysis of stochastic processes. In 1963, engineers designed an original second- generation computer with ferro-transistor elements. Manufacturing this computer continued from 1964 to 1974. The new computer was oriented to solving management problems. The production of computer Ruta 110 designed by local specialists was launched in 1969. The computer was meant for processing large data arrays in automated management systems. Ruta 701 was the first industrial hand-printed optical character reading machine in Europe. Since 1969, the main goal of producers was designing and producing third-generation minicomputers that operated at the rate of 14,000-33,000 instructions per second and had 32-64 kb of RAM. The computer design bureau in Vilnius designed and produced minicomputers CM-1600. Work stations contained two processors and served to professional work with computer network resources and to solving the problems of management and statistics. The computer was manufactured for the

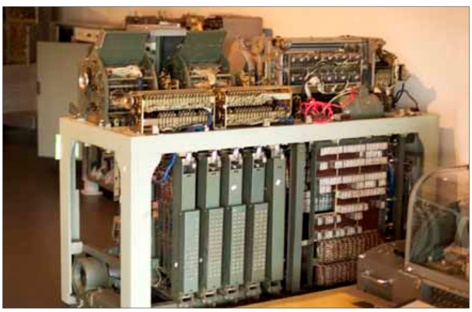


Figure 16.5. Punched-card peripheral device. Museum of Energy and Technology, Vilnius

Figure 16.6. The electronic calculator (1960). Museum of Energy and Technology, Vilnius

period 1982-1988. The machine was capable of 30,000 instructions per second and had a memory of 256 kB expandable to four MB. For the 1986-1990, DEC and VAX 730 clones were produced and used for engineering calculations as well as applied to workstations.

At the end of the 1980s, Lithuanian computer producers united seven plants and several design bureaus. About 18000 people worked in production and management areas, and 2000 were employed in the research and development area. After 1990, the situation in computer industry changed drastically. Local industry was disintegrated or dramatically reduced. About 200 small and medium size companies sell computers and software, prepare computerization projects, propose services and train computer users in Lithuania. At present, the Intel-based PCs are dominating in the market.

CHAPTER 17 | Air and Space Transport

This chapter will help in

- discussing the development of dirigibles – lighter-than-air aircrafts;
- looking at the rise of early helicopters;
- describing early jet aircrafts;
- reviewing turbojet and turbofan engines;
- revising the development of human-powered aircrafts;
- introducing aerobatics and sport aviation;
- representing gliding activities and Lithuanian gliding achievements;
- dealing with unmanned aerial vehicles (drones);
- presenting the early Lithuanian flights (Dobkevičius, Gustaitis, Darius and Girėnas, Vaitkus).

Dirigibles

A dirigible is a type of an aerostat or a lighter-than-air aircraft that can be steered and propelled through the air using rudders and propellers or other thrust mechanisms. Unlike an aerodynamic aircraft such as a fixed-wing aircraft and helicopters that produce lifting by moving a wing through the air, an aerostatic aircraft stay aloft by having a large "envelope" filled with gas that is less dense than the surrounding atmosphere. The first lifting gas used was hydrogen, although this had well-known concerns over its flammability. Helium was rare in most parts of the world, but large amounts were discovered in the USA. This meant that non-flammable gas was rarely used for airships outside the USA. All modern airships, since the 1960s, have used helium. The term *zeppelin* is a trademark that originally referred to the airships manufactured by the German Zeppelin Company that pioneered dirigible design in the early years of the 20^{th} century.

A rigid airship is, in a structural sense, a large, trussed frame covered with fabric. A cross-section through a zeppelin readily shows this. Explosive hydrogen gas, while not limited to causing the Hindenburg crash, was less likely to bring down an airship than structural failure. The design and construction of rigid airships was adequate for static conditions, but the sky is hardly a static environment. Airship frames were stressed in ways their designers could hardly imagine. In addition to bending due to differential loads at various points along their great lengths, there was torque from twisting motions as well as changes in pressure alternately expanding or deflating large gas reservoirs throughout the structure. Structural design that could deal with such a variety of forces had hardly been conceived until rigid airships were born.

The Golden Age of Airships began in 1900 with the launch of the Luftschiff Zeppelin LZ1. Zeppelins were huge – the LZ1 was 128 m long while the last – Graf Zeppelin – was 236 m long. At the beginning of World War I, Zeppelin airships had a framework composed of triangular lattice girders covered with fabric and containing separate gas cells. Multi-plane, later cruciform, tail surfaces were used for control and stability, and two engine/crew cars hung beneath the hull driving propellers attached to the sides of the frame by means of long drive shafts. Additionally, there was a passenger compartment (later a bomb bay) located halfway between the two cars. 1937 started with the most spectacular and widely remembered airship accident. Approaching the mooring mast minutes before landing in New Jersey (USA), on 6 May 1937, the Hindenburg burst into flames and crashed (Figure 17.1). Of 97 people aboard, 36 died (13 passengers and 22 air crew). The Hindenburg disaster shattered public confidence in airships and brought a definitive end to the "golden age".

The prospect of airships as bombers had been recognized in Europe well before they were up to the task. The Italian forces became the first to use dirigibles for a military purpose during the Italo-Turkish War. It was World War I, however, that marked the real debut of the airship as a weapon. Germans, French and Italians all used airships for scouting and tactical bombing roles early in the war, and all learned that the airship was too vulnerable for operations over the front. The decision on ending operations in direct support of armies was made by all in 1917.

The German military believed they had found the ideal weapon for counteracting British naval superiority and striking at Britain itself. Raids on England began in 1915 and were discontinued in 1918. Zeppelins proved to be terrifying but inaccurate weapons. Navigation, target selection and bomb-aiming proved to be difficult under the best of conditions. Darkness, high altitudes and clouds reduced accuracy even further. The physical

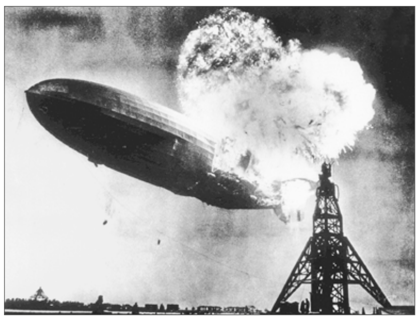


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Figure 17.1. The Hindenburg – moments after catching fire on 6 May 1937

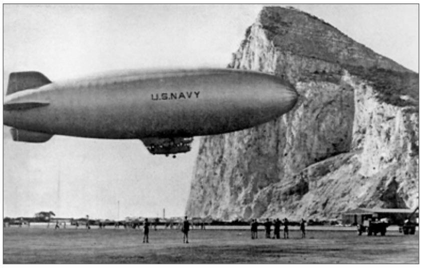

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Figure 17.2. K-class blimps of conducted antisubmarine warfare operations at the Strait of Gibraltar in 1944

damage done by the Zeppelins over the course of the war was trivial, and the deaths that they caused amounted to a few hundred at most. The Zeppelins were initially immune to attack by aircraft and antiaircraft guns: as pressure in their envelopes was only just higher than ambient air, holes had little effect. But once incendiary bullets were developed and used against them, their flammable hydrogen lifting gas made them vulnerable at lower altitudes. Several were shot down in flames. They then started flying higher and higher above the range of other aircrafts, but this made their bombing accuracy even worse and success harder to achieve. Countermeasures by the British included sound detection equipment, searchlights and anti-aircraft artillery. Before the World War, the military was interested in blimps for scouting purposes. British blimps were used for scouting, mine clearance and submarine attack duties (Figure 17.2). Both France and Italy continued airships throughout the war. Aeroplanes had essentially replaced airships as bombers by the end of the war, and the remaining zeppelins were scrapped or handed over to the Allied powers as spoils of war.

The development of airships continued only in the United States, and to a smaller extent, in the Soviet Union that had several semi-rigid and non-rigid airships. The semi-rigid SSSR-V6 was among the largest of these crafts and set the longest endurance flight at the time of over 130 hours. However, it crashed into a mountain in 1938 thus killing 13 of the 19 people on board. While this was a severe blow towards the Russian airship program, they continued to operate non-rigid airships until 1950.

Although airships are no longer used for passenger transport, they are still applied for other purposes such as advertising, sightseeing, surveillance and research.

Helicopters

The first true flight was apparently made by Paul Cornu in 1907 (Figure 17.3) the machine of whose, powered by a small 24 hp engine, could only have been called the “flying bicycle” and consisted as it did of two large spoked wheels on to which short, paddle-shaped wings were splinted

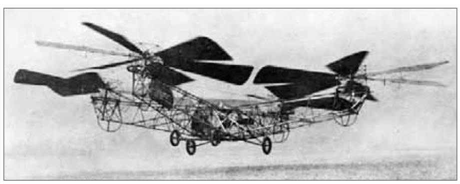

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Figure 17.3. Cornu helicopter (1907)

to form twin two-blade rotors about 6 m in diameter. The rotors were belt-driven and contra-rotating. The central frame supported the engine, pilot seat and fuel tank, and the whole contraption weighed just over 250 kg.

Jet Aircrafts

The early jet. The propeller-driven (pulled) piston engine aircraft reached its apogee in the Second World War. Such designs flew higher and faster but clearly were reaching a limit that had yet to be named. It would only be named "the Sound Barrier" after the development of a revolutionary aircraft utilizing the new means of propulsion. The rocket and jet aircraft of World War II led to the end of the first phase of aeronautical development, subsonic flight and ushered in the next era of supersonic flight. As the propeller-driven aircraft flew faster, they encountered the effects of compressibility that acted on lifting and dragging the characteristics of the airfoil. An increase in the Mach number (speed of sound = 1.0 Mach number) from 0.4 to 0.8 sees an incremental lift coefficient value of only 0.2 together with a large increase in the order of magnitude in drag. The aircraft approaching sound barrier experienced instability and, many times, loss of control.

The Germans flew the first successful jet aircraft in 1939, the Heinkel 178 (Figure 17.4) that was the world's first aircraft to fly under turbojet power and the first practical jet plane. Later, the Me-262, first flown in 1942 and later to become the first operational jet fighter was produced. Powered, by today's standards, feeble axial flow Jumo engines Me-262 could reach 840 kph (Figure 17.5). The technology advanced engine and airframe had primitive armament, fuel and cockpit systems. Fuel had two components: first – a mixture of alcohol, hydrazine hydrate and water, second – hydrogen peroxide. This mixture was highly explosive, corrosive and terribly dangerous to man and machine. Combined with archaic landing skid, a loaded plane could not land without the great peril of a spontaneous explosion. The Messerschmitt Me 163 Komet was a rocket-powered fighter aircraft and the

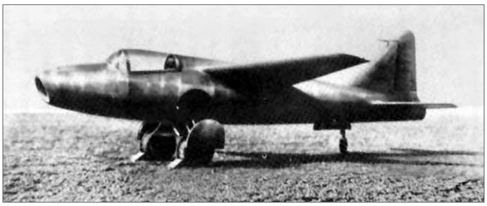

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Figure 17.4. The Heinkel He 178 turbojet plane

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Figure 17.5. The Messerschmitt Me-262A

only that ever to have been operational. The Me 163 in 1944 reached 1,123 km/h. Over 300 aircrafts were built; nevertheless, the aircraft proved ineffective as a fighter.

A jet aircraft is an aircraft propelled by jet engines. The jet aircraft generally flies much faster than a propeller-powered aircraft, and therefore may reach higher altitudes making as high as 10,000–15,000 m. At these altitudes, jet engines achieve maximum efficiency over long distances. The engines in the propeller-powered aircraft achieve their maximum efficiency at much lower altitudes. Some jet aircrafts can move faster than sound. The first turbojet aircraft to fly was the Heinkel He 178, a prototype of the German Air Force in 1939. The first operational jet fighter was the Messerschmitt Me 262 made by Germany during late World War II (Figure 17.5). It was also the fastest conventional aircraft of the Second World War. Although it had first flight in 1941, mass production started in 1944 with the first squadrons operational that year.

In 1944, the UK jet-powered aircraft Gloster Meteor was being committed to the defence of the UK and then ground-attack operations over Europe in the last months of the war. In 1944, Germany introduced into service the Arado Ar 234 jet reconnaissance and bomber plane. Japan developed the jet aircraft in 1945. By the end of 1945, U.S. had introduced their next jet fighter, the Lockheed P-80 Shooting Star into service and the UK – its second fighter design, the de Havilland Vampire. The USA introduced the North American B-45 Tornado, their first jet bomber, into service in 1948, which was capable of carrying nuclear weapons. The first commercial jet service from London to Johannesburg was provided in 1952 with the de Havilland Comet jetliner. The Comet was a four-engine jet aircraft. Small by today's standards, the first Comets carried only 36 passengers. The Comet's flight was its thin aluminum skin 0.028 inches thick. This is about the thickness of a postcard. Further, it used large, square windows that distributed stresses poorly compared to oval or round windows. These elements combined with pressurization needed to fly at higher altitudes. Two aircrafts were lost in 1954. The cause of the crashes was determined to be an explosive decompression of the fuselage after metal failure due to fatigue. The impact of engineering tests on the Comet led to changes in the aircraft design. In U.S., Boeing and Douglas added thicker fuselage skins, triple-strength rounded windows, metal bracing for additional structural strength and, perhaps most importantly, small metal tabs or "stoppers" were placed along the interior of the fuselage to block crack formation.

A *turbofan aircraft* began entering service in the 1950s and 1960s, and this is the most common type of jet in use today. Tu-144 supersonic transport was the fastest commercial jet plane at Mach 2.35 (2,503 km/h). It went into service in 1975 but soon stopped flying. The Mach 2 Concorde aircraft entered service in 1976 and flew for 27 years. Concorde flew regular transatlantic flights from London Heathrow and Paris-Charles de Gaulle Airport to New York JFK profitably flying these routes at record speeds in less than half the time of other airliners. Only 20 aircraft were built (Figure 17.6).

The fastest military jet plane was the SR-71 Blackbird at Mach 3.35 (3,661 km/h). The fastest manned (rocket) aircraft is the X-15 at Mach 6.85 (7,485 km/h).

Jet engines come in several main types: turbojet, turbofan (comes in two main forms – low bypass turbofan and high bypass turbofan) and rocket. These types of the engines are applied to different aircrafts. Turbojets are seldom used but were employed on Concorde. The engine has a high exhaust speed and a low frontal cross-section; thus, it is best suited to a high-speed flight. Low bypass turbofans have a lower exhaust speed than turbojets and are used for transonic and low supersonic speeds. High bypass turbofans are used for a subsonic aircraft, are quite efficient, and therefore widely employed in airliners. Rockets have extremely fast exhaust speeds and are mainly used when high speeds or extremely high altitudes are needed.

The Boeing 787 Dreamliner is a long-range, mid-size wide-body, twin-engine jet airliner developed by Boeing Commercial Airplanes (Figure 17.7). The program was launched in 2004. The plain is designed from 210 to 290 passengers. Boeing states it is the company's most fuel-efficient airliner and the world's first major airliner to use composite materials as the primary material in the construction of its airframe. The 787's design features light-weight construction. The aircraft is 80% composite by volume. Although design changes increased the share of titanium, Boeing actually list its materials by weight to 50% composite, 20% aluminum, 15% titanium, 10% steel and 5% other. Aluminum is used for the wing and tail leading edges titanium – mainly on engines and fasteners, steel – in various areas. Each plain contains approximately 32 tons of *carbon fiber reinforced plastic* (CFRP) made with 23 tons of carbon fiber. The longest-range plain variant can fly 15,000 to 15,700 km, enough to cover Los Angeles to Bangkok or New York City to Hong Kong routes. Cruising airspeed is 900 km/h at typical cruise altitudes.

Figure 17.6. A Concorde aircraft

Figure 17.7. The Boeing 787 Dreamliner

Figure 17.8. A380 in original Airbus livery

The Airbus A380 is a double-deck, wide-body, four-engine jet airliner manufactured by the European corporation Airbus (Figure 17.8). The A380 provides seating to 853 people in all-economy class configurations. The A380-800 has a design range of 15,400 km sufficient to fly from New York to Hong Kong and a cruising speed of Mach 0.85 (about 900 km/h at cruising altitude). The major structural sections of the A380 are built in France, Germany, Spain and the United Kingdom. After assembly (in Toulouse, France), the aircrafts are flown to Hamburg to be furnished and painted. It takes 3,600 litres of paint to cover the 3,100 m^2 exterior of a plane.

By the late 20^{th} century, the development of increasingly more powerful and smaller integrated circuits and advancements in digital technology has eliminated the necessity of *Flight Engineers* on modern airliners. On new generation two-man deck airplanes, sensors and computers monitor and adjust systems automatically. There is no onboard technical expert and the third pair of eyes. If a malfunction, abnormality or emergency occurs, it will be displayed on an electronic display panel and the computer will automatically initiate corrective action to rectify the abnormal condition. One pilot (PF) does the flying, and the other pilot (PNF) will resolve the issue.

Early Lithuanians Flights

The first aircraft with Lithuanian signs flew on 1 March 1919. Military Aviation took part in the battles for independence against Bolshevism and Poland in 1919–1920. The aircrafts of some foreign airlines landed in Kaunas since 1922. Regular inner passenger service was initiated in 1939. Jurgis Dobkevičius built and tried the first original Lithuanian aircraft Dobi-I in 1922. In 1923–1924, he made and tested two more aircrafts. Antanas Gustaitis designed, produced and examined 9 aircrafts of ANBO types in flight for the period 1925-1939. Some of them were built in series. When the Soviet Union annexed all three Baltic States, including Lithuania, Latvia and Estonia in 1940, the inventories of the armed forces of those nations also changed the owner. The designer of the ANBO aircraft Lt.Col. later, Antanas Gustaitis was shot in Moscow. The first of his designs, the ANBO-I, appeared in 1928. It was a single-seat low-wing sports monoplane. This was followed by the first military type ANBO-II, a two-seat primary trainer with a parasol-mounted wing and only modest performance. The next model was the ANBO-III, an advanced trainer that differed from its predecessor externally by its undercarriage and redesigned vertical tail surfaces. The production of the ANBO-II and ANBO-III is supposed to have been 20 machines of each type. In 1932, the ANBO-IV passed its maiden flight. This type soon earned the reputation to be a very reliable aircraft. The formation of ANBO-IVs in 1934 had a tour to different European capitals (Stockholm, Brussels, Copenhagen, London, Paris, Rome, Vienna, Prague, Bucharest, Moscow, etc.). After the tour, the ANBO-IV became relatively well-known outside Lithuania. An improved version with a more powerful engine was the ANBO-41 (about 20 built) that was also equipped with two Lithuanian squadrons. In total, 66 ANBO aircrafts (Figure 17.9) were built.

Transatlantic flights. Steponas Darius and Stasys Girėnas were Lithuanian pilots, emigrants to the United States, who made a significant flight in the history of world aviation. On 15 July 1933, they flew across the Atlantic Ocean, thus covering a distance of 6,411 km without landing, in 37 hours and 11 minutes (172 km/h). In terms of comparison, as far as the distance of non-stop flights was concerned, their result ranked fourth in terms of the duration of flight at the time. Although Darius and Girenas did not have navigational equipment and flew under unfavourable weather conditions, the flight was one of the most precise in aviation history. The aircraft was a single-engine, six-seat, high-wing monoplane. The fuselage was welded chrome-moly steel tubing covered with fabric. The engine was 365 hp (272 kW) Wright Whirlwind J6-9E, radial, air cooled and had 9 cylinders. They were first who officially carried air mail from North America to Europe. After taking off from Floyd Bennett Field in New York, *Lituanica* successfully crossed the Atlantic. The planned route was New York–Newfoundland–Atlantic Ocean–Ireland–London–Amsterdam–Swinemünde–Königsberg–Kaunas (a total of 7,186 km).

With permission from the website *Lietuvos Aviacijos Istorija 1919–1940 m.*

Figure 17.9. The ANBO-VIII aircraft built in 1939

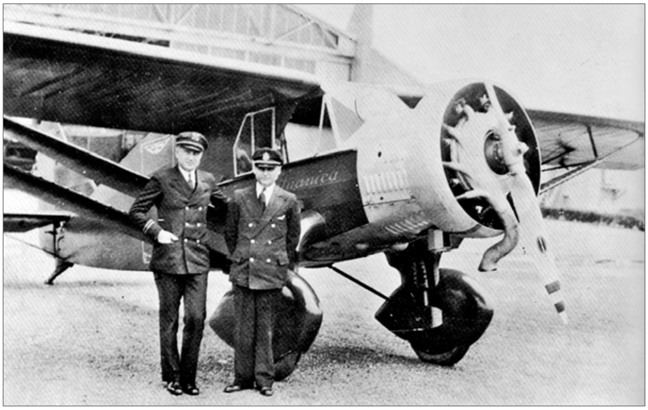

Figure 17.10. Lithuanian pilots Darius and Girenas beside the Bellanca CH-300 Pacemaker aircraft

Figure 17.11. Felix Vaitkus before the transatlantic flight

Due to weather conditions over Ireland, they veered to the north and reached Germany via Scotland and the North Sea. In 37 hours and 11 minutes, until the moment of the crash, they had flown 6411 km (over 7000 km in actual flight path), only 650 km short of their goal – Kaunas (Figure 17.10).

The next transatlantic flight was accomplished by the Lithuanian pilot (US citizen) Felix Vaitkus on *Lituanica II* (Lockheed Vega5B). The aircraft arrived from New York to Kaunas on 21-22 September 1935. He landed in Ireland following 22 h 15 min of flying and covered 5100 km (Figure 17.11).

Human-powered Aircrafts

A human-powered aircraft (HPA) is an aircraft belonging to *human-powered vehicles*. The aircraft is powered by direct human energy and the force of gravity. The thrust provided by human may be the only source. Likewise, the HPA inevitably experiences assistance from thermals or rising air currents. Pure HPAs do not use hybrid flows of energy (solar energy, wound rubber band, fuel cell, etc.) for the thrust. Under nil wind, a flatland-long-gliding aircraft is a form of the HPA where the thrust in nil wind is provided by the running of the pilot. When the pilot loses touch with the ground, his/her thrust ceases to add energy to the flight system and a glide begins.

In 1959, the industrialist Henry Kremer offered the first Kremer Prizes of £5,000 for the first human-powered aircraft to fly a figure-of-eight course round two markers half-a-mile apart. The first officially authenticated take-off and landing of a man powered aircraft SUMPAC was made in Southampton University in 1961 (SUMPAC).

The best flight out of 40 tried was 650 meters. In 1972, a man powered aircraft flew 1,239 metres. In 1977, the *Gossamer Condor 2* flew the first figure-eight, a distance of 2,172 metres winning the first Kremer prize. It was piloted by an amateur cyclist and hang-glider pilot Bryan Allen. Although slow, cruising at only 11 mph, it achieved that speed with only 0.35 hp. The second Kremer prize of £100,000 was won again by Paul MacCready on 12 June 1979 when Bryan Allen flew *Gossamer Albatross* from England to France. A straight distance of 35.82 km took a flight of 2 hours 49 minutes. The Gossamer Albatross was constructed using a carbon fiber frame with the ribs of the wings made with expanded polystyrene. The entire structure was then wrapped in thin, transparent plastic (*mylar aka PET film*). The empty mass of the structure was only 32 kg. The Albatross was powered using pedals to drive a large two-bladed propeller (Figure 17.12).

Figure 17.12. A human-powered aircraft Gossamer Albatross

The third Kremer prize of £20,000 for speed went in 1984 to the design team of the Massachusetts Institute of Technology (MIT) for flying the *Monarch-B* craft on a triangular 1.5 km course in under three minutes (for an average speed of 32 km/h). The current distance record making 115.11 km was achieved on 23 April 1988 by the MIT Daedalus aircraft (Figure 17.13).

The first human-powered helicopter flew for 7.1 seconds and reached a height of 20 cm in 1989 (Figure 17.14).

Figure 17.13. The *Daedalus* light human powered aircraft in flight (1988)

Figure 17.14. A human powered helicopter

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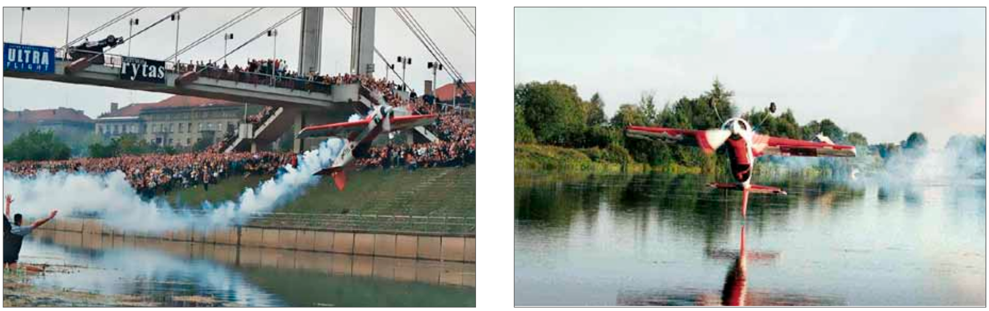
Figure 17.15. A human-powered ornithopter (1902)

An *ornithopter* is an aircraft that flies by flapping its wings. The first ornithopters capable of flight were constructed in France in 1871 Designer used a rubber band to power a small model bird. E. P. Frost made ornithopter powered by steam engine (1870) and in the 1900s powered by internal combustion engine (Figure 17.15). The current world record for the human-powered helicopter equals a height of 20 cm within 19.46 seconds. A human-powered ornithopter flew 145 meters with an average speed of 25.6 km/h (2010). The 32 m wingspan 42 kg aircraft was constructed from carbon fibre, balsa and foam. The pilot sat in a small cockpit suspended below the wings and pumped a bar with his feet to operate a system of wires that flapped the wings up and down in an elegant flapping motion. The aircraft was towed by a car until airborne.

Sport Aviation

Aerobatics is the practice of flying manoeuvres involving aircraft attitudes that are not used in normal flight. Aerobatics are performed in airplanes and gliders for training, recreation, entertainment and sport. Some helicopters are capable of limited aerobatic manoeuvres. Most aerobatic manoeuvres involve the rotation of the aircraft about its longitudinal (roll) axis or lateral (pitch) axis. Other manoeuvres, such as a spin, displace the aircraft about its vertical axis. Manoeuvres are often combined to form a complete aerobatic sequence for entertainment or competition. Aerobatic flying requires a broader set of piloting skills and exposes the aircraft to greater structural stress than for a normal flight (Figure 17.16).

a) b)



With permission from Jurgis Kairys (photo by M. Kairys)

Figure 17.16. Jurgis Kairys flight under a pedestrian bridge in Kaunas, 2000 (a); the flight up wheels above the water (b)

Jurgis Kairys (born on 6 May 1952 in Krasnojarsk, Siberia in the Lithuanian family deported after Soviet occupation) is a Lithuanian aerobatic pilot and aeronautical engineer. He has won many awards for his flying and has invented several manoeuvres, including the "Kairys Wheel". He helped with developing the Sukhoi Su-26, -29 and -31 aerobatic aircrafts and manufactured his own aerobatic aircraft "Juka".

Gliders

Gliding is recreational activity and competitive air sport in which pilots fly an unpowered aircraft known as a glider using naturally occurring currents of rising air in the atmosphere to remain airborne. Gliding as a sport began in the 1920s. When conditions are favourable, experienced pilots can now fly hundreds of kilometres before returning to their home airfields; occasionally, the flights of more than 1,000 kilometres are achieved. One of the measures for a glider's performance is the distance that it can fly for each meter it descends, known as its lift-to-drag ratio (L/D). Depending on the class, this can range in modern designs from 44:1 in the Standard Class up to 70:1 for the largest aircraft. New materials, such as glass fiber and carbon fiber, advances in wing shapes and airfoils, electronic instruments, the Global Positioning System and improved weather forecasting, which allows many pilots to make flights that were once extraordinary. A good gliding performance combined with regular sources of rising air enables modern gliders to fly long distances at high speeds.

In the 1936 Summer Olympics in Berlin, gliding was a demonstration sport and was scheduled to be a full Olympic sport in the 1940 Games. Gliding did not return to the Olympics after the war for two reasons: a shortage of gliders and failure to agree on a single model of a competition glider.

The most commonly used sources of rising air are thermals (updrafts of warm air), ridge lift (found where the wind blows against the face of a hill and is forced to rise) and wave lift (standing waves in the atmosphere, analogous to ripples on the surface of a stream). Ridge lift rarely allows pilots to climb much higher than about 600 m above the terrain; thermals, depending on climate and terrain, can allow climbs in excess of 3,000 m in the flat country and much higher above mountains; wave lift has allowed a glider to reach an altitude of 15,447 m.

Gliding industry in Lithuania is not that large. The company manufactures new LAK-17B winglets of the standard class LAK-19 and the open class glider LAK-20 (length of wings – 23 or 26 m, can be with the reversible engine). The producers have a plan to design the LAK-17B glider equipped with an electric rise engine that will be installed in gliders. It is also planned to manufacture the LAK-17B glider with 15 meter-long wings; at the present time, design works are taking place. The

Figure 17.17. A glider designed for children

Figure 17.18. The plain AN-2 of Vilnius Parachute Training School (Kyviškės airfield, 2014)

LAK-17B glider with 18-meter-long wings is constantly improved – not a long time ago, a new keel and the ends of wings were designed.

Lithuania is the only country in the world where one has the right to become a licensed pilot at the age of nine and to manage an aircraft designed specially for children – a glider that has no analogous all over the world. It is only in Lithuania that such terms as "children aviation" and "training aircraft" have been approved by the law. One-seat gliders LAK-16 of Lithuanian construction are used for training purposes (Figure 17.17).

The Lithuanian Parachute Training Centres trains nonprofessional parachutists and personnel for the Armed Forces (Figure 17.18).

Unmanned Aerial Vehicles

An unmanned aerial vehicle (UAV), commonly known as a *drone*, is an aircraft without a human pilot on board. Its flight is either controlled autonomously by computers in the vehicle or under the remote control of the pilot on the ground or in another vehicle. Predominantly deployed for military they also used in civil applications and may include fire fighting or surveillance of pipelines. The birth of UAVs began in 1959 when United States Air Force officers, concerned about losing pilots over a hostile territory, began planning for the use of unmanned flights. Israel developed the first modern UAV and pioneered the use of UAVs for real-time surveillance, electronic warfare and decoys. The initial generations were primarily surveillance aircraft, but some were armed, which util-

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Figure 17.19. The Armed *Predator* drone firing Hellfire missile (2010)

ized air-to-ground missiles. An armed UAV is known as an unmanned combat air vehicle (UCAV) (Figure 17.19). Drones change the nature of modern aerial combat. The controllers of drones are in no immediate danger, unlike jet pilots. In terms of military logistics, much of the equipment necessary for a human pilot (cockpit, ejection seat, flight and environmental control over pressure and oxygen) can be omitted from an unmanned vehicle thus resulting in a decrease in weight. This may allow greater payloads, range and manoeuvrability.

Rapid advances in technology are enabling more and more capability to be placed on smaller airframes which is spurring a large increase in the number of small unmanned aircraft systems being deployed on the battlefield.

Several UAV types cover:

- target and decoy providing ground and aerial gunnery a target that simulates an enemy aircraft or missile;
- reconnaissance providing battlefield intelligence;
- combat providing attack capability for high-risk missions;
- research and development used to further applications of UAV technologies to be integrated into the field deployed UAV aircraft;
- civil and commercial UAVs specifically designed for civil and commercial applications.

They can also be categorised in terms of range/altitude:

- hand-held 600 m altitude, about 2 km range;
- close 1,500 m altitude, up to 10 km range;
- NATO type 3,000 m altitude, up to 50 km range;
- tactical 5,500 m altitude, about 160 km range;
- MALE (medium altitude, long endurance) up to 9,000 m and range over 200 km;
- HALE (high altitude, long endurance) over 9,100 m and indefinite range;
- HYPERSONIC high-speed, supersonic (Mach 1–5) or hypersonic (Mach 5+) 15,200 m or suborbital altitude, range over 200 km.

UAV remote sensing functions include electromagnetic spectrum sensors, gamma ray sensors, biological sensors and chemical sensors. The aerial surveillance of large areas is made possible with low cost UAV systems (Figure 17.20). Surveillance applications include livestock monitoring, wildfire mapping,

MQ-8B Fire Scout. Author Dammit. Licensed under CC BY-SA-2.5-NL

Figure 17.20. The unmanned rotorcraft design aerial vehicle MQ-8B Fire Scout

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Figure 17.21. The unmanned aerial vehicle *Reaper* during a training mission

pipeline security, home security, road patrol and anti-piracy. UAVs can help with finding humans lost in the wilderness, trapped in collapsed buildings or adrift at sea. The maximum flight duration of unmanned aerial vehicles varies widely. Internal-combustion-engine aircraft endurance depends strongly on the percentage of fuel burned as a fraction of the total weight, and so is largely independent of the aircraft size. Solar-electric UAVs hold the potential for unlimited flight. In 2007, a program of developing technology for a UAV with an endurance capability of over 5 years was revealed.

The MQ-9 Reaper has the following specifications (Figure 17.21):

- General: crew – 0 onboard, 2 in the ground station; length – 11 m; wingspan – 20 m; height – 3.6 m; empty weight – 2,223 kg; max takeoff weight – 4,760 kg; fuel capacity – 1,800 kg; payload – 1,700 kg (internal – 360 kg, external – 1,400 kg);
- Power plant: Honeywell turboprop, 900 hp (671 kW); maximum speed – 482 km/h; cruising speed – 313 km/h; range – 1,850 km; endurance – 14 h fully loaded; service ceiling – 15,240 m; operational altitude – 7.5 km.
- Armament: up to 680 kg on the two inboard weapons stations; up to 340 kg on the two middle stations; up to 68 kg on the outboard stations; up to 14 air to ground missiles can be carried or four air-to-surface missiles Hellfire and two 230 kg laser-guided bombs; testing is underway to support the operation of the Stinger air-to-air missile.

CHAPTER 18 Petroleum and Gas Engineering

This chapter will help in

- discussing petroleum and gas engineering;
- describing offshore oil engineering;
- drawing a timeline for offshore operations;
- introducing oil recovery with reference to water flooding, gas injection, chemical flooding and thermal recovery;
- presenting platforms for deep water (steel, concrete);
- giving the details of shale gas extraction;
- explaining oil extraction in Lithuania;
- getting acquainted with mining engineering.

The 19th century development of petroleum proceeded at a comparatively slow pace compared to other technologies. One principal reason was the continued dominance of coal as principal fuel for industrial processes. In 1850, coal supplied 10% of all energy consumed, but by 1885, this rose to 50%.

Production operations include bringing oil and gas to the surface as well as maintaining, purifying, measuring and testing the obtained products. The early manufacturing of oil in Pennsylvania was done using a modified water well technology because reservoirs were shallow and flowed easily from their own pressure. The behaviour of petroleum reservoirs is completely dependent on their geology. Petroleum occurs in porous rocks such as sandstones, limestones and dolomites. Seals can be shale (fractured shale can be oil-bearing), salt, gypsum, anhydrite, dense limestones and dolomites. Traps are formed by stratigraphy or structure. Lens and unconformes are stratigraphic traps; anticlines, faults and salt domes are structural traps found in porous rocks with or near oil. Most commonly it occurs as a gas cap between oil and the impermeable rock layer. In deep, high-pressure reservoirs, it is mixed or dissolved with oil and found in every rock system down to the Cambrian (Silurian, Permian).

Natural gas sometimes contains a significant amount of methane (~85% CH_4), ethane (~10% C_2H_6) propane (~3% C_3H_8), butane (C_4H_{10}) and pentane (C_5H_{12}). Different types of natural gas include lean or dry (mostly methane), "wet" (higher hydrocarbons), "sour" (hydrogen sulfide) and "sweet" (little or no hydrogen sulfide).

Innovation in drilling technology allowed the development of new fields. The rotary drilling rig was adapted from water-well drillers. "Drilling mud" was pumped down the hole to strengthen walls as the drill bit passed into the soft geologic formations of oil reservoirs. The first oil-burning railway engine was delivered in 1901. Ships began burning oil as tankers started moving crude eastward. Oil occupied about half the space required by coal leaving more room for income producing cargo. Changes in oil gave ships as much as 2000 tons more capacity.

Offshore Oil

Caddo Lake is a freshwater lake straddling the Texas-Louisiana border. This is a shallow lake with water depths rarely over 4 meters. Oil deposits were discovered in the 1900s and the lake became the site of the first offshore oil well in 1910.

The crude derrick was built on a platform with native cypress pilings driven into the sediments of the lake bottom. The well was drilled into sand formation to a depth of over 730 meters. In 1910, the well was completed at 450 barrels a day. Caddo Field is still producing oil and gas today. The lake is studded with small platforms equipped with pumps across the lake.

Louisiana was the site of the next significantly advanced offshore marine environment. In 1938, as fields declined, the onshore exploration of the near shore and then deeper marine waters began with the first offshore well drilled in 5.5 meters of water 19 kilometres offshore in the Gulf of Mexico. The effort mixed the land-based drilling technology with marine systems for the first time creating what is now termed "offshore technology". Thirty years later in 1978, Shell Oil placed the 46,000 ton Cognac platform in water depths of 312 meters drilling 62 separate wells. With a ten-fold increase in crude oil prices after 1973, uneconomic deposits in inaccessible places became economical overnight. One such place is the North Sea while the other is the Arctic. It is with the North Sea and deep-waters of the Gulf of Mexico that offshore engineering designs have become truly heroic in scale. Two principal systems are used today with newer designs be-

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Figure 18.1. A fixed steel platform base under construction

ing developed as depths increase. The steel jacket platform is the tied-beam structure. The great jackets are constructed ashore in dry dock (Figure 18.1) lying on their side resting on flotation tanks. These jackets weigh 25,000 tons or more. They are towed on launch barges weighing up to 10,000 tons and tipped into place while flotation cells are flooded. Dead weights of 110,000 tons on some North Sea platforms resist the enormous sideways forces of waves.

After the well has been completed, hydrocarbons flow from the reservoir to the surface. At the first stage of the reservoir's producing life, pressure from the reservoir forces hydrocarbons from the pores in the formation, moves them to the well and up to the surface. This stage of production is known as *primary recovery*.

The three principal primary recovery drive mechanisms cover *water drive, gas drives* and *gravity drainage*. Water drive uses pressure exerted by water below oil and gas in the formation to force hydrocarbons out of the reservoir. The greater is the depth of the reservoir, the higher is pressure. Water drive is the most efficient natural drive and can be used for producing 50% or more of the oil in the reservoir.

Two types of gas drives are dissolved-gas and gas-cap drives that use the pressure of gas in the reservoir to force oil out of the reservoir into the well. In the dissolved-gas drive, hydrocarbons in oil are light enough to become gaseous when the well releases pressure from the reservoir. This is similar to dissolved carbon dioxide in a soft drink. If a can or bottle is shaken, soda gushes out when the can is opened. When the well is opened, lighter hydrocarbons turn into gases and oil and gas flow up to the surface. The amount of oil recovered from dissolved-gas drives varies from 5 to 30%.

In some reservoirs, gas may be present in space on the top of oil. This gas cap provides pressure to push oil into the well. As the level of oil in the reservoir drops, the gas cap expands and continues to push oil into the well and up to the surface. The more space oil leaves empty in the porous reservoir rock, the more gas expands to take its place. The pressure of a gas-cap drive depletes more slowly than the dissolved gas drive. From 20 to 40% of the oil in the reservoir is recovered with gas-cap drives.

Artificial lifting methods are used when pressure from natural reservoir drive decreases to the point where the well stops producing. Artificial lift uses *pumps and gas injection*. The most common method of pumping oil in land-based wells is beam pumping the unit of which sits on the surface and creates an up-and-down motion to a string of rods called sucker rods. The top of the sucker rod string is attached to the front of the pumping unit and hangs down inside tubing. A sucker rod pump is located near the bottom of the well. The reciprocating action of the walking beam moves the rod string up and down to operate the pump.

Gas lift describes methods in which gas is used for increasing oil well production. Gas lift-dissolved-gas drive or gas-cap drive may provide natural drive to the reservoir. Natural gas can also be injected into the well to lift oil artificially on the same principle.

Natural gas makes oil in the wellbore column much lighter in weight. Because the liquid column is lighter, it exerts less pressure on the bottom of the well. With pressure lower at the bottom, pressure remaining in the reservoir becomes sufficient to push reservoir fluids to the surface through tubing. Gas lift is common when a supply of gas is economical and available and when the amount of petroleum it will lift justifies expenses.

After the well has used up the natural drives of the reservoir and gas lift or pumps have recovered all possible hydrocarbons, statistics show that 25 to 95% of the original oil in the reservoir may still be there. This amount of oil can be worth recovering if prices are high enough. The major methods of improved oil recovery are *water flooding, gas injection, chemical flooding* and *thermal recovery*. These techniques are used when production from the well starts decreasing.

Water flooding is a technique where water is injected into the formation using wells that have ceased production. The injected water enters the reservoir and displaces some of the remaining oil toward producing wells in the same reservoir. The producing wells then pump up oil and water. Several injection wells surround each producing well. Water flooding is the least expensive and most widely used secondary recovery method.

Production can also be increased by *injecting gas*, such as natural gas or nitrogen, into the reservoir. The injected gas expands to force additional volumes of oil to the surface.

Chemical flooding uses special chemicals in water to push oil out of the formation. These chemicals act as surfactants that cause oil and water to mix and breaks oil into tiny droplets that can be more easily moved through the reservoir to the well.

Thermal recovery is used when oil is so viscous, or thick, that it cannot flow through the reservoir and into the well. When oil is heated, its viscosity is decreased and the flow increases. Recovery techniques that use heat are called thermal processes or thermal recovery. *Steam drive* or *steam injection* involves generating steam on the surface and forcing this steam down injection wells and into the reservoir. When steam enters the reservoir, it heats up oil and reduces its viscosity. Heat from steam also causes hydrocarbons to form gases, which also increases the flow. Gases and steam provide an additional gas drive, and hot water also moves thinned oil to production wells.

Another way to use heat in the reservoir is fire flooding, or in situ (in-place) combustion. In fire flooding, the crew ignites a fire in place in the reservoir. They inject compressed air down an injection well and into the reservoir. A special heater in the well starts a fire that burns and begins moving through the reservoir toward production wells. Heat from the fire thins out the oil around it, causes gas to vaporize from it and changes water in the reservoir to steam. Hot water, steam and gas all act to drive oil in front of the fire to production wells.

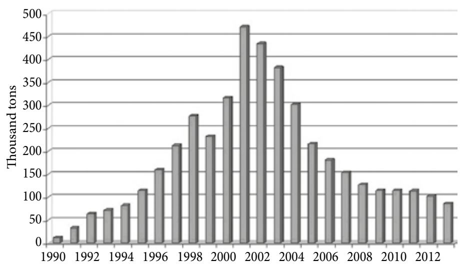
Oil Extraction in Lithuania

Lithuania failed to find oil a long time and only in 1968 the first oil fountain was attained in Western Lithuania. Oil found in Lithuania is extracted from 1850–2000 meters deep, mostly in Cambrian stratum. Due to high pressure, oil raises into surface itself. Lithuania is now operating 12 oil fields. In 2012, active extraction took place in 61 wells. About 0.4 million tons of oil per year are obtained, which makes around 5–13% of the country's needs.

Since the beginning of the industrial extraction of oil in 1990, more than 2.5 million tons of oil have been obtained (Figures 18.2 and 18.3). Scientists predict that Lithuanian oil reserves on land may be as high as 60 million tons. 40–80 million tons can be found in the Lithuanian economic zone in the Baltic Sea. In 2012, crude oil production in Lithuania made 101,880 tons. Up to now, all oil extracted in the oilfields has been exported.

Public Company ORLEN Lietuva (former Mažeikių nafta) is a petroleum refining company operating the only petroleum refinery in the Baltic States along with a crude oil and petroleum product network and marine terminal. The refinery started working in 1972. The design capacity of the refinery was 15 million tons of crude oil per year. Today, the oil refinery processes approximately 10 million tons of crude oil a year (Figures 18.4, 18.5). Until July 2006, the primary feedstock processed by the Mazeikiai refinery was Russian crude oil shipped by the *Druzhba Pipeline* system.

Figure 18.2. An oil extraction site in Western Lithuania



Permission is granted by Ministry of Environment of the Republic of Lithuania.

Figure 18.3. The oil produced in Lithuania

Figure 18.4. The oil refinery in Mažeikiai, Lithuania

Figure 18.5. The transportation of a deep oil processing facility to Mažeikiai refinery

These supplies were supplemented by crude oil and other feedstocks delivered by railway. The refinery also purchases crude oil produced in Lithuania. Currently, the main delivery of crude oil to the refinery comes via the Butinge Oil Terminal in the Baltic Sea. The products made by the refinery include gasoline, diesel fuel, fuel oil, jet fuel JET A-1, liquefied petroleum gas, bitumen and elemental sulphur.

Shale Gas Extraction

Shale gas is natural gas formed as a result of being trapped within shale formations (Figure 18.6). Shale gas was first extracted in 1821, in shallow, low-pressure fractures. Horizontal drilling began in the 1930s, and in 1947 a well was first fracked. Because shales ordinarily have insufficient permeability to allow a significant fluid flow to a well bore, most shales until 2008 were not commercial sources of natural gas.

Shale has low matrix permeability; thus, gas production in commercial quantities requires fractures to provide permeability. Shale gas has been produced for years from shales with natural fractures. The shale gas boom in recent years has been due to modern technology in hydraulic fracturing (fracking) to create extensive artificial fractures around well bores. Hydraulic fracturing is the propagation of fractures in a rock layer by a pressurized fluid (Figure 18.6). The energy from the injection of a highly pressurized

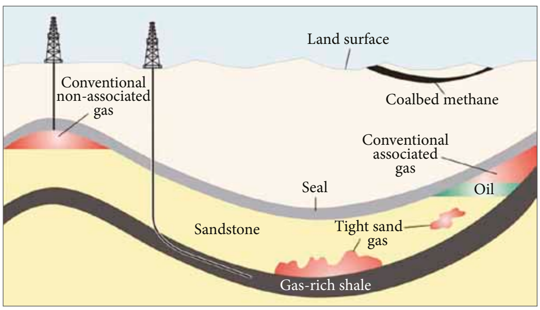


Figure 18.6. Types of gas deposits

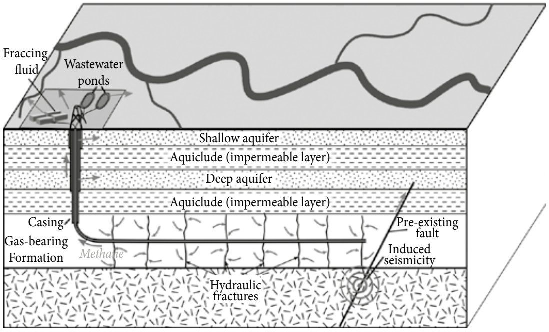


Figure 18.7. Shale gas drilling and fracking

Figure 18.8. Drilling for shale deposits in Western Lithuania

hydraulic fracturing fluid creates new channels in the rock that can increase extraction rates and the ultimate recovery of hydrocarbons. Fracturing equipment operates over a range of pressures and injection rates, and can reach up to 100 MPa and 265 litres per second. Typically, of the fracturing fluid, 90% is water and 9.5% is sand (proppant) with chemicals accounting to about 0.5% (friction reducer, agents countering rust, agents killing microorganism). Proppant types include silica sand, resin-coated sand and man-made ceramics. The most common chemical used for hydraulic fracturing is methanol, isopropyl alcohol, 2-butoxyethanol and ethylene glycol. Chemicals are added to water to facilitate the underground fracturing process that releases natural gas. Since (depending on the size of the area) millions of liters of water are used, this means that hundreds of thousands liters of chemicals are often injected into soil.

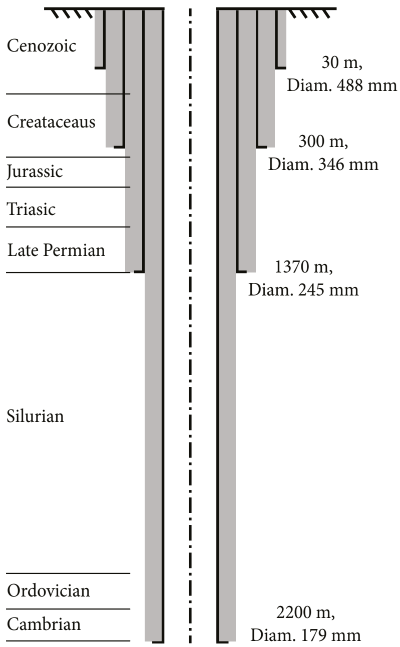


Figure 18.9. A driling site for shale gas in Lithuania

Only about 50% to 70% of the resulting volume of contaminated water is recovered and stored in the above-ground ponds to await removal or cleaning. The remaining “produced water” is left in the earth. Lithuania has about 113 billion cubic metres of shale gas in the deep of about 2000 m. (Figures 18.7–18.9). Natural gas consumption in Lithuania makes about 3 billion cubic meters per year.

Timeline for Offshore Operations

"Trolls" of the North sea. Troll is a natural gas and oil field in the Norwegian sector of the North Sea, 100 kilometres North-West of Bergen. Although Troll is primarily a gas field holding 60% of gas reserves in Norway, it also possesses significant quantities of oil. Gas and oil from the field is extracted via three platforms – Troll A, B and C. The Troll A platform is the largest structure ever to be moved. The total weight is 656,000 tones, the total height is 472 metres and depth is 303 metres. This platform came into production in 1996.

One of continuous-slip-formed concrete cylindrical legs (containing import and export risers) has an elevator that takes over nine minutes to travel from the platform above the waves to the sea floor. The walls of Troll A's legs are over 1 metre thick and are made of steel reinforced concrete formed in one continuous pour. The legs use the groups of six 40-metre anchors holding it fixed in the mud of the sea floor. Gas rises from 40 wells and is exported through a number of pipes. A concrete gravity platform has two parts – the topside deck for production and living quarters and the concrete *gravity base structure* or *GBS*. The fabrication of the GBS started in dry dock with four shafts from 19 skirt cells at the base of the structure. The cells were closed and the shafts were raised to 134 meters. The evolving GBS was then towed to another deeper water site and the shafts rose to 185 meters. At this point, a *riegel* or ring structure was added to the GBS to strengthen the shafts. The height of the riegel is 31 meters. After completion of the riegel, the shafts were topped out at the final 369 meters for the GBS.

The Troll B Platform is a semi-submersible and fabricated from concrete; thought to be the only concrete semisubmersible. This platform came into production in 1995. The Troll C Platform is a conventional steel hull semi-submersible. This platform came into production in 1999. Gulifaks C Platform was launched in the North Sea in 1989, weighed 1.5 million tons and stood 262 meters in height above the sea floor. The structure contained 245,000 tons of concrete and 80,000 tons of reinforcing steel. Steel, alone, is enough for ten Eiffel Towers.

Steel Platforms for Deep Water

Using the lessons learned in designing the previous platforms, engineers built and launched the world's tallest offshore structure, the *Bullwinkle platform* (Figure 18.10) in 1989. *Bullwinkle* is a 529 m tall, pile-supported fixed steel oil platform in the Gulf of Mexico. Of the total height, 412 meters are below the waterline. It is located approximately 260 km southwest of New Orleans. The platform was constructed in 1985–1988. The offshore platform must be fabricated on a shore facility transported and placed on its site. *Bullwinkle* weighs in excess of 75,000 tons. Like all offshore designs, it must resist a 20-meter wave and withstand fatigue stresses caused by millions of small ones. The platform has a period of five seconds and moves almost two meters at the deck level. Under storm conditions, dynamic amplification is about 50%. The footprint of *Bullwinkle* covers 1.8 hectares (124×148 meters) and tapers to 49 by 43 meters at the jacket top and expands upward to an 87 by 56 meter deck. *Bullwinkle* is a single-piece structure erected at one time.

Figure 18.10. A semi-submersible oil platform

Search for oil has led engineers to cope with the harshest of environments, the Arctic. In the mid-1970s, exploratory drilling began in the Beaufort Sea. The design of structures for oil recovery in the Arctic must consider the force of sea ice. Large lateral shear forces produced by ice must be compensated by designs unique to this requirement. Engineers have built artificial earth islands in 20 to 60 meters of water. The typical exploratory island is 100 meters in diameter with production islands of 500–600 meters. They rise 4 to 25 meters above the sea ice.

Concrete Gravity Platforms

A *concrete gravity platform* is the one placed on the seabed, and by its own weight, is capable of withstanding environmental forces (wind and waves) as well as may be exposed to during its lifetime. Gravity-type concrete platforms are used for oil-drilling purpose in oil industry. They are applied when the soil/rock of the seabed is hard and relevant for supporting them since they rest directly on the ocean floor without pile foundation. Concrete platforms are larger and heavier than steel platforms. Although there are various designs of a concrete gravity platform, the base part is usually made of reinforced concrete and consists of huge subsea concrete tanks (Figure 18.11) employed for storing crude oil. For example, the base of a concrete gravity platform built for Mobil Oil in the North Sea consists of 19 hollow concrete cylinders of 60 cm wall thick, 20 m in the inner diameter and 49.5 m in height. The cylindrical concrete storage tanks of the base are built to such a height that when the dock is flooded the base has sufficient freeboard to float on its own buoyancy. The base is then towed out into a deep-water site and tanks are flooded thus causing the base to sink to the seabed.

Figure 18.11. A concrete gravity platform

Figure 18.12. Thick oil washes ashore in Louisiana

The platform rests on several (usually three or four) taller, tapered concrete legs (columns) that rest on several of the tanks. These columns can rise up to 90 m above the top of the tanks. Finally, steel superstructure is placed on the top of the completed concrete substructure. Sometimes grout is pumped beneath the bottom of the tanks to provide a firmer foundation. The platform is heavy enough to remain stable without pile foundation when subject to environmental forces.

On 20 April 2010, an explosion and fire on the drilling rig in offshore Louisiana, beneath about 1,525 m of water, took place. 11 persons were reported missing and approximately 17 injured. A blowout preventer, intended to prevent a release of crude oil, failed to activate. On 22 April 2010, the deepwater rig sank. Remote underwater cameras displayed the well leaking 1000 barrels (1 barrel – 159 L) or 8400 m^3 of crude oil per day. The total oil spill volume was 780,000 m^3 (Figure 18.12). The well was permanently sealed only on 19 September 2010.

Mining Engineering

The mining of gold fields of South Africa began in earnest in 1885. Since 1913, these deposits have produced over 45% of the world's gold annually. The "reefs", as they are commonly called, slope deeply below the surface in complex formations occurring in uniform, microscopic deposits often mixed with pyrites-sulfide minerals rather than oxides. Gold reefs are really seams that are only a few centimetres thick. Indeed, the ratio of the amount of rock mined and crushed to gold recovered is now 40 to 50 tons of rock to 336 g of the precious metal. On the other hand gold reefs are spatially enormous over 160 by 272 km. The gold ore is found in conglomerates, a rock mix of gravels that were laid in a large prehistoric marine basin some 2.7 billion years ago.

The cemented conglomerate of the reefs presented the problems of gold extraction that became next to impossible using the mercury-amalgam process. Mercury dissolves gold and at 10% gold is liquid. The powdered ore is mixed with water to form slurry and passed over copper plates amalgamated with mercury. The plates are scraped yielding amalgams of 40 to 50% gold which are then heated to drive off mercury leaving gold.

The gold extraction problem lay in the nature of the ore that can be a sulphide ore containing iron pyrites (FeS). The pyrites are combined with mercury reducing the extraction of gold to 55 to 65%. The use of a new extraction process, called cyanidation, rose the extraction rate. Cyanidation uses weak solutions of potassium cyanide to precipitate gold onto zinc shavings that could then be separated by heating.

The Western Deep gold mine shaft at Western Deep reached 4,265 meters in 1982 and has since gone further. At a shaft sinking rate of 16 m per day, one can extrapolate the possibility. At these depths, heat is the enemy. The temperature of virgin rock at these levels can approach 60°C. South African mining engineers began pumping down cooled air to lower mine temperatures, but adiabatic heating and that of the rock face kept air temperatures at 30°C. Then, they turned to ice water which is today used on the rock and to chill the air. This has lowered air temperature to 28 degrees, seemingly small but a decrease in two degrees improves the miner's performance and increases productivity as well as improves safety.

Sinking a Western Deep requires great technological skill and money. The shafts are sunk and lateral tunnels or stapes are driven to the reefs. From there, working faces are blasted and the rock raised to the surface in lifts travel at nearly 70 km per hour. Mining operations raise 450,000 tons of rock per month. For every ton of rock, 15 tons of air are pumped down the mine and thousands m^3 of water are out. Since, it takes 50 tons of reef rock to produce 336 g of gold.

The deepest mine in the world, Savuka Mine in the North West Province, South Africa counts 3,774 meters. The Gold Fields Ltd company intends to set a new record by drilling down 4 km (Johannesburg). The estimated 240 million grams of gold is thought to lie at such depths.

The world's biggest hole is a diamond mine located in Russia near the town Mirna and is 525 meters deep and 1.25 km in diameter.

CHAPTER 19 | Bridges

This chapter will help in

- introducing bridge builders of antiquity;
- explaining the structure of a beam, arch and cantilever bridges;
- defining the structure of a suspension, cable-stayed and truss bridges;
- describing moveable bridges;
- presenting achievements in bridge building in Lithuania;
- dealing with military bridges.

A bridge is a structure built to span a valley, road, body of water, or other physical obstacle, for the purpose of providing passage over the obstacle. The design of bridges may vary depending on the function of the bridge and the nature of the terrain where the bridge is constructed. The first bridges were made by nature – as simple as a log fallen across a stream. The first bridges made by humans were probably the spans of wooden logs or planks and eventually stones using a simple support and crossbeam arrangement.

The greatest bridge builders of antiquity were the ancient Romans who built arch bridges and aqueducts that could stand under conditions that would damage or destroy earlier designs. Some of those are still standing today. The Romans also used cement, which reduced the variation of strength found in natural stone. One type of cement, *pozzolana*, consisted of water, lime, sand and volcanic rock. Brick and mortar bridges were built after the Roman era, as the technology for cement was lost and later rediscovered. A rope bridge, a simple type of a suspension bridge, was used by the Inca civilization in the Andes Mountains of South America, just prior to European colonization in the 1500s.

Although large Chinese bridges of wooden construction existed in ancient times, the oldest surviving stone bridge was built from 595 to 605 A.D. which is also historically significant as the world's oldest open-spandrel stone segmental arch bridge. European segmental arch bridges date back to at least the 2nd century A.D. while the enormous Roman era Trajan's Bridge (105 A.D.) featured open-spandrel segmental arches in wooden construction.

During the 18th century, there were many innovations in the design of timber bridges. A major breakthrough in bridge technology came with the erection of the Iron Bridge in England in 1779. It used cast iron for the first time as arches to cross the river Severn. With the Industrial Revolution in the 19th century, the truss systems of wrought iron were developed for larger bridges, but iron did not have tensile strength to support large loads. With the advent of steel, which has high tensile strength, much larger bridges were built. In 1927, welding pioneer Stefan Bryla designed the first welded road bridge in the world near Lowicz, Poland.

A bridge is designed for trains, pedestrian or road traffic, a *pipeline* or *waterway* for water transport or *barge traffic*. An *aqueduct* is a bridge that carries water resembling a viaduct, which is a bridge that connects the points of equal height. A road-rail bridge carries both road and rail traffic. Bridges may be classified by how the forces of *tension, compression, bending, torsion* and *shear* are distributed through their structure.

Beam Bridges

A beam or "girder" bridge is the simplest kind of the bridge (Figure 19.1). In the past, they may have taken the form of a log across a stream but today they are more familiar to us as large box steel girder bridges. There are a number of different types of beam bridges. The beam bridge needs to be stiff and resist twisting and bending under the load. In its most basic form, the beam bridge consists of a horizontal beam that is supported at each end by the piers. The weight of the beam pushes straight down on the piers. Under the load, the top surface of the beam is pushed down or compressed while the bottom edge is stretched or placed under tension. If we imagine that there is an imaginary line running down the centre of the beam, this line remains at its original length while the material above is compressed and the material below is stretched. This line is referred to as the neutral axis.

Figure 19.1. A beam bridge

Arch Bridges

Arch bridges are one of the oldest types of bridges and have been around for thousands of years. Arch bridges have great natural strength. They were originally built of stone or brick but these days are built of reinforced concrete or steel (Figure 19.2). The introduction of these new materials allows arch bridges being longer with lower spans. Weight is transferred to the supports at either end. These supports, called *abutments*, carry the load and keep the ends of the bridge from spreading out. The load at the top of the key stone makes each stone on the arch of the bridge press on the one next to it. This happens until the push is applied to the end supports, or abutments, embedded in the ground.

Figure 19.2. A iron arch bridge

Cantilever Bridges

Cantilever bridges normally use the pairs of cantilevers back to back with a short beam bridge in between the cantilevers (Figure 19.3). Modern motorways have cantilever bridges stretching across them; they have a cantilever coming out from each side and a beam bridge in between them.

Public domain

Figure 19.3. A steel cantilever bridge

Huge pillars take up compression held up by narrow top members. Attached to these are complicated struts and cross bracing that withstand forces causing buckling and twisting. Outer cantilevers have counterweights at the ends to maintain balance.

Suspension Bridges

Figure 19.4. A suspension bridge

Suspension bridges in their simplest form were originally made from rope and wood. Modern suspension bridges use a box section roadway supported by high tensile strength cables (Figure 19.4). Today, the cables are made of thousands of individual steel wires bound tightly together. Steel, which is very strong under tension, is an ideal material for cables. A single steel wire, only 2,5 mm thick, can support over half a ton without breaking. Light and strong suspension bridges can span distances from 600 to 2100 m far longer than any other kind of the bridge. They are ideal for covering busy waterways. Suspension bridges tend to be the most expensive to build. A suspension bridge suspends the roadway from huge main cables that extend from one end of the bridge to the other. These cables rest on the top of high towers and have to be securely anchored into the bank at either end of the bridge. The towers enable the main cables to be draped over long distances. Most of the weight or load of the bridge is transferred by the cables to anchorage systems. These are embedded in either solid rock or huge concrete blocks. Inside anchorages, the cables are spread over a large area to evenly distribute the load and to prevent the cables from breaking free.

A suspension bridge, because of its flexibility, might be thought to present fewer problems than other bridges, but as always, designers have to think carefully. The central part of the deck of one of the longest suspension bridges may vary in vertical position by as much as three metres from summer to winter, as a result of changes in the length of the main cables.

Cable-stayed Bridges

Cable-stayed bridges, like suspension bridges, are held up by cables (Figure 19.5). However, in a cable-stayed bridge, less cable is required and the towers holding the cables are proportionately shorter.

Figure 19.5. A pedestrian cable-stayed bridge in Vilnius

Figure 19.6. An iron truss bridge

Truss Bridges

Truss bridges are composed of connected elements. They have a solid deck and a lattice of pin-jointed or gusset-joined girders for the sides. Early truss bridges were made of wood, and later of wood with iron tensile rods; however, modern truss bridges are made completely of metal such as wrought iron and steel or sometimes of reinforced concrete (Figure 19.6).

Moveable Bridges

A *moveable bridge* is a bridge that moves to allow passage (usually) for boats or barges (Figure 19.7) or for pedestrian (Figures 19.8 and 19.9). It originated in medieval Europe, probably Normandy, as a defensive feature of castles and towns. It was operated by a counterweight and winch. Modern bridges are powered by electric motors,

Author Adrian Pingstone. Public domain

Figure 19.7. A movable Tower Bridge in London

Figure 19.8. A swinging pedestrian bridge in Klaipėda

Figure 19.9. A vertically movable bridge in Klaipėda

whether operating winches, gearing or hydraulic pistons. The world's longest-span movable bridge (340 m) is across the Suez Canal (Egypt) providing road and rail links between the Sinai Peninsula and eastern Nile delta region (2001).

Architecture of Bridges

Wood, brick, stone, cast iron, wrought iron, mild steel, high-tensile steel, alloy steel, aluminium, steel-reinforced concrete, pre-stressed concrete and glass-reinforced plastic are some of the materials available for bridge building.

Bending, compression, impact, oscillation, pressure, tension, torsion, vibration, contraction, corrosion, erosion, expansion, fatigue, friction, rain, river flow, sea-water, scouring, temperature changes, tidal flow, turbulence, waves, wind erosion, wind gusts, wind pressure are some of stresses that bridge materials must withstand in a variety of combinations.

Assembly is more than a simple matter of connecting all parts. Welding produces high temperatures producing expansion and distortion. The management of the cooling process is vital. Poor quality control over welding may allow detrimental changes in the properties of metal. The weight distribution of structure changes during assembly requires precautions such as adjustable jacking. Forcing two parts into alignment produces undesigned and undesirable stresses that can start cracks.

Bridge structures must be maintained. Air contains mainly nitrogen and oxygen as well as much smaller quantities of gases that were generally not present 500 years ago. Sulphur dioxide, sulphur trioxide and nitrogen oxides are some of those. These gases are constantly available on every surface aided on occasion by rain that wets the surface and increases corrosion. Riveted or bolted joints may be penetrated by thin films of liquid that may do its insidious work unseen. There is the entire industry providing solutions such as painting and anodizing.

What are the commonest modes of bridge failures? The straightforward failure of a strut or tie because it is not strong enough is quite rare. Ties rarely snap in two, and struts are seldom crushed by pressure. What is more likely is that fatigue can start cracks resulting in stress concentration, and at a certain moment, the crack can propagate with lightning speed leading to a broken tie. Struts and plates are more likely to fail by buckling than by straight crushing. Joints (including the connection to the ground and river beds) are a frequent source of failure. Some joints consist of a span resting on a support held only by its own weight, and there have been cases where flooding has created enough upthrust for a span to be moved off its pier by flowing water. Some earthquakes are so powerful that no structure can withstand them. Ships have been known to collide with bridges bringing spans crashing into the river.

In the 20th century, bridge design has brought engineering to a level that has been called "structural art." Concrete construction in late 19th century advanced with the first production of artificial cement called "Portland cement". Reinforced concrete construction began in the late 1800s. Designs particularly used reinforcing in the tensile zone of beams. Reinforcing gave concrete always high in resistance to compression as well as great resistance to tensile forces. Joseph Monier, a Paris gardener, built a small reinforced concrete bridge using a very thin arch (1875). Reinforced concrete beams were developed. The beams were correct in the placement of reinforcement steel in the tensile zone bent upwards at the supports and anchored in the compression zone by vertical steel stirrups. Also, an important discovery was the equality of thermal expansion coefficients of iron and concrete that contributes to the high fire resistance of the medium.

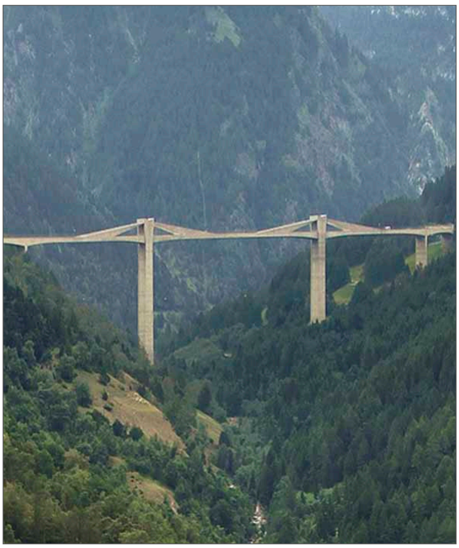


Figure 19.10. The Ganter Bridge in Switzerland rises astonishing 152.4 meters above the ground

The first engineer to examine the theoretical nature of reinforced concrete structures was M. Koenen (Germany) who stated the relationship of tensile and compressive forces in statistical terms. The first successful bridge made of this material was a viaduct with a 100 meter span combined strength (it was a railway bridge) and elegant slenderness (1912).

Pre-stressed structures allowed creating even more dramatic spans such as Ganter Bridge (Figure 19.10). Its unique design (the main span is 174 m with a column height of 150 m) combines the elements of a cable-stayed bridge and a pre-stressed cantilever hollow-box girder bridge with triangular concrete walls above the roadway that contains pre-stressed cable-stays. This hybrid type of a cable-stayed and girder bridge is sometimes referred to as an *extra-dosed* bridge.

Figure 19.11. Žirmūnai bridge across the Neris river in Vilnius, Lithuania

With permission from JSC *Lithuanian Railways*

Figure 19.12. The Lyduvėnai Railway Bridge is one of the longest bridges in Lithuania

In bridge architecture, far more than in engineering, we are highly likely to see geometry as decoration in the form of apparent arches, beams and pillars that actually fail to do something special and are only added for effect. A bridge with long spans may be curved in plan not only to provide for expansion but to reduce the amplitude of oscillations etc. There are places in different cities where huge "arches" lie on their sides, like arch dams; nevertheless, these curves do not resist any forces, so they are aesthetic and of geometrical shapes and not engineering curves, and certainly, not arches. To create a beautiful image, some bridges are built much taller than necessary. This type, often found in east-Asian style gardens, is called a Moon bridge evoking a rising full moon. Other garden bridges may cross only a dry bed of stream washed pebbles intended only to convey an impression of a stream. A bridge in palaces will be frequently built over an artificial waterway as a symbol of a passage to an important place or state of the mind. Elegant appearance is a bridge across the Neris river in Vilnius (Žirmūnų bridge) (Figure 19.11). The Lyduvėnai Railway Bridge is one of the longest bridges in Lithuania (Figure 19.12).

Table 19.1. The leading bridges

Bridge	Main span, m	Completion date	Location
Suspension bridges			
Akashi Kaikyo	3934	1998	Japan
Hull-over Humber	1410	1982	Great Britain
Golden Gate	1280	1937	San Francisco, USA
Bosporus	1074	1973	Istanbul, Turkey
Cantilever bridges			
Quebec (Railway)	549	1917	Quebec, Canada
Delaware River	501	1971	Chester, PA – Bridgeport, USA
Howrah	457	1943	Calcutta, India
Transbay	427	1936	San Francisco, USA
Steel arch bridges			
Bayonne (Kill Van Kull)	504	1931	New York, USA
Sydney Harbor	503	1932	Sydney, Australia
Port Mann	366	1964	British Columbia, Canada
Thatcher	344	1962	Panama Canal, Panama
Continuous truss bridges			
Astoria	376	1966	Columbia River, Canada
Francis Scott Key	360	1978	Baltimore, USA
Tenmo	300	1966	Kumamoto, Japan
Dubuque	258	1943	Mississippi River, USA
Cable-stayed bridges			
Rhine	365	1979	Flehe, Germany
Duisburg-Neuenkamp	350	1970	Duinsburg, Germany
Mesopotamia	340	1972	Corrientes, Argentina
Bratislava	303	1971	Danube River, Czechoslovakia
Concrete arch bridges			
New River	510	1981	Virginia, USA
Krk	390	1980	Adriatic Coast, Yugoslavia
Gladesville	305	1964	Sydney, Australia
Foz do Iguassu	290	1964	Parana River, Brazil–Paraguay
Continuous-plate and box-girder bridges			
Sava I	75–261–75 m	1956	Belgrade, Yugoslavia
Zoobrucke	74–259–145 m	1966	Cologne, Germany
Bonn-Sud	125–230–125 m	1971	Bonn, Germany
San Mateo-Hayward	114–229–114 m	1967	San Francisco, USA

Military Bridges

Military bridges. Far back into history, armies have needed to cross water. The essence of a military bridge is that its parts must be portable using available transport equipment such as trucks, tanks and helicopters. Some specialist tracked vehicles carry folding bridge decks that can be unfolded in front of the vehicle, cantilevering out over the gap, until they rest on the other side, at which point the deck is locked at its mid-point becoming a beam.

Figure 19.13. A military pontoon bridge

A different approach is to assembling a bridge from a set of pre-fabricated standard parts that can be assembled in different configurations.

A pontoon bridge is an ancient form of construction that can be based on boats or on purpose built pontoons which are essentially piers that float instead of resting on the bottom of the channel. Segmental decks can be laid on pontoons and hinged together to form a strong deck. If pontoons are close enough together, the spans can be simple wooden beams. Modern techniques allow for fewer pontoons and longer spans. Military bridges entered a new age with the American Civil War because of the need to carry railroads. In modern times, heavy tanks and artillery pieces have to be carried (Figure 19.13).

CHAPTER 20 Water and Road Transport

This chapter will help in

- defining superships, bulk carrier and container ships;
- introducing ship building in Lithuania;
- looking at the early history of a submarine;
- explaining a sonar and periscope;
- showing a disaster of a modern submarine;
- discussing tunnels for automobile, truck and rail transport (Holland, Gotthard, Kaunas, Channel Tunnel);
- talking about the birth and development of an automobile;
- considering the emergence of an electric vehicle;
- presenting bus, tram and trolleybus transport;
- dealing with bicycle development and manufacturing in Lithuania;
- observing modern road and highway systems.

Super ships

The growing consumption of oil and liquefied natural gas around the world has led to designing the largest ships afloat – supertankers. They fall into two classes: VLCC (*Very-large Crude Carrier*) and ULCC (*Ultra-Large Crude Carrier*). VLCC's range from 250,000 to 275,000 deadweight tons with ULCCs up to 500,000 deadweight tons (Figure 20.1). They are generally powered by steam turbines. This reduces the risk of fire and provides steam to heat oil for pumping. Steam drives pumps, capstans and winches instead of electric motors. Most are single screw ships using variable-pitch propellers to achieve economic performance. Their drafts are so great (18 meters or more) that they can only use the deepest harbours. In shallow areas, offshore facilities have been built to offload them far from land. These large supertankers are all potential environmental disasters by virtue of the spillage of oil. The wrecks of the Torrey Canyon (1967), Amoco Cadiz (1978) and Exxon Valdez (1989) created extensive

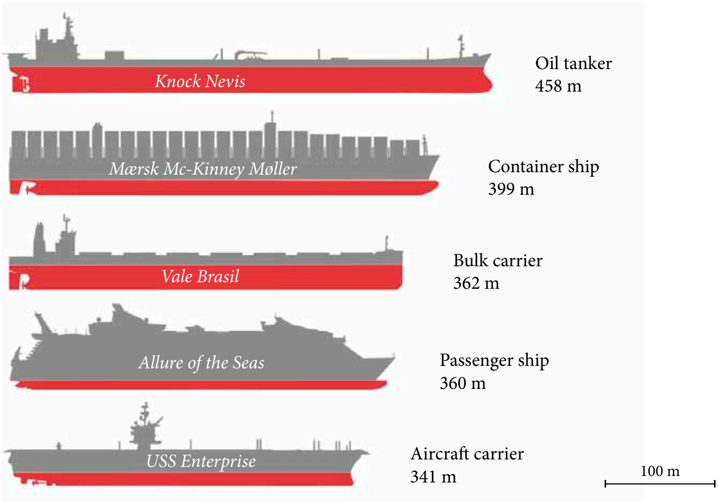


Figure 20.1. A comparison of some longest ships

pollution of nearby coasts. The Torrey Canyon spilled 117,000 tons of crude while the Amoco Cadiz and Exxon Valdez spilled in excess of 200,000 tons each. In 2005, oil tankers made up to 36.9% of the world's fleet in terms of deadweight tonnage. The combined deadweight tonnage of oil tankers and bulk carriers represents 72.9% of the world's fleet. The common rule is that the volume that can be carried in a tanker increases a function of the cube of its length. Oil tankers have a commercial life expectancy of about 30 years. Because of their huge mass, tankers have large inertia making them very difficult to steer. A loaded supertanker could take as much as 4 to 8 kilometres and 15 minutes to come to a full stop and has a turning diameter of about 2 kilometres.

Bulk Carrier and Container Ships

In the late 20th century, seaborne commerce benefited from advances in both ship manufacturing techniques and propulsion systems. These, in turn, have led to the standardization of design for cargo vessels into two main types – a bulk carrier and a container ship (Figure 20.2). The bulk carrier is a vessel that carries nonpetroleum bulk cargos, including grain and economic minerals such as bauxite, kaolin, etc. The

Figure 20.2. A container ship

container ship is specifically designed for transporting a wide range of goods of a varying size that are placed in standardized boxcar-sized containers made of steel. The containers are sized to fit handling gear that is the same throughout the world's major ports. They fit onto trucks or lorries and are, likewise, carried on trains and inland barges. In a word, "one-size fits all".

The size of these ships has followed those of petroleum carriers although not reaching the gargantuan proportions of the largest of that type. Still, it is common to see 250 m cargo vessels of 30,000 ton capacity. These large vessels maintain open ocean speeds of 18 to 24 knots (1 knot = 1.852 km/h). Their average seatime between ports is 7 to 8 days.

For over 25 years after World War II, the geared steam turbine was the chosen power plant for petroleum carriers. Today, it can be found in liquefied natural gas (LNG) carriers only. Oil-diesel exists for marine propulsion rated at 20 to 30 MW.

As regards materials, the greatest change has been in the use of high tensile steel in hulls that have been significantly reduced due to corrosion control costs and consequent decreased use lives of hulls that now contain as much as 65% of mild steel. Naval engineers and architects have revisited the earlier ideas one of which is paired *contra-rotating propellers* (CRP) that have been reported to realize a 14% increase in efficiency. A conventional propeller (CP) system converts energy to thrust by rotating. A swirling flow occurs in the slip stream and its energy of the component of rotating velocity is wasted. As regards the contra-rotating propeller system, the energy of the component of rotating velocity produced by a forward propeller is to be recovered by an aft propeller (swirling flow is to be cancelled with each other), which is additionally installed on the concentric dual shaft and rotates in the opposite direction behind the forward propeller. Since necessary thrust is shared by two propellers in the CRP system, the load of one propeller becomes smaller and improves the efficiency of the propeller.

A newer technological departure in ship design is the "*Fast Ship*". Using new technologies the FastShip is created to reduce time for trans-oceanic voyages by half. A trans-Atlantic voyage that now averages about a week would take only 3,5 days in the FastShip with a round-trip service of eight days. The average service speed of the FastShip would be 78 km/h with a cargo capacity of 8,000 tons. A reduction in capacity compared to the regular container vessel would be offset by the frequency of service. FastShips would be powered by gas turbines using waterjets rather than by propellers to reduce drag. Gas turbines produce up to 5 MW being derived from aircraft designs. The FastShip design calls for a vessel of an overall length of 236 m with all cargo carried within the hull.

Ship Building in Lithuania

The shipbuilding yard in Lithuania was founded in port city Klaipėda in 1952 to produce fishing boats (Figure 20.3). In 1958, the yard started building steel floating docks. JSC "Baltija" Shipbuilding Yard is the only in the Baltic States (Lithuania, Latvia and Estonia) that supplies fleets and marine companies worldwide. "Baltija" builds pontoons, barges, trawlers, floating docks, river ferries, dry cargo ships and container carriers as well as provides ship-repair services.

Figure 20.3. Steel trawl-boat *Dubingiai* ("Baltija" Shipbuilding Yard, 1987)

In 2010, *Western Baltija Shipbuilding* celebrated the onset of the world's largest gas-powered ferry construction. The ferry is able to accommodate 600 passengers and 242 cars, is 129,9 m in length and 19,2 m in breadth, equipped with four ship's azimuth thrusters, a gas-powered electrical system consisting of three 2,430 kW gas engines and a 3,000 kW diesel engine. The speed of the inland waterways ferry is 20 knots. The ferry is called as a very environment-friendly, because the LNG (liquefied gas) engine is based on a "green" concept.

Western Baltija Shipbuilding built the vessel as a fire fighting harbour/escort tug of the ASDT type (Azimuth Stern Drive Tug). The vessel is built with a double hull, and trial speed makes 14,2 kn (Figure 20.4).

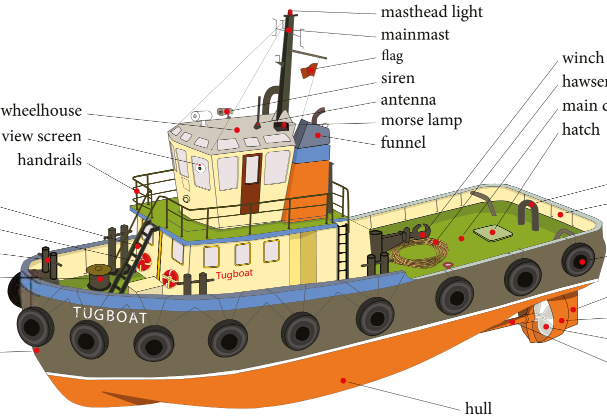


Figure 20.4. A fire fighting harbour/escort tug

Figure 20.5. Barquentine *Meridianas* in Klaipėda port

The Baltic Sea could be described as a cradle of European ship building, as most developments – from the simple boat building of the Vikings through the technically advanced structure of the cogs to the modern shipyards of our days. During the XVIII–XIX century Klaipeda sailing shipping reached the peak of its golden age and outlived the sunset. The sailing ship is a symbol of the nautical city Klaipėda, a modern port of Lithuania (Figure 20.5). Barquentine *Meridianas* was constructed in 1948 (Turku, Finland) and was presented as a training ship of Klaipėda Maritime school.

Early History of a Submarine

In 1776, David Bushnell built the one-man human powered *Turtle* submarine (Figure 20.6). The submarine attempted to sink the British warship with the *Turtle*. The first intended purpose of the submarine was to break the British naval blockade of New York harbour during the American Revolution. With slight positive buoyancy, it floated with approximately six inches of the exposed surface. The *Turtle* was powered by a hand-driven propeller. The operator would submerge under the target, and, while using a screw projecting from the top of the *Turtle*, would attach a clock-detonated explosive charge.

Robert Fulton's cigar-shaped *Nautilus* submarine (1801) was driven by a hand-cranked propeller when submerged and had a kite-like sail for surface power. The *Nautilus* submarine was the first submersible to have separate propulsion systems for surfaced and submerged operations. It also carried the flasks of compressed air that permitted the two-man crew to remain submerged for five hours. The *Nautilus* was 7 meters with a 1.9 meter beam. She was shaped like an ellipsoid. The angle of dive was controlled by horizontal fins on the rudder. Pumping water ballast from the keel allowed her to surface. The *Nautilus* first had an Archimedean screw that Fulton changed to a four-bladed propeller.

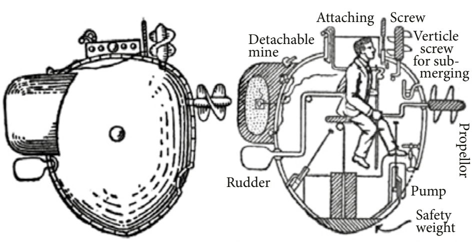


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Figure 20.6. Bushnell *Turtle* submarine

The first successful cigar-shaped 10 meter long with a 1.8 meter beam vessel was built in Holland in 1881. The hull was 0.2 meter flange iron. Holland powered it with 15 to 17 hp and two-cycle Brayton gasoline engine located amidships. Sealed bow and stern compartments contained compressed air for positive buoyancy. A relatively useless pneumatic gun, 3.3 meters long, was the armament of the submarine. The internal combustion engine offered speed and comparative endurance on the surface, but its deadly carbon monoxide exhaust fumes and high oxygen consumption were obstacles to life beneath the surface. In addition to the development of the torpedo, electric batteries and motors were becoming more reliable and offered the submarine designs fifty times the mobility of a hand-powered submersible. By 1900, submarine designers had solved this problem with storage battery and electric motors.

Germany did not build a submarine until 1906. Two reasons are given for this late entry. First, Admiral Von Tirpitz emphasized the development of a surface battle fleet to rival that of Britain, and second, the German Navy objected to the use of steam or gasoline for propulsion.

With the development of the heavy-oil (diesel) engine, the Navy had no technical objection to the submarine. Their designs were similar to Holland derived British series with a displacement of over 500 tons, twin screws, diesels, external main ballast tanks and extended battery power to allow submergence during daylight hours. By 1914, the Germans had 29 "overseas" or long-range vessels, called U-boats, ready for World War I. Still, throughout the period of 1914–1916, German U-boats cruised at will off the English coast. In 1915, the "exchange rate," i.e. ships sunk to U-boats destroyed was 23:1. The other modern innovation in warfare, an airplane, was then used against the submarine in the form of seaplanes and flying boat anti-submarine patrols.

In 1916, a hydrophone (under development since 1915) was developed against submarines. This underwater microphone could hear submarine's machinery and was the first electronic device capable of detecting a completely submerged submarine. In 1914, the most singular innovation in naval warfare to that time was the large scale offensive use of the submarine by the Germans. The threat of U-boats created the whole technology for the detection and destruction of these vessels. In 1914, there were no methods for detecting submerged submarines unless they showed their periscopes or conning towers. Submarine armament was comparatively weak. Torpedoes were fired singularly and were of a short range. Shallow draft vessels and small ships were hard to hit due to depth that most of the early torpedoes ran.

For the period 1916-1918, Germany opened unrestricted submarine warfare. Casualties in Allied merchant shipping rose to over 115 ships a month in 1916. When using U-boats with this extended range and size, the exchange rate rose to 65:1 in 1917. The submarine did not become the new capital ship in that it could not easily destroy other warships. The importance of the submarine as a naval weapon lay in its not being invulnerable but rather being replaceable.

Sonar

Sonar (originally an acronym for **SO**und **N**avigation **A**nd **R**anging). Active sonar is a device that emits bursts of sound in the hope of receiving an echoing answer (Figure 20.9). Echo-ranging sonar was developed a decade or more before World War II. This system utilized a relatively narrow beam (15 to 20°). Coupled with the speed of sound in water of 914 meters/sec, the narrow beam transducer required accurate training of the projector and hence was somewhat time consuming in usage. Since the disadvantage was offset by the fact that electromagnetic waves do not propagate in sea water, hence, sonar is the only reliable means of searching for submerged submarines or objects in general. The shape and intensity of a pulse reflected from a target is dependent upon the aspect and shape of the target. For instance, if a submarine presents only its bow to the impinging pulse, the sonar echo is minimal. The echo from the quarter is stronger with the stern less but greater than the bow due to propeller noise. Most early sonar transducers were piezoelectric devices where changes in the crystal's physical dimensions were induced by electrical to voltage applied across it. Likewise, a returning pulse stresses the crystal generating weak voltage due to compression. The same electrical connections that apply to voltage can be used for indicating the existence of the received pulse. When active sonar is used for measuring the distance from the transducer to the bottom, it is known as *echo sounding*. Modern naval warfare makes the extensive use of both passive and active sonar from water-borne vessels, aircraft and fixed installations. Since active sonar reveals the presence and position of the operator, and does not allow the exact classification of targets, it is used by fast (planes, helicopters) and noisy platforms (most surface ships) but rarely by submarines. When active sonar is used by surface ships or submarines, it is typically activated very briefly at intermittent periods to minimize the risk of detection. Consequently, active sonar is normally considered a backup to passive sonar.

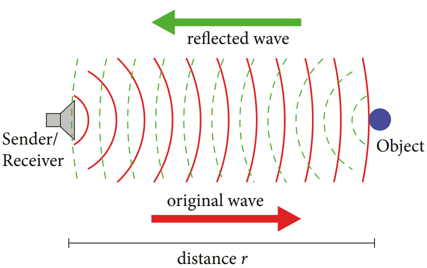

Figure 20.7. The principle of an active sonar

Passive sonar has several advantages. Most importantly, it is silent. If the target radiated noise level is high enough, it can have a greater range than active sonar and allows the target to be identified. Since any motorized object makes some noise, it may in principle be detected, depending on the level of noise emitted and the ambient noise level in the area, as well as the technology used. On a submarine, nose-mounted passive sonar detects in directions of about 270° centred on the ship's alignment, the hull-mounted array of about 160° on each side and the towed array of full 360°. The invisible areas are due to the ship's own interference. Once a signal is detected in a certain direction, it is possible to zoom in and analyze the signal received (narrowband analysis). Since every engine makes a specific sound, it is straightforward to identify the object.

Passive sonar has a wide variety of techniques for identifying the source of a detected sound. For example, U.S. vessels usually operate 60 Hz alternating current power systems. If transformers or generators are mounted without proper vibration insulation from the hull or become flooded, 60 Hz sound from windings can be emitted from the submarine or ship. This can help with identifying its nationality, as most European submarines have 50 Hz power systems. Until fairly recently, an experienced, trained operator identified signals, but now computers may do this. Passive sonar systems may have large sonic databases, but the sonar operator usually finally classifies signals manually. A computer system frequently uses these databases to identify the classes of ships, actions (i.e. the speed of a ship, or the type of the weapon released) and even particular ships.

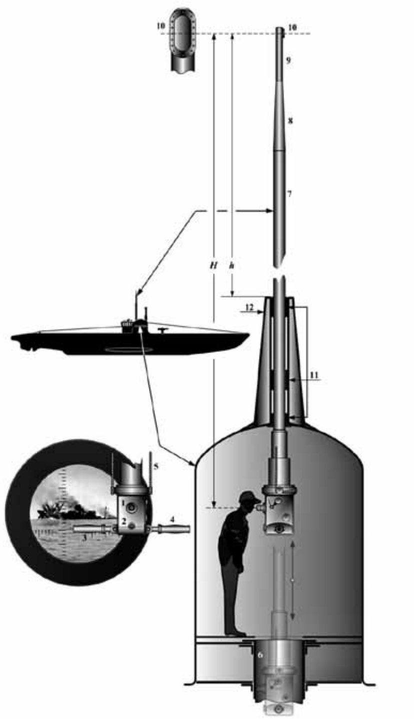

Figure 20.8. The Submarine's monocular attack periscope

The Periscope

A periscope is an optical device for conducting observations from a concealed or protected position (Figure 20.10). The Navy attributes the invention of the periscope (1902) to Simon Lake. Simple periscopes consist of reflecting mirrors and/or prisms at opposite ends of a tube container. The reflecting surfaces are parallel to each other and at an angle of 45° to the axis of the tube.

The design of the periscope has come a long way since this time and they were used extensively on submarines and tanks during both world wars. For more than 50 years, the periscope was the only visual aid of the submarine until underwater television was installed aboard nuclear powered submarines. Today, the periscope is still used in submarines, but is commonly being replaced by a photonics mast. This device is similar to the periscope but uses digital equipment, including digital cameras, to observe above the water.

Disasters of Modern Submarines

K-141 Kursk was a nuclear-powered cruise missile submarine of the Russian Navy lost with all hands when it sank in the Barents Sea on 12 August 2000. At 154 m long and four stories high, it was the largest attack submarine ever built. The outer hull made of high-nickel, high-chrome content stainless steel of 8.5 millimetres thick had exceptionally good resistance to corrosion and a weak magnetic signature that helped with the prevention of being detected by magnetic anomaly detector systems.

On 12 August 2000, at 11:28 local time, there was an explosion while preparing to fire. The only credible report to date is that this was due to the failure and explosion of one of the Kursk's hydrogen peroxide-fuelled (HTP) torpedoes. It is believed HTP, a form of highly concentrated hydrogen peroxide used as propellant for the torpedo, seeped through rust in the torpedo casing. Due to a leaking weld in the torpedo's fuel system, high test peroxide, a form of highly concentrated hydrogen peroxide used as an oxidiser for the torpedo's engine, escaped into the torpedo casing where it catalytically decomposed on the metals and oxides present there, yielding steam and oxygen. The resulting overpressure ruptured the kerosene fuel tank thus causing the first explosion. According to maintenance records, dummy torpedoes, manufactured in the 1990s, had never had their welds checked. Such checks were considered unnecessary as the torpedoes did not carry warheads. The explosive reaction of 1.5 tons of concentrated hydrogen peroxide and 500 kg of kerosene blew off the external torpedo tube cover and the internal tube door. Two minutes and fifteen seconds after the initial eruption, a much larger explosion took place on the submarine combined with rising temperatures due to the initial explosion and caused other torpedoes to explode. The second explo-

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Figure 20.9. The wrecked hull of submarine Kursk moved from water

sion was equivalent to 2–3 tons of TNT, or about 5-7 torpedo warheads. The second explosion ripped a 2-square-metre hole in the hull of the craft (Figure 20.9), which was designed to withstand the depths of 1,000 metres and also ripped open the third and fourth compartments. Water poured into these compartments at 90,000 litres per second killing everything inside. The fifth compartment contained two nuclear reactors of the ship, encased in 13 centimetres of steel and resiliently mounted to absorb shocks in excess of 50g. The bulkheads of the fifth compartment withstood the explosion allowing two reactors to shut down automatically and preventing nuclear meltdown or contamination. The submarine sank in relatively shallow water, bottoming at 108 metres about 135 kilometres off Severomorsk.

A Tunnel for Automobile and Truck Transport

The method of tunnel construction depends on such factors as ground conditions, ground water conditions, the length and diameter of the tunnel drive, the depth of the tunnel, the logistics of supporting tunnel excavation, the final use and shape of the tunnel and appropriate risk management.

There are three basic types of tunnel construction in common use:

- cut and cover tunnels constructed in a shallow trench and then covered over;
- bored tunnels constructed in situ without removing the ground above; they are usually of a circular or horseshoe cross-section;
- immersed tube tunnels sunk into a body of water and sit on, or are buried just under its bed.

Cut-and-cover is a simple method of construction for shallow tunnels where a trench is excavated and roofed over with an overhead support system strong enough to carry the load of what is to be built above the tunnel. Shallow tunnels are often of the cut-and-cover type (if under water, of the immersed-tube type), while deep tunnels are excavated often using a tunnelling shield. Large cut-and-cover boxes are often used for underground metro stations.

An *open building pit* consists of a horizontal and vertical boundary that keeps groundwater and soil out of the pit. The most important difference with cut-and-cover is that the open building pit is muted after tunnel construction, and no roof is placed.

Clay-*kicking* is the process of manually digging tunnels in strong clay-based soil structures. Clay-kicking is relatively silent and hence did not harm soft clay based structures. The clay-kicker lies on a plank at a 45-degree angle away from the working face and inserts a tool with a cup-like rounded end with its feet. Turning the tool with its hands, it extracts a section of soil which is then placed on the waste extract.

It is essential that any *tunnel project starts* with comprehensive investigation into ground conditions. The results of investigation will allow a proper choice of machinery and methods for excavation and ground support as well as will reduce the risk of encountering unforeseen ground conditions. At early stages, the horizontal and vertical alignment will be optimized to make use of the best ground and water conditions.

New York City became a major focus of activity in the construction of transportation structures. With a rapid rise in automobile and truck transport after the turn of the century, Hudson River ferries were carrying 30 million vehicles each year between New York and New Jersey. Manhattan lies on a great river. In New York's case, it is the Hudson and its branches. To connect Manhattan Island to the adjacent shores, the construction of more great bridges was required. In 1905, Alfred Nobel drove rail tunnels under the Hudson River. Two tubes were over seven meters in diameter and 1600 meter long. Under the shorter 1200 meter East River section, four tubes were built. This was completed in 1909. Engineering problems in building tunnels under the Hudson were daunting. The tunnel had to be built through almost liquid mud and silt necessitating careful control over compressed air pressure in the construction. These tunnels provided for rail traffic but did nothing for burgeoning vehicle traffic. Since it would be easier and cheaper to build a bridge than a tunnel, the bridge was initially thought to be a feasible solution. However, there were drawbacks to choosing a bridge crossing. The Hudson River bridge would require a minimum clearance of 60 m for ships to travel to and from the Hudson River ports. Since the Manhattan side of the Hudson did not meet 60 m elevation requirement for the bridge, the final decision on constructing the tunnel was made in 1913.

The construction of two vehicular tunnels was begun in 1920 following Clifford M. Holland‘s (1883–1924) design that included twin steel-tube tunnels just over 9 meters in diameter and 2819 meters long, 1670 meters of which were sub-aqueous.

a) b)

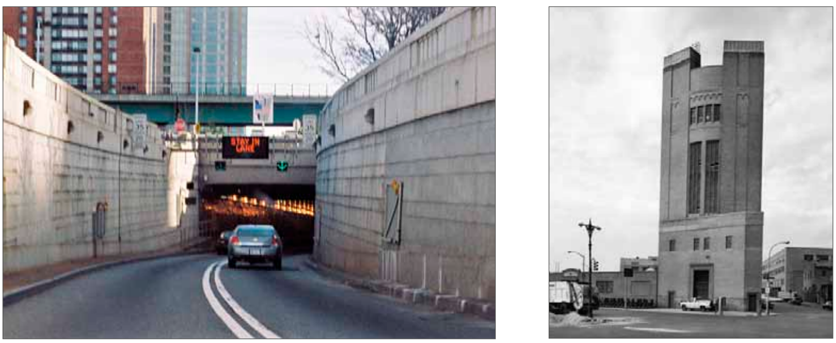

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Figure 20.10. Holland Tunnel Entrance (a) and Ventilation Building (b)

The tunnel tubes were about 4.5 meters apart each with two one-way lanes and a walkway. The design feature is that the ventilation system is the most interesting part (Figure 20.10). It had to be designed to handle air flow over the length of the tunnel and maintain safe levels of carbon monoxide. The design team developed a revolutionary two-duct system – the one that utilized one duct to draw in fresh air, and the other to suck out exhaust air – that was adopted eventually by vehicular tunnels worldwide. To facilitate the exchange of clean and dirty air, the team developed a system of ventilator fans and airshafts to circulate clear air throughout the length of the tunnel. This air is moved by 42 blowing fans and 42 exhaust fans – totalling 6,000 horsepower – arranged in four ventilation buildings. Only 56 out of the total 84 fans are in operation at all times; the other 28 fans are reserved for emergencies. It takes approximately 90 seconds to completely change air in the tunnel.

Tunnel construction workers followed two enormous, 240-ton hydraulically powered shields under the riverbed. Cast-iron shields weighed 400 tons, measured 9 m in diameter, were 4.8 m long and had a forward thrust of 6,000 tons. On a good day, the shield moved about 12 m. On bad days, they did not move at all. The construction of the tunnel began in 1920 and progressed over seven years.

Today, the Holland Tunnel carries approximately 100,000 vehicles per day between Jersey City and Lower Manhattan. In its nearly 80 years of operation, it has carried more than 1.5 billion vehicles.

A *tunnel boring machine* (TBM) is a machine used for excavating tunnels with a circular cross section through a variety of soil and rock strata (Figure 20.11). They can bore through hard rock, sand and almost anything in between. Tunnel diameters can range from a metre (done with micro-TBMs) to almost 16 metres to date. Some types of TBMs, bentonite slurry and earth-pressure balance machines have pressurised compartments at the front end allowing them to be used in difficult conditions below the water table. This pressurizes the ground ahead of the TBM cutter head to balance water pressure. Operators work in normal air pressure behind the pressurised compartment but may occasionally have to enter that compartment to renew or repair cutters. Nowadays, even larger machines exist.

The tunnel boring machine used for digging the 57 km Gotthard Base Tunnel in Switzerland was about 9 m in diameter (Figure 20.12). Up to now, the largest ever built TBM has been 14.87 m in diameter.

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Figure 20.11. The tunnel boring machine

a)

b)

Figure 20.12. The Gotthard Base Tunnel in Switzerland: a – the tunnel boring machine breaking through rocks; b – tunnel junction at a multifunction station

The Zhongnanshan Tunnel in China opened in January 2007 is the world‘s second longest *highway* tunnel and the longest Asia‘s road tunnel making 18 km. The tunnel crosses under the Zhongnan Mountain and the maximum embedded depth of the tunnel is 1640 metres below the surface level. Each 6 m high, 10.92 m-wide tube carries two lanes. The distance between the centre lines of two bores is 30 m. There are also three ventilation shafts.

The Chunnel

In May 1994, the Channel Tunnel or "Chunnel" opened a high-speed rail link between Great Britain and the rest of the European Community. Following two centuries of discussion and two false starts, the construction of a trans-channel tunnel finally became a reality. From 1987 to 1994, the tunnel was completed. The Channel Tunnel is a 50.5 km undersea rail tunnel linking Folkestone, Kent, the United Kingdom with Coquelles, near Calais in northern France. At its lowest point, it is 75 m deep.

The idea for a Channel tunnel originated during the early part of the 19^{th} century in France during the reign of Napoleon Bonaparte. Then, a single, 33 km in length tunnel ventilated through stacks at intervals along the route was proposed. These vents would be secured by artificial islands. A British engineer, William Low suggested a two-tunnel cross-passage approach whereby the piston effect of the trains would replenish air and make the tunnel self-ventilating.

Understanding the geology of the Channel was crucial to any successful drive of a sub-seafloor tunnel. Geology generally consists of north-easterly dipping Cretaceous strata. Characteristics include continuous chalk on the cliffs on either side of the Channel containing no major faulting; four geological strata, marine

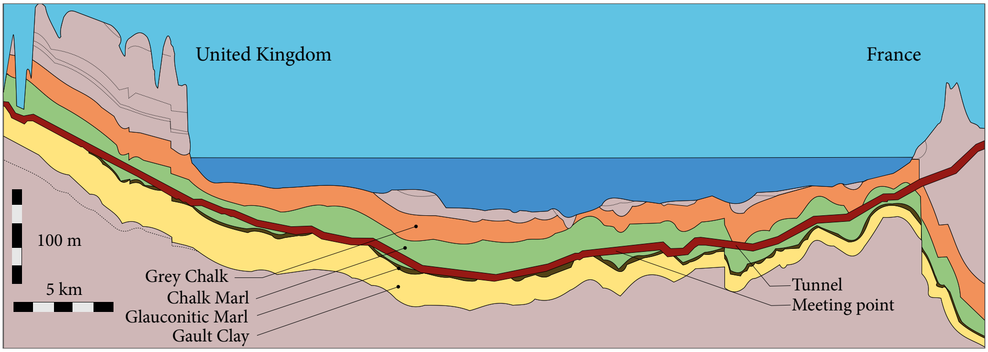

Figure 20.13. Geological profile along the tunnel as constructed

sediments laid down 90–100 million years ago; a sandy stratum, glauconitic marl, is in between the chalk marl and clay. A 25–30 m layer of chalk marl in the lower third of lower chalk appeared to present the best tunnelling medium. Chalk has the clay content of 30–40% providing impermeability to groundwater and yet relatively easy excavation with strength allowing minimal support. For most of its length, the tunnel bores through a chalk marl stratum (layer) (Figure 20.13).

The first effort to build a cross-Channel tunnel begun in 1880 and was abandoned for political more than financial reasons, completing 1,865 m of the English side and 1,840 meters on the French side of the Channel.

The next attempt was begun in 1974 and then unilaterally abandoned by the British government in 1975 for financial reasons. In 1979, the "Mouse-hole Project" was suggested. The concept was a single-track rail tunnel with a service tunnel, but without shuttle terminals. The British government took no interest in funding the project.

Tunnelling between England and France was a major engineering challenge. A serious risk with underwater tunnels is major water inflow due to water pressure from the sea above under weak ground conditions. The objective was to construct two rail tunnels of 7.6 m in diameter (A), 30 m apart and 50 km in length; a service tunnel of 4.8 m in diameter between two main tunnels (B); pairs of cross-passages of 3.3 m in diameter linking rail tunnels to the service tunnel at 375 m spacing (C); piston relief ducts of 2 m in diameter connecting rail tunnels at 250 m spacing (D). Relief

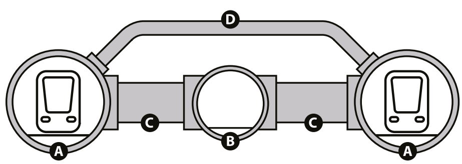

Figure 20.14. A cross-section of the tunnel, including a service tunnel between twin rail tunnels; shown linking the rail tunnels is a piston relief duct necessary to manage pressure changes due to the movement of trains

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Figure 20.15. A car entering a shuttle wagon at the terminal

ducts are necessary to manage pressure changes due to the movement of trains. The service tunnel always preceded the main tunnels by at least 1 km to ascertain the ground conditions (Figure 20.14).

All tunnel services run on electricity shared equally from English and French sources. Power is delivered to locomotives via an overhead line (catenaries) at 25 kV 50 Hz. Diesel locomotives are used for rescue and shunting work. Services offered by the tunnel include Eurotunnel Shuttle roll-on roll-off shuttle service for road vehicles, Eurostar passenger trains and freight trains.

The Chunnel carries two types of high-speed trains – the *Eurostar,* a passenger service only and *Le Shuttle* that carries cars, trucks and buses together with as many as 800 passengers (Figure 20.15). The trains are capable of running the Chunnel in 30 minutes at speeds over 130 km/h. Traffic is computer monitored. High-speed trains generate enough heat to raise air temperature to 50 °C. Water chilled to 3.3 °C is pumped through 480 km of 61 cm diameter pipes.

In 1996, approximately 1,000 passengers became trapped in the Channel Tunnel when two British trains on continent – bound *Eurostar* service broke down owing to failures of electronic circuits caused by snow and ice being deposited and then melting on the circuit boards. In 2007, an electrical failure lasting six hours caused passengers to be trapped in the tunnel on a Eurotunnel shuttle crossing. In 2009, during snowfall, five *Eurostar* trains failed inside the tunnel thus trapping 2,000 passengers. Snow had evaded the winterisation shields of the train, and transition from cold air outside to the tunnel's warm atmosphere had melted snow resulting in electrical failures. The occasion was the first time that a *Eurostar* train was evacuated inside the tunnel. Each wagon has fire detection and extinguishing system with sensing ions, ultraviolet radiation, smoke and gases that can trigger as flame retardant gas to quench a fire. The total cross-tunnel passenger traffic volumes made about 17.0 million in 2010 and freight transportation volumes about 17.7 million tons in 2011.

Kaunas railway tunnel is one of the two tunnels existing in Lithuania and the only railway tunnel operating in the Baltic States. It connects Vilnius and Kaunas. The length of the tunnel is 1.285 km, height – 6.6 metres and width – 8.8 metres. Kaunas railway tunnel was built from 1859 till 1861. The tunnel was dug from both sides in a 30-metre-high hill. The diameter of the tunnel has the shape of a horseshoe. The

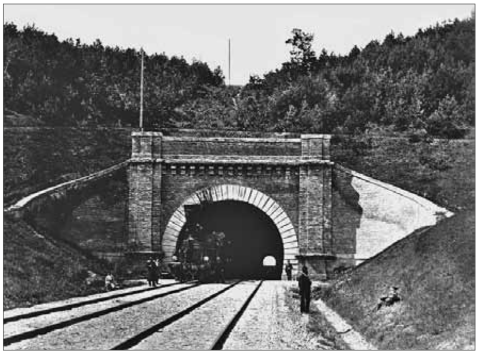

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Figure 20.16. Paneriai railway tunnel (Vilnius)

upper part was made of bricks (stonework), whereas the lower part, used as a basement, of granite stones. The tunnel was reconstructed between 2008 and 2009. Reconstruction works included a fire alarm system. The powerful fire extinguishing system was installed during the reconstruction of the tunnel. The pipes built underneath the tunnel are immediately filled with 35 cubic metres of water which is then sprayed via special sprinklers located along the length of the tunnel. Paneriai railway tunnel was built in 1861 (Figure 20.16). The length of the tunnel is 427 metres, width – 8 metres and height 6.4 metres. In 1960 railway track was dismantled and tunnel was closed.

The Automobile

By definition an automobile or car is a wheeled vehicle that carries its own motor and transports passengers. In 1769, the very first self-propelled road vehicle was a military tractor invented by a French engineer and mechanic Nicolas Joseph Cugnot. His vehicle was used by the French Army to haul artillery at a whopping speed of 4.5 kph on only three wheels. The vehicle had to stop every ten to fifteen minutes to build up steam power. The following year (1770), Cugnot built a steam-powered tricycle that carried four passengers (Figure 20.17). The first automobiles can be characterized as small carriages powered by small engines, using pneumatic tires and built by bicycle mechanics.

Engineering had already produced the diesel and Otto Cycle (1876) engines. Karl Benz, in 1886, made the first commercially feasible automobile. The 1891 vehicle of this firm placed the engine in the front of the chassis with the crankshaft parallel to the long axis of the car. This allowed the use of larger, longer engines. Daimler and Benz invented highly successful and practical gasoline-powered vehicles that ushered in the age of modern automobiles. Daimler and Benz invented cars that looked and worked like those we use today.

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Figure 20.17. The crash of Nicolas J. Cugnot's steam-powered car into a stone wall

Wilhelm Maybach invented the carburetor. The first car to be produced in significant volume was the Oldsmobile (1901–1906) resembling more a horseless carriage than vehicles. Steering wheels replaced the tiller after 1901 and were generally placed on the left side. Other improvements rapidly followed in water-cooled, force-feed lubricated engines, magneto ignitions, shock absorbers, bumpers, acetylene headlamps and vehicle tops.

The 1885 Daimler-Maybach engine was small, lightweight, fast, used a gasoline-injected carburetor and had a vertical cylinder. The size, speed and efficiency of the engine allowed for a revolution in car design. On 1886, Daimler took a stagecoach and adapted it to hold his engine, thereby designing the world's first four-wheeled automobile. Daimler is considered the first inventor to have invented a practical internal-combustion engine.

1908–1927 Ford Model T – the most widely produced and available 4-seater car of the era. It used planetary transmission and had a pedal-based control system. Ford T was proclaimed as the most influential car of the 20th century (Figure 20.18).

On the race tracks of the world names such as France's Bugatti, Germany's Auto Union and Mercedes Benz became legends in the 1930s. In 1934, Dr. Ferdinand Porsche developed a small "people's car" (Volkswagen). The 725 kg vehicle had a flat air-cooled engine in the rear thus producing 36 hp. Over 20 million of those have been built making it a mid-century version of Ford's earlier success.

Chrysler featured four-wheel hydraulic brakes in 1925. Chrysler made the first American aerodynamically designed mass-produced automobile in 1934. This was the Airflow. The automobile had to move efficiently through the air yet stay firmly on the ground. The Airflow did both due in large part to its heavy body and frame. Still, it could cruise at 130 kph in relative comfort. The modern V-8 engine was introduced by Cadillac in 1914. The first practical low-priced V-8 was the 1932 Ford. Chevrolet's 1919-1923 experiment with an air-cooled engine was the last attempt by an American manufacturer to pioneer a major change in engine design.

Figure 20.18. Ford Model T (1915). Museum of Energy and Technology, Vilnius

The pre-war part of the classic era began in 1930, and ended at 1946 (Figure 20.19). It was in this period that integrated fenders and fully closed bodies began to dominate.The new saloon/sedan body style even incorporating a trunk or boot at the rear for storage. By the 1930s, most of the mechanical technology used in today's automobiles had been invented, or "re-invented". For example, front-wheel

a) b)

Figure 20.19. Pre-war classic automobiles: a – Cadillac La Sale (1928, USA), b – Opel (1934–1937, Germany). Museum of Energy and Technology, Vilnius

drive was re-introduced in 1934. The independent suspension was originally conceived in 1873, but not put in production until appearing on the Mercedes-Benz 380 in 1933.

European imports met the demand for fuel economy while quality in U.S. cars fell. Mass production gave advantages to large companies due to economies of scale. By 1970, European and Japanese manufacturers were producing technologically advanced, well-built, fuel efficient, lower-priced cars.

During World War II, gasoline was rationed and in short supply. In Great Britain, United States and Germany, large numbers of gas generators were constructed or improvised to convert wood and coal into fuel for vehicles. A wood gas generator is a gasification unit that converts timber or charcoal into wood gas. Synthesis gas (syngas) consisting of atmospheric nitrogen, carbon monoxide, hydrogen, traces of methane and other gases, which – after cooling and filtering – can be used for powering an internal combustion engine or for other purposes (Figure 20.20).

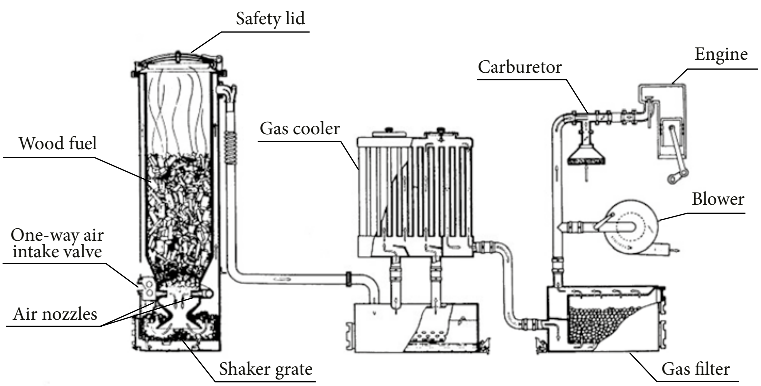

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Figure 20.20. Scheme for a wood gas car

Figure 20.21. A wood gasifier on the Ford truck converted into a tractor

Syngas has less than half the energy density of natural gas. Using modern technologies allow converting 5 kg of wood into 1 litre of fuel. Wood gas generators often use wood; however, charcoal can also be employed as fuel, as it is denser and produces cleaner gas without tarry volatiles and the excessive water content of wood.

Wood gas generators have a number of advantages over the use of petroleum fuels: they can be applied for running internal combustion engines using wood, a renewable resource, have a closed carbon cycle, contribute less to global warming, are sustainable in nature and far produce cleaner burning than a gasoline-powered engine (Figure 20.21). The disadvantages of wood gas generators cover a large specific size, a relatively slow starting speed (time for heating the initially cold batch of wood to the necessary temperature level), stop operation out is difficult as residual heat still produces gas, primary combustible fuel-gas produced during gasification is carbon monoxide, continuous exposure to carbon-monoxide can be fatal to humans even in small to moderate concentrations, the humidity of wood (usually 15 to 20%) condenses during the gas cooling and filtering procedure and yields a liquid, which needs specific waste water treatment that requires about 25 to 35% of the created wood gas energy.

Innovations such as radial tires, independent suspensions, disc brakes, fuel injection and front-wheel drive, tubeless tire, alternator and the torsion-bar suspension were gradually introduced in all automobiles. The Studebaker US6 2.5-ton

a) b)

 Photo by A. V. Valiulis

Figure 20.22. The Studebaker US6 truck (a) and Jeep Willys MB in Museum of Energy and Technology, Vilnius (b)

trucks were manufactured during World War II, produced in the United States for the period 1941–1945 and in the Soviet Union starting from 1942 (Figure 20.22, a). The Willys MB U.S. Army Jeep (manufactured from 1941 to 1945) (Figure 20.22, b) later evolved into the civilian Jeep (Figure 20.22, b).

After the war, the USSR acquired an entire Opel manufacturing line from Germany. A factory in Moscow started in 1947 to manufacture an automobile called *Moskvitch 400* based on the Opel Kadett (Figure 20.23, a). Further models were developed by Soviet engineers (Figure 20.23, b).

a) b)

Figure 20.23. An automobile called *Moskvitch*: a – *Moskvitch 400* (1947–1956), b – *Moskvitch 407* (1958–1963). Museum of Energy and Technology, Vilnius

The car Pobeda first tests were done in 1943 at Gorky Car Plant, when victory in World War II began to seem likely. The first prototype was ready on November 6, 1944. The Pobeda was the first truly "mass market" automobile which appeared in Lithuania after the World War II. Weighing 1,460 kg, the Pobeda has 2.1 litre side-valve straight-4 engine producing 50 hp and top speed of 105 km/h (Figure 20.24).

Similarly, the Europeans and Japanese produced turbochargers, stratified, multi-valve engine systems and rotary engines. It is estimated that over 100,000 patents created the modern automobile. In 1980, Japan replaced the U.S. as the largest car manufacturer in the world. The average weight per car dropped from 1750 kg in 1975 to 1230 kg in 1985. Computer-controlled systems are becoming the rule in engines and suspensions. The potential of alternate fuels such as hydrogen is promising.

Figure 20.24. Pobeda (Victory) car (USSR). Museum of Energy and Technology, Vilnius

On the positive side, automobiles are safer, more fuel efficient, better built and less polluting. The future of the automobile is rosy. Unlike public transportation systems, it is convenient. In many urban areas, it is safer. Goods and services have expanded out of central urban centres in response to this greater consumer mobility. Modern transportation infrastructures reflect this decentralization that is both product of and succour for the automobile. „Super trains" compete with aircrafts for medium distance transit. Still, the car looks to have molded itself into modern society for years to come.

The Electric Vehicle

An electric car is an automobile that is propelled by one electric motor or more and uses electrical energy stored in batteries or another energy storage device. Electric motors give electric cars instant torque creating strong and smooth acceleration.

An electric vehicle is a hybrid technology, which is a combined hybrid vehicle that can be propelled by gasoline and/or electric power. Energy-rich hydrocarbon fuel is replaced by storage batteries dispensing electrical current for the motor (Figure 20.25). All automobiles use storage batteries, but these units are relatively small devices that provide short-lived bursts of electrical current to start the combustion engine. A successful design of a modern electric vehicle is at once both institutional and structural.

The history of the electrical vehicle is perhaps as long, if not longer, than that of a more traditional automobile. In the late 19th century, their development paralleled each other with the electrical vehicle was finally losing out in the early 20th century.

In 1828, Ányos Jedlik, a Hungarian who invented the early type of an electric motor, created a tiny model of a car powered by his new motor. In 1834, Thomas Davenport, the inventor of the first American DC electrical motor, installed his motor in a small model car he operated on a short circular electrified track. In 1838, Robert Davidson built an electric locomotive that attained a speed of 6 km/h. In England, a patent for the use of rail tracks as conductors of electric current was granted in 1840. Similar American patents were issued in 1847.

Figure 20.25. Hybrid car (Toyota) cutaway model showing engine connection and location (2013)

Electric vehicles require batteries and still do require frequent recharging. The ease of recharging in the late 19th century and early 20th century was difficult at best in the cities and totally out of the question in rural areas. The early electrics competed with crude gasoline designs with a host of inconvenient problems of their own not the least of which was the ease of starting. It is ironic that the starting problem of the combustion engine was solved by an electrical engineer Thomas F. Kettering who designed the electric starter that was marketed on the Cadillac in 1912. Overnight, the gasoline engined automobile became both a man's and woman's car.

Despite their relatively slow speed, electric vehicles had a number of advantages over their early-1900s competitors. They did not have vibration, smell and noise associated with gasoline cars. They did not require gear changes, which was the most difficult part of driving for gasoline cars. Electric cars found popularity among well-heeled customers who used them in the city where their limited range was less of a disadvantage. The cars were also preferred because they did not require a manual effort to start, as did gasoline cars that featured a hand crank to start the engine.

The electric vehicle had appealed to urbanites, particularly women, because of the simplicity of operation and starting, as well as due to recharging problem (Figure 20.26). Nonetheless, two other elements – cost and distance – played against the mass acceptance of the electric car. The cost factor, as seen in the last days of mass-produced electric cars, was overcome. Historically, there was a small market for the electric car. From 1910 to 1920, the electric vehicle overcame battery problems, drove down costs and centralized recharging to no avail. The gasoline automobile won the day by simply becoming an urban and rural vehicle. A niche market remained in the cities for the electric, but niche markets are characterized by low production and high cost items which electric vehicles were doomed to remain. Beyond the distance problem – no electric vehicle, even today, will conveniently go beyond a range of a hundred kilometres. Today's electric vehicle makes use of better motors, computer controls, lead-acid batteries and simplified, faster recharging and boost systems. In 1994, a battery-powered pick-up truck covered 1420 km in 24 hours using recharges of only 18 minutes at a time. Before the pre-eminence of internal combustion engines, electric automobiles held many speed and distance records. Among the most notable of these, was the breaking of 100 km/h in 1899. Before the 1920s, electric automobiles were competing with petroleum-fuelled cars for the urban use of a quality service car.

Figure 20.26. German electric car with the chauffeur (1904)

Electric vehicles have several advantages over those with internal combustion engines :

- *Energy efficient*. Electric vehicles convert about 59–62% of electrical energy from the grid to power at the wheels-conventional gasoline vehicles only convert about 17–21% of the energy stored in gasoline to power at the wheels.
- *Environmentally friendly*. EVs emit no tailpipe pollutants, although the power plant producing electricity may emit them. Electricity from nuclear-, hydro-, solar-, or wind-powered plants causes no air pollutants.
- *Performance benefits*. Electric motors provide quiet, smooth operation and stronger acceleration and require less maintenance than internal combustion engine.
- *Reduce energy dependence*. Electricity is a domestic energy source.

Electric vehicles face significant battery-related challenges:

- *Driving range*. Most electric vehicles can only go about 160–320 km before recharging – gasoline vehicles can go over 500 km before refuelling.
- *Recharge time*. Fully recharging the battery pack can take 4 to 8 hours. Even a "quick charge" to 80% capacity can take 30 min.
- *Battery cost*: Large battery packs are expensive and may need to be replaced one or more times.
- *Bulk & weight*: Battery packs are heavy and take up considerable vehicle space.

If electric vehicles do become a larger segment of the total vehicle population, they will cause lead-acid battery problems. The environmental consequences of producing and reprocessing large numbers of these batteries are of some concern. It is projected that a 1998 electric car model releases 60 times more lead per kilogram relative to a conventional automobile burning leaded gasoline. Lead releases are not from electric vehicle operation but in the manufacture (smelting) and disposal of used batteries. The picture is not much improved by the replacement of the lead-acid battery with nickel-cadmium or nickel metal hydride designs. The electrolytes of these designs are even more toxic. Sodium-sulphur or lithium-polymer technologies are attractive in this regard but less far along in development.

While most current highway-speed electric vehicle designs focus on lithium-ion and other lithium-based variants, a variety of alternative batteries can also be used. Lithium based batteries are often chosen for their high power and energy density but have a limited shelf-life and cycle lifetime that can significantly increase the running costs of the vehicle. Variants such as *Lithium iron phosphate* and *Lithium-titanate* attempt to solve durability issues with traditional *lithium-ion* batteries. Other battery technologies include:

- *lead acid* batteries are still the most used form of power for most of the electric vehicles employed today. The initial construction costs are significantly lower than for other battery types, and while power output to weight is poorer

than other designs, range and power can be easily added by increasing the number of batteries;
- *NiCd* – largely superseded by NiMH;
- *Nickel metal hydride* (NiMH);
- *Nickel iron battery* – known for its comparatively long lifetime and low power density

Several battery technologies such as *zinc-air battery* or *molten salt battery* are also in development.

Zinc-bromine flow batteries or *Vanadium redox* batteries can be refilled, instead of recharged, for saving time. The depleted electrolyte can be recharged at the point of exchange or taken away to a remote station.

Different sources and reports (2011) have found that *hybrid electric vehicles*, *plug-in hybrids* and all-electric cars generate more carbon emissions during their production than the current conventional vehicles, but still have a lower overall carbon footprint over the full life cycle. The initial higher carbon footprint is due mainly to battery production. As an example, the study estimated that 43 percent of production emissions for a mid-size electric car were generated from battery production.

Range and refuelling time. Most cars with internal combustion engines can be considered to have an indefinite range, as they can be refuelled very quickly almost anywhere. Electric cars often have less maximum range on one charge than those powered by fossil fuels, and therefore can take considerable time for recharging. The Tesla Roadster can travel 394 km per charge. The Roadster can be fully recharged in about 3.5 hours from a 220-volt, 70-amp outlet installed at home. However, using the European standard 220-volt, 16-amp outlet charging will take more than 15 hours.

Automakers can extend a short range of electric vehicles by building them with battery switch technology, which along with a 160 km driving range, will be able to go to a battery switch station and switch a depleted battery with a fully charged one in seconds. The process is cleaner and faster than filling a tank with gasoline and the driver remains in the car the entire time; however, because of high investment cost, its economics are unclear.

Another way is the installation of direct current *Fast Charging* stations with high-speed charging capability from three-phase industrial outlets so that consumers could recharge the 170 km battery of their electric vehicle to 80% in about 30 minutes (Figure 20.27).

Electric vehicles can also use a direct motor-to-wheel configuration that increases the amount of available power. Having multiple motors connected directly to the wheels allows for each of those to be used for both propulsion and as braking systems, thereby increasing traction. In some cases, the motor can be housed

Figure 20.27. A fast charging station of electric cars in Vilnius (2014)

directly in the wheel. When not fitted with an axle, differential or transmission, electric vehicles have less rotational inertia of drivetrain (the drivetrain of a motor vehicle is the group of components that deliver power to the driving wheels). On the other housing, the motor within the wheel can increase the unsprung weight of the wheel, which can have an adverse effect on handling the vehicle.

On the other hand, electric motors are more efficient in converting stored energy into driving a vehicle, and electric drive vehicles do not consume energy while at rest or coasting. Some of the energy lost when braking is captured and reused through regenerative braking, which captures as much as the one fifth of the energy normally lost during braking. Typically, conventional gasoline engines effectively use only 15% of fuel energy content to move the vehicle or to power accessories, and diesel engines can reach on-board efficiencies of 20% while electric drive vehicles have on-board the efficiency of around 80%.

The question that remains for the electric vehicle is whether or not its time has really come. The future of battery electric vehicles depends primarily upon the cost and availability of batteries with high specific energy, power density and long life, as all other aspects such as motors, motor controllers and chargers are fairly mature and cost-competitive with internal combustion engine components. Solar cars are electric cars that derive most or all of their electricity from built in solar panels. After the 2005 World Solar Challenge established that solar race cars could exceed highway speeds, vehicles after slight modification could be used for transportation.

Trams and Trolleybuses

Street tramways were popular as public transport for passengers in the 19th and early 20th centuries. Trams originated in the USA. The first horse-drawn tram was built in New York in 1831 and was opened in 1832. The horse-drawn tram in Lithuania (Vilnius and Kaunas) operated in 1893–1925 (Figure 20.28). The first steam tram in England was launched in 1873. Early steam trams had the engine and boiler mounted within passenger cars. Dirt, smoke and heat were in the vicinity of passengers, and soon the engine and boiler were mounted on a separate frame. Such trams were really like miniature trains, with a small locomotive and one or (usually) more trailers. Electric

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Figure 20.28. The horse-drawn tram in Kaunas 1892 (closed in 1929)

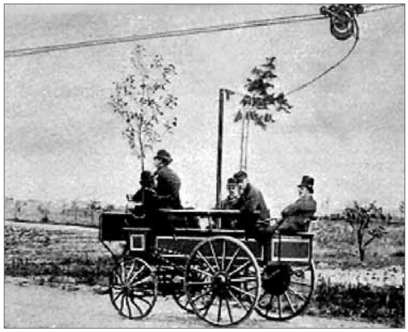

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Figure 20.29. The world's first trolleybus (1882)

traction was the optimum solution to tramways but, in most cases, had to wait until a suitable source of supply was available. The first such vehicle, the Electromote, was made in 1882 in Halensee, Germany (Figure 20.29). The main disadvantage of street tramways was that halts could only be made by the car pulling up in the middle of the roadway with consequent danger to ascending and descending passengers from being struck by other vehicles in spite of preventive legislation. The first tram line in the current Lithuanian territory was set in operation in Klaipėda in 1904. Thirty years later, it was replaced by busses.

A trolleybus was an alternative. This means of transport could collect current from the same overhead wires and yet, running on wheels as opposed to rails, could draw up close to the kerb. Developed in the USA before the First World War and having pneumatic tyres, the trolleybus became popular city transport. The trolleybus has another disadvantage – one trolleybus could not overtake another on the same line. Trolleybuses are used in towns where streets are wide and there is relatively little other conventional traffic. The first trolleybus network in Lithuania was opened in Vilnius in 1956 (Figure 20.30). Later, low-floor trolleybuses (Figure 20.31) appeared.

With permission from JSC *Vilniaus viešasis transportas* (photo by E. Sisko)

Figure 20.30. The first trolleybuses in Vilnius, (1955)

Figure 20.31. A modern low-floor trolleybus in Vilnius (2014)

Buses

Horse buses were chiefly used from the 1820s until the early 20th century. A horse-drawn 'omnibus' was first put into service in London in 1829 (Figure 20.32). Horse buses in London were replaced by the buses equipped with the internal combustion engine only in 1911.

The first known public bus line moving passengers within Paris in a carriage with many seats was launched by Blaise Pascal in 1662. Services ceased after 15 years and no further similar means of transport were known until the 1820s. Horse buses operated in many cities during the later part of the 1800s and early 1900s. The last horse bus in London stopped operation in 1914. Early horse-drawn buses were a combination of a hackney carriage and a stagecoach. From the 1830s, steam powered buses existed (Figures 20.33, 20.34). The first internal combustion engine buses were developed along with the automobile (Figure 20.35).

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Figure 20.32. The first Omnibus

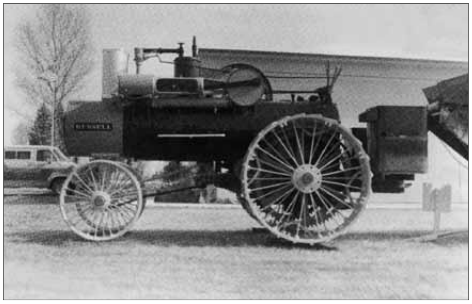

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Figure 20.33. A steam-powered vehicle

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Figure 20.34. French steam bus (1911)

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Figure 20.35. A double-decker bus of the 1906

Author Carsten Pietzsch. Licensed under CC0 1.0

Figure 20.36. A 30,7-meter long bus

Models expanded in the 20th century thus leading to the widespread introduction of the contemporary recognizable form of full size buses from the 1950s. Bus types include *single-decker bus*, *double-decker bus*, *triple-decker bus*, *articulated and bi-articulated bus* and *bus trailers* (Figure 20.36).

Smaller midi-buses have a lower capacity, and *open-top bus*es are typically used for leisure purposes. Many new buses have an increasing shift to *low-floor buses* for easier accessibility. *Coaches* are designed for longer-distance travel and fitted with seat-belts, toilets, audio-visual entertainment systems and can operate at higher speeds. Coaches may be single- or double-deckers, articulated and often include a separate luggage compartment under the passenger floor. *Guided buses* are fitted with technology to allow them to run in designated guide-ways. The diesel engine is the most common power source since the 1920s.

Lithuania had facilities for bus production. Kaunas Bus Factory was the one operating in Lithuanian SSR. The factory produced more than 12,000 buses for the period from 1956 to 1979 (Figure 20.37) and was established in an old Ford workshop in the 1930s.

Author Kirill Borisenko. Licensed under CC BY-SA 3.0

Figure 20.37. A bus for 19 passengers produced in Kaunas Bus Factory (1956)

Bicycles

A bicycle is a human-powered, pedal-driven, single-track vehicle, having two wheels attached to a frame, one behind the other. The oldest forerunner of the bicycle – *a celerifere* (running machine) was riding in Paris 1792 by pushing against the ground with feet. This heavy wooden machine was without a handlebar and had just an armrest. Bicycles were introduced in Europe in the 19th century (Figure 20.38). They are the principal means of transportation in many regions of the world. The first chain-driven model was developed around 1885. The invention has had an enormous effect on society in terms of culture and of advancing modern industrial methods. Several components that eventually played a key role in the development of the automobile were invented for the bicycle and included pneumatic tires, ball bearings and chain-driven sprockets.

A penny-farthing bicycle, a type of the bicycle with a large front wheel and a much smaller rear wheel, was popular until the development of the safety bicycle in the 1880s (Figure 20.39). Larger front wheels, up to 1.5 m in diameter, enabled higher speeds on bicycles limited to direct drive.

Materials used in bicycles have followed a combination of high strength and low weight. Since the late 1930s, alloy steels have been used for frame and fork tubes. In the 1980s, aluminum alloy frames and other components became popular due to their light weight, and most mid-range bikes are now principally the aluminum alloy of some kind. More expensive bikes use carbon fibre due to its significantly lighter weight and profiling ability, allowing designers to make a bike both stiff and compliant by manipulating the lay-up. Other exotic frame materials include titanium and advanced alloys.

The great majority of today's bicycles have a frame with upright seating. These upright bicycles almost always feature a frame consisting of two triangles: the front triangle and the rear triangle (Figure 20.40). The front triangle consists of the head tube, top tube, down tube and seat tube. The rear triangle

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Figure 20.38. A velocipede (1869)

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Figure 20.39. A penny-farthing bicycle

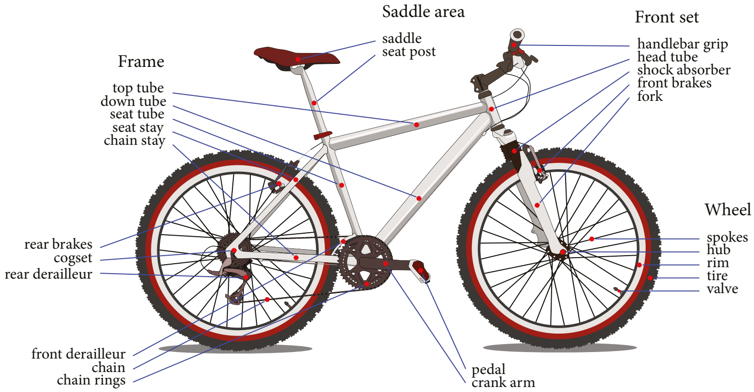

Figure 20.40. Diagram of a bicycle

includes the seat tube, paired chain stays and seat stays. The chain stays run parallel to the chain connecting the bottom bracket to the rear dropout where the axle for the rear wheel is held. Most bicycles use a chain to transmit power to the rear wheel. A very small number of bicycles use a shaft drive to transmit power, or special belts. A variable gear ratio helps a cyclist with maintaining optimum pedalling speed while covering varied terrain. Some, mainly utility, bicycles use hub gears with between 3 and 14 ratios. Modern bicycle brakes may be *rim brakes*, in which friction pads are compressed against wheel rims, *internal hub brakes*, in which friction pads are contained within wheel hubs or *disc brakes*, with a separate rotor for braking. Bicycle *suspension* serves two purposes: to keep wheels in continuous contact with the ground thus improving control, and to isolate the rider and luggage from jarring due to rough surfaces thus improving comfort. Road bicycles use the tires of 18 to 25 millimetres wide, most often completely smooth, or slick. Off-road tires are usually between 38 and 64 mm wide and have treads for gripping under muddy conditions.

Special bikes have more unusual abilities such as the tandem (Figure 20.41). It can carry two riders: one in the front and one in the back. Each rider has separate handlebars and pedals, but there is only one chain. The riders can pedal together for extra speed.

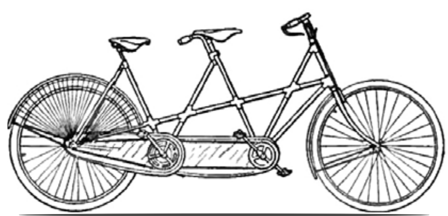

Figure 20.41. A tandem bicycle

Early bicycles were adapted to the fashionable elites. Subsequently, bicycles helped with creating new kinds of businesses such as bicycle messengers, travelling seamstresses, riding academies and racing rinks. The bicycle allows people travelling into the country, since it is three times as energy efficient as walking and three to four times as fast. There are special bicycles designed as cargo, physical fitness, track racing, artistic cycling bicycles, etc. A trailer bike is a bicycle designed to carry one or more children in positions that closely resemble that of the bicycle rider. A trailer bike was patented in Canada and has been sold from 1990s (Figure 20.42). A number of cities around the world have implemented schemes known as *bicycle sharing systems* (Figure 20.43). These public transport systems offer an alternative to motorized traffic to help in reducing congestion and pollution.

Figure 20.42. A trailer bike

Figure 20.43. A bike-sharing station in Vilnius (2014)

Bicycle Production in Lithuania

Between World Wars, the most popular bicycles in Lithuania were those made in England and Germany. The first Lithuanian bicycles "Ereliukas" and "Kregždutė" were made in Šiauliai Bicycle and Motor Factory *Vairas* in 1951 and were suitable both for girls and boys. Bicycles "Venta" and "Rambynas" produced from 1983 had a demountable frame. After the re-organization of the factory and following the establishment of joint Lithuanian and German enterprise *Baltik-Vairas* in 1993, bicycles "Panther" have been manufactured. About 75% of the output is mountain, trekking and touring bicycles (Figure 20.44). The rest is simple trekking 1-speed bicycles that are still very popular among people.

Figure 20.44. Bicycles from *Baltik-Vairas* factory (2013)

Roads and Highways

Historically many roads were simply recognizable routes without any formal construction or maintenance. The world's oldest known paved road was laid in Egypt sometime between 2600 and 2200 B.C. Modern roads are normally smoothed, paved or otherwise prepared to allow easy travel. The road is a line of communication (travelled way) using a stabilized base other than rails open to public traffic, primarily for the use of road motor vehicles running on their own wheels. The road includes bridges, tunnels, supporting structures, junctions, crossings, interchanges and toll roads. Roads consist of one, or sometimes two, roadways each with one or more lanes and also any associated sidewalks and road verges.

A superhighway was a creation of the European militaristic regimes of Italy and Germany in the 1930s. The first autostrade opened in 1924 but was quickly surpassed in scale and number by the *autobahnen* of Germany. The superhighway in Europe and the United States had an explicit military purpose. In the US, the label of "National System of Interstate and Defense Highways" can be traced to 1935. The German Reichautobahn plan called for 4000 km of superhighways. By 1942, a total of 2096 km of autobahn had been built.

The German superhighway became the model for all succeeding designs. They commonly had a central divider or median, the use of landscaping, the use of feeders and exits that carefully merged with the main lanes. Another element that still causes drivers problems today was the unrelenting straightness of the route. Fatigue and boredom are a result of this sameness of the horizon and driving environment. The German system evolved into the E-system of highways in use throughout Europe today.

The US Interstate system exceeds 60,000 km with the five longest east-west routes. Each is in excess of 3200 km in length. Like their European predecessors, the "Interstates", as they are called, were like no other highways in America. Old-style highways had pavements 10 to 12.5 cm thick, whereas the Interstate pavements were 22.5 to 25 cm. All at-grade highway and railroad intersections were eliminated (Figure 20.45).

Highway curvature, elevation, sight distance and grade were standardized to specific design speeds making 112 kph for flat areas, 96 kph – for rolling terrain and 80 kph – for hills. Gradients were 3, 4 or 5% accordingly. The width of the minimum lane was 3.65 m. Today, commerce, commuters and travellers take the superhighway for granted. They have revolutionized interstate and, perhaps more importantly, inter-urban travel. The linkage of metropoles has had dramatic effects on the flow, distribution and volume of goods and services. Personal mobility and

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Figure 20.45. The Pan-American Highway

traffic safety have increased. The growth of suburbs has been a creation of the superhighway and the increased number of personal vehicles. Economic corridors have grown up around the superhighway. Less positive impacts of the superhighway are environmental. Noise and air pollution have increased with the spread of the superhighway. Likewise, opening up isolated areas have endangered indigenous species of plants and animals. By the same token, accessibility to parks and wildlife areas have allowed more people to visit what had been difficult to appreciate before the coming of superhighway systems.

The United States has the world's largest network of highways – approximately 6,430,366 km of highway within its borders. China's highway network is the second most extensive in the world with a total length of about 3.573 million km. China's expressway network is quickly expanding, stretching some 85,000 km at the end of 2011.

The longest international highway *Pan-American Highway* connects many countries in the Americas and is nearly 25,000 km long. The longest national highway (point to point) is the Trans-Canada Highway, which makes 7,821 km long. It runs east-west across southern Canada. The highway begins on the east coast in Newfoundland and reaches Vancouver on the Pacific Coast. The longest national highway (circuit) is Australia's Highway 1 at over 20,000 km. It runs almost the entire way around the continent's coastline. The busiest highway Highway 401 in Ontario, Canada, has volumes surpassing an average of 500,000 vehicles per day. The widest highway (maximum number of lanes) *Katy Freeway highway* in Houston, Texas, has a total of 26 lanes in some sections. The highest international highway the *Karakoram Highway*, between Pakistan and China, is at an altitude of 4,693 m.

Roads are the lifeblood of European trade and social utility (Figure 20.46). Despite the increasing focus on the use of other modalities like railway, shipping and all kinds of public transport, roads carry by far the majority of land freight transport and passenger traffic. Reducing travel times relative to city or town streets, modern highways with limited access and grade separation create increased opportunities for people to travel for business, trade or pleasure and also provide trade routes for goods. Modern highways reduce commute and other travel time, but additional road capacity can also create new induced traffic demand. If not accurately predicted at the

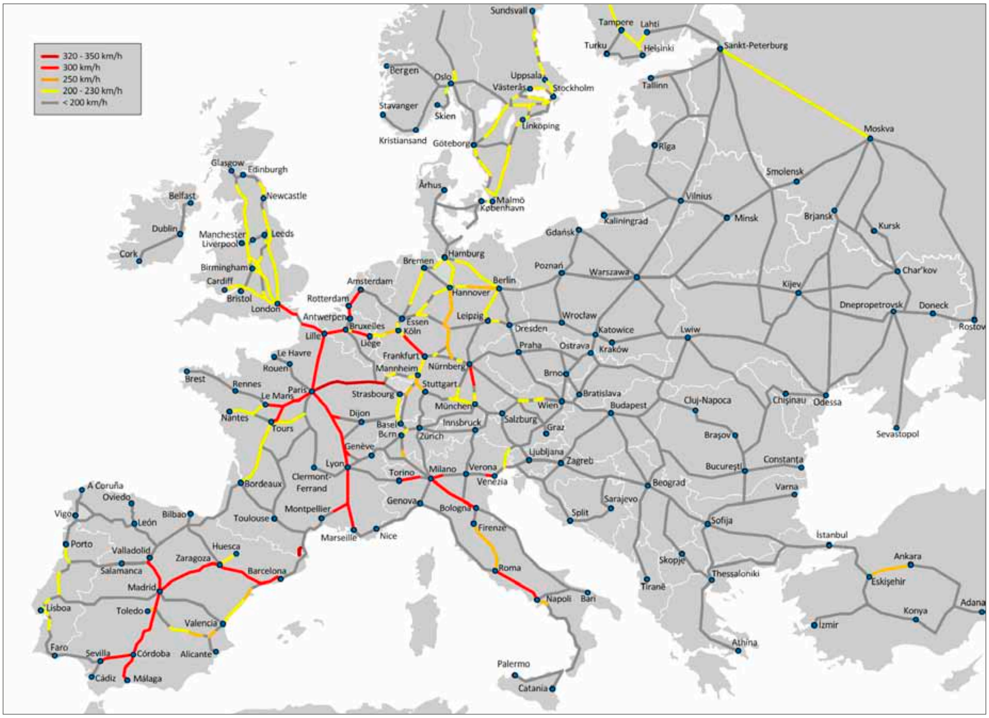


Figure 20.46. The approximate extent of the completed motorway network in Europe (2010)

planning stage, this extra traffic may lead to the new road becoming congested sooner than anticipated. Economics and society depend heavily on efficient roads. Building new roads or expanding square metres of asphalt might seem the obvious way to do this. However, the demand for 'traffic space' will always exceed the supply, traffic jams are a fact of life. In the European Union, (EU) 44% of all goods are moved by trucks over roads and 85% of all people are transported by cars, buses or coaches on roads.

Roads can be and are designed for a variety of lives (8-, 15-, 30-, and 60-year designs). Like all structures, roads deteriorate over time. Deterioration is primarily due to accumulated damage from vehicles; however, environmental effects such as frost heaves, thermal cracking and oxidation often contribute. It is empirically determined that the effective damage done to the road is roughly proportional to the fourth power of axle weight. A typical tractor-trailer weighing 36.287 t with 3.629 t on the steer axle and 16.329 t on both of the tandem axle groups is expected to do 7,800 times more damage than a passenger vehicle with 0.907 t on each axle. Potholes on roads are caused by rain damage and vehicle braking or related construction works.

Today, the road sector faces huge challenges, including ambitious demands such as better, quicker and cheaper production, construction and maintenance. To minimize the downtime of roads for maintenance activities, the overall quality of constructions has to be upgraded. Time slots available for repair and rehabilitation works are becoming tighter and tighter, which means that maintenance techniques have to be speeded up. Furthermore, environmental regulations with respect to air pollution and noise emissions by traffic and the use of natural raw materials are becoming more and more stringent.

New or future European road construction concepts are sown today. High priority will be attached to the environmental friendliness of road transport. New transport systems such as road trains combined with advanced traffic management systems provide efficient, smooth and low energy transport of goods. Zero emission vehicles with silent tyres, combined with new noise absorbing road surfaces will reduce air and noise pollution. Underground or covered roads will improve the aesthetic features of infrastructure and create space for new, non-transport related functions.

Special attention is paid to public security. As a result of this concept, suburban roads are transformed into multi-usable streets serving the safety of all kinds of the users of public space. An economic interest in road transport will be served by infrastructure that is reliable and available around the clock. New construction and maintenance techniques have been introduced to upgrade and rehabilitate the old (existing) network and to build new roads to complete the networks fast and cost-effectively. Roads are built to high quality and durable standards resulting in low maintenance. The need for low maintenance helps with minimizing 'downtime' and optimizing the availability of the road network. New intelligent in-car techniques as well as smart road and travel management systems will increase the capacity of roads as well as reduce the number of casualties. Dedicated lanes have been introduced on a wide scale to give priority to certain types of vehicles, e.g. long distance transport lanes (interurban) and separate lanes for buses and bicycles in urban and suburban areas.

Access to convenient transportation for people of all ages and physical abilities is the ultimate requirement for responding to mobility demands for transport communication in the future.

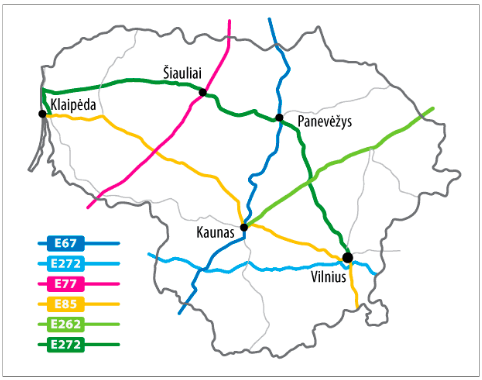


Figure 20.47. The major highways in Lithuania

Highways in Lithuania connect the capital city of Vilnius with the port city of Klaipėda via the second largest city Kaunas, as well as Vilnius and Panevėžys (Figure 20.47). The length of Vilnius-Klaipėda highway is over 300 kilometres, that of Vilnius-Panevėžys highway – 135 km. The highways have four lanes (two lanes each way) separated by grass lines and safety rails. Most of the sections from Kaunas to Klaipėda and from Vilnius to Panevėžys have a motorway status of 130 km/h speed limit during summer and 110 km/h during winter. The section from Kaunas to Vilnius has an expressway status of 110 km/h speed limit during summer and 100 km/h during winter.

CHAPTER 21 | Rail Transport

This chapter will help in

- discussing the rise and development of rail transport;
- explaining steam locomotives, diesel engines, high-speed rail and magnetic levitation transport;
- introducing the construction and functioning of subways;
- showing the development of Lithuanian railways and cable railway (funicular) transport.

Roads of rails called *Wagonways* were being used in Germany as early as 1550. These primitive railed roads consisted of wooden rails over which horse-drawn wagons or carts moved with greater ease than over dirt roads. Wagonways were the beginnings of modern railroads.

By 1776, iron had replaced wood in rails and wheels on carts. Wagonways evolved into tramways and spread throughout Europe. Horses still provided all pulling power. In 1789, the first wagons with flanged wheels were designed. The flange was a groove that allowed wheels to better grip the rail, which was an important design that carried over to later locomotives.

The invention of the steam engine was critical to the invention of modern railroad and trains. Richard Trevithick built the first steam engine tramway locomotive. On 22 February 1804, the locomotive hauled a load of 10 tons of iron, 70 men and five extra wagons 9 miles in the town of Merthyr Tydfil (Wales). It took about two hours.

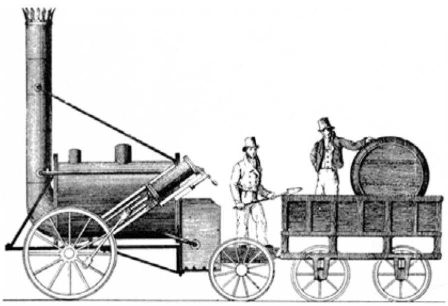
In 1825, the first railroad to carry both goods and passengers on regular schedules using locomotives began functioning. Stephenson's locomotive pulled six loaded coal cars and 21 passenger cars with 450 passengers over 9 miles in about one hour.

Stephenson is considered to be the inventor of the first steam locomotive engine for railways. In 1825, Stephenson moved to the Liverpool and Manchester Railway, where together with his son Robert, built (1826–29) the Rocket (Figure 21.1). The train travelled through the nation with its 4.5 ton weight and the top speed of 40 km/h. Although it did its job of transportation, the "Rocket" was highly unpopular to people. Those who had the job of carrying freight on wagons thought that the locomotive would replace them. There were also safety and reliability problems with the locomotive. It often broke down thus delaying delivery and arrival times of packages and people. Poor roads and flimsy bridges frequently resulted in accidents and injury. Another problem was the way that railroads were constructed. Only one track was laid instead of two for each direction. This increased the possibility of a head-on collision. Even when trains were functional, their smokestacks spewed black smoke and burning embers that often scorched passengers and their clothing or set proximate buildings on fire. The locomotive was essential to transportation but disliked by most because of a number of limitations it had.

Colonel John Stevens is considered to be the father of American railroads. The first railroad charter in North America was granted to John Stevens in 1815. Grants to others followed, and work soon began on the first operational railroads. Designed and built by Peter Cooper in 1830, the Tom Thumb was the first American-built steam locomotive to be operated on a common-carrier railroad.

The Pullman sleeping car was invented by George Pullman in 1857. Pullman's railroad coach or sleeper was designed for overnight passenger travel. Sleeping cars were being used on American railroads since the 1830s, however, early sleepers were not that comfortable and the Pullman sleeper was.

a)



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b)

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Figure 21.1 Stephenson's 'Rocket' locomotive: a – Rocket drawing, b – closer view

After the First World War, the expansion of railroads slowed and was in retreat ever since due to the competition of the modes of transportation. In Europe, trains have maintained a fairly consistent infrastructure for both passenger and freight traffic, whereas in the United States, only a semblance of passenger service remains. This situation has resulted from the expansion of highways for motor vehicles that carry people and goods. With the success of the Douglas DC-2/3 airplane in commercial aviation, the economics of carrying passengers over long distances shifted from railroads for good.

A steam locomotive had evolved in a steady progression to a relatively simple, powerful machine for either tractive effort (freight) or speed (passenger-service). In 1938, the British engine, *Mallard*, set the world record for the steam rail locomotive at 206 km/h. This represented the culmination of improvement in the efficiency of steam locomotive design that began with the 'Rocket' in 1829. Thermal efficiency, as measured in thermodynamic terms, is simply difference in the temperature of the working substance (steam, air, etc.) before and after its use. As Carnot discovered, it is therefore desirable to start with the highest initial temperature and end with the lowest final temperature. The efficiencies of rail locomotives never reached much above 8%. By 1935, the steam successor of American railroads was in service. The *Zephyr* was a hundred-mile-per-hour streamliner powered by a revolutionary marriage of prime movers – the diesel and electric motor.

Diesel engines were poorly suited to rail use. The early models had poor efficiencies and power-to-weight ratios. The General Motors build the first successful rail diesel engine. Kettering invented the two-cycle diesel. This diesel design combined the four cycle intake and exhaust strokes into the power stroke making a smoother, faster engine. He combined his lighter design with the fuel injector that removed heavy fuel system plumbing by separating each injector for each cylinder. Kettering's engines were built of new chromium, manganese and silicon-alloy steel called *Cromonsil* that further reduced weight. His eight-cylinder two-cycle diesel produced 75 hp per cylinder at a ratio of one horsepower per 10 kg. The diesel did not drive wheels as steam engines did. Instead, rotational energy was taken from the engine's shaft and converted into electrical power by a direct-current (DC) generator. This current fed series-wound, DC traction motors that could be reversed on downgrades providing dynamic braking and reducing wear on brakes. Combined with automatic air brake, the diesel locomotive replaced heavier, less efficient steam engines.

During World War II, the railway system played a vital role in the war effort transporting military personnel, equipment and freight to the front lines and often evacuating the entire factories and towns. The progress of the Second World War reached a definitive point when the Soviet army succeeded in pushing back Germany advance. In the wake of their retreat, the Germans had left large numbers of BR-52

(Kriegslok 2-10-0) locomotives. The BR-52 was a unique locomotive that came into existence as a result of German transportation needs during World War II (Figure 21.2).

Figure 21.2. The BR-52 steam locomotive

The BR-52 was comprised of about 5,000 separate parts, 2,000 of which were carefully designed for manufacturing simplicity. The BR-52 frames were fabricated from 30 mm steel plate. It required 11,650 man hours to build a BR-52. The weight of a completed BR-52 and its tender was 139 tons. About 6,000 BR-52s were manufactured during the war, but design was so successful that several hundred more were constructed after the cession of hostilities thus representing the largest production run in German railway history. About 2700 remaining were in the Soviet Union after the war ended (Figure 21.3).

Figure 21.3. Steam locomotive El-2500 of the armoured train from World War II (Sevastopol museum)

Some of those had a standard gauge and were used in the Baltic and border republics, whereas the rest were equipped with a broad gauge. To make up for the severe damage caused by fighting in the war, the USSR needed more effective freight power. Between 1945 and 1955, a total of 4200 units of class L locomotives were delivered (maximum speed 90 km/h). These were popular freight locomotives and survived into the strategic reserve era. The star of post-war locomotive designs was undoubtedly the express passenger locomotive. This not only looked right, it performed well and was a big step forward from the earlier passenger classes, well able to handle heavy express trains in the world's largest country in terms of land mass. A total of 251 locomotives were built between the prototype in 1950 and the last one in 1956. The locomotive weighed 133 tons excluding its twelve-wheeled tender. It had a great area of 6.75 m^2, a heating surface of 243m^2, a super-heater surface of 132 m^2, boiler pressure of 15 atmospheres, coupled wheels of 1850 mm in diameter and two (only) cylinders of 575x800 mm. Thus, its power output was 2265 kW (3000 hp) and its maximum speed – 125 km/h.

By the 1950s, steam engines were no longer manufactured in the U.S. Still, Europe had transportation infrastructure which it never lost in place after World War II. Post-war developments continued in passenger train design but it was for the Japanese to bring passenger rail service back as a viable alternative to air or highway travel for short-to-medium distances (100 to 500 km). The development of high-speed rail by Japan National Railways (JNR) in the 1960s spurred a new interest in passenger trains. The new trains, called "Bullet Trains" or "Shinkensen", were highly aerodynamic. High technology designs combined the roadway and train into a system. Powerful electric motors are driven by electrical overhead grids removing heavy diesels from the vehicle while the rails and roadbed are specially designed and monitored by computer. In 1981, the French "TGV" reached a maximum speed of 380 kph. By 1989, this top speed was raised by a new TGV to over 400 kmh. Prototypes have already exceeded 500 km/h over short distances by removing frictional resistance that plagued railroad engineering since the 'Rocket'.

As an alternative to high-speed rail based on traditional flanged-wheel vehicles, the technology of magnetic levitation, or *maglev*, has received considerable attention and research, though its practical applications have been limited by cost, safety concerns and satisfaction with traditional high-speed systems. A maglev vehicle rides on an air cushion created by electromagnetic reaction between an on-board device and another embedded in its guide-way. Propulsion and braking are achieved by varying the frequency and voltage of a linear motor system embodied in the guideway and reacting with magnets on the vehicles.

The idea that a train could be raised above a track and propelled by magnetic means was German in origin harking back to the 1930s where German engineers wrote papers on the theory of "attractive levitation". World War II interrupted this development but the idea remained. However, it was not for Germany to build the first successful maglev. This honour fell to the United States. Henry H. Kolm built the Magneplane, a scale model maglev in 1973. Germany had several maglev prototypes by the late 1970s and tested a full-scale, passenger carrying maglev the Transrapid 05 in 1979. The JNR engineers produced a 500 km/h maglev (without passengers) in 1978. By 1981, the Japanese had a maglev that combined passenger capacity and speed. The Germans were building Transrapid 06, a 500 km/h maglev capable of carrying 200 people.

Lithuanian Railways

The main network of Lithuanian Railways consists of 1749 km of 1,520 mm broad gauge railway 122 km of which are electrified (Figure 21.4). They also operate 22 km of standard (1,425 mm) gauge lines. A 179 km 750 mm narrow gauge network is listed as an object of cultural heritage.

In February 1851, the Tsarist Government of Russia made a decision on building the St. Petersburg – Warsaw railway line with a length of approximately 1,250 kilo-

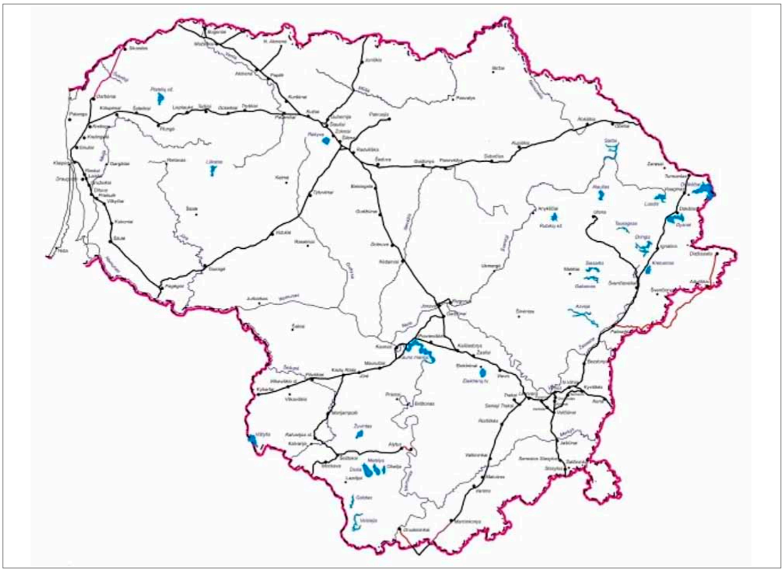

Figure 21.4. Lithuanian railway network (2012)

metres. Construction was completed in 1862. In May 1858, construction started near Vilnius on the first section of 19 kilometres. The first train from Daugavpils (Latvia) arrived in Vilnius on 16 September 1860. In 1861, this branch was completed to the Prussian border. The construction of the section from Lentvaris to Warsaw was completed on 15 December 1862. The railway was partly destroyed during both world wars. The Vilnius–Minsk railway was open to public traffic on 14 January 1873. In 1872, the construction of Mažeikiai–Jelgava railway section started, and in 1873, public road traffic was launched. Several other important lines such as Liepaja–Kaišiadorys (opened in 1871), Radviliškis–Daugavpils (1873), Vilnius–Lydia (1884) and Varėna–Alytus–Šeštokai–Suwalki (1899) were built.

Figure 21.5. A building of the railway station in Marijampole

During World War I (1915–1918), German military authorities completed Lauksargiai–Šilėnai line (124.4 km) in 1916 and ordered the construction of the road to launch Šiauliai–Joniškis–Jelgava the works of which were completed in 1916.

With permission from JSC "Lithuanian Railways"

Figure 21.6. The freight railway station in Vilnius (2013)

Figure 21.7. The production of straight and curved switches in the *Voestalpine VAE Legetecha* factory (2013)

During the period of Independent Lithuania (1918–1940) the state owned 943 km of roads, six locomotives, 15 passenger and 57 freight wagons (leased from the Germany). In 1923, railway section Šeštokai–Kazlu Rūda was completed, and in 1923, Kazlu Rūda was connected with Mariampolė (Figure 21.5). In 1926, railroads Kužiai–Telšiai and in 1932 Telšiai–Kretinga were finished.

One of the most important works in the development of Lithuanian railways was the electrification of lines Vilnius–Kaunas and Lentvaris–Trakai. In 1975, the first electric train arrived from Vilnius to Kaunas. During the Soviet period, other important works were carried out: for the period 1972–1975, the line Švenčionėliai–Utena was reconstructed from a narrow gauge to 1520 mm. In 1982, a new railroad Radviliškis–Pakruojis was built, and the railway from Adutiškis to Didžiasalis (currently disassembled) was extended. In 1979, the replacement of steam locomotives by diesel locomotives was finished. 1995 in Vilnius was established new modern company for production of various types of track intersections (curve and straight) Voestalpine VAE Legetecha (Figure 21.7).

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Figure 21.8. A railcar in Lithuania

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Figure 21.9. The 750 mm narrow gauge railroad

Railcar type trains were launched in 2008 for passenger transportation in less overcrowded routes (Figure 21.8).

There is a 179 km 750 mm narrow gauge network listed as an object of cultural heritage. The network was split into a separate company *Aukštaitijos Siaurasis Geležinkelis* in 2001. The same year, 68 km long narrow gauge *Aukštaitijos Siaurasis Geležinkelis* serving five stations was established (Figure 21.9).

Funicular Railway

A funicular is a cable railway in which a cable attached to a pair of tram-like vehicles on rails moves them up and down a steep slope. The basic idea of funicular operation is that two cars are always attached to each other by a cable, which runs through a pulley at the top of the slope. Counterbalancing of the two cars, with one going up and one going down, minimizes energy needed to lift the car going up. Early funiculars used two parallel straight tracks, four rails, with separate station platforms for each vehicle. Two-rail configurations avoid the need for switches and crossings; the three-rail layout is wider than the two-rail layout, but the passing section is simpler to build. Some four-rail funiculars have the upper and lower sections interlaced and a single platform at each station. (Figure 21.10).

The oldest funicular first documented in 1515 in Austria. The line originally used wooden rails and a hemp haulage rope and was operated by human or animal power. The first railway in England with wooden rails was used for delivering coal

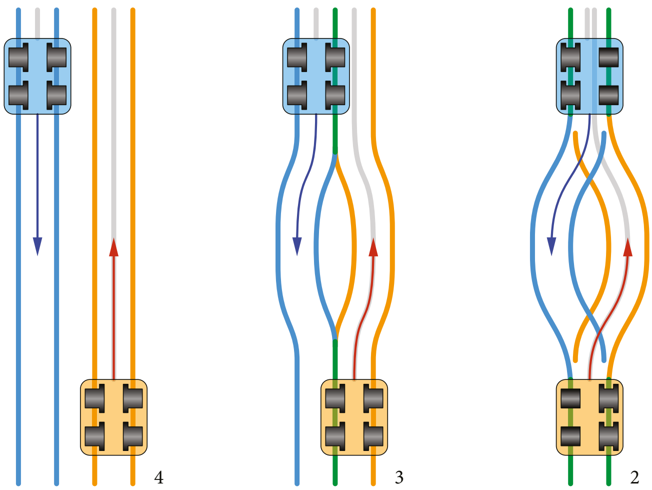

Figure 21.10. A funicular four-rail, three-rail and two-rail layouts

Figure 21.11. The funicular railway in Vilnius

to barges on the River Severn by 1606. In the 18th century, funiculars were used to allow barge traffic on canals to ascend and descend steep hills.

Modern funicular railways operating in urban areas date from the 1860s. The first line in Lyon opened in 1862, Budapest Castle Hill Funicular was built in 1869 and Istanbul the first underground funicular – in 1875. The Mount Beacon Incline Railway (New York) has an average gradient of 64% and a maximum gradient of 74%.

The first funicular railway in Lithuania was built in Kaunas in 1931 and was made of a wood-panelled coachwork that climbed 142 metres (gradient 25.9%). Aleksotas Funicular Railway (located in Kaunas) was officially opened in 1935. The length of the funicular is 142 metres and gradient – 29.5%. In 2003, a funicular railway started operating in Vilnius, covers a 71-meter long distance and takes people from the foot to the top of Gediminas Hill (Figure 21.11).

Subways

A subway is a train system that is located entirely underground. The first rapid transit system was the London Underground opened in 1863. Its oldest sections completed 150 years of operation in 2013, and its tunnels were in reality cut-and-cover construction with much of the line above ground as motive power was provided by locomotives used on the British Surface railways. Many cities followed London when they realized how useful such a train system could be. Paris followed the British lead beginning the construction of "Le Metro" in 1898. Its first sections were built just under the existing streets by a modified version of the cut-and-cover method.

The London subway sought to relieve traffic congestion. The Metro in France was built to solve a problem of sanitation. Garbage men could not get through. The Metro was electrified from its inception. Since subways are on a different grade than general traffic, they can operate entirely autonomously, free of traffic and other problems on surface streets. They use dedicated tracks, so they can also run rapidly and frequently, making them a valuable addition to a public transit network.

Typically, a subway runs on electricity often provided through the so-called "third rail". Trains run on two traditional tracks, just as above-ground trains do, and power themselves through paddles attached to the third rail. In some cases, a subway system may integrate above ground or elevated tracks in addition to an underground system. Commonly, the subway remains underground in the downtown section of the city, while elevated tracks branch out into less busy areas. These tracks may also network with other transit systems such as commuter rail bringing people in from suburbs or a city bus system. When it is well run and networks with other transit, it can play a significant role in reducing general traffic thus making a city more pleasant to live, work and play in.

In many cities, a subway includes multiple track systems that spread out underneath a city for wide coverage. Frequent stations along the tracks ensure that people will be able to quickly reach their destinations. The train company usually makes an effort to make the system easy to navigate and hospitable for visitors (Figure 21.12).

Figure 21.12. A subway train

Figure 21.13. A subway station in Moscow

New York's subway was begun in 1900 with 34 kilometres complete by 1904, eight of which were electrified. The year 1931 saw the start of one of the most important subways of the 20th century – the Moscow Subway. The soil under the Soviet capital was so unstable that it was nearly impossible to make a drive even with shields. Engineers utilized a chemical process for hardening the soil. A grout of silicate of soda ("water glass") was injected into the ground up to 10 meters using perforated pipes. This hardened and facilitating tunnelling that could then be lined with concrete. The advance was rapid (1 meter/shift), but costly. The first line opened on 15 May 1935 (Figure 21.13). The Moscow subway stations have their air changed four times an hour, cleaned each morning, and the tunnels are washed with high pressure hoses once a month. High speed escalators (1 m/s) were first used in this system.

The biggest rapid transit system in the world by length of routes (including non-revenue track) and by number of stations is the New York City Subway. By length of passenger lines, the largest are the Seoul Metropolitan Subway, Beijing Subway, Shanghai Metro and London Underground. The busiest metro systems in the world by daily and annual ridership are the Tokyo subway, the Seoul Metropolitan Subway and the Moscow Metro. As of May 2012, 184 cities have built rapid transit systems.

CHAPTER 22 Engineering of Space Flights

This chapter will help in

- introducing rocket birth and grown-up (fuel, specifications);
- explaining manned space flights (Gagarin, Grissom, White, Chaffee, Komarov, Armstrong, Aldrin);
- presenting the Apollo and Soviet spaceflight programs;
- defining the space shuttle program;
- pointing out Lithuanian print in space programme;
- describing international space programmes (International space station, Constellation, the Mangalyaan mission, Aurora programme);
- identifying the main probes into the planets (Mariner, Venera, Pioneer, Voyager).

Grown-Up of the Rocket

Only American R.H. Goddard successfully flew a liquid-fuelled rocket (O_2 and gasoline). This took place in March 1926. It took 56 meters in 2.5 seconds. Two principal missile weapons were developed at Peenemunde (Germany) - the V-I, a pulse-jet glider bomb and the A-4 (later, the V-2), a true rocket. The V-I had the fuselage of 7.7 meters long with a 1000 kg warhead. It flew on 80-octane gasoline with a fuel tank capacity of 567 litres. Fuel consumption was one gallon per mile with a speed of 580 kph. It had an automatic pilot mounted in the tail with controls driven by compressed air. Its wing span was 5.1 meters. The A-4 rocket consisted of four main sections: warhead, instrument compartment, fuel tank and a tail section with a motor (Figure 22.1). Fuel, alcohol and liquid O_2, were pumped into the motor by a centrifugal pump driven by steam generated by potassium permangenate ($KMnO_4$) and hydrogen peroxide (H_2O). The rocket was fired at three stages: the electrical ignition of a pinwheel, a preliminary stage (gravity flow of fuel only) and the main stage (pumped fuel with 27 tons of thrust). The "burn" lasted about one minute with the speed of sound reached 25 seconds. The maximum velocity was around 1500 m/sec.

The liquid-propellant rocket was the world's first long-range combat-ballistic missile and the first known human artifact to enter outer space. The ballistic missile that was developed at the beginning of the Second World War in Germany was specifically targeted at London. At the beginning of September 1944, over 3,000 V-2s were launched as military rockets during the war.

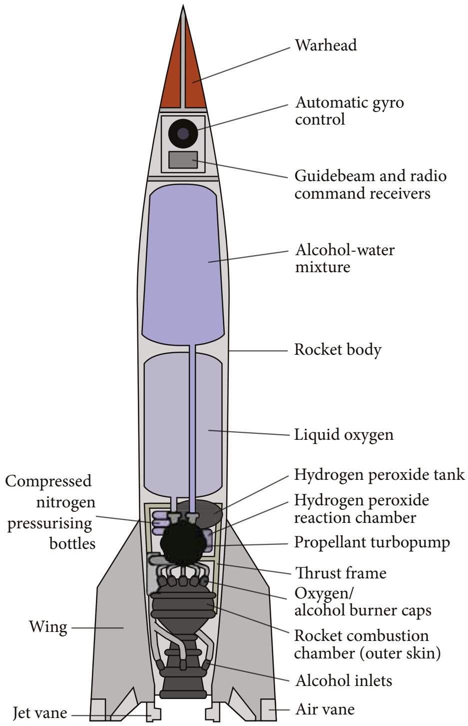

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Figure 22.1. Diagram of the V-2 rocket

The A-4 used a 75% ethanol/water mixture for fuel and liquid oxygen (LOX) for oxidizer. At launch, the A-4 propelled itself for up to 65 seconds on its own power, and a program motor controlled the pitch to the specified angle at engine shutdown, from which the rocket continued on a ballistic free-fall trajectory. The rocket reached a height of 80 km after shutting off the engine. Fuel and oxidizer pumps were steam turbines, and steam was produced by concentrated hydrogen peroxide with sodium permanganate catalyst. Both alcohol and oxygen tanks were an aluminium-magnesium alloy. The combustion burner reached a temperature of 2500–2700 °C. The specifications of the rocket include 12,500 kg in weight, 14 m in length, 1.65 m in diameter, a speed of 5,760 km/h, an operational range of 320 km, the maximal flight altitude of 88 km and the warhead of 1,000 kg.

Programs for the development of rocketry and the instrumented exploration of space had proceeded in the United States and the Soviet Union since the end of World War II. Both nations had benefitted from the captured German plans and launched vehicles as well as the incorporation of the actual persons into their space programs. Both nations were developing scientific satellites but no one could have gauged an impact of the launch of the Soviet satellite Sputnik I in 1957 (Figure 22.2), which did demonstrate the Soviet Union's superiority in a long-range, "heavy-lift" rockets.

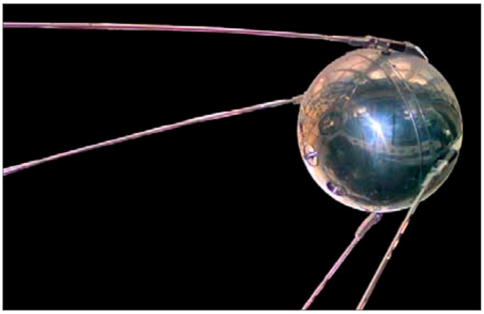

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Figure 22.2. The first artificial satellite Sputnik 1

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Figure 22.3. The dog that became the first living creature sent into space (1957)

Dog *Laika*, the first living creature to be put into the Earth orbit, was launched into space in 1957 (Figure 22.3). Eight more Sputnik missions with similar satellites carried out experiments on a variety of animals to test spacecraft life-support systems; they also tested re-entry procedures and furnished data on space temperatures, pressures, particles, radiation and magnetic fields.

In the forty-odd years of the existence of the U.S. and Soviet space programs, crowning achievements for each have covered landing humans on the moon and deep-space missions (U.S.), launching the first satellite, the first manned flight and developing the sustained habitation of space (Soviet). The major project sequences for each manned program have been Mercury/Gemini/Apollo/STS (U.S.) and Vostok/Voskhod/Soyuz/Mir (Soviet).

Manned Space Flights

In 1961, the USSR announced the successful launch and return of its first cosmonaut Yuri Gagarin who made a single orbit aboard Vostok 1. On 5 May 1961, the US launched its first Mercury astronaut, Alan Shepard, but on a suborbital flight. The first woman in space was Valentina Tereshkova who entered orbit on 16 June 1964 aboard the Soviet mission Vostok 6.

The "secret" of the Soviet success laid principally in the multi-engine booster that was little more than a package of smaller, less-powerful rockets combined into larger, more capable design. The vehicle the Soviets built was the SS-6 multi-engine workhorse that launched Sputnik and Vostok 1 spacecrafts. The SS-6 consisted of a long central core stage and four strap-on boosters. In fact, the 1964 Voskhod, a second-generation Soviet spacecraft replacing the one-man Vostok, weighed in at 5.300 kilograms almost the weight the U.S. Apollo spacecraft of five years later.

To fly in space, engineers had to contend with deficiencies in computer capability, a poor understanding of the exact orbital motion (particularly that of the lunar orbit), unsophisticated guidance and tracking systems, materials for the thermal protection of the spacecraft and landing systems. New materials such as Teflon and thermal resistant ceramics were discovered. In tracking, the development of global communications employing microwave and satellite technology led to networks like the Deep Space Network. New technology for liquid and solid fuel booster rockets led to the great lift capacities of lunar rockets. Hydrogen-oxygen liquid propellant technology is based on the equation

$$O_2 + 4H_2 - 1.98\ H_2O + 1.97\ H_2O + 0.06\ H + 0.03\ OH + \text{Heat}$$

that was combined with advances in variable thrust motors. The launch of a rocket has been termed a "controlled explosion." Ballistic missiles since the German V-2s were really atmospheric devices. Future designs required that would never again see the atmosphere after launch and would survive for extended periods in deep space. Advances in solid state propellants required a great deal of engineering devoted to the shape of grain, the configuration of the combustion chamber and the control of burning the propellant. Modern solid propellants belong to one of two classes, double base or composite grain propellants. Typical compounds that make up the first are nitroglycerin ($C_3H_5(ONO_2)_3$) and nitrocellulose ($C_6H_7O_2(ONO_2)_3$). Nitrocellulose adds strength to the grain and nitroglycerin, a high-performance burning component. The composite grain propellant is a mixture of a rubbery polymer mixture of fuel (polybutadiene) and a crumbly oxidizer such as potassium nitrate (KNO_3) or ammonium perchlorate (NH_4CrO_4).

As both nations rushed to get their new spacecraft flying with men, the intensity of the competition caught up to them in early 1967 when they suffered their first crew fatalities. On January 27, the entire crew of Apollo 1, Gus Grissom, Ed White and Roger Chaffee, were killed by suffocation in a fire that swept through their cabin during a ground test. Then on April 24, the single pilot of Soyuz 1, Vladimir Komarov, was killed in a crash when his landing parachutes tangled after a mission cut short by electrical and control system problems. Both accidents were determined to be caused by design defects in the spacecraft and were corrected before the manned flights resumed. In 1969, with the landing of Apollo 11, Neil Armstrong and Buzz Aldrin became the first men to set foot on the Moon (Figure 22.4).

Apollo went on to put 12 astronauts on the lunar surface (first manned landing in 1969 and last landing in 1972). Apollo succeeded in achieving its goal of manned lunar landing despite the major setback of a 1967 Apollo 1 cabin fire

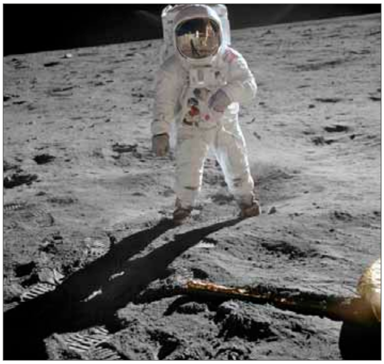

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Figure 22.4. The first man to set foot on the Moon

that killed the entire crew during a pre-launch test. Not all accidents were fatal. En route to the Moon of Apollo 13 expedition (1970), approximately 320,000 km from Earth, a mechanical shock ruptured an oxygen tank. Because fuel cells combined hydrogen and oxygen to generate electricity and water, the remaining fuel cells were finally shut down and left the Command Module (CM) on limited-duration battery power. The crew was forced to shut down the CM completely and to use the Lunar Module (LM) as a "life boat". Damage to the Service Module made safe return from lunar landing impossible. The lunar module consumables were intended to sustain two people for a day and a half, not three people for four days. Unlike the CM, that was powered by fuel cells producing water as a by-product, the LM was powered by silver-zinc batteries, and thus electrical power and water (used for equipment cooling and drinking) were critical consumables. To keep LM life support and communication systems operational until re-entry, the LM was powered down to the lowest levels possible. As Apollo 13 neared the Earth, the crew first jettisoned the Service Module. Further complication was the fact that the reduced power levels in the LM caused internal temperatures to drop to as low as 4 °C. The un-powered CM got so cold that water began to condense on solid surfaces causing concern that this might short out electrical systems. Finally, on 17 April 1970, Apollo 13 splashed down in the South Pacific (Figure 22.5). The crew jettisoned the Lunar Module *Aquarius*, leaving the Command Module *Odyssey* to begin its lone re-entry through the atmosphere.

Apollo used Saturn family rockets as launch vehicles and succeeded in achieving its goal of manned lunar landing. After the first landing, sufficient flight hardware remained for nine follow-on landings with an ambitious plan for extended lunar geological and astrophysical exploration.

Both nations went on to fly relatively small, non-permanent manned space laboratories *Salyut* and *Skylab* using their *Soyuz* and *Apollo* crafts as shuttles. Manned reconnaissance stations were found to be a bad idea, since unmanned satellites could do the job much more cost-effectively. The U.S. Manned Orbital Laboratory was cancelled in 1969, whereas the Soviets cancelled Almaz in 1978. In a season of *detente*, the two competitors declared the end to the race. On 17 July 1975, the Apollo-Soyuz Test Project was the case when two crafts docked and the crews exchanged visits.

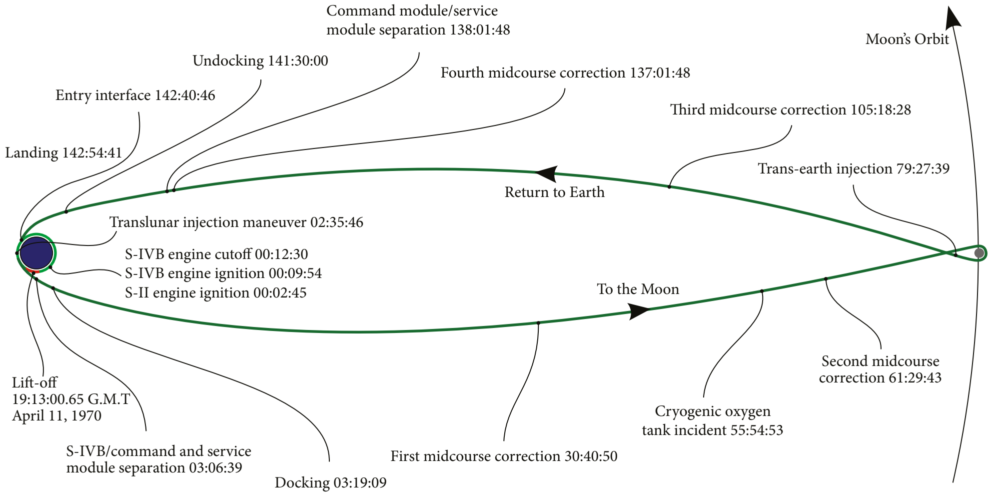

Figure 22.5. The flight trajectory followed by Apollo 13

The Soviet Lunar program never succeeded for two reasons: first, the lack of a reliable, giant booster like the *Saturn* V and second, vehicle endurance. The Soviets did attempt to launch a *Saturn-sized* booster, but the three tested failed beginning with the first in 1969 and ending with the last two in 1971 and 1972. Vehicle endurance was a problem for the Soviets as well. *Soyuz-derived* vehicles, although this type was the backbone of Soviet Russian flights in the latter 20th century, were incapable of a manned lunar-loop mission of the same kind as *Apollo 8*. The lunar-loop route required the vehicle endurance of seven days and the longest *Soyuz* flights to that time were Cosmos 212 and Cosmos 213 of five days each.

Another difference in the U.S. and Soviet lunar programs were the approaches taken in the missions. The U.S. successfully used a rendezvous and docking type of an approach where the Apollo flight module was coupled and decoupled to a landing module. The Soviets never demonstrated this type of technology. The "direct launch" approach of the Soviets was that used in the early Luna missions and would have placed the flight lander module into a direct lunar injection path. To do so, a more powerful booster would have required. The *Mir* space station program differed from *Salyut* in conception. The *Mir*

Figure 22.6. The first reusable spacecraft, the *Space Shuttle*

was designed as a modular station built around an orbital core launched in 1986. *Kvant* modules were launched and attached to the *Mir* core module to be served by crews ferried in new Soyuz-TM vehicles and resupplied by Progress unmanned "space tugs" developed for the *Salyut* program. With the collapse of the Soviet Union and its subsequent impact on the finances of the space program, the *Mir* has been the Russian program's single, long-running success story.

Although its pace slowed, space exploration continued after the end of the Space Race. The United States launched the first reusable spacecraft, the *Space Shuttle*, 1981 (Figure 22.6). In 1988, the Soviet Union duplicated this with an unmanned flight of the *Buran shuttle*, its first and only reusable spacecraft.

Lithuanian Print

Rimantas Stankevičius was a Lithuanian cosmonaut who tested Soviet space shuttle *Buran*. In 1979, he was assigned for preparing *Buran*, the USSR space shuttle (Figure 22.7). In 1982, he passed all required exams and became the first Lithuanian cosmonaut. After 1984, he trained to fly the *Buran* space shuttle. He accomplished 14 test-flights with *Buran's* counterpart BTS-02 aircraft and 6 taxi tests with *Buran*. He was both the pilot and the commander of the space shuttle. He was killed in a crash of his Su-27 fighter plane during an air show in Italy in 1990. The only orbital launch of Buran occurred on 15 November 1988 from the Baikonur Cosmodrome. It was lifted into orbit, on an unmanned mission, by the specially designed Energia rocket. After completing two revolutions around the Earth, space shuttle Buran began descent into the atmosphere and did automated landing.

The Buran Space shuttle was never used again after the first flight. Instead, the Soviet Union continued with developing space stations using the Soyuz craft as a crew shuttle.

Figure 22.7. Soviet An-225 Mechta aircraft with the Space Shuttle Buran on its back

Space Shuttle Program

The American space program concentrated on the construction of a reusable spaceplane or "shuttle". The U.S. Skylab (1973–1974) logged 513 days during three missions. The 56,700 kilogram Skylab brought the two space powers to parity in lift

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Figure 22.8. A drag chute deployed by Space Shuttle landing

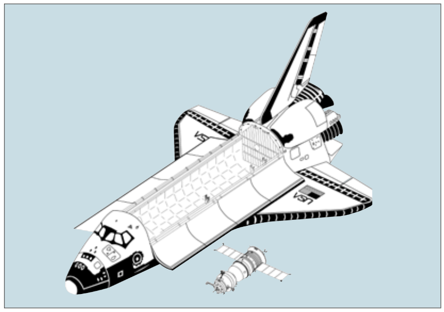

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Figure 22.9. Cargo delivering by Space Shuttle

capacity. The Soviets went on to build *Energia* now the successor to the *Saturn* in terms of thrust.

NASA's Space Shuttle program was the U. S. manned launch vehicle program from 1981 to 2011. The winged Space Shuttle orbiter was launched vertically, usually carrying four to seven astronauts. The Shuttle consists of three elements – the orbiter, an external liquid fuel tank and two solid-fuel booster rockets. It is a part aircraft, part missile (Figure 22.6).

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Figure 22.10. Space Shuttle transportation by the Boeing 747 Shuttle Carrier Aircraft

The Shuttle is the only winged manned spacecraft to have achieved orbit and land, and the only reusable space vehicle that has ever made multiple flights into orbit (Figure 22.8). Its missions involved carrying large payloads to various orbits (including segments to be added to the International Space Station), providing crew rotation for the International Space Station and performing service missions (Figure 22.9). The orbiter also recovered satellites and other payloads from orbit and returned them to Earth. Each vehicle was designed with a projected lifespan of 100 launches, or 10 years' operational life.

The Shuttle Carrier Aircraft was used for ferrying space shuttles from landing sites back to the Shuttle Landing Facility at the Kennedy Space Centre and to and from other locations too distant for the orbiters to be delivered by ground transport (Figure 22.10).

Flown successfully in 1981 it was the complex vehicle design that produced the worst manned-flight tragedy. The first U.S. in-flight fatality was the Space Shuttle *Challenger* destruction 73 seconds after lift-off (1986). Investigation found that a faulty O-ring seal allowed hot gases from the shuttle solid rocket booster to impinge on the external propellant tank and booster strut. The Space Shuttle Columbia was lost as it returned from a two-week mission (2003). Damage to the thermal protection system of the shuttle led to structural failure of the left wing of the shuttle. Investigation revealed damage wing panel resulted from the impact of a piece of foam insulation that broke away from the external tank during the launch. Shuttle vehicles would suitable fly only to low-earth orbit. The mission of the vehicles is to carry the components of a permanent manned space facility. From this platform, missions to the moon and Mars are contemplated.

International Space Programmes

The "space race" has produced supercomputers and materials such as new polymers, metals and ceramics as well as fine technology for micro-size and digital processing. It is difficult to gauge the impact of a space flight on engineering as a whole at this juncture.

Recent space exploration has proceeded, to some extent in worldwide cooperation, the high point of which was the construction and operation of the *International Space Station* (ISS). The United States continued other space exploration, including major participation with the ISS along with its own modules. It also planned a set of unmanned Mars probes, military satellites and more. The Constellation Program, begun in 2004, was aimed at launching the next-generation multifunction Orion spacecraft by 2018. A subsequent return to the Moon by 2020 was to be followed by manned flights to Mars. Russia offers a wide commercial launch service while continuing to support the ISS with several of their own modules. They also operate a manned and cargo spacecraft that continues after US Shuttle program is ended. They are developing a new multi-function manned spacecraft for use in 2018 and also have plans to perform manned moon missions. The program focuses on putting a man on the moon in 2020 thus becoming the second country to do so. The European Space Agency (ESA) has taken the lead in commercial unmanned launches since the introduction of the *Ariane 4* spacecraft in 1988. The Agency has launched various satellites, has utilized the manned Spacelab module aboard US shuttles and has sent probes to comets and Mars. It also participates in ISS with its own module and the unmanned cargo spacecraft. Currently, the ESA has a program for the development of an independent multi-function manned spacecraft CSTS scheduled for completion in 2018.

The mission of *Mangalyaan* is a planned Mars orbiter to be launched in 2013 by the Indian Space Research Organisation (ISRO). This will be India's first mission to Mars. In case of success, India will become the third nation in the world to reach Mars before Asian powers China and Japan the earlier attempts of which had failed. NASA selected MAVEN, an orbiter mission to be launched in 2013 to study the atmosphere of Mars. NASA selected *InSight*, a lander mission for 2016, with a drill and seismometer to determine the interior structure of Mars.

The exploration of Mars has taken place over hundreds of years, beginning in earnest with the invention and development of the telescope since the 1600s. Since 6 August 2012, there have been two rovers beaming signals back to Earth on the surface of Mars (*Opportunity* and *Curiosity* of the *Mars Science Laboratory* mission) and three orbiters currently surveying the planet (Figure 22.12). The *Curiosity* mission is the most courageous and expensive attempt to answer the grand question: is there life on Mars? To date, no sample return missions have been attempted for Mars, and one attempted return mission for Mars' moon Phobos has failed.

Thus, when people come to walk on Mars, it will take many years. NASA specialists prepare and launch Curiosity Rover as a part of a long-term plan at the end of which astronauts will walk on Mars, as they did on the Moon. When NASA's next Mars rover touches down in 2021, its six-wheeled sister, *Curiosity*, may still be chugging around the Red Planet. NASA's goals at Mars combine both scientific discovery and human exploration (Figure 22.11). The agency wants to return pieces of the Red Planet to Earth so that researchers can analyze them for signs of life. NASA is also working toward sending astronauts to the vicinity of Mars by the mid-2030s.

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Figure 22.11. *Curiosity* rover on Mars surface

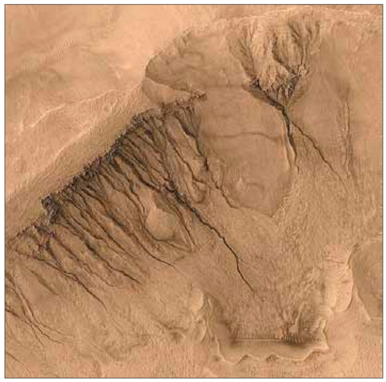

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Figure 22.12. Gullies, similar to those formed on Earth, are visible on this image from Mars Global Surveyor (1999)

Figure 22. 13. Chinese Shenzhou 5 re-entry capsule

Further goals include an ambitious plan called the *Aurora* programme that intends to send a human mission to Mars soon after 2030. The People's Republic of China has achieved a manned space flight and operates a commercial satellite launch service. China's manned space program is entering with ambitious plans, including a permanent space station, manned lunar missions and a possible manned mission to Mars in 2040–2060. The successful China's first crewed mission in space in 2003 by *Shenzhou 5* spacecraft (Shenzhou program) made China the third nation to launch a human into orbit missions (Figure 22.13). Chinese moon lander Chang'e 3 is currently (2014) on the lunar surface. The rover has instruments to survey the lunar subsurface and is built to function for one year. In 2007, China tested a ballistic missile designed to destroy satellites in orbit.

The Indian human spaceflight programme is a proposal by the Indian Space Research Organisation (ISRO) to develop and launch the ISRO Orbital Vehicle, which is to carry a two-member crew to Low Earth Orbit. The recent reports indicate that the human spaceflight will be after 2017.

Japan has deployed a module in the ISS and operates an unmanned cargo spacecraft. Japan has plans to launch a Mars fly-by probe. Their lunar probe, SELENE, is touted as the most sophisticated lunar exploration mission in the post-Apollo era. Japan's *Hayabusa* probe was mankind's first sample return from an asteroid, and *IKAROS* was the first operational solar sail. Although Japan developed the manned capsule spacecraft, none of them have been launched. Japan hopes to be launching astronauts aboard a manned capsule or space plane by 2022. The capsule or mini-shuttle would accommodate a crew of three and carry up to 400 kilograms of cargo. Japan's current ambition is to establish a Moon base by 2030.

As of 2007, citizens from 33 nations (including space tourists) have flown in space aboard Soviet, American, Russian and Chinese spacecraft.

Probes to the Planets

Since the early 1960s, there has been great activity directed towards sending space probes to either make close encounters with, or land on, the planets. The first successful fly-by of Venus took place in 1962 by the *Mariner 2* craft. *Mariner 10* took the first pictures of the cloud tops of the dense atmosphere of Venus, after which it went on to fly-by Mercury. The Soviet Union concentrated on soft landings on Venus. *Venera 4*, launched in 1967, released a spherical capsule (1m in diameter). A parachute system permitted a slow descent through the atmosphere. During the 94-minute descent, it recorded the temperatures of 370 °C and sent back data on the pressure, density and chemical composition of the atmosphere. A later spacecraft, *Venera 7*, was the first to survive the journey to the planet's surface relaying data such as pressures in excess of 90 atmospheres and a temperature of 475 °C. Later, a more sophisticated spacecraft succeeded in sending back pictures of the rocky surface of Venus. The *Viking* spacecrafts in 1976 returned excellent pictures of the surface as well as analyzed the Martian soil. No organic compounds were found in the samples examined. Two American *Pioneer* spacecrafts were launched in 1972 and 1973 on trajectories that would enable both crafts to fly-by Jupiter. After a journey of 21 months, they sent back the first detailed images of the planet. *Pioneer 11* was re-positioned on a flight path to intercept Saturn. Again, close-up images of Saturn and its extensive ring system were relayed back to Earth.

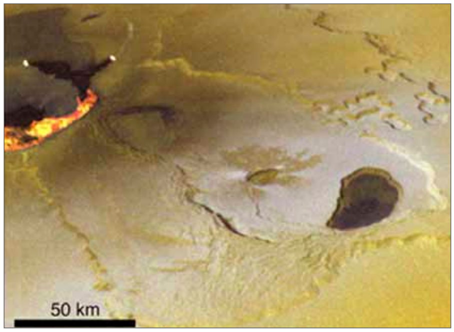

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Figure 22.14. Volcanic eruption in action on Jupiter's moon Io

The *Voyager* spacecrafts had a mass of over 800 kg with a high-gain antenna 3.7 m in diameter and a sophisticated array of detectors and high and low resolution TV cameras. 1979, *Voyager 1* made the closest approach (278,000 km) to Jupiter and returned fascinating detailed images of the clouds in the Jovian atmosphere. The craft made close-up fly-bys of the four Galilean satellites Io, Ganymede, Callisto and Europa, discovering, rather surprisingly, that Io showed the signs of active volcanism (Figure 22.14). Next, *Voyager 1* began travelling out of the solar system. *Voyager 2* encountered Uranus in 1986 and Neptune in 1989.

CHAPTER 23 Nuclear Power Birth and Development

This chapter will help in

- describing the birth of nuclear power;
- explaining the operation of nuclear fission reactors;
- introducing the development of nuclear fission reactors in Lithuania;
- discussing the development of Soviet nuclear power;
- showing the disasters of nuclear power reactors (Three Mile Island, Chernobyl, Fukushima);
- defining the operation of nuclear fission reactors (International Thermonuclear Experimental Reactor);
- pointing out the development of nuclear fission reactors (DEMO, PROTO nuclear fusion power plants) in the future.

It is in the 1930s, that the world saw the most profound innovation of the 20th century – nuclear power. Based on Otto Hahn and G. Strassmann's discovery of nuclear fission in 1938, Niels Bohr and A. Wheeler (1939) completed a detailed model of fission where the nucleus splits with the release of a neutron which can, in practice, start a nuclear chain reaction providing adequate amounts of rare isotope U^{235}. The first nuclear reactor was built in Chicago in 1942. Enrico Fermi was the designer of this device. Briefly, a mass containing uranium is spread in some suitable arrangement throughout a block of graphite. Whenever fission takes place in this system, an average number of neutrons, v, is emitted at a continuous distribution of energy of the order of 1 MeV. A neutron is a non-charged nuclear particle. Since interacting protons have mutual electromagnetic repulsion stronger than their attractive nuclear interaction, neutrons are often a necessary constituent within the atomic nucleus that allows a collection of protons to stay atomically bound. Neutrons bind with protons and one another in the nucleus via nuclear force thus effectively stabilizing it.

There is a small probability, ε, that the neutron will be absorbed before its energy has been appreciably decreased. This leads often to the fission of U^{238}. This fission type has a very low absorption probability. In the majority of neutron emissions, the energy of the "fast neutron" is dissipated due to collision with carbon atoms in graphite. It takes 6.3 collisions with carbon atoms to reduce energy by a factor of *e*. It takes ca.110 collisions to reduce the energy value from 1MeV to the thermal value of 0.025 eV. This neutron can be absorbed by the resonance process in uranium.

$$U^{238} + n \rightarrow U^{239}$$

$$\text{or } U^{235} + n\ (\text{slow}) \rightarrow \text{Fission}$$

The reactor at Oak Ridge (1943) was the first to produce an appreciable power level. Its core was 7,25 m^3 of graphite blocks; 1248 fuel channels, each 4.4 cm square, was located in a 17 cm rectangular lattice. Fuel was 35 tons of uranium metal, 2.8 cm in diameter, 10 cm long and jacketed with aluminum to prevent oxidation. The reactor was air cooled with 47,200 cubic meters of air/min flowing through the core channels (fuel was round in the square holes of channels). At 3800 KW, fuel temperature was 245 °C and that of graphite – 130 °C. The concrete shield was 2.1-meters thick.

Power is generated, today, in high pressure reactors generating steam for turbines thermodynamically and economically much like conventional power plants. More controversial is the so-called "Fast" or "Breeder" reactor. There is 77% probability of the absorption of a neutron by U^{238}. This process yields highly-fissile Pu^{239} by a (n, γ) reaction. Further, Th^{232} can be converted to fissionable U^{233} by the same process. These reactors produce more fuel than they use in a "breeder blanket." They use liquid sodium metal (Na, liquid at 90°C) in a closed loop to heat water for steam. It is projected that these reactors can increase the amount of fission fuel by a factor of 150. Nuclear power has created two parallel problems for today's engineers:

- how to develop a viable alternative;
- how to dispose of radioactive waste being produced.

The first problem is being addressed to both near and long-term research. Clearly petroleum and coal will be the primary source of energy for the next quarter of the century. Nuclear power may develop into a 25–40% component of energy supply. Without designs like the breeder reactor being built (on-line), this projection may not be valid. Long-term solutions hinge on the development of thermonuclear or fusion power and gas technologies such as hydrogen.

The second problem in nuclear power comes from the proper disposal of its waste by-products which appear to be long lived (>10^5 years). Ceramic and glass technology has been examined for containment and long-term storage of more dangerous nuclides (Ni, Co, Sr).

Nuclear Fission Reactors

Nuclear technology involves the reactions of atomic nuclei. Notable nuclear technologies consist of *nuclear power, nuclear medicine* and *nuclear weapons*. Nuclear power is a type of nuclear technology covering the controlled use of nuclear fission to release energy for work, including propulsion, heat and the generation of electricity. Nuclear energy is produced by a controlled nuclear chain reaction that creates heat used for boiling water, producing steam and driving a steam turbine used for generating electricity and/or to doing mechanical work. Currently, nuclear power provides approximately 15.7% of the world's electricity (in 2004) and is applied for propelling aircraft carriers, icebreakers and submarines.

In natural nuclear radiation, by-products are very small compared to the nuclei they originate. Nuclear fission is the process of splitting a nucleus into roughly equal parts, and releasing energy and neutrons in the process. If these neutrons are captured by another unstable nucleus, they can fission as well thus leading to a chain reaction. A mass of a fissile material large enough (and in a suitable configuration) to induce a self-sustaining chain reaction is called a *critical mass*.

When a neutron is captured by a suitable nucleus, fission may occur immediately, or the nucleus may persist in an unstable state for a short time. If there are enough immediate decays to carry on the chain reaction, the mass is said to be *prompt critical*, and energy release will grow rapidly and uncontrollably, usually leading to an explosion. Nuclear *fission* is splitting a massive nucleus into photons in the form of gamma rays, free neutrons and other subatomic particles (Figure 23.1). In a typical nuclear reaction involving ^{235}U and a neutron

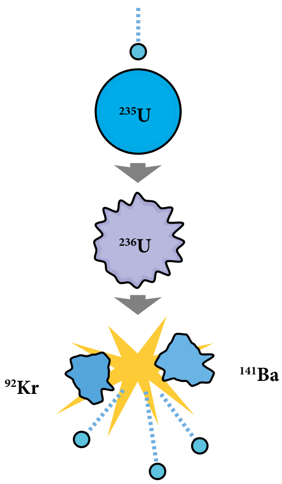

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Figure 23.1. Fission reaction

$$^{235}_{92}U + n = ^{236}_{92}U$$

followed by

$$^{236}_{92}U = ^{144}_{56}Ba + ^{89}_{36}Kr + 3n + 177 \text{ MeV}.$$

On earth, the most likely *fusion reaction* is Deuterium-Tritium reaction. Deuterium and Tritium are both isotopes of hydrogen.

$$^{2}_{1}\text{Deuterium} + ^{3}_{1}\text{Tritium} = ^{4}_{2}He + ^{1}_{0}n + 17.6 \text{ MeV}$$

In most electric power plants, water is heated and converted into steam driving a turbine-generator to produce electricity. Fossil-fuelled power plants produce heat by burning coal, oil or natural gas. In a nuclear power plant, the fission of Uranium atoms in the reactor provides heat to produce steam for generating electricity. The most widely used design consists of a heavy steel pressure vessel surrounding a reactor core that contains Uranium fuel formed into cylindrical ceramic pellets sealed in long metal tubes called fuel tubes. Pins are arranged in groups to make a fuel assembly. A group of fuel assemblies forms the core of the reactor.

Heat is produced in a nuclear reactor when neutrons strike Uranium atoms causing them to fission in a continuous chain reaction. Control elements, which are made of materials that absorb neutrons, are placed among fuel assemblies. When control elements, or control rods as they are often called, are pulled out of the core, more neutrons are available and the chain reaction speeds up, producing more heat. When they are inserted into the core, more neutrons are absorbed, and the chain reaction slows or stops, reducing heat. Most commercial nuclear reactors use ordinary water to remove the heat created by the fission process. These are called *light water reactors*. Water also serves to slow down, or "moderate", neutrons. In this type of the reactor, the chain reaction will not occur without water to serve as a moderator. Two types of different light-water reactor design are currently in use – the *Pressurized Water Reactor* (PWR) (Figure 23.2) and the *Boiling Water Reactor* (BWR).

In a PWR, heat is removed from the reactor by water flowing in a closed pressurized loop. Heat is transferred to the second water loop through a heat exchanger. The second loop is kept at lower pressure allowing water to boil and create steam used for turning the turbine-generator and producing electricity. Afterward, steam is condensed into water and returned to the heat exchanger. In a BWR, water boils inside the reactor itself, and steam goes directly to the turbine-generator to produce electricity. In the latter case, steam is condensed and reused.

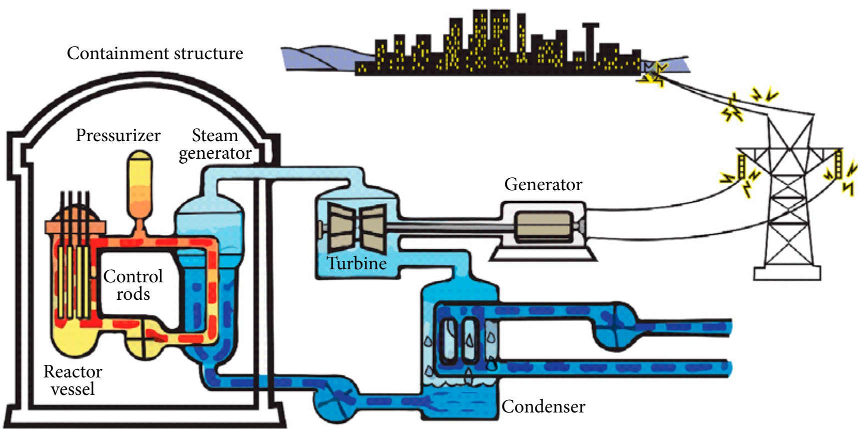

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Figure 23.2. A pressurized water reactor

Development of Nuclear Fission Reactors in Lithuania

For 26 years (1983–2009), Lithuania was a nuclear power (Figure 23.3) state. Unit 1 of the Ignalina Nuclear Power Plant came online in 1983 and was closed in 2004. Unit 2 came online in 1987 and was closed on 31 December 2009. Each unit of the Power Plant was equipped with two turbines for rotating 800 MW generators. The construction of Unit 3 started in 1985 but was suspended in 1988; its demolition began in 1989 and dismantling was completed in 2008. The construction of Unit 4 has never started because of the public backlash against nuclear power following the Chernobyl disaster. Lithuania agreed to close the Plant as a part of its accession agreement to the European Union.

The RBMK is a water-cooled power reactor based on graphite-moderated plutonium production (*Russian: Reaktor Bolshoy Moshchnosti Kanalnyy, "High power Channel-type Reactor"*). The RBMK design was built primarily to be powerful, quick to build and easy to maintain. Full physical containment structures for each reactor would have more than doubled the cost and construction time of each plant.

Figure 23.3. The Ignalina Nuclear Power Plant

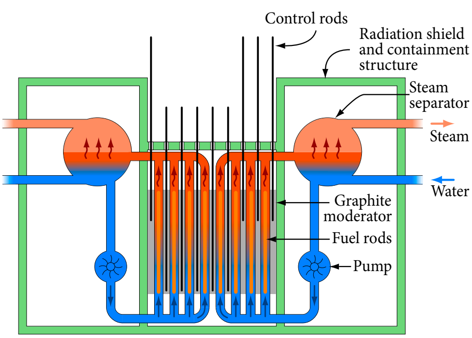

Figure 23.4. Diagram of the RBMK nuclear reactor

Additionally, RBMK reactors were designed to allow fuel rods to be changed without shutting down both for refuelling and for plutonium production (for nuclear weapons). The Ignalina Nuclear Power Plant contained two RBMK-1500 water-cooled graphite-moderated channel-type power reactors. The RBMK-1500 reactor was originally the most powerful reactor in the world with an electrical power capacity of 1,500 megawatts (MW) and thermal power capacity of 4,800 MW (Figure 23.4).

The reactor pit is made of reinforced concrete and has dimensions making 21.6×21.6×25.5 meters. It houses the vessel of the reactor made of a cylindrical wall and top and bottom metal plates. The vessel contains a graphite stack and is filled with a helium-nitrogen mixture for providing an inert atmosphere for graphite and for the mediation of heat transfer from graphite to coolant channels. The moderator blocks (250×250×500 mm) are made of nuclear graphite. There are holes with 11.4 cm in diameter through the longitudinal axis of the blocks for fuel and control channels. The blocks are stacked inside the reactor vessel into a cylindrical core. The maximum allowed temperature of graphite is less or equal to 730 °C. The reactor vessel is a steel cylinder with the wall thickness of 16 mm. The moderator is surrounded by a cylindrical water tank. Water is supplied to compartments from the bottom and removed from the top. The tank, sand layer and concrete of the reactor pit serve as additional biological shields. The top of the reactor is covered with the upper biological shield. It is penetrated by standpipes for fuel and control channel assemblies. The top and bottom are covered with 4 cm thick steel plates welded to be helium-tight. The space between the plates and pipes is filled with *serpentinite*, rock containing a significant amount of bound water. There is a lower biological shield below the bottom of the reactor core.

The fuel channels consist of welded zircaloy pressure tubes led through the channels in the centre of graphite moderator blocks. The top and bottom parts of the tubes are made of stainless steel and joined with the central zircaloy segment with zirconium-steel alloy couplings. The tubes are welded to the top and bottom metal plates of the reactor vessel. There are 1661 fuel channels and 211 control rod channels in the reactor core (Figure 23.5).The fuel assembly is suspended in the fuel channel on a bracket with a seal plug. Small clearance between the pressure channel and the graphite block makes the graphite core susceptible to damage. If the pressure channel deforms, e.g. by too high internal pressure, the deformation or rupture can cause significant pressure loads to the graphite blocks and lead to their damage, and possibly propagate to neighbouring channels.

Fuel pellets are made of uranium dioxide powder, sintered with a suitable binder into barrels 1.15 cm in diameter and 15 mm long. To reduce thermal expansion issues and interaction with cladding, the pellets have hemispherical indentations. The enrichment level is 2%, the maximum allowable temperature of the fuel pellet is 2,100 °C.

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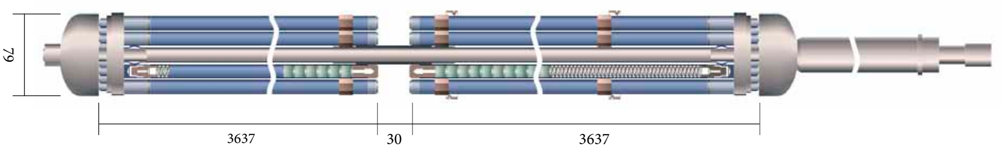
Figure 23.5. A reactor core at the Ignalina Nuclear Power Plant in Lithuania



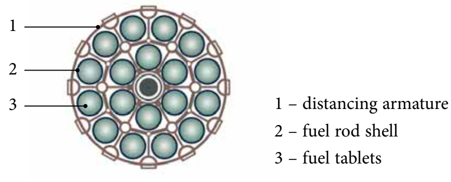


Figure 23.6. The fuel rod holder of the RBMK reactor

Fuel rods are zircaloy (1% Nb) tubes 1.36 cm in outer diameter and 0.825 mm thick. The rods are filled with helium at 0.5 MPa and hermetically sealed. Each rod contains 3.5 kg of fuel pellets. Fuel rods are 3.64 m long, with 3.4 m of that being active length. The maximum allowed temperature of the fuel rod is 600 °C. Fuel assemblies consist of two sets of 18 fuel rods. The total mass of uranium in the fuel assembly is 114.7 kg. Fuel burn-up is 20 MW·d/kg. The RBMK fuel assembly is cylindrical to fit round pressure channels (Figure 23.6). The refuelling machine is mounted on a gantry crane and remotely controlled. Fuel assemblies can be replaced without shutting down the reactor. When a fuel assembly has to be replaced, the machine is positioned above the fuel channel, mates to it, equalizes pressure within, pulls the rod and inserts a fresh one. The spent rod is then placed in a cooling pond. The capacity of the refuelling machine with the reactor at a nominal power level is two fuel assemblies per day. Spent fuel was placed in storage casks.

Most of the control rods of the reactor are inserted from above. Control rods have a 4.5 m long graphite section at the end, are separated by a telescope (creates a water-filled space between graphite and the absorber) of 1.25 m long and a boron carbide neutron absorber section. When the control rod is fully retracted, the graphite displacer is located in the middle of the core height, with 1.25 m of water at each of its ends. Control rod channels are cooled by an independent water circuit and kept at 40–70 °C. The narrow space between the rod and its channel is the primary cause of their slow insertion time (nominally about 0.4 m/s). Additional static boron-based absorbers are inserted into the core when it is loaded with fresh fuel. About 240 absorbers are added during the initial core loading. These absorbers are gradually removed with an increase in fuel burn-up.

The reactor has two independent cooling circuits. Cooling water is fed to the reactor through lower water lines. Each feeding pressure channel passes through the core where feed water boils. The mixture of steam and water is led to steam separators. Steam is taken from the separators and led to two turbo-generators in the turbine hall, then to condensers, reheated to 165 °C and pumped to deaerators where the remains of gaseous phase and corrosion-inducing gases are removed. From steam separators, feed-water is led back into the reactor. The nominal coolant flow through the reactor is 46,000–48,000 m^3/h. Steam flow at full power is 5,440–5,600 t/h. The nominal temperature of cooling water at the inlet of the reactor is about 265–270 °C and outlet temperature is 284 °C at the pressure of 6.9 MPa. The level of water in steam separators, the percentage of steam in the reactor pressure tubes, the level at which water begins to boil in the reactor core, neutron flux and power distribution in the reactor and feed-water flow through the core have to be carefully controlled.

Soviet Nuclear Power and Chernobyl

Soviet military planners saw nuclear weapons as a guarantee against disastrous invasions such as the one Russia and her allies suffered in World War II. The threat of nuclear attack and the Soviet Union's penchant for secrecy led to the creation of "secret cities". At least nine secret nuclear sites were built. Within these cities laboured over 700,000 people the Soviet nuclear program was created. The Soviets were the first to build a 5000 kilowatt power reactor in 1954. The first American power reactor went into operation in 1957. The same year, the Soviets launched the first nuclear-powered surface ship, the icebreaker *Lenin*. The addition of nuclear power to the Soviet electrical grid would allow Siberia and the Soviet Far East to develop as earlier Soviet engineers had originally hoped. None of the downsides of nuclear power such as waste disposal and safety were mentioned in these early days. Least of all was any discussion of nuclear accidents.

Soviet nuclear reactor design followed a course different than western reactor types. The basic and most fundamental differences lay not in the principles of reactor design but in operation and containment the structures of which, in western reactors, are made of thick reinforced concrete that will in almost all instances "contain" a nuclear accident. Of 62 Soviet designed power reactors, 53 supply 10% of the total electricity for Russia. This compares with 103 U.S. reactors supplying between 16 to 19% of electrical supply. 47 of the Soviet designs are pressurized water reactors called VVERs that are similar to such designs elsewhere. Different VVERs models do not have concrete containment. The remaining designs are graphite-moderated, water cooled boiling water reactor designs called RBMK. Moderators are materials that dampen or absorb neutron emissions by reactors. In the VVERs and western designs, water is used for moderating neutron "flux" keeping the reactor within prescribed power levels. The RBMK design is uncommon and served two purposes – generate electricity and weapons-grade plutonium. The reactor known as Chernobyl Unit 4 was a RBMK design.

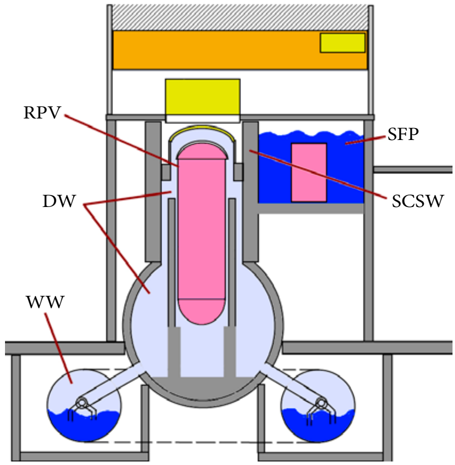
Chernobyl is by far the most widely-known nuclear accident. Before 26 April 1986, there had been perhaps larger releases of radioactivity from secret city sites such as Kyshtym near Chelyabinsk in the 1950s. These earlier incidents were not reactor-related. The Chernobyl accident was an explosion that vaporized 50 tons of uranium fuel and resulted in the remainder of the core melting into a glassy mix of reinforced concrete and fuel underneath the former reactor. Only at the Three Mile Island accident in 1979 had been a meltdown. The difference between Three Mile Island and Chernobyl was the lack of a concrete containment building in Chernobyl. The difference was catastrophic.

The details of the Chernobyl disaster are given in many excellent accounts of the accident. At 1:24 A.M., on a Saturday, reactor operators were running an unauthorized safety test to determine an RBMK's response to a power outage that would stop steam from flowing to turbo generators. A series of mistakes, any of one which by itself would not have been fatal, when taken together led to the first of two explosions. The most telling mistake was a decision on disabling the cooling system of the reactor to allow the test to continue. When the reactor was forced to increase its power level by removing its control rods – internal moderators located within the core – the reactor overheated as its fission rate increased. The Chernobyl operators attempted to manually reinsert control rods but heat had deformed metal guides for the rods and the reactor went out of control. The parts of the core melted down mixing with cooling water producing explosive hydrogen gas. The first explosion was clearly of this mixture, whereas the second may have been a nuclear one.

It has been a mantra in the design of domestic power reactors that such purely nuclear explosions cannot occur. This is likely true for most designs. Again, the RBMK is different. Theoretically, an RBMK can, and at Chernobyl, likely did. Fallout from the blasts spread nuclear radioactive waste downwind across the Ukraine and Western Europe. Thyroid cancer incidence and leukaemia in children has increased dramatically in Belarus, Ukraine and the Russian Federation. The increase is seen out to 500 kilometres from Chernobyl.

Today, Unit 4 is entombed in "The Sarcophagus". The concrete and steel structure is deteriorating due to the haste of its construction and intense radioactivity still present at the site. In 1997, Ukraine requested 3 billion dollars in aid to construct a permanent tomb for the Chernobyl reactor. Chernobyl's ultimate damage was to nuclear power industry itself. It remains to be seen what the future of nuclear power will be into the millennium.

The Fukushima nuclear disaster on 11 March 2011 was the largest nuclear disaster since the Chernobyl disaster of 1986. The disaster was evoked by the earthquake and tsunami and emerged from a series of equipment failures. The plant comprises six separate boiling water reactors. At the time of the quake, Reactor 4 had been de-fuelled while 5 and 6 were in cold shutdown for planned maintenance. Immediately after the earthquake, the remaining reactors 1–3 shut down automatically, and emergency generators came online to power electronics and coolant systems. However, the tsunami following the earthquake quickly flooded low-lying rooms in which emergency generators were housed. The flooded generators failed, cutting power to critical pumps that must continuously circulate coolant water through a nuclear reactor for several days in order to keep it from melting down after being shut down. As the pumps stopped, the reactors overheated, and, due to normal high radioactive decay, heat was produced in the first few days after the shutdown of the nuclear reactor. As water boiled away in the reactors and water levels in fuel rod pools dropped, reactor fuel rods began to overheat severely and melt down (Figure 23.7). In the intense heat and pressure of the melting reactors, a reaction between nuclear fuel metal cladding and the remaining water surrounding them produced explosive hydrogen gas. As workers struggled to cool and shut down the reactors, several hydrogen-air chemical explosions occurred. At this point, only prompt flooding of the reactors with seawater could have cooled the reactors quickly enough to prevent meltdown. Sea water that had been exposed to the melting rods was returned to the sea radioactive in large volumes for several months until recirculating units could be put in place to repeatedly cool and re-use a limited quantity of water for cooling.



DW – drywell, WW – wetwell, SFP – spent fuel pool, SCSW – secondary concrete shield wall, RPV – reactor pressure vessel

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Figure 23.7. Cross-section sketch of a typical BWR Mark I containment

The total amount of radioactivity released into the atmosphere was approximately one-tenth as much as was released during the Chernobyl disaster. Significant amounts of radioactive material have also been released into ground and ocean waters. A few of the plant's workers were severely injured or killed by disaster conditions (drowning, falling equipment damage, etc.) resulting from the earthquake. There were no immediate deaths due to direct radiation exposures.

Dangerous Neighbourhood

The Belarusian Nuclear Power Plant will be built at the Ostrovets site, 45 kilometres from Vilnius, the capital city of Lithuania. The project foresees the construction of two nuclear reactors between 2016 and 2020, and probably two more reactors by 2025. The pressurised water reactors *Water-Water Power Reactor* type will be installed.

Kaliningrad Nuclear Power Plant will consist of two VVER-1200/491 pressurized water reactors. The reactors have a capacity of 1150 MWe each. The first reactor is planned to be operational by 2017 with onsite construction started in 2010, and the second build-out will be from 2012 to 2018. The construction site is only 14 kilometres from Lithuanian border. Kaliningrad Nuclear Power Plant will be isolated from electricity supplies from Russia if the Baltic States de-synchronize themselves from the Russian electricity grid and join the synchronous grid of Continental Europe.

Shift to Fusion

As carbon-based fuels grow increasingly scarce in the face of ever-growing demand, new and more sustainable sources of energy will be necessary to meet global energy needs. Fusion power has the potential to provide sufficient energy to satisfy mounting demand, and to do so sustainably, with a relatively small impact on the environment.

Nuclear fusion is a nuclear reaction in which two or more atomic nuclei collide at very high speed and join to form a new type of an atomic nucleus (Figure 23.8).

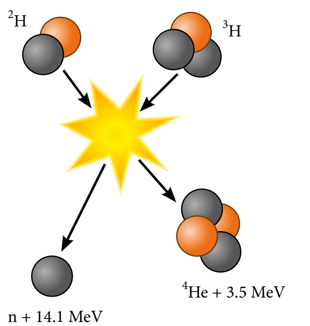

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Figure 23.8. A nuclear fusion reaction – fusion of deuterium with tritium creating helium-4, freeing a neutron, and releasing 17.59 MeV of energy

If the matter is sufficiently heated (hence being plasma), the fusion reaction may occur due to collisions with extreme thermal kinetic energies of particles. This process is called *thermonuclear fusion*, and is the only one which seems to be useful for obtaining fusion energy. It takes considerable energy to force nuclei to fuse, because all nuclei have a positive charge due to their protons, and as like charges repel, nuclei strongly resist being put close together. Accelerated to high speeds, nuclei can overcome this electrostatic repulsion and be forced close enough for attractive nuclear force to be sufficiently strong to achieve fusion. High temperatures give nuclei enough energy to overcome their electrostatic repulsion.

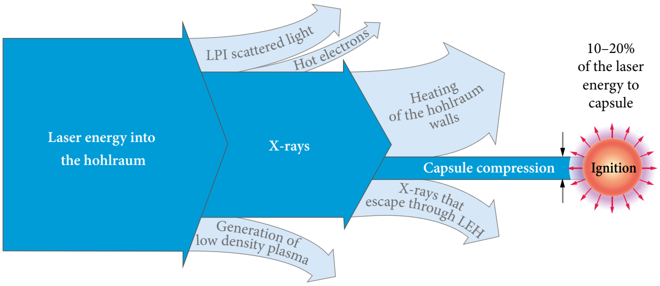

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Figure 23.9. Laser energy to hohlraum x-ray to target capsule energy coupling efficiency

The goal of “igniting” of the fusion fuel is the releasing more fusion energy than the laser delivers to the target. The laser beam heats a small metal cylinder. The heat causes the cylinder, known as a hohlraum (German for “hollow room”, or cavity), to re-emit the energy as intense X-rays, which are more evenly distributed and symmetrical than the original laser beams. The laser will generate 3 megajoules of infrared laser energy. About 1.5 megajoules of this is left after conversion to UV, and about 15 percent of this is lost in the x-ray conversion in the hohlraum. About 15 percent of the resulting x-rays (or about 150 kilojoules) will be absorbed by the outer layers of the target (Figure 23.9).

The resulting inward directed compression is expected to compress the fuel in the center of the target to a density of about 1,000 g/cm^3. For comparison, lead has a normal density of about 11 g/cm^3. It is expected this will cause about 20 MJ of fusion energy to be released. Improvements in both the laser system and hohlraum design are expected to improve the energy generation up to 100–150 MJ of fusion energy. An economical fusion reactor would require that the fusion output be at least an order of magnitude more than this input.

For deuterium and tritium, optimal reaction rates occur at temperatures on the order of 100,000,000 K. In 1992, the Joint European Torus (JET tokamak achieved the ignition of its fuel. Ignition is considered a key requirement if fusion power is to ever become practical. “Ignition” refers to the point at which the energy given off in the fusion reactions currently underway is high enough to sustain the temperature of the fuel against all losses of energy, so that fusion reactions can continue. This was an important milestone for the international fusion research community. The fusion of lighter nuclei, which creates a heavier nucleus and often a free neutron or proton, generally releases more energy than it takes to force nuclei together. This is an exothermic process that can produce self-sustaining reactions. Energy released in

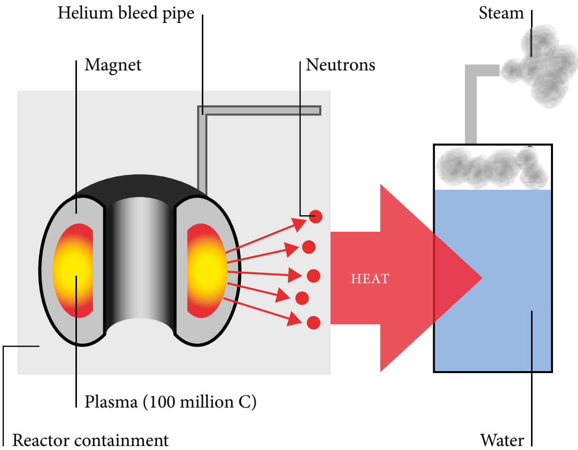

Figure 23.10. Scheme for a fusion reactor

most nuclear reactions is much larger than in chemical reactions, because binding energy that holds a nucleus together is far greater than that holding electrons to the nucleus (Figure 23.10). For example, ionization energy gained by adding an electron to a hydrogen nucleus is 13.6 eV, i.e. less than the one-millionth of the 17 MeV released in the deuterium-tritium reaction.

The ITER (*International Thermonuclear Experimental Reactor*) is an international nuclear fusion research and engineering project currently building the world's largest experimental tokamak nuclear fusion reactor at the Cadarache facility in the south of France. The ITER project aims to make the long-awaited transition from experimental studies on plasma physics to full-scale electricity-producing fusion power plants. The project is funded and run by seven member entities – the European Union, USA, India, Japan, China, Russia and the South Korea.

The ITER fusion reactor is expected to demonstrate the principle of producing more energy from the fusion process than is used to initiate it, something that has not yet been achieved with the previous fusion reactors. The construction of the facility began in 2007, and the first plasma is expected to be produced in 2020. The predicted start of deuterium-tritium operation is planned in 2027.

The ITER is designed for producing approximately 500 MW of fusion power sustained for up to 1,000 seconds by the fusion of about 0.5 g of the deuterium/tritium mixture in its approximately 840 m^3 reactor chamber. The ITER is expected to produce (in the form of heat) 10 times more energy than the amount consumed to heat up plasma to fusion temperatures.

At such high temperatures, particles have vast kinetic energy, and hence velocity. The hydrogen bomb was an unconfined, runaway fusion reaction. Producing a confined, controlled fusion reaction has proven to be the hard part. If unconfined, particles will rapidly escape thus taking energy with them and cooling plasma to the point where net energy is no longer produced. A successful reactor would need to contain particles in a small enough volume for a long enough time for much of plasma to fuse. In magnetic confinement reactors, plasma, the gas of charged particles, is confined using magnetic fields. A solid confinement vessel is also needed both to shield magnets and other equipment from high temperatures and energetic photons and particles and to maintain a near-vacuum for plasma to populate. The containment vessel is subjected to a barrage of very energetic particles, the place where electrons, ions, photons, alpha particles and neutrons constantly bombard it and degrade the structure. Engineering task is to reduce that interaction and to develop materials that can stand up to both neutron bombardment and great heat (millions of degrees) generated by plasma. Beyond the inner wall of the containment vessel, one of several test blanket modules will be placed. These are designed to slow and absorb neutrons limiting damage to the rest of the structure and breeding tritium for fuel from lithium and the incoming neutrons. Energy absorbed from fast neutrons is extracted and passed into the primary coolant. This heat energy would then be used for powering an electricity-generating turbine in a real power plant. The external diameter of the ITER vacuum vessel will measure 19.4 metres while the internal one – 6.5 metres. Once assembled, the whole structure will be 11.3 metres high and 5,116 tons weight. The vacuum vessel is the central part of the ITER machine in which plasma is contained by means of magnetic fields. The main vessel is a double walled structure with poloidal and toroidal stiffening ribs between 60 millimetres thick shells to reinforce the vessel structure. These ribs also form flow passages for cooling water. The space between the double walls will be filled with shield structures made of stainless steel. The inner surfaces of the vessel will be covered with blanket modules. These modules will provide shielding from high-energy neutrons produced by fusion reactions and some will also be used for tritium breeding concepts.

After the expected success of the ITER experimental nuclear fusion reactor, the development will be continued referring to the DEMO (DEMOnstration Power Plant) Nuclear Fusion Power Plant. The goal of the DEMO will be to produce at least four times that much fusion power on a continual basis than the ITER fusion reactor. Moreover, while the goal of the ITER is to produce 10 times as much power as is required for breakeven, the goal of the DEMO is to produce 25 times as much power. 2 to 4 GW of thermal output produced by the DEMO will be on the scale of a modern electric power plant. Also, notably the DEMO is intended to be the first fusion reactor to generate electrical power. Earlier reactors, such as the ITER, merely dissipate thermal power they produce into the atmosphere as steam.

To achieve its goals, DEMO plasma density will be about 30% greater than that of the ITER. As a prototype, the commercial fusion reactor DEMO could make fusion energy available by 2033.

In fusion reactors, some of the components will become radioactive due to neutrons impinging upon them. It is hoped that plasma facing materials will be developed so that wastes produced in this way will have much shorter half lives than waste from fission reactors, with the waste remaining harmful for less than one century. The process of manufacturing tritium currently produces long-lived waste, but both the ITER and DEMO will produce their own tritium, dispensing with the fission reactor currently used for this purpose.

PROTO is a beyond DEMO experiment, a part of the European Commission long-term strategy for research on fusion energy. PROTO would act as a prototype power station, taking in any remaining technology refinements and demonstrating electricity generation on a commercial basis. The expected timeline is post-2050. This might possibly make PROTO the first commercial nuclear fusion power plant in the world.

Fusion reactors confront numerous technically challenging issues. Nobel laureate in physics, Pierre-Gilles de Gennes, said of nuclear fusion, "We say that we will put the sun into a box. The idea is pretty. The problem is, we don't know how to make the box". A technical concern is that the neutrons produced by fusion reactions will damage the materials from which the reactor is built. Research is in progress to determine how reactor walls can be designed to long and intense neutron bombardment. The damage is primarily caused by high energy neutrons knocking atoms out of their normal position in the crystal lattice. A related problem is that neutron bombardment will induce radioactivity in the reactor material itself. Maintaining and decommissioning a commercial reactor may thus be difficult and expensive. Another problem is that superconducting magnets are damaged by neutron fluxes.

Proponents believe that the amount of radioactive waste produced should be hundreds of times less than that of a fission reactor, produce no long-lived radioactive waste and be impossible for any such reactor to undergo a large-scale runaway chain reaction. In addition, the amount of fuel contained in a fusion reactor chamber (one half gram of deuterium/tritium fuel) is only sufficient to sustain the fusion burn pulse from minutes up to an hour at most, whereas a fission reactor usually contains several years' worth of fuel. In the case of an accident, it is expected that a fusion reactor might release far less radioactive pollution than would an ordinary fission nuclear plant. The ITER fusion reactor does not produce fissile materials necessary for the construction of a weapon.

Nuclear accidents, because of the powerful forces involved, are often very dangerous. Unlike convention weapons, intense light, heat, explosive force and radiation exposure are the deadly components of a nuclear weapon. Civilian nuclear and radiological accidents primarily involve nuclear power plants. Most common are nuclear leaks that expose workers to hazardous materials and a more serious hazard of releasing a nuclear material into the surrounding environment. The most significant accidents were meltdowns occurred at Three Mile Island in the USA and Chernobyl in the Soviet Ukraine. The earthquake and tsunami in 2011 caused serious damage to three nuclear reactors and a spent fuel storage pond at the Fukushima Daiichi Nuclear Power Plant in Japan. Military accidents usually involve the loss or unexpected detonation of nuclear weapons. The USA nuclear bomb tested in 1954 produced a larger yield than expected, which contaminated nearby islands, a Japanese fishing boat (with one fatality) and raised concerns about contaminated fish in Japan. In the 1950s through 1970s, submarines and aircrafts lost several nuclear bombs some of which have never been recovered.

CHAPTER 24

Chemical Engineering: Fuel, Materials and Fertilizers

This chapter will help in

- introducing chemical engineering in the 18^{th} – 19^{th} centuries;
- discussing the manufacturing of synthetic petroleum in the 20^{th} century;
- explaining the development of chemical engineering in Lithuania (fertilizers, PET resin, paints, biotechnological pharmacy);
- presenting the process of recycling polymers.

Soap and Synthetic Petroleum Production

Chemical engineering was developed in the late 19^{th} century. Before the Industrial Revolution (18^{th} century), industrial chemicals and other consumer products such as soap were mainly produced through batch processing, which is labour intensive, and individuals mix predetermined amounts of ingredients in a vessel, heat, cool or pressurize the mixture for a predetermined length of time. The product may then be isolated, purified and tested to achieve a saleable product. Batch processes are still performed today on products such as pharmaceutical intermediates, perfumes, paints, pure maple syrups, etc. The Industrial Revolution led to an unprecedented escalation in demand, both with regard to quantity and quality, for bulk chemicals such as sulfuric acid and soda ash. This meant two things: first, the size of activity and the efficiency of operation had to be enlarged, and second, serious alternatives to batch processing such as continuous operation, had to be examined. Today, commodity chemicals and petrochemicals are predominantly made using continuous manufacturing processes, whereas speciality chemicals, fine chemicals and pharmaceuticals are made using batch processes.

By 1840, Germany had advanced industrial chemistry to the modern stage of development. The variety of chemicals ranged from dyes to drugs. The U.S.

began exploiting its own vast resources of coal, petroleum, phosphate and sulfur to quickly offset the German advantage in explosives technology, acids and "fine chemicals".

Coal and petroleum provided raw materials and power, phosphate provided fertilizer and nitric acid and sulfur provided all important acid, H_2SO_4. Coal tar derivatives, including aspirin (acetyl-salicylic acid), novocaine and phenacetin, were duplicated by the U.S. Artificial fibres like acetylated cellulose produced "rayon." Cellulose and coal tar, subjected to heat and pressure, are the raw materials of plastic of Bakelite. Further research led to long-chain polymers or giant molecules such as nylon.

Synthetic fuels were advanced by German petrochemical engineers. Forced to rely on its highly developed chemical industry, Germany in World War II, pioneered in the production of synthetic fuels and gas from coal. The U.S. has coal reserves that include coal and lignite. In wartime, Germany utilized its "brown coal" for power generation and chemistry. This "poor" coal is very similar to lignite first discovered to have fuel properties by the Greeks of Antiquity who noted that the addition of fuel oil allowed it to support combustion and burn itself. Germany knew this and developed coal-fired plants that were virtually undetectable due to pollution control systems such as electrostatic separators and chemical scrubbers.

German chemists succeeded in producing synthetic petroleum from coal in 1913. A German industrial giant, I.G. Farben, increased the yield and quality of the gasoline produced. The manufacturing process was separated into two steps: the hydrogenation of coal to a liquid phase and the decomposition of heavy oil to lighter, gasoline-size molecules in a vapour phase over catalyst beds. Germany's synthetic petroleum industry produced 128 million barrels for the period 1938-1945 reaching a peak of 25.5 million barrels in 1944. After Germany's defeat, the bulk of synthetic petroleum plants were dismantled.

U.S. experiments were successful in producing coal-derived petroleum that was unable to economically compete with natural petroleum. As the world's petroleum reserves decline, the point should be reached where coal-derived petroleum oil will become a viable alternative. It may never power the family vehicle but could provide lubricants and other distillates of strategic and industrial value.

Development of Lithuanian Chemical Engineering

The first small soap factories in Lithuania appeared in the beginning of the 19th century. In 1900, about 4 000 t of soap were produced. During the 3rd decade of the 20th century, soap factories in Kaunas, Klaipėda and Panevėžys were built. In 1937, about

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Figure 24.1. *Lifosa*, a producer of nitrogen-phosphorus fertilizer (Jonava city, Lithuania)

2 800 t of soap were made. In 1979, soap manufacturing reached 22 000 t of soap per year. The production of chemicals and plastics was commenced only on the second half of the 20th century. Today, chemicals and plastics, the two closely related sectors, stand out as some of the best-established industrial activities in Lithuania. The production of the main chemicals, fertilizers and primary forms of plastics accounts for over 80 percent of the country's chemical industry turnover (Figure 24.1).

Photo courtesy of PKN *Orlen*

Figure 24.2. View of the refinery

Key manufacturers include the producers of nitrogen and nitrogen-phosphorus fertilizers and the makers of *polyethylene terephthalate* (PET) resin, paints and biotechnological pharmacy. Mažeikiai refinery has a design processing capacity of 15 million tons of crude oil per year. However, it is more efficient to process around 8 million tons of crude oil while using the remaining capacity for processing other feedstock. Mažeikiai refinery is the only in the Baltic States (Figure 24.2).

Cement Production

Cement is a binder, a substance that can bind other materials together. Cement used in construction can be characterized as being either *hydraulic* or *non-hydraulic*, depending upon the ability to be used in the presence of water. Non-hydraulic cement will not set under wet conditions or underwater and is made by the cement mix with activated aluminium silicates such as fly ash. This allows setting in wet condition or underwater. Portland cement is the most common type of cement in general use around the world. This cement is made by heating limestone (calcium carbonate) with small quantities of other materials such as clay to 1450 °C in a kiln. In the process known as *calcination,* a molecule of carbon dioxide is liberated from calcium carbonate to form calcium oxide which is then blended with the other materials that have been included in the mix. The resulting hard substance, called "*clinker*", is then ground with a small amount of gypsum into a powder to make ordinary Portland cement.

At the Northern edge of Lithuania, the town of Naujoji Akmenė, the only cement manufacturer in Lithuania is located. In 1952, the operation of the first rotary kiln burning the clinker started. In 1970, clinker burning kilns of 185 m in length and 5 m in diameter started operating. The maximum capacity of these kilns was 3, 4 mil. t of cement per year. The preparation stage of the raw material in the wet manufacturing process results in the production of a *raw mix* that is in a suitable state for feeding the kiln in which it is converted by heat into clinker (Figure 24.3, b). This is a chemical transformation. The raw mix consists of a mixture of materials that will react together to form calcium silicates that confer on clinker its strength-giving properties. Mineral particles in the raw mix usually consist principally of calcium carbonate from the limestone component of the mix and alumina-silicates from clay or shale components, together with a certain amount of quartz (silicon dioxide). Slurry moisture about 35%. In 1930, Germany made first attempts to redesign the kiln system to minimize the waste of fuel. This led to the development of *raw mix preheaters*. The *gas-suspension preheater* is a conical vessel where a dust-bearing gas-stream produces a vortex within the vessel. If the entire feed of the raw mix is encouraged to pass through the cyclone, it is found that a very efficient heat exchange takes place and the raw mix is efficiently heated. The number of cyclone stages used in practice var-

a)

Photo by A. V. Valiulis

b)

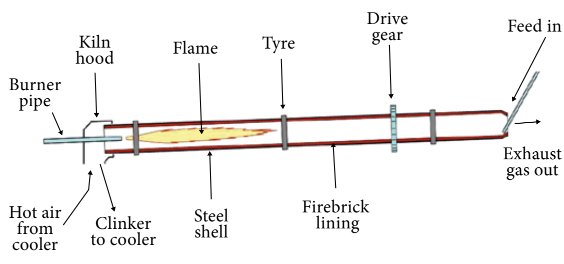

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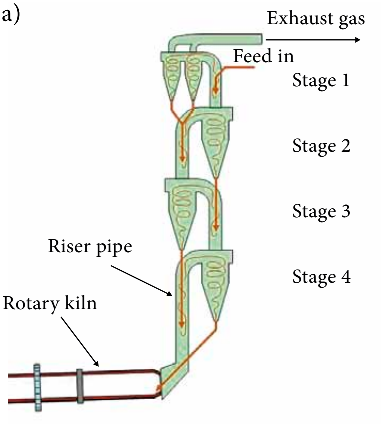
Figure 24.3. The cement manufacturer in Naujoji Akmenė: a – circular slurry mixer basins of plants using the wet process, b – general layout of a rotary kiln

a)



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b)

Photo by A. V. Valiulis

Figure 24.4. Conversion from the wet production process: a – 4-stage pre-heater, showing the path of feed; b – rotary kiln of the dry production process

ies from 1 to 6 (Figure 24.4). Typical cyclones have 6 m in diameter. The most commonly encountered suspension preheaters have 4 cyclones. The raw mix preheating in the upper cyclone reach 300 °C, whereas that in the lower one – 1000 °C. As the hot feed that leaves the base of the preheater string is typically 20% calcined, the kiln has less subsequent processing to do, and therefore can achieve a higher specific output.

Conversion from the wet production process (used for more than 50 years) to the dry one started in Naujoji Akmenė in 2006 (Figure 24.4, b) The new technology significantly increases the efficiency of the plant and reduces power consumption. The line should operate for around 50 years.

Plastic Recycling

The history of Vilnius factory "Plasta" dates back to 1961. The factory started manufacturing household articles of moulded plastic and various elements of pressed plastic. During 1961–1965, producing high density polyethylene pipes and low density polyethylene films was commenced. In 1966–1970, the moulding-blowing method of manufacturing polyethylene packages was introduced. In the beginning of the 1970s, the company started launching polyethylene recycling technologies that made the utilisation of the existing resources of raw materials more efficient.

In 1995–1996, the factory was equipped with a modern productions-line for garbage bags. In 1997–1998, "Plasta" launched the production of new generation polypropylene (PP) indoor drainage systems.

Plastic is a substance which, due to its peculiar chemical composition, hardly dissociates – the process may last up to several decades. Even though the recycling technology of plastics is very complicated and expensive, recycling plays an important part in environmental protection – it helps to save thousands of tons of primary raw material produced using exhaustible natural recourses like oil.

Recycling secondary raw material starts from conscious handling with worn out materials that have already served their function – these are mainly packages. If they go to the common stream of garbage, there is a risk that will end as waste at a disposal site. However, if worn out packages go to special dumpsters or other collection sites – they become a valuable raw material (Figure 24.5). Following a number of technological processes, the collected PE secondary raw material turns into a semi-manufactured product – granules. Thanks to the complex technological process, factories can convert various worn out polyethylene packages – bags, hoods, canisters, etc. – into valuable products: garbage bags, films for various purposes, pipes, canisters, buckets, barrels, roof tiles, plastic sheets. Choosing this way has several advantages: environment is not polluted, and exhaustible natural resources are preserved.

a)

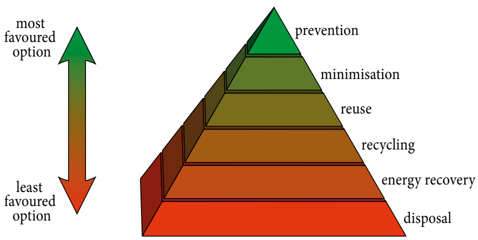
b)

Figure 24.5. Recycling plastics: a – plastic scrap, b – converted material – PE granules

Incineration of Waste

There is plenty of garbage on our surrounding. It spreads pollutions and diseases. Municipal solid waste to a large extent is biogenic, e.g. paper, cardboard, wood, cloth, food scraps. Energy-from-waste is the process of generating energy in the form of electricity and/or heat from the incineration of waste (Figure 24.6, b). Consequently, this energy is often recognised as renewable energy.

a)



b)

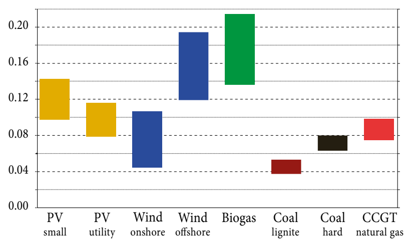


Figure 24.6. Diagram of the waste hierarchy (a). Levelized cost of electricity in euro per kWh. PV – photovoltaic, CCGT – integrated coal-gassification combined cycle (b)

a) b)

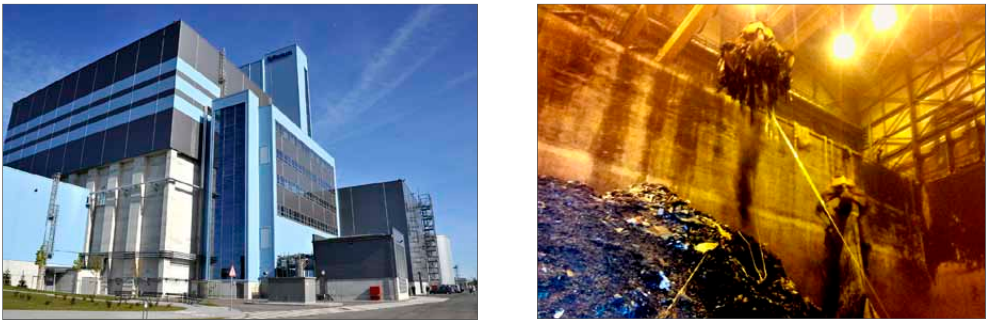


Figure 24.7. An incineration plant in Klaipėda (Western Lithuania) (a). Waste transfer to an incinerator (b)

Incineration is the combustion of organic material such as waste with energy recovery (Figure 24.7, a). Incineration generally entails burning waste to boil water which powers steam generators that make electric energy and heat to be used in homes and industries. Typically half of the energy content in municipal solid waste is from biogenic material. In the event that the waste was landfilled, 1 metric ton of municipal solid waste would produce approximately 62 cubic metres methane. The methane has more than twice the global warming potential than the CO_2 gas. The system works is this: a dump truck drops the municipal waste into a warehouse-sized pit. Then a giant claw grabs nearly a truckload of garbage and dumps it into an incinerator (Figure 24.7, b). Incinerator is burning garbage at more than 1000 °C. Incineration has a number of outputs such as the ash and the emission to the atmosphere of *flue gas*. The flue gases may contain significant amounts of particulate matter, heavy metals, dioxins, furans, sulfur dioxide, methane, and hydrochloric acid. Incineration plants must be designed to ensure that the flue gases reach a temperature of at least 850 °C in order to ensure proper breakdown of toxic organic substances. If plants have adequate controls, these outputs can't add a significant pollution component to stack emissions.

CHAPTER 25 | Biomedical Engineering

This chapter will help in

- explaining activity at the interface of medicine and engineering (electrocardiogram, electroencephalogram, electromyogram, thermography);
- describing radiology, nuclear medicine and radiation;
- introducing ultrasound, magnetic resonance imaging and computed tomography;
- considering the application of radioisotopes (positron emission tomography, tomotherapy, radiation therapy);
- defining the operation of pacemakers and defibrillators;
- presenting the creation of artificial organs (artificial heart, artificial kidney).

Biomedical engineering is one of the newest areas of modern engineering and describes activity at the interface of medicine and engineering. Biomedical engineering (BME) is the application of engineering principles and design concepts to medicine and biology. This field combines the design and problem solving skills of engineering with medical and biological sciences to advance healthcare treatment, including diagnosis, monitoring, treatment and therapy. Prominent biomedical engineering applications include the development of biocompatible prostheses, various diagnostic and therapeutic medical devices ranging from clinical equipment to micro-implants, common imaging equipment, regenerative tissue growth, pharmaceutical drugs, etc. This dispassionate characterization of the physiological process into hydraulic (heart, circulatory system), pneumatic (respiratory system) and communication (nervous system) systems allows a biomedical engineer to design monitoring and control systems taking into account the human “hemodynamic system” and other organs.

One of the first attempts to monitor body functions involved the recognition of bioelectric potentials associated with nervous activity, heartbeat, muscular activity, etc. Luigi Galvani, in 1786, recognized that electricity was generated by the body. To measure these potentials a transducer capable of converting ionic potential and currents into electric counterparts is necessary. The first ones being developed were a simple two-pronged electrode that recorded characteristic wave forms associated with electric fields generated by ionic currents. The examples of the latter ones are those of the electrocardiogram and electroencephalogram.

The earliest electrocardiogram (EKG) had bioelectric potentials made with immersion electrodes-buckets of salt water in which the subject placed one hand and one foot. This principle was developed in 1903. The EKG was built in 1912 and embodied most of the functions familiar in today's machines.

The electroencephalogram (EEG) measures neuronal activity of the brain. The correlation of EEG patterns with specific brain activity is only a relatively general level at present-delta, theta, alpha and beta rhythms. Other types of bioelectric potential devices include the EMG (electromyogram), EGG (electroretinogram) and EOG (electrooculogram). These devices are used for recording activity patterns of the eye and muscles. Modern intensive care units utilize variations in these types of monitoring systems with the addition of instrumental observations of respirations and temperature.

Monitoring skin temperature by infrared devices has led to the development of the *thermograph* – a scanning device that maps infrared energy onto photographic paper producing a thermogram. Mapped variations in temperature over the body can determine the registration of the areas of anomalous thermal activity. Temperature measurement is done for diagnostic rather than monitoring purposes.

Biotelemetry is the measurement of biological parameters at a distance. A Dutch doctor W. Einthoven invented the electrocardiogram in 1903. The first distant transmission of electrocardiograms from a hospital to a laboratory was provided in 1903. Modern biotelemetry uses radio telemetry. As diagnostic devices, X-ray and radioisotopic systems, ultrasonic systems and particle beam accelerators were created. As control systems, pacemakers, defibrillators and artificial organs were designed. A medical device is intended for use in establishing the diagnosis of a disease or other conditions, or for cure, mitigation, treatment or prevention of the disease. Some examples include pacemakers, infusion pumps, heart-lung machine, dialysis machines, artificial organs, implants, artificial limbs, corrective lenses, cochlear implants, ocular prosthetics, facial prosthetics, somato prosthetics (may include custom made breast prostheses, fingers, partial/total hands, toes and partial/total feet) and dental implants. Prostheses are generally used for replacing parts lost by injury (traumatic) or missing from birth or for supplementing defective body parts. Inside the body, artificial heart valves are in common use with artificial hearts and lungs seeing less common use but under active technology development.

Biomedical imaging is a major segment of medical devices. This area deals with enabling clinicians to directly or indirectly "view" things not visible in plain sight. This can involve utilizing ultrasound, magnetism, UV, other types of radiology or means. Imaging technologies are often essential to medical diagnosis and are typically the most complex equipment, including fluoroscopy, magnetic resonance imaging (MRI), nuclear medicine, positron emission tomography (PET), projection radiography such as X-rays and computed tomography (CT) scans, ultrasound, optical microscopy and electron microscopy found in a hospital.

Radiology

Radiology uses an array of imaging technologies such as X-ray radiography, ultrasound, computed tomography, nuclear medicine, positron emission tomography and magnetic resonance imaging to diagnose or treat diseases. Plain radiography was the only imaging modality available during the first 50 years of radiology. X-rays were discovered by Conrad Roentgen in 1895. Radiographs (or roentgenographs) are produced by transmitting X-rays through a patient. This penetrating radiation was quickly recognized for its importance to medical diagnoses. Later, it was utilized for therapy. X-rays are generated when fast-moving electrons suddenly decelerate upon striking a target. With wave-lengths in the 0,1–10 nanometers range they have little difficulty passing through most matter. X-rays strike an undeveloped film, which is then developed chemically, and an image appears on the film (Figure 25.1). Film-screen radiography, in the early 1960s, was being replaced by digital radiography (DR) where X-rays strike a plate of sensors that convert the signals generated into digital information transmitted and converted into an image displayed on a computer screen.

The images produced by X-rays are due to different absorption rates of varying tissues. Calcium in bones absorbs X-rays the most, so bones look white on a film recording of the X-ray image while fat and other soft tissues absorb less, and therefore look gray. Air absorbs the least, so lungs look black on a radiograph. Air is used as a contrast medium for examining the brain while barium sulfate gives greater contrast to the gastrointestinal tract.

Public domain

Figure 25.1. Bone radiology

Nuclear Medicine and Radiation

The widespread clinical use of nuclear medicine began in the early 1950s, as knowledge about radionuclides, the detection of radioactivity and using certain radionuclides to trace biochemical processes expanded. In nuclear medicine imaging, radiopharmaceuticals are taken internally, for example intravenously or orally. Then, external detectors (gamma cameras) capture and form images from radiation emitted by radiopharmaceuticals. By the 1970s, most organs of the body could be visualized using nuclear medicine procedures. In the 1980s, radiopharmaceuticals were designed for use in the diagnosis of heart disease. The development of single *photon emission computed tomography* (SPECT), around the same time, led to the three-dimensional reconstruction of the heart and establishment of the field of nuclear cardiology. More recent developments in nuclear medicine include the invention of the first *positron emission tomography* scanner (PET). The concept of emission and transmission tomography, later developed into single photon emission computed tomography (SPECT), was introduced in the late 1950s. PET imaging is now an integral part of oncology for diagnosis, staging and treatment monitoring.

Radionuclide therapy can be used for treating conditions such as hyperthyroidism, thyroid cancer and blood disorders. In nuclear medicine therapy, the radiation treatment dose is administered internally (e.g. intravenous or oral routes) rather from an external radiation source. Radiopharmaceuticals used in nuclear medicine therapy emit ionizing radiation that travels only a short distance, thereby minimizing unwanted side effects and damage to non-involved organs or nearby structures. Commonly used radiation sources may include Caesium-137, Cobalt-60, Iridium-192, Iodine-125, Palladium-103 and Ruthenium-106.

In biology and agriculture, radiation is used for inducing mutations to produce new or improved species. Another use in insect control is the sterile insect technique where male insects are sterilized by radiation and released, so they have no offspring to reduce the population. In industrial and food applications, radiation is used for the sterilization of tools and equipment. An advantage is that the object may be sealed in plastic before sterilization. An emerging use in food production is the sterilization of food using irradiation.

Ultrasound

Medical ultrasonography uses ultrasound (high-frequency sound waves) to visualize soft tissue structures in the body in real time (Figure 25.2). No ionizing radiation is involved, but the quality of the images obtained using ultrasound is highly dependent on the skill of the person performing the exam. *Ultrasonography* (sonography) uses a probe containing multiple acoustic transducers to send pulses of sound into a material. A sound wave is typically produced by a piezoelectric transducer. The frequencies can be anywhere between 2 and 18 MHz. Typical frequencies used include 2.25 MHz for general purpose studies, 3.5 MHz for studies of children and 5.0 MHz for paediatric echo cardiography.

Author Dr. Wolfgang Moroder. Licensed under CC BY-SA 3.0

Figure 25.2. Medical ultrasound visualization

Whenever a sound wave encounters a material with a different density (acoustical impedance), a part of the sound wave is reflected back to the probe and is detected as an echo. The time it takes for the echo to travel back to the probe is measured and used for calculating the depth of the tissue interface causing the echo. The greater is the difference between acoustic impedances, the larger is the echo. If the pulse hits gases or solids, the difference in density is so great that most of acoustic energy is reflected and becomes impossible to see deeper. Larger patients may have a decrease in image quality due to sound wave absorption in the subcutaneous fat layer. This results in less sound waves penetrating to organs and reflecting back to transducer ultimately causing a poorer quality image. Ultrasound is also limited by its inability to image through air (lungs, bowel loops) or bone. Its use in medical imaging has developed mostly within the last 30 years. Because ultrasound does not use ionizing radiation, unlike radiography, CT scans and nuclear medicine imaging techniques, it is generally considered safer.

Magnetic Resonance Imaging

Magnetic resonance imaging (MRI) examination employs strong magnetic fields to align atomic nuclei (usually hydrogen protons) within body tissues, then uses a radio signal to disturb the axis of the rotation of these nuclei and observes the radio frequency signal generated as the nuclei return to their baseline states. Radio signals are collected by a small antennae, called coils, placed near the area of interest. An advantage of MRI is its ability to produce images in axial, coronal, sagittal and multiple oblique planes with equal ease. MRI scans give the best soft

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Figure 25.3. A MRI scanner

tissue contrast of all imaging modalities. One disadvantage is that a patient has to hold still for long periods of time in a noisy, cramped space while imaging is performed (Figure 25.3). MRI has great benefit in imaging the brain, spine and musculoskeletal system.

Computed Tomography

Computed tomography (CT) imaging uses X-rays, in conjunction with computing algorithms, to image the body. The first commercially viable CT scanner was introduced in Great Britain in 1972. An X-ray tube opposite an X-ray detector (or detectors) rotates around a patient in a ring-shaped apparatus and produces a computer-generated cross-sectional image (tomogram). Detail images can be reconstructed into three-dimensional (3D) images. Radio-contrast agents are often used for the enhanced delineation of anatomy. Although radiographs provide higher spatial resolution, CT can detect more subtle variations in the attenuation of X-rays. CT exposes the patient to more ionizing radiation than a radiograph. CT scanning is used for diagnosing some urgent and emergent conditions such as cerebral hemorrhage, pulmonary embolism, tearing of the aortic wall, appendicitis, kidney stones.

Radioisotopes

Medicine has used radiation therapy as a treatment for cancer for more than 100 years. The earliest roots traced from the discovery of x-rays in 1895 by Röntgen. The field of radiation therapy began to grow in the early 1900s largely due to the work of Marie Curie who discovered radioactive elements polonium and radium in 1898. Radium was used in various forms until the mid-1900s, when cobalt therapy and caesium units came into use. Medical linear accelerators have been also used as the sources of radiation since the late 1940s. With the invention of computed tomography (CT) in 1971, three-dimensional planning became a possibility. CT-based planning allows physicians to more accurately determine dose distribution using axial the tomographic images of the patient's anatomy. The advent of new imaging technologies, including *magnetic resonance imaging* in the 1970s and *positron emission tomography* in the 1980s, has moved radiation therapy from 3-D conformal to intensity-modulated radiation therapy and to image-guided radiation therapy – *tomotherapy*.

While X-rays are radiation actively generated by devices, radioisotopic techniques rely on the inherent capacity of radioactive elements such as Phosphorus-34, Iodine-131 and Molybdenum-99 to generate emissions detectable by counting or scanning devices. By introducing these elements, specific areas of the body can be evaluated for physical variations in blood flow, organs and physiological processes of extra or intracellular respiration. These radioisotopes or elements can be counted in vitro in or outside the body. This is important because of the deleterious effect of this ionizing radiation. All natural radiation, X-ray, gamma, alpha or beta causes the ionization of surrounding tissues. The amount of tissue damage is directly proportional to the kind and amount of the specific radiation type with alpha particles (helium nuclei) being the most damaging to the biological tissue. This component of radiation effect has a useful role and has been utilized in the therapy of tumours. This form of radioisotope utilization is broadly termed *radiation therapy*.

Radiation therapy is commonly applied to the cancerous tumour because of its ability to control cell growth. Ionizing radiation works by damaging the DNA of the exposed tissue leading to cellular death. To spare normal tissues (such as skin or organs which radiation must pass through to treat the tumour), shaped radiation beams are aimed from several angles of exposure to intersect at the tumour, providing a much larger absorbed dose in that particular place rather than in the surrounding, healthy tissue. Gamma ray sources, such as Cobalt-60, have been engineered into devices to generate emissions. Even more advanced engineering is involved in the increasing use of linear accelerators to produce high energy X-rays for tumour therapy. Solid targets of various materials are placed in beams of neutrons, electrons or mesons that produce high-energy penetrating radiation of very high flux (cross-sectional intensity). This high density of radiation reduces the total patient's exposure while maximizing that of the tumour.

Pacemakers and Defibrillators

A *pacemaker* is a medical device that uses electrical impulses delivered by electrodes contacting the heart muscles to regulate the beating of the heart. The primary purpose of a pacemaker is to maintain an adequate heart rate. Sometimes a pacemaker and defibrillator are combined in a single implantable device.

The first experiments on an electrical impulse to regulate the rhythm of the human heart were conducted in 1899. An external pacemaker constructed in 1958 had 45 kg in weight, was powered by a 12 volt auto battery and was connected by electrodes attached to the heart. The implantable devices long time suffered from the unreliability and short lifetime of the mercury battery. From 1971, lithium-iodide cells have become the standard for future pacemaker designs. Two varieties of pacemakers – internal (implanted) and external have come into usage since the 1960s. Internal pacemakers are implanted with the pulse generator placed in a surgically formed pocket. Electrodes, with leads, are attached to the myocardium. External pacemakers are used on patients with temporary heart irregularities.

The pulse generator is worn on the belt or wrist with the electrodes placed over the heart area. A modern pacemaker generator is in a hermetically sealed metal case, using titanium as encasing metal.

A *defibrillator* is used when the heart's activity becomes wholly unsynchronized. This disorder is characterized by rapid irregular contractions of the myocardium. Ventricular fibrillation is the most dangerous with death occurring in minutes. The most successful development in countering fibrillation is the defibrillator. Defibrillators were first demonstrated in 1899 when small electrical shocks induced ventricular fibrillation in dogs.

The first use of the defibrillator for a human was in 1947. The first device applied a brief (0.25-1 s) burst of 60 Hz alternate current of 6 ampere counter shock that resynchronizes the heart. This technique is termed alternate current defibrillation. In 1962, a new method of defibrillation was developed and included a capacitor charged to high voltage and rapidly discharged through electrodes across the chest of the patient. The amount of energy applied to the heart varies between 100-400 Watt second or joules. The discharge is for 5 microseconds.

Today, portable defibrillators are among the many very important tools carried by ambulances. Gradual improvements in the design of defibrillators have led to the availability of automated external defibrillators. These devices can themselves analyze the rhythm of the heart, diagnose shockable rhythms and charge to treat. This means that no clinical skill is required in their use, which allows lay people to respond to emergencies effectively.

Artificial Organs

In the world of medicine, one of the greatest advancements has been the ability to create artificial organs that are able to restore the proper function of a patient's body. They can be used both for functions that are essential to life and for purposes that are not related to survival but do improve the person's quality of life. The organs that can be replaced artificially are quite numerous and may include ears, ovaries and even the heart and lungs. Perhaps the most common manifestation of an artificial organ is found with mechanical aids that are used for improving the person's ability to hear and distinguish sounds. As cochlear implants have become less expensive and more available, it is thought that the worldwide incidence of deafness will decrease dramatically. One case of a survival situation where an artificial organ will make the difference between life and death is in a heart transplant. If a patient is awaiting a new heart, an artificial heart can be temporarily used for keeping the person alive until the new heart becomes available. In recent times, the models that can stand alone and provide a permanent replacement for a heart that has functional impairment have been created. The availability of artificial organs can do much to improve a person's life, from providing the essential bodily functions for survival to improving sensory capabilities such as sight and hearing.

The decades of the 1960s and 1980s saw the most dramatic biomedical engineering creations in the form of artificial devices that can substitute for natural organs performing life-sustaining processes. The three most prominent devices of this type include the artificial kidney or dialysis machine, the heart-lung machine and the artificial heart.

The first dialyzer used as an artificial kidney was built in 1913. Haemodialysis refers to the removal of undesirable molecules from blood through a semi-permeable membrane into a dialyzing bath. Artificial kidneys are then hemodialyzers. The first successful artificial kidney was designed in Netherlands in 1940 and used a rotating coil where blood was pumped through cellophane tubing round about a drum bathed in osmotic volute to remove water and associated impurities. Further improvements have involved parallel flow arrangement allowing ultra-filtration by the increased hydrostatic pressure of blood on the filter membrane. Still, the present artificial kidneys are hardly portable organs. Patients must spend at least 12 hours per week on these external machines (Figure 25.4).

An artificial kidney is a medical device that performs the function of a missing or damaged kidney, filtering blood to remove waste products and returning purified blood to the body. The technology behind artificial organs is constantly improving and being refined. The recent machines of dialysis are far from perfect.

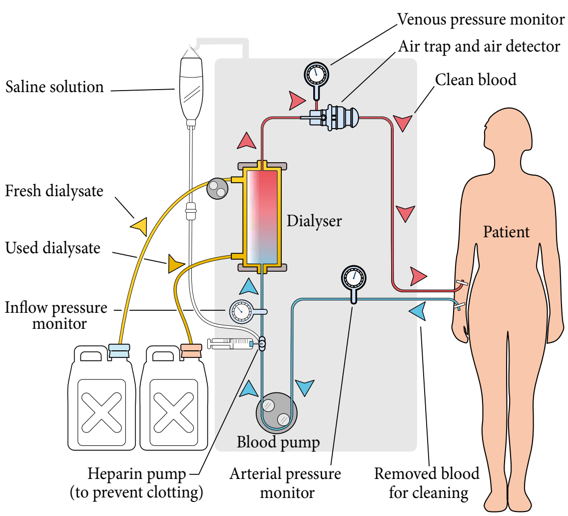


Figure 25.4. Dialysis (artificial kidney)

Their efficiency is only around 10% of that of a functioning kidney, and when used three times per week, they are incapable of controlling unhealthy fluctuations in the concentrations of metabolites such as urea in the blood.

Nanotechnology offers enticing possibilities in this area. Thin nanomembranes would be more permeable to solutes in blood than thicker conventional membranes. Such nanomembranes could be carefully engineered to contain highly selective pores instead of the randomly sized pores present in membranes used today.

Artificial Heart

An artificial heart is a device that replaces the heart. Artificial hearts are typically used for bridging the time to heart transplantation, or for permanently replacing the heart in case heart transplantation is impossible. The first implantation of an artificial component of the circulatory system was made in 1951 when a ball check-valve was placed in the descending aorta. Clinical success was tempered in subsequent cases by failures due to embolism distal to the valve or a rupture of the vessel at valve junctions. Leaflet valves were tried in the late 1950s to replace the critical aortic valve, but fatigue of material, fracture and stiffening negated early work. Better ball-valve designs of the 1960s have led to longer term success. The principal failures in late post operative stages (2 years) were generally associated with improper tissue healing around the valve implant and blood clot formation. Experimentation with partial or whole cardiac replacement began in 1967. The first artificial device was transplanted into a patient in 1981. This design functioned successfully as a heart replacement until a suitable human transplant could be found. Although other similar inventions preceded going back to the late 1940s, the first artificial heart to be successfully implanted in a human appeared only in 1982 (Figure 25.5). The first two patients to receive these hearts survived 112 and 620 days beyond their surgeries.

The challenge of designing an implantable heart substitute has evolved into a material science question as well as into a mechanical one. Newer designs have moved away from the pulsing action of the natural heart. The body's circulation does not require this cyclic mode, and non-pulsating devices show promise with low infection rates and greater ease of rate variability compared to the earlier designs. A centrifugal pump or an axial-flow pump can be used as an artificial heart resulting in the patient being alive without a pulse.

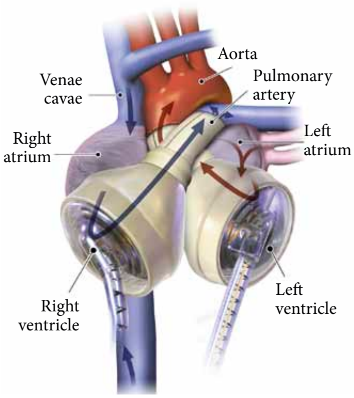


Figure 25.5. Total artificial heart

An artificial heart is made out of metal, plastic, ceramic and animal parts. A titanium-aluminium-vanadium alloy is used for the pump and other metal parts because it is biocompatible and has suitable structural properties. Titanium parts are cast at a specialized titanium processor. Except for blood-contacting surfaces, titanium is machined to a specific finish. Blood-contacting surfaces receive a special coating of titanium microspheres that bond permanently to the surface. With this coating, blood cells adhere to the surface and create a living lining. A blood-contacting diaphragm within the pump is made from a special type of polyurethane that is also textured to provide blood cell adherence. Two tubular grafts are made from polyester (used for attaching the device to the aorta) and the valves are actual heart valves removed from a pig. Other parts that make up the motor are made from titanium or other metals and ceramics.

There are several critical issues when designing an artificial heart. The biocompatible materials must be chosen; otherwise, the pump could fail. The efficiency of the motor must be optimized so that minimal heat is generated. Because of possible rejection, the total volume and surface area of the entire device should be kept as small as possible. A typical artificial heart weighs around 1,200 g. In 1996, a 46-year-old man received the world's first *total artificial heart* and kidney (Taiwan).

CHAPTER 26 Development of Public Services and Domestic Technology

This chapter will help in

- briefly describing the creation of basic public services, including water supply, treatment and distribution, sewage removal;
- explaining gas and electricity supply;
- describing waste disposal, heating and lighting;
- briefly describing the creation of the main domestic technologies (washing, bathing, toilet facilities, cleaning, laundering, ironing, home spinning and grain grinding);
- introducing the development of kitchen facilities (refrigerators, cooking employing microwave ovens, gas and electricity).

All public services in cities have been created by the cooperation of many individual technologies developed over a number of years. The services considered fall into three main categories: water supply (sources, treatment, distribution, drainage and sewage disposal), power supply (gas, electricity, hydraulics and pneumatics) and waste disposal.

Water Supply

Water supply is the basic public service. No town can survive for long without a supply of water for drinking, washing and various industrial processes. Substantial resources of materials, man-power and technological skill have thus been required in order to ensure that the need of a population for a constant supply of water can be fulfilled. Egypt and Mesopotamia civilizations, Greek city-states and Romans established intricate irrigation works, transporting water to their fields, cities and constructed aqueducts to bring water to their citizens. The remains of many of these aqueducts are evidence of the engineering skill of designers and builders. All water can be obtained from rivers, lakes, springs and simple or artesian wells. Normally, some form of a pumping or water-raising apparatus is required. All rivers with a reliable flow of water in many parts of the world have been liable to have a proportion of their volume extracted for public supply. Efforts for water supply, river control and irrigation are very ancient. One of the oldest civil engineering structures in the world is a dam built for such purposes at Sadd el-Kafara, 32 km south of Cairo, dating from the 3rd millennium B.C. Most early dams were the gravity ones. The arch dam has proved particularly suitable for large modern dams in mountainous terrain where concrete has been used as the main structural material. Techniques for dam construction improved substantially in the second half of the 19th century. In those places where surface water has not been easily available, towns and cities have exploited the potential for underground sources. In some dry regions, such supplies are vital to any permanent settlement. The construction of wells and boreholes encouraged the use of powerful pumping engines to maintain a continuous flow and thus provided the primary need for large steam engines.

Treatment. Before water can be distributed for public consumption, various processes of treatment are required to ensure that it maintains wholesome quality. The first requirement is filtration to remove the coarser forms of pollution, but the chemical treatment of dangerous organisms is also now obligatory. Slow sand filtration processes are the most popular but need substantial space for filtering beds. The impact of cholera, however, strengthened the requirements for the filtration and introduction of chlorination to kill off undesirable bacteria. The further addition of chemicals such as fluoride to harden children's teeth against decay is a matter of continuing controversy about the medication of water supply.

Distribution. The distribution of treated water to its consumers has called on the skills of engineers. In case when reservoirs were comparatively remote from the towns (100–150 km) where water supply was needed and culverts and aqueducts were necessary, the need for siphons, tunnels and pumping engines increased (Figure 26.1). In most cases, the level of reservoirs did not permit a siphon action necessary to convey water to the taps. Such systems were supplemented with water towers. Powerful steam engines were frequently installed by waterworks in the 19th century to perform this function. Water, when put under pressure by pumping or passing it through a siphon, or being conveyed underground, has to be contained in strong piping.

Water pipes made of elm trunks were quickly changed by cast-iron pipes, flanged and bolted together. For larger pipes carrying water over-ground, wrought-iron pipes were used and well suited for aqueducts carrying water across valleys and other obstacles on the routes from reservoirs to its urban consumers. Subsequently,

Figure 26.1. An 800 length aqueduct in Vilnius built in 1930

wrought-iron was replaced by mild steel which has remained in the widespread use for conveying water although concrete piping is now used for some functions. In some parts of the world with even scarcer resources of fresh water, the desalination of sea water is the sole means of maintaining essential water supplies (Kuwait, United Arab Emirates).

Drainage. In the first place, drains are required to remove surface water, and second, to allow sewage removal. With an increase in urban building and paving streets, natural means of allowing surface water to percolate sub-soil are blocked, and therefore it becomes necessary to provide artificial drains for this purpose. Rain drains protect settlements from the hazard of serious flooding.

Removal of sewage. Water supply has been an essential factor in developing methods for disposing large volumes of organic waste. As the settlement acquired reliable sources of fresh water, a proportion of this became available to flush water closets and to maintain a continuous flow of sewage waste through a network of well constructed main and tributary sewers. Most of the early systems of sewers were aimed at discharging their effluent into a river downstream from the town being served, but gradually it was recognized this was not acceptable for inland towns where other settlements depended upon the same river for their water supplies. Sewage treatment (before discharging) involved some form of the sedimentation process, with the addition of chemicals as appropriate to speed bacteriological decomposition and to remove unpleasant odours. Sewage treatment did much to remove the curse of cholera and other fever diseases from cities. Today, concrete and plastics instead of traditional constructional materials of brick and iron are used for sewers.

Gas and Electricity Supply

Gas. Coal-gas had been recognized as a by-product of coal before the 18th century at the end of which was first discovered how to apply it as a public service. Coal distillation product *coal-gas* was used for lighting and heating, and also for inflating balloons. The equipment employed was crude, and burning gas generated smoke and smells. But, by the end of 1815, some 40 km of gas mains had already been laid. For cooling and washing gas, cast-iron retorts were used. For gas purification, lime was employed. Some companies introduced oil-gas to replace coal-gas in the 1820s, using a process based on the distillation of whale oil. However, it proved to be an expensive alternative. Gas had been used mainly for illuminating streets and public places. The introduction of an atmospheric burner (around 1840) mixing air with gas just before combustion greatly enhanced its utility, and the subsequent applications in the Bunsen burner (1855) and gas-ring (1867) increased its performance as a heater in boilers and ovens. Gas cooking, however, did not become common until the 1870s. Dry gas meters also became widespread as a means of registering the amount of fuel supplied to the domestic consumer. Thanks to the incandescent gas mantle patented in 1885, gas industry continued to provide an efficient system of illumination into the 20th century. The distillation of coal in retorts underwent little change until the 1960s, although traditional horizontal retorts were gradually replaced by vertical retorts more suitable for continuous production. From the second half of the 20th century, coal-gas retorts disappeared and were replaced by gas supplied from gas or oil-gas fields.

Electricity. While a scientific knowledge of electric power increased rapidly in the 19th century, with the discovery of both chemical and mechanical means of producing a continuous current, electricity supply industry was remarkably slow. The outbreak gave a brilliant light of arc lamps and the development of electric traction in the form of the tramcar and underground train. Incandescent electric lamps in Britain and first public electric tramcar service in Berlin appeared in 1881. A number of problems regarding engines, power stations generating sufficient electricity, the power transmission system and the design of cars had to be overcome before the electric tramcar could offer a really efficient system of transport. Edison (USA) supplied direct current from his power stations. The most subsequent electrical engineers adopted alternating current at high voltage, which had substantial advantages over direct current for long-distance transmission. Gradually, electric power stations merged and by the 1920s, were being combined into national grids with fewer but larger power stations producing more and more electricity. Since the 1950s, a significant number of nuclear-fuelled power stations have also come online. The most conspicuous public service of electricity has been in the provision of street lighting. With the introduction of the gas-discharge lamp, filament bulbs were replaced by neon lights and by mercury and sodium-vapour discharge lamps, which have become generally used in the streets of human settlements all over the world. The sodium lamp, in particular, has become very popular because of its remarkable cheapness and efficiency in giving diffused yellow light. This universality of electric power has been a powerful factor in increasing the mobility of industry and population. Electric power, in short, has become the largest single public service in modern societies.

Waste Disposal

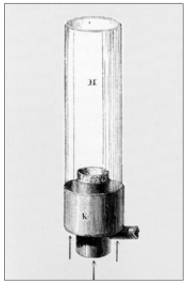
Modern life generates many forms of waste, and important public services are devoted to disposing them. The water-borne removal of organic waste is closely related to the provision of fresh water supply. Industrial and domestic garbage is collected regularly in modern societies and subjected to various forms of processing to recover metals, glass and other reusable materials. Many countries have installed expensive plants necessary to burn such garbage, using it as a fuel to heat water and generate power. The production of dangerous fumes after burning garbage is not problem-free. Some of modern industrial waste can present considerable health hazards (for example, from nuclear energy installations and certain chemical works). Normally, producers are prevented from emitting these effluents into the air and local water reservoirs by legislative controls. The maintenance of surveillance in these situations is an important public service, and modern technological equipment such as that used for monitoring the presence of radioactive materials and for protecting personnel against dangerous materials, performs a vital part of this service.

Heating and Lighting

People at home tried to keep at least one fire, usually in the kitchen, perpetually burning. At night, it was damped down, ashes drawn over embers and then covered with a curfew for safety's sake. In the morning, the curfew was removed and fresh life was blown into the embers with the assistance of a pair of bellows. From primitive times until the late 18^{th} century, there were two chief ways of creating flame by wood friction and by a tinder box. The metal box contained a piece of hard stone (flint), one of metal, which was an iron and sulphur compound, and tinder, which was of a dry, flammable material such as charred linen, dried fungi, feathers or moss. Flint was used for striking metal to pare off a tiny fragment and so create heat. The fragment fell upon tinder, ignited it and, by blowing on it, flame could be induced. The tinder box continued to be the usual way to obtain flame until matches became fairly cheap after 1850. A number of patents followed in the early 18th century. Unfortunately, yellow phosphorus-tipped matches ignited more readily. In 1890, the Salvation Army exposed that yellow phosphorus, used then in the manufacture of matches, was a massive health hazard. Amorphous red phosphorus led to the development of the safety match created in Sweden in 1855.

Lighting. Before the 1830s, the only artificial light came from candles and oil lamps. Candles were also generally home-made from tallow poured into moulds. In the late 18th century, sperma whale oil began to be used to make candles and, in the following century, paraffin wax and stearin candles were employed.

a)



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b)

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Figure 26.2. A argand lamp with a circular wick and glass chimney (a) and a modern kerosene lamp for open air (b)

Oil lamps made used fuel derived from various sources – olive oil, animal or vegetable oils, rape seed, whale oil, paraffin, etc. In 1784, a tubular wick, which was a hollow cylinder enclosed in a glass chimney, was devised (Figure 26.2 a). The resulting current of air so improved combustion that light from a single wick could be increased up to ten times. 20th century paraffin oil lamps give much brighter light because fuel is vaporized before it reaches flame, so less energy is wasted.

In 1816, London streets and public buildings obtained gas illumination. The refinement of the gas mantle revolutionized gas illumination. It gave far better light and did away with much of dirt and smell. In the Bunsen burner invented in 1855, air was mixed with gas before ignition thus giving a hotter flame (Figure 26.3, a).

a)

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b)

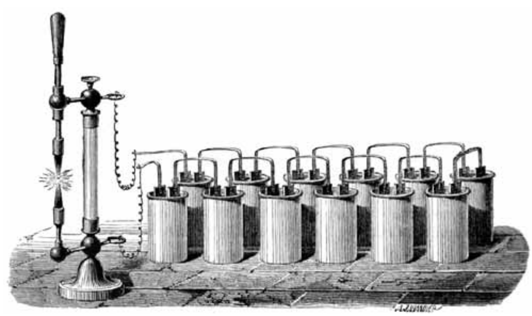

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Figure 26.3. The Bunsen burner from 1855 (a) and arc lamp from 1871 (b)

In 1871, the *arc lamp* was installed for use in streets, factories and railway stations. However, the arc lamp was totally unsuitable for domestic illumination. Light was too costly and difficult to adjust (Figure 26.3, b).

Electric lighting at home had to await the filament lamp. A successful carbon filament lamp was produced almost simultaneously (1879) by Thomas Alva Edison in America and Joseph Wilson Swan in the UK. The carbon filament lamp dominated the market until about 1910.

Meanwhile, better metal filaments, including osmium (1902), tantalum and then tungsten (1906), were devised. The tungsten filament with its high melting point of 3410 °C was very durable, but high temperatures blackened inside the bulb. In 1913, the argon-filled incandescent lamp with a coiled-coil filament was developed. Today's electric lamp gives about four times as much light as a carbon filament lamp for the same consumption of electricity. The *fluorescent lamp*, introduced in the 1930s, was developed from the Hg vapour discharge lamp. The lamps are coated inside with several different phosphorescent compounds (phosphors). Ultra-violet light produced by discharging electricity through mercury vapour in the tube is absorbed by phosphors emitting light. The colour of such lamps can be altered by using different types of phosphors. Fluorescent lamps use little electrical power, and tubes last much longer than the equivalent filament lamps.

Heating. It is well known that the ancient Romans employed a method of central heating by *hypocaust* (Figure 26.4). Such systems were used in all types of buildings, including homes. This is a heating method whereby hot air from a basement furnace is passed under the floor and through the wall flues to heat all rooms of a dwelling. The hypocaust was an under floor chamber.

With the collapse of the Roman Empire in the west, the hypocaust idea in this part of Europe fell into disuse, and it was many centuries before central heating systems were once again devised.

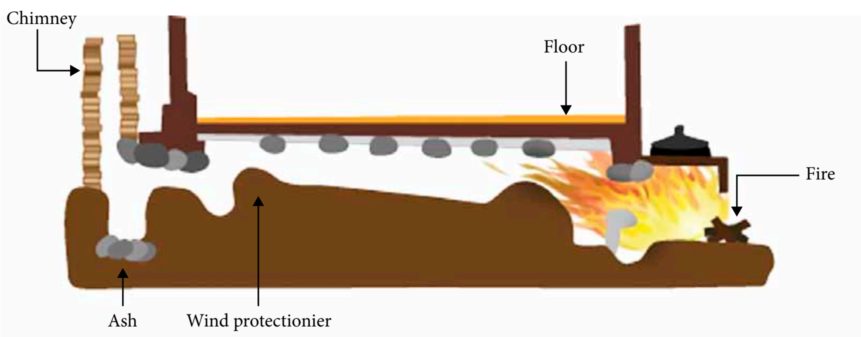

Figure 26.4. Korean hypocaust heating

In Korean traditional architecture underfloor heating (*ondol*) use direct heat transfer from wood smoke to the underside of a thick masonry floor (Figure 26.4). In modern usage it refers to any type of underfloor heating, or a hotel or sleeping room.

Until the beginning of the 17th century, most rooms were heated by wood, peat or charcoal burning on the open hearth. The hearth surface was of brick or stone. Gradually, coal began to replace timber as the normal domestic fuel. The modern open fire burns smokeless fuel; convection heating and provision of hot water may be incorporated. Paraffin heating stoves became available in the 1860s. Their use was chiefly in rural areas where gas, and later electricity, were very late in coming. Several methods were developed in the 18th and 19th centuries to use steam for central heating in factories and other large buildings. The use of electricity for heating lagged behind its adoption for lighting. The earliest heaters (1894) consisted of cast-iron radiant panels in which heating wires were embedded in enamel. A scheme for exploiting off-peak electricity in heaters containing concrete blocks that would retain heat when switched off during intervening hours was begun in the 1930s.

Washing, Bathing and Toilet Facilities

Washing and bathing. The problems of washing or bathing the person chiefly emerged from the problems of heating sufficient water, the unreliability of domestic water supply and the inadequacy of sewage disposal facilities. The Romans attached great importance to adequate water supply. All over Europe, they built great aqueducts to bring water from the mountain and hill regions to the cities on plains. There had been adequate supplies of fresh water for washing and sanitation purposes and heating water by the hypocaust method. The Roman communities were provided with public (or private in larger houses) latrines, well equipped with running water and washing facilities.

After the collapse of the western part of the Roman Empire, in the medieval monasteries, cold water was provided in basins for daily washing and warmed for less frequent bathing. It was carried in the wooden bathtubs. From the 16th until mid-19th century, personal cleanliness lapsed to a very low level, mainly due to a lack of domestic facilities. People washed in the kitchen or in their bedrooms. Water supply was intermittent until about the mid-18th century. Heated baths, using charcoal or gas for fuel, were expensive but became popular in well-to-do households where baths were fitted with taps connected to a cold water supply. From the 1880s, bathrooms were designed in new houses containing a bath, often with shower fitting, a bidet and a washbasin. Soon after 1900, a more expensive porcelain-enamel finished bath began replacing the painted cast iron one.

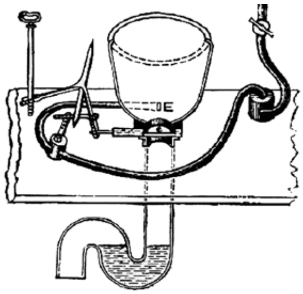

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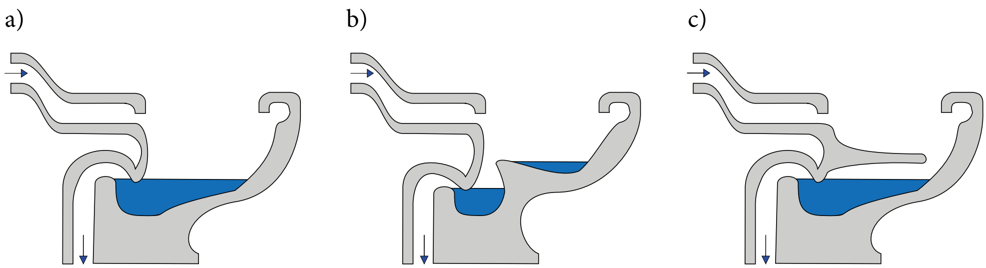
Figure 26.5. Cumming's valve closet patented in 1775

Until the late 19th century, water was still heated over the kitchen fire and carried upstairs in jugs. Electric instantaneous water heaters were marketed from the 1920s. In more modern times, this has given way to steel enamelled baths and a plastic bath.

Sanitation and toilet facilities. During the Middle Ages, public latrines were provided in towns. In castles, the garderobe was built into the thickness of the great walls and drained into the moat. In Medieval times, faeces were collected in pots and content thrown on the street (some cities issued a law that prescribed a cry of warning before emptying the pot in order to prevent unhappy accidents). In the 18th century, sewage disposal was still primitive and insanitary. The usual method was a privy outside the back of a house. These privies had wooden seats and were built over small pits. In many cases, the pits drained through into drinking water supply with serious results. Indoors, chamber pots were used. The invention that abolished the privy was the water closet. The original invention is from 1596. Unfortunately, owing to the lack of reliable water supplies and drainage systems, 179 years were to pass before the first WC was patented in London by watchmaker, Alexander Cumming (Figure 26.5). This closet had a sliding valve that was inefficient and remained unsatisfactory until, in 1778, a WC that was the best standard for a hundred years was produced.

The major breakthrough came in around 1820 when the flush-type toilet was invented in Britain. The design is basically the syphon toilet with a water reservoir to flush the load away with one swift tug on the rope. In the 1870s, new designs of WCs were invented, which gradually replaced the 100-year old closet type (Figure 26.6).



Author Egmason. Licensed under CC BY 3.0

Figure 26.6. The styles of toilet: a – the washdown style; b – he wash-out style; c – the reverse bowl or shelf style

In this closet, water was always present in the pan, but a strong force of water was needed to flush it. In 1857, personal hygiene was developed further by the invention of toilet paper.

Cleaning

Before 20th century technologies brought labour-saving appliances into the house, cleaning was extremely hard work. All equipment included brushes made from twigs, hair, wool, cotton or linen, dusters of soft cloth or feathers. Until the mid-19th century, almost all cleaning equipment and agents were made at home. Even soap was home-made, which also was very unattractive and smelled unpleasant.

Mechanical cleaning devices. A machine for sweeping floors was patented as early as at the beginning of the 19th century. These were not very efficient as consisted of two rollers and contained roller brushes rotated by a belt pulley attached to the back roller. The first really satisfactory and effective carpet sweeper was patented in 1876 (Figure 26.7). The brush was rotated by means of the friction of four rubber-treaded carrying wheels against the drum to which it was fitted. Frictional carpet resistance was minimized by setting the brush spirally in tufts round the drum so that only a few bristles at a time could be in contact with the carpet.

The suction cleaner. The vacuum cleaner was one of "machine age" marvels of the early twentieth century. This invention is older than automobile. The vacuum cleaner replaced a traditional broom. Vacuum suction is caused by difference in air pressure. An electric fan reduces pressure inside the machine. Atmospheric pressure then pushes air through the carpet and into the nozzle, and dust is literally pushed into the bag. *Suction* is the maximum pressure difference that the pump can create. For example, a typical domestic model has the suction of about negative 20 kPa. This means it can lower pressure inside the hose from normal atmospheric pressure (about 100 kPa) by 20 kPa. The higher is suction rating, the more powerful is the cleaner.

The first successful powered suction cleaner was produced in 1901 for cleaning railway carriages. This blew dirt from one end of the carriage to the other by means of compressed air. It was clear from the beginning that sucking to be better than blowing. The vacuum suction cleaner was patented in 1901. It wasn't exactly portable – the first vacuum was a wagon and long hoses were pushed into a house to suck up dirt. But, although his machine worked very well

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Figure 26.7. An early carpet sweeper

Figure 26.8. An early hand-pumped vacuum cleaner

and was in demand, it suffered from the lack of a small electric motor to power it. It was powered by a piston engine and transported through the streets on a horse-drawn van.

Nearly all early domestic cleaners worked by means of one person powering bellows by hand or foot and another operating the suction device on the end of a long handle (Figure 26.8).

The first suction cleaner with a small electric motor was demonstrated in 1907. It consisted of a sweeper inside a tin canister and a dust bag supported on a long handle. For many years after their introduction, vacuum cleaners remained a luxury item, but after World War II, they became common among the middle classes. Modern vacuum cleaners, which are used in homes as well as in industry, exist in a variety of sizes and models – small battery-operated hand-held devices, domestic vacuum cleaners, huge stationary industrial appliances that can handle several hundred litres of dust before being emptied and self-propelled vacuum trucks for the recovery of large spills or removal of contaminated soil. There are also turbo-powered cleaners and, since 1980, many machines are electronically controlled. In 2004, suction cleaner, a vacuum cleaner that floats on a cushion of air, appeared in the market.

Exhaust filtration. Typically, the filter is positioned so that incoming air passes through it and then through the motor for cooling purposes (Figures 26.9, 26.10). For air filtration, several methods are used. The *bag method* involves a paper or fabric bag that allows air to pass through but attempts to trap all dust and

Figure 26.9. A *cyclonic separation* vacuum cleaner

a) b)

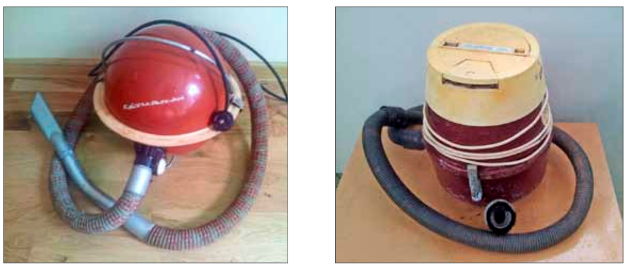

Figure 26.10. Lithuanian vacuum cleaners manufactured by Vilnius Electric Welding Equipment Factory: a – *Saturnas* manufactured in 1963; b – *Audra* manufactured in 1976

debris in the bag that may be disposable or designed to be cleaned and re-used. In the *bagless method,* the role of the bag is taken by a container and reusable filter equivalent to a reusable fabric bag. In a *cyclonic separation* vacuum cleaner, the bagless method is also employed, which causes intake air to be cycled or spun so fast that dust is forced out of air and falls into a storage bin. Operation is similar to that of a centrifuge. A *water filtration* vacuum cleaner uses water as a filter. It forces intake air to pass through water before it is exhausted. The ultra fine air filtration method is used as a secondary filter after air has passed through the rest of the machine. It is meant to remove any remaining dust that could harm the operator. Some vacuum cleaners also use a charcoal filter to remove odours.

Laundering

Washing. Cleaning fabrics was a great problem. The traditional washing method, in use from Roman times, was to agitate fabrics in hot water and soap in immense iron cauldrons; water might be heated in the kitchen and carried out in ewers or fire was lit under the cauldron. It was very hard work to heat enough water and to carry it to fill and empty washing tubs and cauldrons. Soap was very expensive, and therefore mostly home-made. Washdays lasted two or more days. The earliest special-purpose washing device was the scrub board invented in 1797 (Figure 26.11). Better facilities and equipment to help in the washing process gradually became available from the late 18th century onwards.

During the 19th century, many attempts were made to design a washing machine. The first patent for a machine to ‘wash, press out water and to press linen and wearing apparel’ was taken out as early as 1780. By the 1880s, a few machines that could heat water in the tub had been made; for this purpose, a gas jet, alternatively a coal-fired boiler was used. The first electrically-powered machines were made in the USA soon after 1900 (Figure 26.12). There were two chief methods of creating the washing action. One agitated water by a revolving disc fitted with fins mounted in the base of the tub

a)

b)

Figure 26.11. A washboard (a) and a paddle – a wooden table with a stick to beat clothes during washing (b)

Figure 26.12. *Woman's Friend* washing machine from 1890

and operated by a driving mechanism beneath. The other was a perforated cylinder driven to rotate, first in one direction, then in the other. The rotary washing machine was patented in 1858. As electricity was not commonly available until at least 1930, some early washing machines were operated by a low-speed single-cylinder hit and miss gasoline engine. Early automatic washing machines were usually connected to water supply via temporary slip-on connectors to sink taps. Later, permanent connections to both hot and cold water supplies became the norm, as dedicated laundry water hook-ups became common. Most modern front-loading European machines now only have cold water connection ("cold fill") and rely completely on internal electric heaters to raise water temperature.

On early electromechanical timers, the motor ran at a constant speed throughout the wash cycle. However, by the 1950s, demand for greater flexibility in the wash cycle led to the introduction of more sophisticated electrical timers that enabled greater variation in functions such as washing time. With this arrangement, the electric timer motor is periodically switched off to permit clothing to soak, and is only re-energized just prior to a micro-switch being engaged or disengaged for

Figure 26.13. A washing machine from 1950

Figure 26.14. Washing machines produced in Lithuania (2013)

the next stage of the process. The automatic washing machine with programmed cycles of washing, rinsing and spin-drying became a widespread success in the 1960s (Figures 26.13 and 26.14).

Many models now have microprocessor control. In 2008, Xeros Washing Machine created a washing machine that uses only a cup (less than 300 ml) of water and 20 kg of re-usable plastic chips to carry out a full wash. The machine leaves clothes virtually dry and uses less than 2% of water and energy otherwise used by a conventional machine. As such, it could save billions of litres of water each year.

Ironing

Ironing is the use of a heated tool (an iron) for removing wrinkles from fabric. The heated iron, to smooth damp fabric, was introduced in Europe in the late 16th century. The ancestor of such heated smoothing irons was an oriental pan iron, in use in the Far East since early medieval times and made to the traditional pattern ever since. This design of brass or bronze resembled a small saucepan and had a wooden handle: the pan, containing burning charcoal or coal, was rubbed over fabric.

Ironing uses heat energy, chemical energy, electrical energy and mechanical energy. The first known use of heated metal to "iron" clothes is known to have occurred in China. Until the emergence of the self-heating iron in the mid-19th century, there were two chief types of the heated iron: the charcoal iron and the sad iron. The charcoal iron was hollow and designed to contain a heated cast-iron slug or burning charcoal. The slug iron had a gate at the rear to be lifted up for the insertion of a red-hot slug; the charcoal iron is a top plate that was removed to put in the fuel. Charcoal irons also had top chimneys for the escape of smoke and holes inserted along the sides to aid combustion (Figure 26.15, b). The sad iron, or flat iron, was made of solid cast iron (Figure 26.15, a). The later 19th century saw the rapid development of many forms of a self-heating iron. Gas-heated

a)

Photo by A. V. Valiulis

b)

Figure 26.15. A flat sad iron (a) and a charcoal iron (b)

With permission from JSC Machinery Plant *Astra*

Figure 26.16. An ironing press produced in Lithuania (2013)

irons, from the 1850s, were the first to be satisfactorily operated. Then, other types of fuel such as oil, paraffin, naphtha, methylated spirits, carbide, acetylene and petrol came in use. The first electric iron appeared in the USA in the early 1880s, but this was not connected to the electric supply and had to be heated frequently on a special stand. The electric iron, however, made little progress until the late 1920s, because comparatively few households were wired for electric current. Thermostatic control came in the late 1930s. The steam-or-dry iron appeared in the 1950s. Modern versions are now much more controllable and streamlined. Modern commercial dry cleaning and full-service laundry providers usually use a large appliance called a steam press (Figure 26.16) and a rotary iron.

Home Spinning and dyeing

Spinning is an ancient textile art in which plant, animal or synthetic fibbers are drawn out and twisted together to form yarn. Hand-spinning was cottage industry in medieval Europe where wool spinners (often women and children) would provide enough yarn to service the needs of the family. A spinning wheel is a device for spinning thread or yarn from natural or synthetic fibres. Hand powered spinning wheels are powered by a spinner turning a crank for a flywheel with their hand, as opposed to pressing pedals. A single drive wheel has one drive band that goes around the flywheel and bobbin or the flyer. Most of the drive bands for single drive wheels are made from synthetic cord which is elastic and does not slip easily on the wheel. The spinning wheel in Lithuania appeared in first half of the 18th century (Figure 26.17).

Figure 26.17. Lithuanian spinning wheel around 1940

One of the interesting old survived technology is leather dyeing (Figure 26.18). The tannery dates back at least nine centuries. The tannery is composed of numerous stone vessels filled with a vast range of dyes and various odorous liquids. The tannery processes the hides (skins) of sheep and goats turn into high quality leather products. This is all achieved manually, without the need for modern machinery. The workers stand in the stone vessels arranged like honeycombs, filled with different dyes. When the dying process has been completed the hides are dried on the roofs.

Figure 26.18. The leather tannery in Fes (Marroco)

Grain Grinding

Quern is a primitive, hand-operated mill for grinding grain. Rotary quern has been found in almost every Iron Age farm and village. They were used for grinding grains of wheat, barley or rye into flour to make bread and other foods. Rotary quern consisted of two quern stones, one on the top of the other. The lower stone did not move; the top stone was turned around a wooden axle that passed up through the hole in its centre. The top stone was fitted with a wooden handle used for turning the stone around. The hard, rough surfaces of the quern stones moving against each other ground grains into flour. Rotary quern was an important new technology that probably transformed daily life of people in the Iron Age. The idea of rotary quern arrived in Europe in the middle of the Iron Age (about 400–300 BC) and quickly spread. Rotary quern in Lithuanian territory appeared at about the 10th century and were used till middle of the 20th century (Figure 26.19).

a)

b)

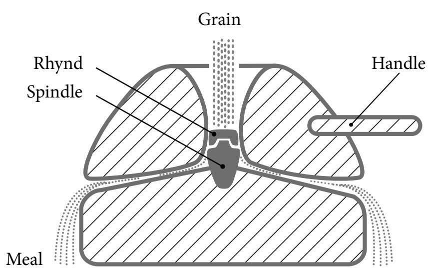

Figure 26.19. Rotary quern: a – rotary hand quern; b – sketch of rotary quern

Kitchen

A kitchen is a room or part of a room used for cooking and food preparation. A modern residential kitchen is typically equipped with hot and cold running water, a refrigerator, a microwave oven, a dishwasher and other electric appliances. The main function of the kitchen is cooking or preparing food, but it may also be used for dining, food storage, dishwashing and laundry.

In the Roman kitchen, the layout and facilities were simple but adequate. During the earlier years, the kitchen was situated in the atrium, but in larger homes, under the Empire, it became a separate room at the back of the house. Roman cuisine was sophisticated and varied. A rainwater cistern supplied piped water for preparing food and washing up at a mortar-covered stone sink; a drain carried water away. Wood or stone tables were provided, shelves and racks for utensils lined walls, and great storage jars stood upon the stone or brick floor. Very slowly, after about 1500, kitchens became better equipped and furnished and were more comfortable. By the 17^{th} century, arrangements for feeding, even in larger houses, were less communal, and kitchens were smaller, less lofty and draughty.

Early medieval European longhouses had an open fire. The "kitchen area" was between the entrance and fireplace. These early buildings had a hole in the roof through which some smoke could escape. Besides cooking, fire also served as a source of heat and light in the room. Until the 18^{th} century, food was cooked over an open fire. With the advent of the chimney, the first brick-and-mortar hearths were built. Pots made of iron, bronze or copper started replacing the pottery used earlier. Temperature was controlled by hanging the pot higher or lower over the fire (Figure 26.20). The medieval smoke kitchen remained common, especially in rural farmhouses and generally in poorer homes until the middle of the 20^{th} century.

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Figure 26.20. Medieval cooking

Technological advances during industrialization brought major changes to the kitchen. Around 1800, for large kitchens, more energy efficient stoves appeared. They used one fire for heating several pots hung into holes on the top of the stove and were thus heated from all sides instead of just from the bottom. These stoves were still fired with wood or coal. The gas stove appeared in 1825. In the second half of the 19^{th} century, cities began building water distribution pipes into homes, and sewers to deal with waste water and gas pipes were laid. At the turn of the 20^{th} century, electric stoves appeared, but like the gas stove, the electric stove had a slow start.

a)

Photo by A. V. Valiulis

b)

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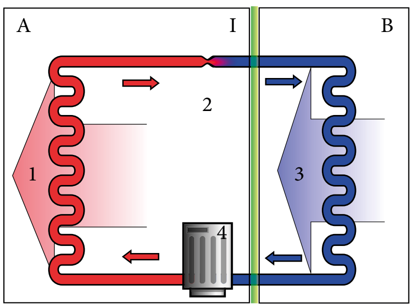
Figure 26.21. A typical kitchen corner during a Soviet period (1960–1980) (a) and a modern modular kitchen (b)

Brick-and-mortar stoves fired with coal remained the norm until well into the second half of the century. Pots and kitchenware were typically stored on open shelves. Gas and water pipes were first installed in big cities; small villages were connected only much later. By the 1950s, the 'unit kitchen' was built into new houses and the kitchen became a bright and colourful place. From the 1980, the development of plastics and microelectronics has revolutionized kitchen design as well as all its functions (Figure 26.21).

A refrigerator is a common household appliance that consists of a thermally insulated compartment and a heat pump (mechanical, electronic, or chemical) that transfers heat from the inside of the fridge to its external environment so that the inside of the fridge is cooled to a temperature below the ambient temperature of the room (Figure 26.22). In 1963, Alytus fridge factory produced the first domestic refrigerator in Lithuania. During the first year, 25 refrigerators were made. In 1975, over 1 million refrigerators per year were manufactured. In 2006, the factory exported its products to more than 40 countries worldwide.



A – hot compartment (kitchen)
B – cold compartment (refrigerator box)
I – insulation,
1 – condenser
2 – expansion valve
3 – evaporator unit
4 – compressor

Author Ilmari Karonen. Licensed under GFDL

Figure 26.22. Vapor-compression refrigeration cycle

Cooking. Charcoal was the preferred fuel in the Roman kitchen. Cooking vessels, generally iron, bronze, brass or copper, were supported upon iron tripods over charcoal fires.

a)

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b)

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Figure 26.23. Cooking on open fire (a) and in a rural house (b)

From the early Middle Ages until the development of the kitchen range in the late 18th century, all cooking of food, with the exception of baking carried out in ovens, was done above or in front of the open hearth (Figure 26.23). There was also a mechanism to raise and lower the height of pots thus moving them along or away from the heat source. In the 19th century, ovens were generally built into the wall by the side of an open hearth, so sharing the same chimney flue. They were brick-lined, with the brick or stone floor and could be closed with a metal door.

Over many years, it became understood that a further enclosure of the fire would be more economical in use of fuel as well as giving better heat control. Late 19th century ranges were large and much more efficient. They were fitted with one or two ovens, four to six hobs, a hot water tank, a warming closet for plates and many gadgets. Heat and smoke were carefully controlled and fuel consumption was much more economical. In the early 19th century, the supply of gas was extended and many designs of gas cooking ovens appeared. From the 1890s, gas largely took over from coal as a cooking fuel. In the 1930s,

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b)

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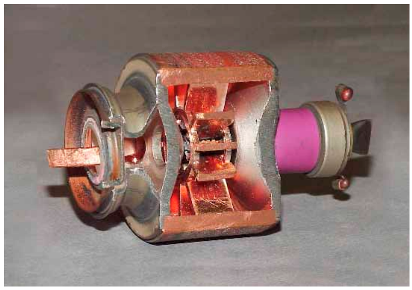
Figure 26.24. Kitchen cookers:
a – electric cooker; b – gas cooker

cookers appeared to be clad in coloured vitreous, easy-to-clean and enamel. In 1923, the thermostatic control of oven, and later, the automatic ignition of burners were invented. Since 1965, conversion to natural gas, automatic oven timing, electric spark ignition for burners, etc. has come (Figure 26.24).

Cooking by electricity. The first electric cooker was demonstrated in 1891. Most of the early cookers consisted of a number of separate appliances. It was heated by elements at the top and bottom as well as controlled by switches; the elements consisted of cylindrical ceramic formers coiled with wire. Like gas cookers, electric cookers were for many reasons slow to attract custom. People distrusted them. Heat was not visible, and therefore it was easy to burn oneself. The elements were as yet unreliable and expensive. It was not until the 1930s that electricity began seriously to compete with gas for cooking. Since 1950, electric cookers have been recognized by customers in all modern advances (Figure 26.24, a). With the 1970s came the ceramic hob and the cool-top hob, both products of modern technology. Fast-heating tubular-sheathed radiant rings replaced the earlier burners and later solid hot-plates, which had been slow to heat, and ovens were thermostatically controlled. The new cookers have become fast, clean and efficient in the conservation of energy.

Microwave ovens. Microwave cooking makes use of short radio waves. A microwave cooker, as a method, was introduced in Britain in 1959. As regards traditional cooking methods, the surface of food is heated from an exterior source (solid fuel, gas or electricity) and the centre of food is heated after a considerable period of time has passed. In microwave cooking, radio waves penetrate very quickly to the centre of food and raise its temperature. Heat is produced when water molecules in food vibrate (at a rate of 2,450,000,000 times per second) when food absorbs microwave radiation. The movement of molecules produces friction, which causes heat that cooks or warms up food. Such cooking is extremely clean as well as inexpensive. The output power of most microwave ovens

a)



b)

Figure 26.25. Magnetron (a) and a microwave oven with a metal shelf (b)

is thyristor-controlled and microprocessor-controlled. In a microwave oven, food is cooked by exposing it to microwave radiation. Most household microwave ovens operate on a frequency of 2450 megahertz (MHz or million cycles per second) in a continuous wave mode. The source of radiation in a microwave oven is a *magnetron tube* (Figure 26.25, a).

The magnetron, basically, converts 60 Hz power line electric current to the electromagnetic radiation of 2450 MHz. High 50 (60) voltage (typically 3,000 to 4,000 volts) that powers the magnetron tube is produced by a step-up transformer rectifier, and the filter converts the alternating current or 50 Hz line voltage to direct current. Microwave energy from the magnetron is transferred to the oven cavity through a waveguide section. A mode stirrer spreads microwave energy more or less evenly throughout the oven (Figure 26.25, b).

CHAPTER 27 Assault and Defence Weapons and Technologies

This chapter will help in

- introducing prehistoric weapons and armament until the 20th century (catapult, howitzer, machine gun, shrapnel shell, smokeless powder);
- describing the weapons of the 20th century (poisonous gas, tank, mines, torpedoes, panzerfaust, multiple-barrelled rocket launchers);
- showing the creation of nuclear weapons;
- discussing the manufacture of battle tanks and artillery;
- defining nuclear submarines, aircraft carriers and navy frigates;
- looking at the production and application of fighter aircrafts, ballistic missiles and unmanned aerial vehicles;
- explaining the development of military attack helicopters and robots;
- dealing with surveillance devices, airborne early warning and control;
- presenting the Defence Satellite Communications System.

Military technology is the collection of equipment, vehicles, structures and communication systems designed for use in warfare. It comprises the kinds of technology that are distinctly military in nature and not civilian in application, usually because they are impractical in civilian application, have no legal civilian usage, or are dangerous to use without appropriate military training. The power of the state traditionally based on military capacity and economic weight (without a strong economic base military power cannot be sustained) must now share its space with technology becoming an essential component in assessing military and economic capabilities. Technology is therefore power, and all states with ambition to play a part on the international stage make great efforts to ensure they are in possession of it. In other words, the strategic positioning of a nation in this century will depend on a combination of its economic and military influence. Towards 2025, the population of the planet will have increased by more than 1000 million inhabitants, which means considerable additional pressure on the resources of the planet. Around that time, the existing energy model will foreseeably have entered into crisis. The backdrop of these changes is the irrefutable evidence of an increase in the average temperature of the planet, including consequences that we still do not completely understand but which inevitably will have an effect on the production of food, transport and certain parts of the world becoming uninhabitable and which will lead to massive movements of the population. New threats and awareness on the part of society have contributed to the Armed Forces requiring new capabilities based on technological development in order to take on new missions and responsibilities that provide protection against nuclear, biological, chemical and radiological attacks.

Prehistoric Weapons

Among the earliest and most widespread weapons of the man was a spear – a wooden pole with one end sharpened with a stone or bone. From a very early period, there were two types of spearing, thrusting and throwing. A throwing spear, or javelin, tended to be lighter. Two of man's other original weapons are a club and an axe. The club was originally made of hardwood. The original axe was made entirely of stone, simply an almond-shaped head sharpened by flaking. A bow was first developed by the Mediterranean civilizations (around 15000 B.C.). Most bows were man-sized. A string was made with plaited leather strips. The discovery of metal (about 2000 B.C.) had a dramatic impact on weapons. Daggers had existed in stone, but, using bronze and copper, a sword could now be made. With the discovery of iron (around 1500 B.C.), copper, bronze and iron weapons coexisted for a long time. For personal protection, two basic types of early armour – scaled and lamellar were used. The former consisted of a short tunic on which were sewn overlapping bronze scales, whereas lamellar armour had pliable metal plates, or lames, laced together in slightly overlapping horizontal rows. Helmets were originally made of cloth, but this gave a way to leather, metal, or a mixture of the two. The third major item of personal protection was a shield (about the 2nd millennium B.C.). Shields existed in several different shapes – round, rectangular and oval and were made of leather, leather-covered wood and had overlaid thin strips of metal. From about 3500 B.C., the dominant weapon was a chariot. Chariots were used for making frontal charges on the enemy in order to create panic, and their crews were equipped with both javelins and spears. At the end of the 2nd millennium B.C., horse cavalry began to appear. Early cavalry were armed with bows and spears. To counter strong city defence, the Assyrians introduced battering rams designed to break down the main gates to the city. By 500 B.C., the Greeks invented early artillery in the shape of a *catapult*. Catapults were used for projecting arrows, javelins and smaller stones to a range of 230 m. The elephant was another one new weapon in the Roman Empire. The general Hannibal took elephants on his march across the Alps, which led to the defeat of the Romans in 216 B.C. By 400 A.D., Vandals, Goths and then Huns appeared from the East. These races relied almost entirely on the horse. Horses could travel up to 160 km a day. The Byzantines built their armies around the cavalryman having taken the Hun model to heart. There were two types of Byzantine infantry, light, which mainly consisted of archers, and heavy, armed with a lance, a sword and an axe. During this period, the first chemical weapon made its appearance. This was *Greek fire* – in essence, liquid fire and a forerunner of today's napalm. The main ingredients were sulphur, pitch, nitre and petroleum that were boiled together. By the 11th century, a crossbow had reached Western Europe, but, because of its lethality when compared with other weapons of the time, the Church considered it an unsuitable weapon of war. During the 14th century, complete plate armour built round the breastplate or cuirass was developed, and it was now that the mediaeval craftsman began demonstrating a high degree of engineering skill.

Gunpowder is generally considered to be the most significant technological development in the history of warfare. The first recorded use of gunpowder as a propellant in a firearm was at Metz in 1324. The gun was casted iron. These early guns were of only limited range and were very inaccurate. A hand-gun also made a significant advance during the 15th century with the invention of the arquebus. For the first time, a firer could use both hands to steady the weapon while firing. By the mid-16th century, the cavalryman was able to regain some of the power he had lost to the arquebus through the pistol. During the latter half of the 15th century, there were also significant improvements to artillery guns. Another new infantry weapon, which appeared during the 17th century, was a *hand grenade* and the *grenade fired from artillery* weapons (mortars). A new type of the gun, which appeared at the end of the 17th century, was a *howitzer* having a comparatively short barrel and a marked parabolic trajectory. The howitzer was mounted on a gun carriage, and did have advantages over the mortar in longer range and greater mobility.

In the late 18th and the first half of the 19th centuries, significant changes in the mechanization of manufacturing processes and inter-changeability of parts were brought. By the end of the 18th century, the art of boring out the barrels of cannon had reached a level where, for the first time, true standardization of calibre could be guaranteed. In 1857, Smith and Wesson patented the first revolver. The first *machine gun* was built in Belgium in the 1850s. A more successful Gatling model was one of the first machine guns to use metal cases – fed in by means of a hopper the whole system worked employing a hand crank that operated an endless screw. Later, the hopper was replaced by a drum magazine, and then, by a straight horizontal ammunition strip. The first truly automatic machine gun was built by Hiram Maxim (USA) in 1884. At the beginning of the 18th century, the *shrapnel shell* (after British inventor Henry Shrapnel) was adopted. It consisted of a hollow shell filled with musket balls and fitted with a bursting charge and fuse. The next radical change occurred in the middle of the century when round shot was replaced by the elongated conical projectile. The problem of round shot, in terms of accuracy, was that the centre of gravity seldom coincided with the geometric centre, which meant that trajectories would vary considerably. In 1888, Nobel produced *smokeless powder*, a mixture of nitroglicerine and nitrocotton.

20th Century Weapons

The most horrific weapon to be used during the 1st world war was poisonous gas. The first significant use of it came in the German offensive against the Russians at the end of January 1915. During the battle of Ypres in April 1915, Germans used chlorine gas that attacks the lungs. Later, lachrymatory gas appeared, followed by phosgene, mustard gas and other gases. Proper gas masks were quickly produced. By mid-summer 1916, the Mark 1 tank emerged in production lines (Figure 27.1).

The Mark 1, with a crew of eight, weighed 28.5 tons and was powered by a 6-cylinder 78 kW engine with a top speed of 6 km/h. The first tank against tank action took place between three British Mark IVs and three German tanks in 1918 in France. The next stage was the anti-tank rifle that appeared in 1917.

The initial role of the aircraft was reconnaissance, and aircrafts were used as bombers in 1914. The first arms of the aircraft were rifles, pistols, grenades and even darts. The main technical problem was how to achieve a forward-firing machine gun that would not damage propeller blades. The breakthrough came after the development of the interrupter gear by the Dutch aircraft designer Anthony Fokker. This synchronized the firing of the machine gun with the position of the propeller blades – it only fired when they were at the horizontal. The threat of the aircraft brought about the need for air defence weapons. To provide early warning of the aircraft's approach, sound detectors were developed, and by 1918, the aircraft could be detected over 8 km away.

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Figure 27.1. The Mark I tank

In 1914, the war at sea was still primarily a matter of bringing the enemy's fleet to battle and destroying it. For the period 1914–1918, two naval weapons – the torpedo and the mine come. The war quickly showed it was the submarine rather than the surface launched torpedo which was the greater threat. The torpedo had become more efficient as a result of the adoption of a gyroscope for directional control. The submerged submarine still had to be detected, however, a solution to this was not found until 1918.

In 1935, work on applying radiolocation to military purposes was started. Very soon, this method was able to detect an aircraft and represent its presence by 'blips' on a cathode ray tube. It soon became clear that radar could be applied to many other uses. It was installed in night fighters in order to detect a hostile aircraft when airborne. It became a vital weapon in the war against the U-boat complementing sonar using which submarines on the surface could be detected. The main problem with radar was that it worked as an 'active' system, which meant it could be detected and jammed. Radar as a navigation aid only came into service during the second half of the war. The most significant antisubmarine weapon, developed by the Americans in 1942, was an aerially launched acoustic homing torpedo.

Mines, like torpedoes, became increasingly sophisticated as the war progressed. At the outbreak of the war, the Germans began using an ingenious parachute mine. The parachute was used for controlling the speed of descent. As the mine sank, a spring-loaded button in the side operated as a hydrostatic switch, which stopped the clock when it reached a certain depth. As the mine touched the bottom, an electrical circuit was completed, and the mine came alive ready to detonate under the influence of a ship's magnetic field.

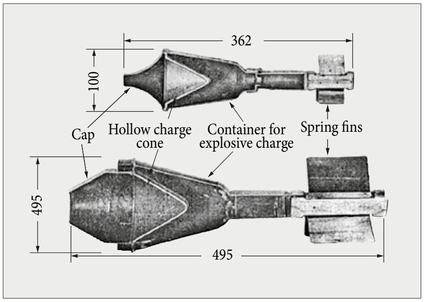
For the period 1919–1939, progress in tank development was monitored. The Germans began the war with light tanks armed only with machine guns, and with medium 50 mm and 75mm guns. The French and the Russians had light, medium and heavy tanks (Figure 27.2). Most tank ammunition was solid shot, using the tungsten carbide core. By the end of 1942, the Germans had begun introducing a new heavy tank with 88 mm gun. Tanks like the Tiger and Panther were heavily armoured. Lighter tanks, like the American Sherman, British Cromwell and Russian T-34, were more manoeuvrable (Figure 27.2). The tanks often were developed for specialized roles-mine clearing, bridging, flame-throwers, swimming tanks, air portable tanks.

Then, during the latter half of the war, a family of hand-held recoilless hollow charge round projectors like German Panzerfaust (Figure 27.3, a), American Bazooka and British FIAT were introduced.

Public Domain

Figure 27.2. Soviet heavy tank T-35

a)



b)

Figure 27.3. German Panzerfaust (a); Volkssturm soldiers with Panzerfaust in Berlin, 1945 (b)

Figure 27.4. "Katyusha" multiple rocket launcher

The Germans (Nebelwerfer) and Russians (Katyusha) used the concept of the multiple-barrelled rocket launcher (Figure 27.4), but more dramatic was the German development of free flight rockets.

By 1944, there were two types in existence. The first, the V-1, was in essence a flying bomb fitted with wings, tail plane and a ramjet motor driven by low grade aviation spirit. It carried 850 kg of explosive in its warhead and had a range of 400 km. The V-2 was a rocket bomb, fin stabilized, carrying a 1000 kg warhead and 9 tons of liquid fuel. The most ominous weapon that appeared during the Second World War was the atomic bomb. On 6 August 1945, came the dropping of the atomic bomb on Hiroshima, and three days later, on 9 August, on Nagasaki. Ever since 1945, the world has lain under the shadow of the nuclear weapon. The main effects of a nuclear explosion are blast, heat and immediate and residual radiation.

Nuclear Weapon

Nuclear weapon is the one of mass destruction and can kill and bring significant harm to a large number of humans (and other life forms) and/or cause great damage to man-made structures (e.g. buildings), natural structures (e.g. mountains) or biosphere in general. A nuclear weapon is an explosive device that derives its destructive force from nuclear reactions, either fission or a combination of fission and fusion. Both reactions release vast quantities of energy from relatively small amounts of matter (Figure 27.5).

A modern thermonuclear weapon weighing little more than 1,100 kg can produce explosive force comparable to the detonation of more than 1.1 million tons of TNT. The design of the nuclear weapon is more complicated than it might seem. Such a weapon must hold one or more subcritical fissile masses stable for deployment, then induce criticality (create a critical mass) for detonation. It is also quite difficult to ensure that such a chain reaction consumes a significant fraction of fuel before the device flies apart. A uranium bomb by USA air forces was dropped on the Japanese city Hiroshima on 6 August 1945, and three days later the plutonium-based bomb on Nagasaki (Figure 27.6). These two bombings resulted in the deaths of approximately 200,000 people – mostly civilians – from acute injuries sustained from explosions. Approximately a half of the deaths from Hiroshima and Nagasaki died two to five years afterward from radiation exposure. Throughout the Cold War, the opposing powers had huge nuclear arsenals sufficient for killing hundreds of millions of people. The only countries known to possessing such weapons are the United States, the Soviet Union (succeeded as a nuclear power by Russia), the United Kingdom, France, India, Pakistan, North Korea and the People's Republic of China. In addition, Israel is also widely believed to possess nuclear weapons, though it does not acknowledge having them. There are more than 17,000 nuclear warheads in the world (as of 2012), with around 4,300 of them considered "operational" and ready for use.

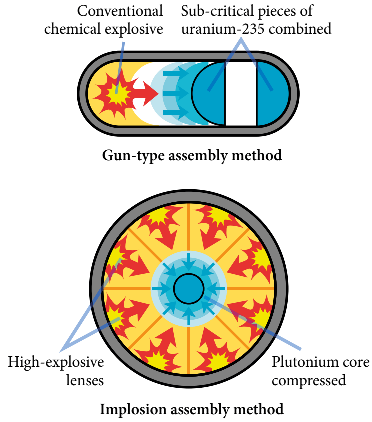


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Figure 27.5. Methods of assembling fission nuclear weapons

Public Domain

Figure 27.6. A plutonium implosion-type fission bomb exploded over Nagasaki

Battle Tanks, Missiles and Artillery

Americans and Russians made used the German wartime expertise to develop ballistic missiles. Both liquid and, now more usually, solid fuels are used. Two other categories of nuclear weapons exist today. The first is missiles able to carry about 200 kiloton (1 kiloton = the destructive force of 1000 tons of TNT) warhead (atomic bombs dropped on Japan were 12 kiloton) to a distance of 1600 km. A new type

Figure 27.7. A modern tank

is the *cruise missile* designed for flying low enough to pass under radar coverage can be air, ground and sea-launched. Missiles are also used in a large variety of conventional roles such as *air-to-air*, *surface-to-air*, *anti-tank* and *air-to-surface*. Crucial to all air defence systems is the incorporation of an *Identification Friend-or-Foe* (IFF) device that interrogates the aircraft, usually through the form of a radio signal to which the built-in responder of the aircraft will reply if it is friendly.

Modern main battle tanks (Figure 27.7), like the American M-1 Abrams, Russian T-80, British Challenger and German Leopard 2 weigh 45–60 tons and have a 120 mm gun. New types of ceramic armour specifically designed to increase protection against hollow charge projectiles without increasing weight, are also being introduced.

Modern *artillery* is most obviously distinguished by its large calibre, firing an explosive shell or rocket and being of such a size and weight as to require a specialized carriage for firing and transport (Figure 27.8, a). However, its most important characteristic is the use of indirect fire, whereby firing equipment is aimed without seeing the target through its sights. Artillery is used in a variety of roles depending on its type and calibre. The general role of artillery is to provide fire support coordinated with the manoeuvre of forces to destroy, neutralize or suppress the enemy. Weapons covered by the term „modern artillery" include "cannon" artillery (such as howitzer, mortar and field gun) and rocket artillery (Figure 27.8, b).

a)

b)

Figure 27.8. A military artillery system (a) and guided artillery shell (b)

An anti-tank missile (ATM) is a guided missile primarily designed for hitting and destroying heavily-armoured military vehicles (Figure 27.9). Most modern ATMs have shaped charge high explosive warheads designed specifically for penetrating armour. Top-attack weapons such as the Indian *Nag*, American *Javelin* and the Swedish *Bill* are designed to focus the explosion down through an armoured fighting vehicle's thinner turret-roof or upper-hull armour. Armour systems have continued in development alongside ATMs, and the most recent generations of armour are specifically tested to be effective against ATM strikes, either by 'tricking' the missile into not detonating against the armour itself (such as in Slat armour) or using some form of reactive armour to "attack" the missile upon impact, disrupting shaped charge that makes the warhead effective.

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Figure 27.9. Man-portable fire-and-forget anti-tank missile Javelin

Surveillance Devices

Surveillance devices have undergone notable improvements during the past forty years. High-flying „spy planes" like the US Lockheed SR-71 Blackbird carry long-focus cameras that can cover an area of 325 km^2 in one exposure from a height of 40 km with sufficient resolution (Figure 27.10). Another class of the surveillance device is the remotely piloted vehicle (RPV) that can be directed on photographic flight missions with pictures being instantaneously reproduced on a television screen at the ground control station.

Attention is turned to the threat of war in space, as a network of military intelligence and communication satellites already exists, which is very likely that anti-satellites (ASATs) will soon make an appearance. War in space is likely to use directed energy the weapons of which comprise lasers that use photons of electro-magnetic radiation and particle beams made up of sub-atomic particles which have mass and destroy a target through nuclear reaction, sonic and radio frequency weapons. Emerging military technologies include ballistic missile defence, long-range conventional strike weapons and space warfare systems.

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Figure 27.10. A Lockheed U-2 Dragon Lady spy plane in flight

Nuclear Submarines

For at least a thousand years, dominating the oceans has been a key step toward dominating the world. Britain, France, Spain, Portugal and the Netherlands waged war on high seas for centuries in a global competition to control commerce, protect shipping and spread influence in the form of colonies. Nuclear submarines are incredible pieces of technology. A nuclear submarine is the one powered by a nuclear reactor, is almost undetectable and can launch a devastating surprise attack on almost any country in the world (Figure 27.11).The roles of military submarines are almost limitless, but the most technologically advanced ones are used for military purposes.

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Figure 27.11. Modern nuclear submarine Typhoon (*Akula*, Russia)

Aircraft carrier and Navy Frigates

Aircraft carriers and their embarked air wings are the most important weapons systems in the Navy (Figure 27.12). These large ships never operate alone but as the core of a carrier battle group that consists of cruisers, destroyers, frigates and submarines that safeguard the carrier.

*Navy frigates***.** Frigates fulfil a Protection of Shipping (POS) mission as Anti-Submarine Warfare combatants for amphibious expeditionary forces, underway replenishment groups and merchant convoys (Figure 27.13).

Public Domain

Figure 27.12. Aircraft carriers

Public Domain

Figure 27.13. Frigates of the UK Navy

Fighter Aircrafts

A fighter aircraft is a military one designed primarily for air-to-air combat with other aircraft, as opposed to a bomber designed primarily to attack ground targets by dropping bombs (Figure 27.14).

US Air Force is developing a hypersonic stealthy aircraft capable reach 6400 km/h speed and flying altitudes of about 30 km with transcontinental range. In comparison, the state-of-the-art SR-71 stealth reconnaissance plane, in use from 1964 to 1998, topped out at about an air speed of 3540 km/h and 25,7 km, while the Concorde, the fastest commercial jet ever built, flew at a maximum speed of about 2200 km/h and an altitude of 18 km (Figure 27.15).

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Figure 27.14. A fighter aircraft

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Figure 27.15. B-2 Spirit stealth bomber

Ballistic missiles

Ballistic missiles are powered by rockets initially but then follow an unpowered, free-falling trajectory toward their targets. They are classified by the maximum distance they can travel, which is a function of how powerful the engines of the missile (rockets) are and the weight of the payload of the missile. To add more distance to the range of a missile, rockets are stacked on the top of each other in a configuration referred to as staging. A total of 30 nations have deployed operational ballistic missiles (Figure 27.16).

Public Domain

Figure 27.16. An intercontinental ballistic missile

Unmanned Aerial Vehicles

Unmanned Aerial Vehicle (UAV) is an aircraft with no pilot on board (Figure 27.17). UAVs can be a remote controlled aircraft (e.g. flown by a pilot at a ground control station) or can fly autonomously based on pre-programmed flight plans or more

a)

Author Brazilian Air Force. Attribution

b)

Public domain

Figure 27.17. Unmanned Aerial Vehicles: a – Hermes 450; b – RQ-7 Shadow

complex dynamic automation systems. UAVs are currently used for a number of missions, including reconnaissance and attack roles. For the purposes of this article, and to distinguish UAVs from missiles, a UAV is defined as being capable of controlled, sustained level flight and powered by a jet or reciprocating engine. In addition, a cruise missile can be considered to be a UAV, but is treated separately on the basis that the vehicle is the weapon. The acronym UAV has been expanded in some cases to UAVS (Unmanned Aircraft Vehicle System).

Military Attack Helicopters

Just as jeeps and trucks replaced the horse for military mobility, a military helicopter has come a long way toward replacing motor transport. Originally used for only high value missions such as medical evacuation, helicopters are now the primary vehicle for air assault and many other military tasks (Figure 27.18).

a)

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b)

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Figure 27.18. Attack Helicopters: a – AH-64 Apache; b – Mi-35

Military Robots

Military robots are autonomous and remote-controlled robots, or robotically enhanced systems designed for military applications: powered exoskeletons, bomb disposal robots, unmanned ground vehicles, unmanned combat air vehicles and autonomous underwater vehicles (Figures 27.19, 27.20). Military robots date back to World War II and the Cold War in the form of the German *Goliath* tracked mines and the Soviet teletanks. Some experts and academics have questioned the use of robots for military combat, especially when such robots are given some degree of autonomous functions.

Public domain

Figure 27.19. British soldiers with captured German *Goliath* remote-controlled demolition vehicles (1944)

Public domain

Figure 27.20. A modern military robot

Airborne Early Warning and Control

An airborne early warning and control (AEW&C) system is an airborne radar system designed for detecting an aircraft (Figure 27.21). Used at a high altitude, radars allow operators to distinguish between a friendly and hostile aircraft hundreds of miles away. AEW&C aircrafts are used for defensive and offensive air operations.

a)

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b)

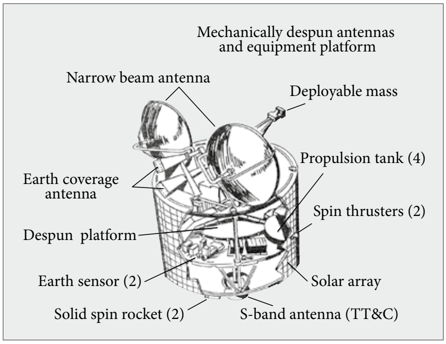
Public domain

Figure 27.21. An Early Warning and Control aircraft: a – Boeing E-3 Sentry; b – E-2 Hawkeye

Defence Global Satellite Communications System

A defence satellite communications system (DSCS) provides military communications to soldiers in the field as well as commanders (Figure 27.22). DSCS supports the defence communications system, the Army's ground mobile forces, the Air Force's airborne terminals, Navy ships at sea, the White House Communications Agency, the State Department and special users. Overall DSCS responsibility resides in the United States Strategic Command.

a)



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b)

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Figure 27.22. A defense satellite communications system (a) and the image of the earth surface (b)

CHAPTER 28 Development of the Nearest Future Technologies

This chapter will help in

- describing the creation of the nearest future technologies;
- explaining the development of recovery, recycling and re-use of post-consumer waste, a reduction in fossil feedstock, a decrease in the CO_2 footprint, a replacement for the current inefficient processes for more energy and resource efficient processes, reinvention of materials and products with significantly increased material properties.

In a world with growing pressures on resources and the environment, people have no choice but go for the transition to a resource-efficient and ultimately regenerative circular economy. Experts agree that unless we cut greenhouse gas emissions – especially carbon dioxide (CO_2) – by 50–80% (compared with today) by 2050, the impact on global warming will be disastrous. For example, cement industry is one of two primary producers of carbon dioxide, creating up to 5 percent of worldwide emissions of this gas, of which 50% is from the chemical process, and 40% – from burning fuel. Cement manufacture contributes greenhouse gases both directly through the production of carbon dioxide when calcium carbonate is heated, producing lime and carbon dioxide, and indirectly through the use of energy, particularly if it is sourced from fossil fuels. But with world energy demand expected to double by this date, the challenge will be enormous. It means we must act fast, using a portfolio of solutions, since no single solution will be capable of reducing CO_2 emissions on the massive scale required. This includes renewable energies, energy efficiency and CO_2 capture and storage (CCS). Indeed, if deployed in all industry sectors, CCS has the potential for reducing CO_2 emissions by over 50% by 2050.

Novel and radically improved production processes are the key to increase the efficiency of energy, resource and CO_2 in industrial value chains. Addressing these challenges requires appropriate technologies, processes and products with intelligent product design as well as smart processes over the value chain to:

- use energy and resources more efficiently (reduce) within the existing installed base of industrial processes;
- re-use waste streams and energy within and between different sectors, including recovery, recycling and re-use of post-consumer waste.
- replace the current feedstock by integrating novel and renewable feedstock (such as bio-based) to reduce fossil feedstock and mineral raw materials dependency while reducing the CO_2 footprint of processes or increase the efficiency of primary feed stock; replace current inefficient processes for more energy and resource efficient processes when sustainability analysis confirms benefits (Figure 28.1).
- reinvent materials and products to have a significantly increased impact on resource and energy efficiency (integration of recycled materials, easy recyclability and re-usability as well as improved material properties such as light weight for lower consumption vehicles) and improved insulation properties (for energy efficient buildings).

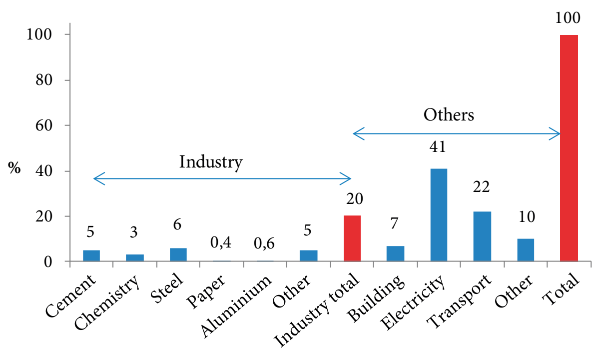

Figure 28.1. Worldwide CO_2 emissions by sectors in %

The two key resource and energy efficiency targets within a time horizon of 2030:

- The world energy market is driven by a growing number of factors ranging from the impacts of climate change, the depletion of oil and gas, high costs, the unpredictable supply of fuel and the price of CO_2 allowances. A strong wind energy sector does not only mean reduced CO_2, cleaner air and secure biodiversity but also sustainable economic growth, reduced dependence on energy imports, high quality jobs, technological development and global

competitiveness as well as a reduction in fossil energy intensity of up to 30% from the current levels by 2030 through a combination of, for example, cogeneration-heat-power, process intensification, the introduction of novel energy-saving processes, energy recovery and the progressive introduction of alternative (renewable) energy sources within the process cycle.
- Mineral resources are crucial for the competitiveness and innovation of economy. Mineral industry comprising the producers and users of industrial minerals and metals aggregates ornamental or dimensional stone, oil, gas and derivates as well as coal and by-products, provides vital inputs to the economy and social well-being of the countries. By 2030, up to 20% reduction in non-renewable, primary raw material intensity compared to the current levels, by increasing chemical and physical transformation will have yield and/or use secondary (through optimised recycling processes) and renewable raw materials. This may require more sophisticated and more processed raw materials from raw material industries.

There is the need to implement more means of cost efficient energy storage, which is essential to providing flexibility in adopting more use of renewable energy sources as well as benefit from harvesting waste energy. These could be through, e.g. battery technology, fuel cells, super capacitors as well as through novel thermo-chemical solutions to local storage. The generation of renewable energy through improved photovoltaic technology is also a must for success. Some of these technologies will also find application in other parts of the value chain (energy, transport and construction sector) through higher resource efficient electrification of society (smart grids, building heating/cooling, transport propulsion, etc.).

Lithium ion batteries are a popular power source for portable electronic devices, but not yet sufficient for industrial or automotive goals. Other energy storage solutions, e.g. chemical storage, fluid transportation, etc., will require the development of novel materials (i.e. inorganic *phase change materials* (PCM) for high temperatures, high heat capacity (e.g. >800°) and PCM materials based on organic and polymeric molecules) as well as their production processes. Regarding H_2 storage materials (HSM), research is still open in order to find materials meeting the above mentioned needs (e.g. hydrides) as well as routes to process them. Other examples could be more energy and resource efficient processes to develop inorganic *light emitting diodes* (LEDs) as well as *organic light emitting diodes* (OLEDs). Additional areas could include new sustainable processes for novel composites (bio-based and non-bio-based), new thermo-electrical materials or coated materials with new functionalities as well as the materials of potential future use in the construction sector (i.e. thermo-electrical roof tiles, energy storing clay blocks, ultra-high-performance concrete, lightweight construction composites, etc.) and, for example, sustainable processes for materials used in the energy sector such as cost-efficient windmills having superior energy conversion capacity and novel materials with applicability in energy efficient power plants.

The two clear impact goals could be defined particularly from the environmental and sustainability perspective:

1. A reduction in fossil energy intensity of up to 30% from the current levels by 2030.
2. By 2030, up to 20% reduction in non-renewable, primary raw material intensity versus the current levels.

Several major impact opportunities are identified for the process industry in order to contribute to energy efficiency. Just as an illustration it is worthwhile to mention:

- In primary steel production, efficiency improvements on the order of 20 to 30% are available based on the existing technology.
- Improvements to steam supply systems and motor systems offer efficiency potentials on the order of 15 to 30%.
- Combined heat and power generation can bring 10 to 30% fuel savings over separate heat and power generation.
- The production of paper from pulp could substantially reduce energy needs with more efficient drying and gasification technologies.
- Using biomass feedstock and recycling more plastic waste could reduce life-cycle CO_2 emissions substantially.
- CO_2 capture and storage (CCS) could be applied to several industries on a gigatonne-scale, especially in the production of chemicals, iron, steel and cement. It will be most effective when combined with *sustainable utilisation technology* (CCU), converting CO_2 into chemical fuel feedstock and therefore creating value rather than generating waste.

The following examples illustrate the potential impact of some selected innovations:

- Enough biomass in the right density and composition in a sustainable way would lead to > 50%–70% or even higher reduction in greenhouse gas (GHG) emissions compared to fossil-based;
- ULCOS Blast Furnace top gas recycling with carbon-capturing and sequestration (CCS) might represent a 50% CO_2 reduction and a 15% energy reduction compared to the average blast furnace.
- ULCOS new direct reduction and smelting processes with or without CCS might represent 80% CO_2 reduction compared to the average blast furnace with CCS and 20% energy and CO_2 reduction without CCS as compared to reference technology nowadays.

- Magnesium based clinker might represent over 100% CO_2 reduction compared to the average kiln (sink). The avoidance of process emissions, product carbonisation (no CCS required).
- Biomass and natural gas utilization might represent a 35% CO_2 reduction compared to the average kiln.
- Electrolysis might represent the total independence of fossil fuel in the production process. The net effect of successfully deploying inert anodes for aluminium production could represent a reduction in electricity consumption of 10 to 20% compared to advance Hall-Héroult smelters (from 13 to 11 kWh/kg aluminium). Apart from electricity savings, oil and coal consumption would be reduced by 18 GJ per ton of aluminium, because the use of carbon anodes would be avoided.

The annual cost of corrosion for US industry in various industrial processes totals $3.7 billion of dollars in petroleum refining industry, $1.7 billion in the chemical industry and $5.9 billion in pulp and paper production and processing. A 10% reduction in corrosion costs in these three industries alone using improved materials could save $1.1 billion each year. Today, there are 693.8 mln tons of aluminium in use and 11 mln tons of old scrap are recycled in one year. Using secondary aluminium as feedstock reduces energy production by 95% in addition to conserving resources.

Competition for water is set to be one of the biggest challenges for the global community in the coming years. This competition will be particularly fierce in the areas of high urbanisation or industrialisation. Membrane technology could be extremely helpful in providing a new source of water that has never been available before. Now, researchers have shown that water can be successfully extracted from plant chimneys and that it is ready for large scale industrial use. Since at least 40% of water in the flue gases of an industrial plant can be recovered using membrane technology, food, paper, cement, steel and petrochemical producers could benefit from it. The recovered water could be used directly as demineralised water for industrial applications or in agriculture, or even for consumption, after mineralisation. The value of energy savings obtained by using this technology could be enormous.

The construction sector is faced with ever growing challenges. Societal developments, such as an ageing and growing population, lead to demands for more comfort, better mobility and more safety and security. Buildings account for 40% of the EU energy demand. Construction uses more raw materials than any other sector; the creation and operation of the built environment accounts for an important consumption of natural resources.

The competitiveness of future product generations will be based, to a large extent, on 'smart integrated systems' converging a whole range of technologies and improving the characteristics of the entire product in which they are incorporated.

These systems are often very complex, networked, energy-autonomous, miniaturised and reliable. The integration of smart systems addresses the trend toward miniaturised multifunctional devices and specialised connected and interacting solutions. In this context, the ability to miniaturise and integrate intelligence and new functionalities into conventional and new components and materials are among the most ambitious challenges facing industry.

Materials are everywhere; they make modern life possible and constitute the basis of the welfare of mankind. Today advanced engineering materials and the industry of technologies face a number of important challenges, including environmental legislation, recycling requirements and increasing the costs of energy and the commodity of raw materials.

An economy without steel products is hardly imaginable. Machinery in manufacturing and processing industry, petroleum refineries and chemical plants for producing basic materials or pharmaceuticals are constructed of steel. Planes, trains, ships, automobiles and many other things needed on a daily basis could not be manufactured unless there were structural parts and components made of steel. Today, the steel sector is under increasing pressure to address a number of challenges like the growing impact of globalisation and worldwide competition, population growth, increased urbanisation, the depletion of some natural resources and climate change. To maintain its competitiveness, steel industry will have to meet the combined challenging targets of environmental friendliness and economic growth.

Zentallium® represents a new mechanically alloyed and thus very powerful composite material that consists of an aluminium-based material and carbon nanotubes (CNTs) and is manufactured by powder metallurgy route. The material is characterized by a nanostructured Al-matrix microstructure with grain sizes well below 100 nm. Zentallium® exceeds the strength of stainless steel and achieves the level of structural steel (700-1000 MPa). In comparison to Al-based materials, the elastic modulus appears to be 20–30% higher. CNTs stabilize this alloy system proven to 240 °C, which, for commercial Al-alloys, also represents an unmatched value.

Self-healing materials are those able to repair themselves after damage caused by, e.g. impact, abrasion, corrosion, wear, fire and ice. Self-healing materials are of great value for society and economy, most certainly for industrial sectors such as chemistry, high-tech and energy. They can increase the reliability and durability of a material and enhance safety when used in vehicles, machinery and transportation infrastructure. They can reduce maintenance, incidents, injuries and permanent damage, environmental pollution and urban noise. The development of the self-healing capacity of materials such, including asphalt, concrete and coatings, is a great technological innovation. The development and encapsulation of viscoelastic polymers that undergo spontaneous self-healing and non-elastic repair materials such as calcium carbonate precipitated by bacteria or new cement hydrates the formation of which is stimulated by the presence of hydrogels are forecast.

Graphene is a material with an extraordinary combination of physical and chemical properties: it is a very thin (two-dimensional) material, conducts electricity better than copper, is stronger than steel and has unique optical properties. The availability of graphene, i.e. of ultimate carbon films of monoatomic thickness, opens a strong opportunity to develop a new family of nanostructured materials. Radical advances in their engineering would make them very promising, for example, for the development of post-silicon electronics.

Fibre and textile-based materials and products have always played a vital role in human life and there is no reason to believe that their importance will diminish in the 21st century. With a growing world population and a rapid increase in textile consumption in developing countries, a whole range of new application areas for textiles and constantly rising user requirements (in terms of functionality, variety, precision, performance, reliability, user and environmental friendliness of textile products) is set to rise.

Some 98% of computing devices are now embedded in all kinds of equipment. Computers are found in everyday devices such as mobile phones, credit cards, domestic TVs, multimedia equipment and washing machines as well as in cars, planes, offices and factories. In the coming years, the share of the value of embedded electronic components in the value of the final product is expected to reach significant percentages in the areas such as industrial automation, telecommunications, consumer electronics and intelligent homes and health/medical equipment. In our digital era, demand for data storage capacity is growing exponentially. Scientists have experimentally demonstrated the recording and retrieval of five-dimensional digital data by ultrafast laser nanostructuring in glass. The potential advantages of this new technology are outstanding: 360 TB/disc data capacity, thermal stability up to 1000 °C and practically unlimited lifetime. These could make the 5D optical memory perfect for organisations with huge numbers of documents, like national archives or museums.

An ageing population, expectations for a better quality of life and changing lifestyles call for improved, more efficient and affordable healthcare. A better understanding of the functioning of the human body in the molecular and nanometre scale as well as the ability to intervene at pre-symptomatic, acute or chronic stages of an illness are of utmost importance to meet these expectations. Nanomedicine, the application of nanotechnology to medical problems, has the potential for enabling early detection and prevention and for essentially improving the diagnosis, treatment and follow-up of many illnesses, including cancer, cardiovascular diseases, diabetes, AIDS, Alzheimer's and Parkinson's disease as well as any kind of inflammatory and infectious diseases.

The term 'photonics' is used for describing a new field of study and enterprise that emerged as the synthesis of a number of disciplines all involved in the mastery of the photon: optics, material science, electrical engineering, nanotechnology, physics and chemistry. Its importance can be seen from the multitude of application sectors where it is increasingly seen to be driving innovation, including information, communication, imaging, lighting, displays, manufacturing, life sciences and healthcare, safety and security.

Research is ongoing into nanotechnology, quantum computers, bioengineering/biotechnology, gene therapy, stem cell treatments, vaccine development, bionic body parts, cloning, nuclear fusion, advanced materials (e.g. graphene), scramjet, drones, railguns, high-energy laser beams, superconductivity, green technologies such as alternative fuels (e.g. fuel cells, self-driving electric & plug-in hybrid cars), augmented reality devices and wearable electronic devices, artificial intelligence, the simulation of the human brain, more efficient and powerful LEDs, solar cells, integrated circuits, wireless power devices, engines, batteries, etc.

Theoretical physics currently investigates quantum gravity proposals such as M-theory, superstring theory and loop quantum gravity. Dark matter is also in the process of being detected via underground detectors. Laser Interferometer Gravitational-Wave Observatory (LIGO) is trying to discover gravitational waves.

The European Organization for Nuclear Research, known as CERN, located in the suburbs of Geneva, Switzerland, is the world's largest particle physics laboratory where the world's best physicists and engineers use advanced particle accelerators to help with solving age old questions about the universe like what it's made of and how it started. The *Large Hadron Collider* (LHC) is the world's largest and highest-energy particle accelerator the aim of which is to allow physicists to test predictions for different theories of particle and high-energy physics (Figure 28.2). The collider is contained in a circular tunnel of 3.8-metre in diameter with a circumference of 27 kilometres and at a depth ranging from 50 to 175 meters underground. It crosses the border between Switzerland and France at four points. In total, over 1,600 superconducting magnets are installed. Approximately 96 tons of superfluid helium-4 is needed to keep the magnets, made of copper-clad

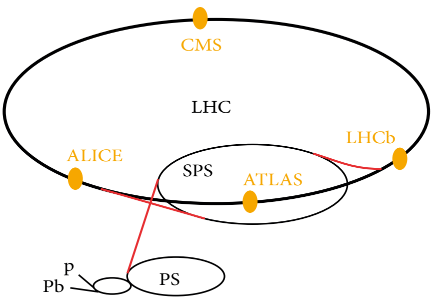

Figure 28.2. Large Hadron Collider: ATLAS (*A Toroidal LHC Apparatus*) detector; CMS (*Compact Muon Solenoid*) particle physics detector; LHCb (*Large Hadron Collider beauty*) detector; ALICE (*A Large Ion Collider Experiment*) detector

niobium-titanium, at their operating temperature of 1.9 K (−271.25 °C) thus making the LHC the largest cryogenic facility in the world at liquid helium temperature. When running at full design power of 7 TeV per beam, the total collision energy of particles is 14 TeV. Under this energy, protons move at about 0.999999991 *c*, or about 3 metres per second slower than the speed of light. Prior to being injected into the main accelerator, the particles are prepared by a series of systems that successively increase their energy. The first system is the linear particle accelerator generating 50 MeV protons, which feeds the Proton Synchrotron Booster (PSB). The protons are accelerated to 1.4 GeV and injected into the Proton Synchrotron (PS) where they are accelerated to 26 GeV. Finally, the Super Proton Synchrotron (SPS) is used for a further increase in their energy to 450 GeV before they are at last injected into the main ring where the proton is accelerated to their peak 4 TeV energy.

At about 2020, the construction of *Very Large Hadron Colliders* in the U.S.A. (California) and EU (that would pass under Lake Geneva) can be started. The machine could be completed around 2035. It would require a tunnel of 80–100 kilometres around and would collide protons at energies around 100 teraelectronvolts (TeV), compared with the 14 TeV of the LHC at CERN. To build a 100 TeV machine, physicists will need to develop superconducting magnets that can operate at higher fields than the current generation, perhaps 20 instead of 14 tesla. One leading candidate material for such magnets is niobium tin (Nb_3Sn) that can withstand higher fields (up to 30 tesla) but is expensive and must be cooled below 18 kelvin.

Spacecraft designs like the *Multipurpose Crew Vehicle* (ORION) are also being developed (Figure 28.3). A new vehicle will be able to carry astronauts to Earth orbit, to the moon, asteroids, and eventually to Mars. The component parts of the Orion have been designed to be as generic as possible, so that between the first test flight of the craft in 2014 and its projected Mars voyage in the 2030s, the spacecraft can be upgraded as new technologies become available. Whereas the *James Webb Space Telescope* will try to identify early galaxies as well as the exact location of the Solar System within our galaxy in 2018, the *Advanced Technology Large-Aperture Space Telescope* has orders-of-magnitude better resolution and sensitivity than its predecessors and will try to find biosignatures of terrestrial exoplanets (planned to be launched

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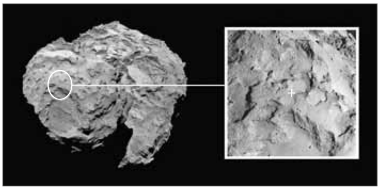
Figure 28.3. ORION docked to Mars Transfer Vehicle



Courtesy NASA/JPL-Caltech

Figure 28.4. The comet and landing site

Author DLR. Licensed under CC-BY-3.0-de

Figure 28.5. Probe touchdown on the comet

in 2030). The International Space Station will provide an intermediate platform for space missions and zero gravity experiments. NASA and ESA plan a manned mission to Mars in the 2030s. The *Variable Specific Impulse Magneto-plasma Rocket* (VASIMR) is an electro-magnetic thruster for spacecraft propulsion and is more than five times faster than traditional propulsion technology.

On 12 November 2014, the European Space Agency's Rosetta spacecraft successfully landed on the surface of the comet (Figures 28.4 and 28.5). If successful, landing will mark the first time humans have soft-landed a probe on a comet. Landing the probe on the comet whizzing through deep space isn't easy. Landing is a risky operation. The spacecraft and comet are racing through space at over 60 000 km/h. The aim of the rendezvous is to unlock the secrets hidden within the icy treasure chest before 4.6 billion years. Rosetta is the first spacecraft to orbit a comet nucleus, and is the first spacecraft to fly alongside the comet as it heads towards the inner Solar System. It is planned to be the first spacecraft to examine at close proximity how a frozen comet is transformed by the warmth of the Sun.

The time between a forecast or idea and implementation becomes shorter and shorter. Cognition has not limits and many forecasts or facts from fiction today are as an ordinary constituent of surrounding reality.

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Appendixes

Appendix I.
Geologic and Biological Timeline of the Earth

Astronomical and geological evidence indicates that the Universe is approximately 13,700 million years old, and our solar system is about 4,567 million years old. Nearly 3,000 million years ago, the Earth was cool enough for land masses to form. The supercontinent *Rodinia* was formed about 1100 million years ago and broke into several pieces that drifted apart 750 million years ago. Those pieces came back together about 600 million years ago thus forming the Pan-African Mountains in a new supercontinent called *Pannotia* that started breaking up 550 million years ago to form *Laurasia* and *Gondwana*. *Laurasia* included what are now North America, Europe, Siberia and Greenland. *Gondwana* included what is now India, Africa, South America and Antarctica. Laurasia and Gondwana rejoined approximately 275 million years ago to form the supercontinent *Pangea* the breakup of which still goes on today and has contributed to the formation of the Atlantic Ocean.

Table A1. Geologic and Biological Timeline of the Earth

Precambrian Time – 4567 to 542 mya (mya = million years ago)
Hadean Eon (4567 to 3800 mya). There is no geologic record for the Hadean Eon – 4500 mya: Formation of the Earth. – The Earth's original hydrogen and helium atmosphere escapes Earth's gravity. – 3900 mya: Cataclysmic meteorite bombardment. – The Earth's atmosphere becomes mostly carbon dioxide, water vapour, – methane and ammonia. – Formation of carbonate minerals starts reducing atmospheric carbon dioxide. Archean Eon (3800 to 2500 mya). First known oxygen-producing bacteria *cyanobacteria* (blue-green algae) form stromatolites. – 3000 mya: The atmosphere has 75% nitrogen and 15% carbon dioxide.
Proterozoic Eon (2500 to 542 mya)
Paleoproterozoic Era (2500 to 1600 mya) Siderian Period (2500 to 2300 mya) – Stable continents first appear. – 2500 mya: First free oxygen is found in the oceans and atmosphere.

Continued Table A1

Proterozoic Eon (2500 to 542 mya)
Rhyacian Period (2300 to 2050 mya) – 2200 mya: Organisms with mitochondria capable of aerobic respiration appear.
Orosirian Period (2050 to 1800 mya) – Intensive mountain development. – Oxygen starts accumulating in the atmosphere.
Statherian Period (1800 to 1600 mya) – Complex single-celled life appear. – Abundant bacteria and archaeans.
Mesoproterozoic Era (1600 to 1000 mya)
Calymmian Period (1600 to 1400 mya) – Photosynthetic organisms proliferate. – Oxygen builds up in the atmosphere above 10%. – Formation of the ozone layer starts blocking ultraviolet radiation from the sun. – 1500 mya: Eukaryotic (nucleated) cells appear.
Ectasian Period (1400 to 1200 mya) – Green (Chlorobionta) and red (Rhodophyta) algae abound.
Stenian Period (1200 to 1000 mya) – 1200 mya: Spore/gamete formation indicates the origin of sexual reproduction. – 1100 mya: Formation of the supercontinent Rodinia.
Neoproterozoic Era (1000 to 542 mya)
Tonian Period (1000 to 850 mya) – 1000 mya: Multicellular organisms appear. – 950 mya: Start of Stuartian-Varangian ice age.
Cryogenian Period (850 to 630 mya) – 750 mya: Breakup of Rodinia and formation of the supercontinent Pannotia. – 650 mya: Mass extinction of 70% of dominant sea plants due to global glaciation.

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The Supercontinent of Rodinia

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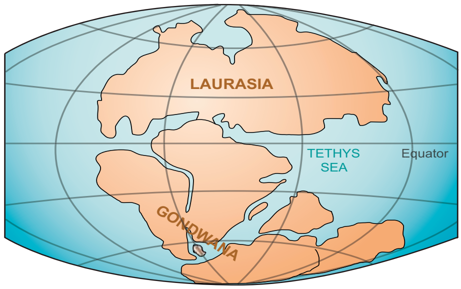
The Supercontinent of Pannotia

Continued Table A1

Proterozoic Eon (2500 to 542 mya)

Ediacaran (Vendian) Period (630 to 542 mya)
- 580 mya: Soft-bodied organisms develop. Jellyfish, Tribrachidium and Dickinsonia appear.
- 570 mya: Shelled invertebrates appear.
- 550 mya: Pannotia fragments into Laurasia and Gondwana



Laurasia and Gondwana 200 million yers ago

Phanerozoic Eon (542 mya to present)

Cambrian Period (542 to 488.3 mya)
- Abundance of multicellular life.
- Most of the major groups of animals first appear.

Tommotian Stage (534 to 530 mya)
- 510 mya: Vertebrates appear in the ocean.

Ordovician Period (488.3 to 443.7 mya)
- Diverse marine invertebrates, such as trilobites, become common.
- First green plants and fungi on land.
- Fall in atmospheric carbon dioxide.
- 443 mya: Glaciation of Gondwana. Mass extinction of many marine invertebrates. 49% of genera of fauna disappear.

Silurian Period (443.7 to 416 mya)
- 420 mya: End of Andean-Saharan ice age.
- Stabilization of the Earth's climate.
- Land plants and coral reefs appear.
- First fish with jaws – sharks.
- Insects (spiders, centipedes) and plants appear on land.

Devonian Period (416 to 359.2 mya)
- Ferns and seed-bearing plants (gymnosperms) appear.
- Formation of the first forests.
- 400 mya: Land animals and wingless insects appear.
- 375 mya: Vertebrates with legs appear
- First amphibians appear.

Continued Table A1

Phanerozoic Eon (542 mya to present)

- 374 mya: Mass extinction of 70% of marine species. Such extinction occurs over 20 million years. Surface temperatures drop from about 34 °C to about 26 °C.
- 370 mya: First trees appear.

Carboniferous Period (359.2 to 299 mya)

Mississippian Epoch (Lower Carboniferous) (359.2 to 318.1 mya)
- 350 mya: Beginning of the Karoo ice age.
- Large primitive trees develop. Forests consist of ferns, club mosses, horsetails and gymnosperms.
- Vertebrates appear on land.
- First winged insects.
- Animals laying amniote eggs appear (318 mya).

Pennsylvanian Epoch (Upper Carboniferous) (318.1 to 299 mya)
- 300 mya: First reptiles.
- Giant arthropods populate the land.
- Transgression and regression of the seas caused by glaciation.
- Deposits of coal form in Europe, Asia and North America.

Permian Period (299 to 251 mya)
- 275 mya: Formation of the supercontinent Pangea.
- Conifers and cycads first appear.
- Earth is cold and dry.
- 260 mya: The end of the Karoo ice age.
- Period of great volcanism in Siberia releases large volumes of gases (CO_2, CH_4, and H_2S). Oxygen (O_2) levels drop from 30% to 12%. Temperatures reach 50–60 °C on land and 40 °C at the sea-surface. The Earth's worst mass extinction eliminates 90% of ocean dwellers and 70% of land plants and animals.

The Supercontinent of Pangea

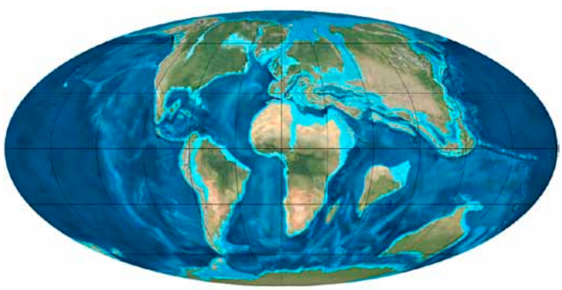
Mesozoic Era (251 to 65.5 mya)

Triassic Period (251 to 199.6 mya)
- Break-up of Pangaea starts.
- Survivors of P-T extinction spread and recolonize.
- Reptiles populate the land.
- 240 mya: Sea urchins (*Arkarua*) appear.
- 235 mya: Evolutionary split between dinosaurs and lizards.
- Giant marine *ichthyosaurs* and *plesiosaurs* populate the seas.
- First small dinosaurs such as *coelophysis* appear on land.
- 225 mya: *Adelobasileus* proto-mammal emerges.
- 205 mya: First evidence of mammals.
- 201 mya: Volcanism in Central Atlantic Magmatic Province kills 20% of all marine families.

Continued Table A1

Mesozoic Era (251 to 65.5 mya)
Jurassic Period (199.6 to 145.5 mya) – The Earth is warm. There is no polar ice. – Cycads, conifers and ginkgoes are the dominant plants. The age of dinosaurs. Giant herbivores and vicious carnivores dominate the land. – Flying reptiles (Pterosaurs) appear. – 180 mya: North America separates from Africa. – 166 mya: Evolutionary split of monotremes from primitive mammals. – 150 mya: First birds like *Archaeopteryx* appear.
Cretaceous Period (145.5 to 65.5 mya) – Period of Active Crust Plate Movements. – 125 mya: Africa and India separate from Antarctica. – Global warming starts (120 mya). – Flowering plants (angiosperms) appear. – 110 mya: Crocodiles appear. – 105 mya: South America breaks away from Africa. – Formation of the Atlantic Ocean. – The Earth has no polar ice. – Birds and the oldest group of living placental mammals develop. – 100 mya: The Earth's magnetic field is 3 times stronger than today. – 90 mya: Global warming ends. – Western Interior Seaway separates North America into Laramidia (west) and Appalachia (east). – 68 mya: *Tyrannosaurus rex* thrives. – 67 mya: Eruptions start in India and produce a great volume of lava and gases. Meteor impact, 170 km crater in Yucatan, Mexico. Mass extinction of 80–90% of marine species and 85% of land species, including dinosaurs.

Cenozoic Era (65.5 mya to today)
The supercontinent of Pangea that existed during the time of the dinosaurs had split apart by the Cenozoic, and the continents were on their paths to where they are today.
Paleogene Period (65.5 to 23.03 mya)
Tertiary Period (65.5 to 2.58 mya)



The Cenozoic Era

Continued Table A1

Cenozoic Era (65.5 mya to today)
Paleocene Epoch (65.5 to 55.8 mya) – 63 mya: End of volcanic eruptions in India. – Flowering plants become widespread. – Social insects achieve ecological dominance. – Appearance of placental mammals (marsupials, insectivores, lemuroids, creodonts). – Formation of the Rocky Mountains. – 55.8 mya: Major global warming episodes. North Pole temperature averages 23°C.
Eocene Epoch (55.8 to 33.9 mya) – 50 mya: India meets Asia forming the Himalayas. – 45 mya: Australia separates from Antarctica. – Modern mammals (rhinoceros, camels, early horses) appear. – 35.6 mya: Meteor impacts (90 and 100 km craters) in the USA and Russia create global cooling on the permanent Antarctic ice sheet.
Oligocene Epoch (33.9 to 23.03 mya) – Appearance of many types of grass. – First elephants with trunks.
Niogene Period (23.03 mya to today)
Miocene Epoch (23.03 to 5.3 mya) – African-Arabian plate joins to Asia. – 14 mya: Antarctica separates from Australia and South America. Circum-polar ocean circulation builds up the Antarctic ice sheet. – Warmer global climate, first raccoons appear. – Drying of continental interiors, forests give way to grasslands – 6 mya: Upright walking (bipedal) hominins appear.
Pliocene Epoch (5.3 to 2.58 mya) – 4.4 mya: Appearance of Ardipithecus, an early hominin genus. – 4 mya: North and South America join at the Isthmus of Panama. Animals and plants cross the new land bridge. Ocean currents change in the newly isolated Atlantic Ocean. – 3.9 mya: Appearance of Australopithecus, the genus of hominids. – 3.7 mya: Australopithecus hominids inhabit Eastern and Northern Africa. – 3 mya: Formation of the Arctic ice cap. – Accumulation of ice at the poles. Climate becomes cooler and drier. – Spread of grasslands and savannas. Rise of long-legged grazing animals.  Licensed under CC BY-SA 3.0 Australopithecus afarensis
Quaternary Period (2.58 mya to today)
Pleistocene Epoch (2.58 mya to 11,400 yrs ago) – Several major episodes of global cooling, or glaciations. – 2.4 mya: *Homo habilis* appear. – 2 mya: Tool-making humanoids emerge. Beginning of the Stone Age. – 1.7 mya: *Homo erectus* first moves out of Africa.

End of Table A1

Cenozoic Era (65.5 mya to today)
– Presence of large land mammals and birds. – 700,000 yrs ago: Human and Neanderthal lineages start diverging genetically. – 530,000 yrs ago: Development of speech in Homo Heidelbergensis. – 400,000 yrs ago: Hominids hunt with wooden spears and use stone cutting tools. – 370,000 yrs ago: Human ancestors and Neanderthals are fully separate populations. – 300,000 yrs ago: Hominids use controlled fires. – 230,000 yrs ago: Neanderthal man spreads through Europe. – 160,000 yrs ago: *Homo sapiens* appear. Origin of human female lineage (Mitochondrial Eve). – 125,000 yrs ago: Eemian stage or Riss-Würm interglacial period. Hardwood forests grow above the Arctic Circle. Melting ice sheets increase sea level by 6 meters.  Author Lillyundfreya. Licensed under CC BY-SA 3.0 Homo habilis – 105,000 yrs ago: Stone age humans forage for grass seeds such sorghum. – 80,000 yrs ago: Non-African humans interbreed with Neanderthals. – 74,000 yrs ago: Toba volcanic eruption releases a large volume of sulfur dioxide. *Homo sapiens* are reduced to about 10,000 individuals. – 60,000 yrs ago: The oldest male ancestor of modern humans. – 46,000 yrs ago: Australia becomes arid, bush fires destroy habitat and mega-fauna die off. – 40,000 yrs ago: *Cro-Magnon* man appears in Europe. – 28,000 yrs ago: Neanderthals disappear from fossil record. – 22,000 yrs ago: *Tioga* glacial maximum sea level is 130 meters lower than today. – 20,000 yrs ago: Invention of the first fired pottery vessels. – 19,000 yrs ago: Antarctic sea ice starts melting. – 15,000 yrs ago: Bering land bridge between Alaska and Siberia allows human migration to America. – 12,900 yrs ago: The explosion of a comet over Canada causes the extinction of American mega-fauna such as the mammoth and sabretooth cat. – 11,400 yrs ago: End of the Würm/Wisconsin glacial period. Sea level rises by 91 meters. Holocene Epoch (11,400 years ago to today) – Development of agriculture. Domestication of animals. – 9,000 yrs ago: Metal smelting starts. – 5,500 yrs ago: Invention of the wheel. – 5,300 yrs ago: The Bronze Age. – 5,000 yrs ago: Development of writing. – 4,500 yrs ago: Pyramids of Giza. – 3,300 yrs ago: The Iron Age. – 2,230 yrs ago: Archimedes advances mathematics. – 250 yrs ago: Start of the Industrial Revolution. – 50 yrs ago: Space travel: artificial satellite orbits the earth (1957), humans walk on the surface of the moon (1969).

The Near-Term Future of the Earth

Human industrial activity that relies on burning fossil fuels, such as coal and petroleum products, has been generating greenhouse gases carbon dioxide (CO_2), methane (CH_4), and nitrous oxide (N_2O) in large quantities since about 1750 (Figure A1). The chart below shows the levels of atmospheric carbon dioxide during the last millennium and its sharp rise during the last century. Atmospheric models predict that elevated greenhouse gases will cause global warming and influence weather patterns that will melt polar ice and destroy the habitat of animals such as the polar bear. The increase of global temperatures will also reduce the amount of snow deposited on mountains thus decreasing the flow of water in rivers which are now used for navigation, irrigation and as sources of potable water. Carbon dioxide will also increase the acidity of sea water and threaten coral reefs and shell-building oceanic life forms.

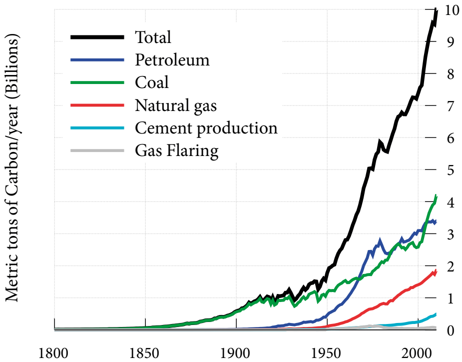


Figure A1. Atmospheric concentration of carbon dioxide

Today, the concentration of atmospheric carbon dioxide is 380 parts per million (ppm) and the North Pole's mean annual temperature is –20 °C. The analysis of core sediments in the Arctic Circle indicates that 55 million years ago, carbon dioxide concentration was 2,000 ppm and the North Pole's temperature averaged 23 °C. Satellite images show approximately a 20% reduction in the Earth's minimum ice cover between 1979 and 2003. Arctic perennial sea ice has been decreasing at a rate of 9% every ten years. At this rate, the summertime Arctic Ocean will be ice-free before the year 2100.

There is a large amount of water stored as ice over the landmasses of Greenland and Antarctica. If the ice sheets melt, the resulting rise in the global sea level will flood many coastal areas around the world. The Greenland ice sheet contains enough water to increase the global sea level by 7.3 meters, the West Antarctic ice sheet could raise sea level by 5.8 meters and the East Antarctic ice sheet could raise the sea level globally by 51.8 meters. The combined effect of melting all the ice on Greenland and Antarctica would result in a sea level rise of 65 meters.

The Long-Term Future of the Earth

The future of the Earth is linked to the fate of the Sun that is halfway through its life cycle and will exhaust its supply of hydrogen fuel in around 4,000 million years. As the Sun cools, its core will collapse and its atmosphere will expand transforming the Sun into a red giant star. The swelling Sun will engulf the planets closest to it, and the Earth will be completely vaporized. The Sun will die in several stages. When its core crashes inwards, it will start fusing helium atoms into carbon. When helium supply runs out, the centre will collapse again and form a white dwarf star that will become dimmer until its light finally fades. The final collapse of stars that are a few times larger than the Sun ends in a massive supernova explosion that leaves behind a rapidly spinning neutron star.

Long before the Sun becomes a white dwarf, 2,000 million years from now, our Milky Way Galaxy is predicted to collide with the Andromeda Galaxy. The collision will take place for several million years and result in one combined super galaxy named *Milkomeda*. The sun may become a part of the Andromeda system during the collision and could eventually end up far away from the new merged galactic centre.

Long-Term Future of the Earth (my – millions of years)

+200 years: Possible global warming event caused by anthropogenic carbon dioxide (CO_2).
+1500 my: The sun is about 6000 million years old and 15% brighter than today.
+2000 my: The Milky Way Galaxy starts colliding with Andromeda Galaxy.
+3000 my: The Solar system becomes a part of the new Milkomeda Galaxy.
+4000 my: The sun is about twice as bright as today and its radius is 40% greater. The sun starts exhausting its supply of hydrogen.
+5000 my: The sun starts changing into a red giant star, 3 times its present size. The Earth is engulfed by the red giant Sun.
+10000 my: The red giant Sun collapses and becomes a white dwarf.
+20000 my: The white dwarf Sun becomes a black dwarf.

Appendix II.
Timelines of Technologies Development

Prehistoric technology is the one that predates recorded history. History is the study of the past using written records. Anything prior to the first written accounts of history is prehistoric, including earlier technologies. About 2.5 million years before writing was developed, technology began with the earliest hominids that employed stone tools that may have been used for starting fires, hunting and burying their dead. The following outline is provided as an overview of and topical guide to prehistoric technology.

The Stone Age is a broad prehistoric period during which stone was widely used for manufacturing implements with a sharp edge, a point or a percussion surface. The period lasted roughly 2.5 million years, from the time of early hominids to Homo sapiens in the later Pleistocene era, and largely ended between 6000 and 2000 B.C. with the advent of metalworking. The Stone Age lifestyle was that of hunter-gatherers who travelled to hunt game and gather wild plants, with minimal changes in technology. As the last glacial period of the current ice age neared its end (about 12,500 years ago), large animals like the mammoth and bison antiquus became extinct and climate changed. Humans adapted by maximizing resources in local environments, gathering and eating a wider range of wild plants and hunting or catching smaller game. The domestication of plants and animals at early stages in the Old World (Afro-Eurasia) Mesolithic and New World (American continent) Archaic periods led to significant changes and reliance on agriculture in the Old World Neolithic at a New World Formative stage. The agricultural life led to more settled existences and significant technological advancements.

The Lower Paleolithic is the earliest subdivision of the Paleolithic or Old Stone Age. It spans the time from around 2.5 million years ago when the first evidence of craft and use of stone tools by hominids appears in the current archaeological record, until around 300,000 years ago, spanning Oldowan and Acheulean lithic technologies. Stone tools were applied by the early human (hominid) who used stone tool technology such as a hand axe, similar to that of primates, which is found to be limited to the intelligence levels of modern children aged 3 to 5 years. The first "homo" species began with Homo habilis about 2.4 to 1.5 million years ago. The intelligence and use of technology did not change much for millions of years. *Homo habilis* ("handy man") lived from approximately 2.3 to 1.4 million years ago in Africa and created stone tools called Oldowan tools. *Homo ergaster* lived in eastern and southern Africa about 2.5 to 1.7 million years ago and used more diverse and sophisticated stone tools than its predecessor, Homo habilis, including having refined the inherited Oldowan tools, developed the first Acheulean bifacial axes. *Homo erectus* ("upright man") lived about 1.8 to 1.3 million years ago in West Asia and Africa and is thought to be the first hominid to hunt in coordinated groups, employ complex tools and care for infirm or weaker companions. *Homo antecessor* is the earliest hominid in Northern Europe lived from 1.2 million to 800,000 years ago and used stone tools. *Homo heidelbergensis* lived between 600,000 and 400,000 years ago and applied stone tool technology similar to Acheulean tools used by Homo erectus. Control over fire by early humans – European and Asian sites dating back 1.5 million years ago seem to indicate the controlled use of fire by Homo erectus. Homo heidelbergensis may have been the first species to bury their dead about 500,000 years ago.

The Middle Paleolithic period in Europe and the Near East covers a time period during which the Neanderthals lived (c. 300,000–28,000 years ago). The earliest evidence of settlement in Australia dates to around 40,000 years ago when modern humans likely crossed from Asia by island-hopping. Rock shelters exhibit the earliest traces of human life in India, some of which are approximately 30,000 years old. Homo neanderthalensis used stone tools that date back to around 300,000 years ago and include smaller, knife-like and scraper tools. Homo neanderthalensis buried their dead, doing so in shallow graves along with stone tools and animal bones. Homo sapiens are the only living species in the Homo genus originated in Africa about 200,000 years ago. Greater mental capability and ability to walk erect provided freed hands for manipulating objects, which allowed for far greater use of tools. The earliest undisputed human burial so far dates back 130,000 years. Human skeletal remains stained with red ochre and with a variety of grave goods.

The Upper Paleolithic period is the one when advancements in human intelligence and technology changed radically between 60,000 and 30,000 years ago. Homo sapiens reached full behaviour modernity around 50,000 years ago due to the highly developed brain capable of abstract reasoning, language, introspection and problem solving. Tools were stone bladed, made of antlers or bones. Human beings may have begun wearing clothing as far back as 190,000 years ago. Sewing needles possibly appeared around 40,000 years ago, which indicates people were wearing clothes at this time. Art included cave painting, sculpture, carvings and engravings of bone and ivory as well as musical instruments such as flutes.

The Mesolithic period is the transitional period between Paleolithic hunter-gatherers, beginning with the Holocene warm period around 11,660 BP and ending with the Neolithic introduction of farming, the date of which varied in each geographical region. Adaptation was required during this period due to climate changes that affected environment and the types of available food. Spears or arrows were found in the earliest known Mesolithic battle site in the Sudan area.

The Neolithic period was the first agricultural revolution representing the transition from hunting and gathering nomadic life to agriculture existence. It evolved independently in six separate locations worldwide circa 10,000–7000 years BP (8,000–5,000 B.C.). The introduction of agriculture is a defining characteristic of Neolithic societies, which resulted in a swing from a nomadic lifestyle to one that was more sedentary, the use of agricultural tools such as a plough, digging stick and hoe (tool), the domestication of animals, including dogs and pottery that emerged as a defining characteristic of the Neolithic period. Architecture included houses and villages built of mud-brick and wattle and daub and the construction of storage facilities, tombs and monuments. Metalworking embraced copper use and began as early as 9000 B.C. in the Middle East.

A copper pendant found in northern Iraq is dated to 8700 B.C. Numeric counting as record keeping evolved from a system of counting using small clay tokens that began in Sumer about 8000 B.C. Stone tools – ground and polished ones – were created during the Neolithic period and the wheel – in the late Neolithic period when was introduced for making pottery.

Table A2. Timelines of Tool and Technology Development by Years

Years	Events
1.8 million years ago	Fire and then cooking
500 thousand years ago	Shelter construction
400 thousand years ago	Pigments in Zambia. Spears in Germany
200 thousand years ago	Glue in Italy
160–40 thousand years ago	Burial
64 thousand years ago	Arrowhead in South Africa
61 thousand years ago	Sewing needle in South Africa
60 thousand years ago	Bow
36 thousand years ago	Cloth woven from flax fibre in Georgia
35 thousand years ago	Flute in Germany
28 thousand years ago	Twisted rope
16 thousand years ago	Pottery in China
6000 B.C.	Kiln in Mesopotamia
5000–4500 B.C.	Rowing oars in China
3000 B.C.	Bronze in Mesopotamia
3000 B.C.	Papyrus in Egypt
	1st millennium B.C.
7th-mid–7th century B.C.	Two-masted ships (foresail) by Etruscans in Italy
6th century B.C.–515 B.C.	Crane in Ancient Greece
5th century B.C.	Cast iron in Ancient China; crossbow in Ancient China and Ancient Greece
5th–3rd century B.C.	Cupola furnace in Ancient China; cupola furnace were used for remelting most, if not all, iron smelted in the blast furnace.
5th–4th century B.C.	Trebuchet in Ancient China
Before 421 B.C.	Catapult in Ancient Greece
3rd century B.C.	The papermaking process in China; canal lock in Ancient Suez Canal in Hellenistic Egypt

Continued Table A2

Years	Events
	1st millennium B.C.
Early 3rd century B.C.	Waterway connecting two seas (Ancient Suez Canal) by Greek engineers
3rd century B.C.	Raised-relief map in China; water wheel in Hellenistic kingdoms
3rd–2nd century B.C.	Blast furnace in Ancient China
After 205 B.C.	Dry dock in Hellenistic Egypt
2nd century B.C.	A spritsail used on a Roman merchant ship; finery forge in China (finery forges used for making wrought iron from pig iron); the pulp papermaking process in China
1st century B.C.	Segmental arch bridge, arch dam, watermill (grain mill) in the Roman Republic
	1st millennium A.D.
1st century	Multiple arch buttress dam in the Roman Empire
1st–2nd century	Buttress dam in the Roman Empire
2nd century	Seismometer in China; crankshaft and lateen sail in the Roman Empire
2nd–3rd century	Arch-gravity dam in the Roman Empire
3rd century	Roman Hierapolis sawmill (the earliest known machine to incorporate a crank and connecting rod mechanism)
4th century	Field mill, fishing reel in Ancient China
4th–5th century	Paddle wheel boat in the Roman Empire
5th century	Horse collar in China
5th–6th century	Pointed arch bridge in Cappadocia, the Eastern Roman Empire
6th century	Pendentive dome (Hagia Sophia) in Constantinople, toilet paper in China
7th century	Banknote; porcelain in China
9th century	Gunpowder in China; numerical zero in Ancient India
10th century	Fire lance that shot a weak gunpowder blast of flame and shrapnel, fireworks in China
	2nd millennium A.D.
1000	Chinese invent gunpowder
1041	Movable type printing press invented in China
1050	The decimal number system is brought to Spain by Arabs
1280	Eyeglasses are invented
1350	Suspension bridges built in Peru
1450	Alphabetic, movable type printing press invented by Johann Gutenberg

Continued Table A2

Years	Events
	2^{nd} millennium A.D.
1450	Alphabetic, movable type printing press invented by Johann Gutenberg
1500	Ball bearing invented by Leonardo Da Vinci together with flying machines, including a helicopter; the first mechanical calculator and one of the first programmable robots
1510	Pocket watch invented by Peter Henlein (Germany)
1576	Ironclad warship invented by Oda Nobunaga (Japan)
1581	Pendulum invented by Galileo Galilei (Italy)
1603	Thermometer invented by Galileo Galilei
1608	Telescope invented by Hans Lippershey (German-Dutch lens-maker)
1609	Microscope invented by Galileo Galilei
1614	Logarithms developed by John Napier (Scotland)
1642	Adding machine invented by Blaise Pascal (France)
1643	Barometer invented by Evangelista Torricelli (Italy)
1645	Vacuum pump invented by Otto von Guericke (Germany)
1657	Pendulum clock invented by Christiaan Huygens (The Netherlands)
1679	Pressure cooker invented by Denis Papin (France)
1683	Bacteria discovered by Anton van Leeuwenhoek (The Netherlands)
1687	"Principia". Newton's physics forms the foundation of modern science (UK)
1698	Steam engine invented by Thomas Savery (England)
1671	Infinitesimal calculus developed by Gottfried Leibniz (Germany)
1705	Steam piston engine invented by Thomas Newcomen (England)
1708	Mechanical (seed) sower invented by Jethro Tull (England)
1710	Thermometer invented by Rene Antoine Ferchault de Reaumur (France)
1733	Flying shuttle invented by John Kay (Germany)
1742	Franklin stove invented by Benjamin Franklin (USA)
1752	Lightning rod invented by Benjamin Franklin
1767	Spinning jenny invented by James Hargreaves (England)
1769	Steam engine invented by James Watt (England)
1774	Oxygen isolated by J. Priestly (England)

Continued Table A2

Years	Events
	2^{nd} millennium A.D.
1779	The weaving process initiated by first steam-powered mill automation
1781	The planet Uranus (Germany) discovered by William Herschel
1787	Lithographic printing process (called polyautography) invented in Germany
1783	Hot air balloon invented by Montgolfier brothers (France)
1791	Steamboat invented by John Fitch (USA)
1798	Vaccination invented by Edward Jenner (England)
1799	Nitrous oxide (laughing gas) discovered by Humphrey Davy (England); The conveyer belt invented by Oliver Evans (USA).
1800	First practical batteries
1804	Locomotive invented by Richard Trevithick (England)
1814	Steam locomotive invented by George Stephenson (England)
1816	Miner's safety lamp invented by Humphry Davy (England). The first bicycle developed by Karl D. Von Sauerbronn (Germany). The first photographic negative developed by a chemist Joseph Niepce (France)
1816	Stethoscope invented by Rene Laennec (France)
1820	The *Arithmometer*, first mass-produced calculator, invented by Charles Xavier Thomas de Colmar (France)
1821	Electric motor invented by Michael Faraday (England)
1822	First mechanical computer designed by Charles Babbage (England)
1823	The electromagnet developed by William Sturgeon (England)
1824	Portland Cement invented by Joseph Aspdin (England)
1826	Photography invented by Joseph Nicephore Niepce (France)
1831	Chloroform discovered by J.von Liebig (Germany); M. Faraday (England) discovers electro-magnetic current thus making possible generators and electric engines
1834	Braille alphabet invented by Louis Braille (France)
1834	Refrigerator invented by Jacob Perkins (USA)
1834	Combine harvester invented by Hiram Moore (USA)
1835	Morse code invented by Samuel Morse (USA). Revolver invented by Samuel Colt (USA)
1838	Electric telegraph invented by Charles Wheatstone (England) (also Samuel Morse). Louis Daguerre (France) perfects the Daguerrotype

Continued Table A2

Years	Events
	2^{nd} millennium A.D.
1839	Vulcanization of rubber invented by Charles Goodyear (USA)
1842	Anaesthesia invented by Crawford Long (USA)
1843	Typewriter invented by Charles Thurber (USA)
1846	Sewing machine invented by Elias Howe (USA). Rotary printing press invented by Richard M. Hoe (USA).
1849	Reinforced concrete developed by Joseph Monier (France)
1859	Charles Darwin (England) publishes *The Origin of Species*. Etienne Lenoir (Belgium) demonstrates the first successful gasoline engine
1862	Revolving machine gun invented by Richard J. Gatling (USA). Mechanical submarine invented by Narcís Monturiol i Estarriol (Spain)
1866	Dynamite invented by Alfred Nobel (Sweden). The methods for refining aluminium using electrolytic action developed by Charles Hall (USA)
1869	The Periodic Table (Russia) produced by Dimitri Mendeleev
1873	The Remington typewriter invented by Christopher Sholes (USA). James Clerk Maxwell (Scotland) states the laws of electro-magnetic radiation. The most popular form of barbed wire developed by Joseph Glidden (USA).
1876	Gasoline carburettor invented by G. Daimler (Germany)
1877	Phonograph invented by Thomas Alva Edison (USA). Microphone invented by Emile Berliner (USA).
1878	Cathode ray tube invented by William Crookes (England). The incandescent lamp invented by A. Edison (USA).
1883	The first skyscraper built in Chicago (ten stories). The machine gun invented by S. Maxim (UK)
1885	Motor cycle invented by Gottlieb Daimler and Wilhelm Maybach (Germany). The first automobile to run on the internal-combustion engine developed by Karl Benz (Germany)
1888	Radio waves produced by Heinrich Hertz (Germany). Eiffel Tower built in (France)
1891	Zipper invented by Whitcomb L. Judson (USA)
1893	Wireless communication invented by Nikola Tesla (Austrian Empire). Diesel invented by Rudolf Diesel (Germany)
1895	Radio signals invented by Guglielmo Marconi (Italy). Cinematograph developed by Auguste and Louis Lumiere (France). X-rays discovered by Wilhelm Roentgen (Germany)
1896	Typesetting machines were developed. The dishwashing machine was patented

Continued Table A2

Years	Events
	2^{nd} millennium A.D.
1898	Liquefied hydrogen produced by James Dewar (UK). The hydrogen was cooled to -205 °C under pressure of 180 atmospheres
1900	Quantum theory developed by Max Planck (Germany). The first rigid airship Zeppelin is built (Germany). The tractor developed by Benjamin Holt USA)
1901	Vacuum cleaner invented by Hubert Booth (UK). The first safety razor for shaving developed by King C.Gillette (USA). Guglielmo Marconi (Italy) used radio waves to transmit the letter "S" in Morse code
1903	Powered airplane invented by Wilbur Wright and Orville Wright (USA)
1905	The Theory of Relativity (Germany) written by Albert Einstein
1907	Colour photography invented by Auguste and Louis Lumiere (France). Helicopter invented by Paul Cornu (France). Radio amplifier invented by Lee DeForest (USA). Bakelite invented by Leo Baekeland (USA)
1911	The model of the atom proposed by Ernest Rutherford (UK)
1919	The first electric typewriter developed by James Smathers (USA)
1923	Sound film invented by Lee de Forest (USA). Television Electronic invented by Philo Farnsworth (USA). The first workable TV camera developed by Vladimir Zworykin (Russia)
1924	Electro Mechanical television system invented by John Logie Baird (Scotland)
1926	The first successful liquid fuelled rocket (USA) launched by Robert Goddard. Aerosol can invented by Erik Rotheim (Norway)
1928	Antibiotics, penicillin invented by Alexander Fleming (Scotland). The first electric razor developed by Jacob Schick (USA)
1931	Iconoscope invented by Vladimir Zworykin (USSR)
1937	Jet engine invented by Frank Whittle (UK) and Hans von Ohain Alan (Germany). The concept of a theoretical computing machine (UK) developed by Alan Turing
1938	Ballpoint pen invented by Laszlo Biro (Hungary)
1942	Enrico Fermi (USA) lead a large team that developed the first nuclear fission reactor
1943	Alan Turing (UK) develops the code-breaking machine that breaks Germany encryption machine Colossus. Aqua-lung invented by Jacques-Yves Cousteau and Emile Gagnan (France)
1944	Wernher von Braun (Germany) led a team of engineers to develop a liquid fuelled rocket that could deliver a war head 320 km. This was the first ballistic missile used in the war
1945	The first atomic bomb detonated at the Trinity NM test site (USA) on 16 July 1945

Continued Table A2

Years	Events
	2^{nd} millennium A.D.
1946	Microwave oven invented by Percy Spencer (USA). ENIAC (Electronic Numerical Integrator and Calculator) was developed (USA)
1948	Bell Laboratories (USA) announced the invention of the transistor
1951	Nuclear power reactor invented by Walter Zinn (USA). Universal Automatic Computer (UNIVAC) invented by John Presper Eckert and John Mauchly (USA)
1956	Optical fibre invented by Basil Hirschowitz, C. Wilbur Peters and Lawrence E. Curtiss (USA). Videocassette recorder invented by Alexander M. Poniatoff (USA)
1957	Sputnik I and Sputnik II launched by the USSR. IBM (USA) shipped its first copies of Fortran to customers
1958	The first integrated circuit, or silicon chip, produced by the Jack Kilby & Robert Noyce (USA). The theoretical work for lasers developed by Charles H. Townes, Arthur L. Schawlos (USA), Nicolay Basov and Aleksandr Prokhorov (USSR).
1959	The first push button plain xerographic paper copier (Xerox Corporation, USA)
1960	Laser invented by Theodore Harold Maiman (USA).
1961	Juri Gagarin (USSR) is the first man in space. Alan Shepard (USA) made the first sub-orbital flight. Optical disc invented by David Paul Gregg (USA)
1963	Computer mouse invented by Douglas Engelbart (USA). Cassette tapes and players introduced by Phillips Electronics (The Netherlands).
1967	Automatic Banking Machine (cash machine) invented by John Shepherd-Barron (UK). Hypertext invented by Andries van Dam and Ted Nelson (USA)
1968	Computer mouse demonstrated by Telefunken company (Germany). Video game console invented by Ralph H. Baer (USA).
1969	Moon landing – Neil Armstrong and Buzz Aldrin set feet on the moon (USA)
1971	E-mail invented by Ray Tomlinson (USA). Liquid Crystal Display invented by James Fergason (USA). Pocket calculator invented by Sharp Corporation (Japan). Floppy Disk invented by David Noble with IBM (USA). The first commercial microprocessor product from Intel (USA)
1972	The first E-mail message was sent over the ARPANET by Ray Tomlinson (USA)
1976	Viking 1 (USA) landed on Mars and was the first craft from Earth to land on another planet
1981	The first space shuttle was launched. The ship was piloted by John Young (USA)
1986	Shuttle rocket *Challenger* (USA) explodes during the launch killing the entire crew. The Chernobyl nuclear disaster (Ukraine)

End of Table A2

Years	Events
	2^{nd} millennium A.D.
1991	Human stem cells for blood in bone marrow discovered by Irving Weissman (USA)
1992	The Internet was opened up to commercial business
1993	Intel introduces the Pentium processor (USA). CERN donated the WWW technology to the world
1994	The Global Positioning System (GPS) with 24 satellites became fully operational (USA)
1995	Microsoft introduces Windows 95 for the Intel PC family of computers (USA)
1996	Java programming language introduced by Sun Microsystems (USA). The first time computer defeats the world chess champion, Gary Kasparov in a match. The implantation of the first total artificial heart (Taiwan)
1997	Blogging begins
1998	John Glenn returned to Earth orbit for 7 days as a part of a US Space Shuttle Mission. Google opens as the major Internet search engine and index (USA)
1999	Identified stem cells in the adult brain that grow into nerve cells
2000	Detailed mapping of the human genome was completed. NASA announced that by studying high resolution photographs from the Mars Global Surveyor spacecraft, scientists spotted gullies and trenches and fanlike deltas that could have been carved by fast-moving water. Intel and AMD both started shipping PC CPU chips running at 1 GHz processor speed. The explosion and sink of the nuclear submarine Kursk (Russia)
2001	Digital satellite radio (USA). *Wikipedia*, the online encyclopaedia and world's largest wiki (Hawaiian word *wiki*, meaning "quick"), launched by Jimmy Wales and Larry Sanger (USA).
2003	First cloning of a mammal (Dolly the sheep) (Scotland)
2004	*Facebook*, another social networking website, was started
2005	*YouTube* began storing and retrieving videos
2006	*Twitter* was launched as a social networking and micro-blogging site
2008	The first full transplant of a human organ grown from adult stem cells carried out by Paolo Macchiarini at the Hospital Clínic of Barcelona (Spain)
2010	Two-dimensional material *graphene* produced by Andre Geim and Konstantin Novoselov at the University of Manchester (UK).
2011	Fukushima Daiichi nuclear disaster (Japan)
2012	The first complete sequences of individual human genomes (USA)

Appendix III.
Emerging technologies

Emerging technology	Status	Potential applications
	Electronics	
Spintronics	Working prototypes	Data storage. Mechanical magnetic hard disk drives
Electronic nose	Research, commercialization	Detecting spoiled food, chemical weapons and cancer
Thermal copper pillar bump	Working prototypes in discrete devices	Conventional thermal systems, heat sinks, bulk thermo-electrics. Electric circuit cooling, micro-fluidic actuators, small-device thermoelectric power generation
Flexible electronics	Research, development, some prototypes	Flexible and folding electronic devices (such as smart phones), flexible solar cells that are lightweight, can be rolled up for launch and are easily deployable
Memristor	Working prototype	Some current integrated circuits, many other electronics devices. Smaller, faster, lower power consuming storage, analogue electronics. Artificial intelligence
	Energy	
Airborne wind turbine	Research	Producing electricity
Artificial photosynthesis	Research, experiments growing interest in a macro-science global project	Improve natural photosynthesis, convert sunlight and water into hydrogen and carbon dioxide into carbohydrates
Biofuels	Diffusion	Energy storage, fuels for transport
Concentrated solar power	Growing markets in California, Spain, Northern Africa	Photovoltaics. Producing electricity
Fusion power	Hypothetical, experiments; for 60+ years	Fossil fuels, renewable energy, nuclear fission power. Producing electricity, heat, fusion torch, recycling with waste heat
Lithium-air battery	Research, experiments	Energy storage methods: hydrogen, chemical batteries, some uses of fossil fuels. Laptops, mobile phones, long-range electric cars; storing energy for an electric grid

Continued Appendix III

Emerging technology	Status	Potential applications
	Energy	
Nanowire battery	Experiments, prototypes	Energy storage methods: hydrogen, chemical batteries, some uses of fossil fuels. Laptops, mobile phones, long-range electric cars; storing energy for an electric grid
Thorium fuel cycle	Research started in the 1960s, still ongoing	Uranium based nuclear power. Producing electricity, heat
Wireless energy transfer	Prototypes, diffusion, short range consumer products	Power cords, plugs, batteries. Wirelessly powered equipment: laptop, cell phones, electric cars, etc.
Energy harvesting	Experiments	Batteries. Constant energy source for mobile, wearable and ubiquitous devices
	Manufacturing	
3D printing	Commercial production	Manual prototype production, some methods of mass production that lack customizability. Rapidly prototyping and producing plastic objects and multi-material items with the potential for significantly customizing products for individual consumers
Molecular assembler	Hypothetical, experiment	3D printing, traditional manufacturing methods and tools
	Materials science	
Quantum dots	Research, experiments, prototypes	LCD, LED, Quantum dot laser, future use as programmable matter in display technologies, optical high-speed data transmission, laser scalpel
Aerogel	Hypothetical, experiments, diffusion, early uses	Traditional insulation, improved insulation, insulated glass in case it can be made clear, sleeves for oil pipelines, aerospace, high-heat and extreme cold applications
Conductive polymers	Research, experiments, prototypes	Conductors, lighter and cheaper wires, antistatic materials, organic solar cells
Femtotechnology, picotechnology	Hypothetical	Present nuclear, new materials, nuclear weapons, power
Graphene	Hypothetical, experiments, diffusion, early uses	Silicon-based integrated circuits, components with higher strength to weight ratios, transistors that operate at higher frequency, storing hydrogen for fuel cells, sensors to diagnose diseases

Continued Appendix III

Emerging technology	Status	Potential applications
	Materials science	
High-temperature superconductivity	Cryogenic receiver front-end RF and microwave filter systems for mobile phone base stations; hypothetical and experiments for higher temperatures	Copper wire, semiconductor integral circuits. No loss conductors, frictionless bearings, magnetic levitation, lossless high-capacity accumulators, electric cars, heat-free integral circuits and processors
High-temperature superfluidity	Superfluid gyroscopes already exist but work at very low temperatures	Mechanical gyroscope, high-precision measure of gravity, navigation and manoeuvre devices, possible devices to emit the gravitomagnetic field, frictionless mechanical devices
Metamaterials	Hypothetical, experiments, diffusion	Classical optics. Microscopes, cameras, metamaterial cloaking, cloaking devices
Programmable matter	Hypothetical, experiments	Coatings, catalysts. Claytronics, synthetic biology
Nanomaterials: carbon nanotubes	Hypothetical, experiments, diffusion, early uses	Structural steel and aluminium. Stronger, lighter materials, space elevator
Multi-function structures	Hypothetical, experiments, some prototypes, few commercial	Composite materials. Self health monitoring, self healing materials
	Robotic	
Unmanned vehicle, underwater vehicle, autonomous underwater vehicle	Research and development, some diffusion	Manned vehicles, human spying. Surveillance, oceanography, commercial aerial surveillance, wildfire mapping, road patrol and anti-piracy. Search and rescue, Unmanned space and surface vehicle
Swarm robotics	Hypothetical, experiments	Distributed computing, autonomous construction, space construction
Powered exoskeleton	Research, development, prototypes, diffusion, commercializing	Electric wheelchairs, forklifts. Heavy lifting, muscle related diseases, warfare, care for the elderly and disabled.
Molecular nanotechnology, nanorobotics	Hypothetical, experiments	Producing products and parts, retail distribution. Machines (desktop, industrial) that can make anything given the materials
	Transport	
Airless tire	Research, development, early prototypes	Conventional tire, safer tires
Alcubierre drive	Attempting to create a warp field in lab tests	Conventional space propulsion. Interstellar exploration and colonization

Continued Appendix III

Emerging technology	Status	Potential applications
	Transport	
Alternative fuel vehicle	Commercialization, diffusion	Internal combustion engine. Electric vehicle, hydrogen vehicle, compressed air vehicle. Reducing air pollution, decreasing oil consumption
Flexible wings (Active Aeroelastic Wing, Adaptive Compliant Wing), fluidic flight controls	Experiments, prototypes	Flight control systems: ailerons, elevators, elevons, flaps, flaperons. Controlling aircraft, ships
Flying car	Early commercialization, prototypes	More effective road transportation
Maglev train, vactrain	Research, early commercialization	Trains with higher speed
Scramjet	Research, development	Hypersonic aircraft jet engine
Solar sail	In 2010, IKAROS was the world's first spacecraft designed to use solar sailing propulsion to be successfully launched	Space travel
Hyperloop	Research, development	High-speed rail. Faster way to get somewhere
Nuclear launch cannon	Hypothetical	Very cheap launch for cargo
	Military	
Airborne laser	Research and development, trials, discontinued	Missile defence. Tracking and destruction of tactical ballistic missiles
Directed energy weapon	Research, development, some prototypes	Warfare. Firearms. Laser weapon system
Electromagnetic weapons	Research and development	Warfare (coilgun, railgun)
Particle beam weapon	Research and development	Warfare (Strategic Defense Initiative)
Pure fusion weapon	Hypothetical	
Stealth technology	Research and development	Camouflage. Electronic countermeasures (Stealth aircraft, radar-absorbent material)

Continued Appendix III

Emerging technology	Status	Potential applications
	Others	
Closed ecological systems	Research and development, working demonstrators	Agriculture, scientific research, space colonization
Genetically modified food	Research and development, commercialization	
Meat incubator	Research and development	Creating synthetic meat products
Body implants, prosthesis	Trials, from animal (e.g. brain implants) to human clinical (e.g. insulin pump implant)	Various fields of medicine. Brain implant, retinal implant, prosthetics
Genetic engineering	Research, development, commercialization	Animal husbandry, plant breeding. Creating and modifying species, eliminating genetic disorders, genetically modified food
Stem cell treatments	Research, experiments	Other therapies, treatment for a wide range of diseases and injuries
Tissue engineering	Research, diffusion	Organ printing
Artificial brain	Research	Treating neurological diseases; artificial intelligence
Machine translation	Diffusion	Human translation of natural languages in the areas where the language is formalized
Machine vision	Research, prototyping, commercialization	Biotic vision and perception, including humans. Biometrics, controlling processes (e.g. in a driverless car, automated guided vehicle), detecting events, robot vision
Optical computing	Hypothetical, experiments; some components of integrated circuits have been developed	Many electronics devices, integrated circuits. Smaller, faster, lower power consuming computing
Quantum computing	Hypothetical, experiments, commercialization	Atomtronics, electronic computing, optical computing, quantum clock. Much faster computing. Hypothetical high-temperature superconductivity and superfluidity
Speech recognition	Research, development, commercialization	Mechanical input devices

End of Appendix III

Emerging technology	Status	Potential applications
	Others	
Atmospheric carbon dioxide removal	Research and development	Fossil carbon sources. Climate change mitigation, input for producing carbon-neutral fuel
Cloak of invisibility	Successful experiments cloaking small objects under some conditions	Camouflage, cloaking microscope tips at optical frequencies
Magnetic levitation	Research, development, commercialization (Maglev train)	Wheels, tires, conventional transportation systems. High temperature superconductivity, cryogenics, low temperature refrigerators, superconducting magnet design and construction, precision manufacturing, maglev car, maglev based spacecraft launch
Magnetic refrigeration	Already used for achieving cryogenic temperatures in the laboratory setting (below 10K)	Refrigeration without the need for compression and with more energy efficient that may appear in refrigerators, air conditioners and spacecrafts
Magnetorheological fluid	Developed and researched	Shock absorber. Magnetorheological damper, heavy motor damping, seismic dampers, body armour fluid bullet resistant, electrorheological fluid

Appendix IV.
The main directions to activities performed in materials science and engineering for achieving strategic aims concerning society development

Main directions of activity	Evaluation of the Current Situation and Plans for the Future
Materials design	The subject of contemporary materials science and engineering is the adjustment of materials, beginning from their chemical composition, constituent phases and microstructure up to the set of properties required for a particular application. Computer simulation is employed in certain cases, which will provide a possibility of improving the properties along with their prediction even before manufacturing materials and with a significant reduction in expenditures and time required for investigating and implementing them.
Computational materials science	Soon, computer modelling will become an indispensable tool in materials science and engineering. Computer strategy provides the description of materials from the chemical and physical points of view in a broad scale of both dimension and time.
Advanced analytical techniques	The development of new engineering materials in the future and the discovery of new phenomena making decisions on their properties call for the development, introduction and dissemination of new and more efficient research techniques thus making the examination of materials possible in the atomic scale like high-resolution transmission electron microscopy, scanning probe microscopy, X-ray and neutron diffraction as well as various types of spectroscopy integrated with more powerful computers facilitating fast visualisation.
Synthesis and processing	The goal of manufacturing and processing techniques for the future is to design engineering materials from the complex arrangements of atoms and particles with the same accuracy and control as is currently used in semiconductor materials, and including, e.g. chemical conversion from simple precursor units, fast prototyping of ceramic and metal components using streaming technique, microwave sintering and deposition methods for gas phases (CVD, PVD) to form thin films and make the infiltration of composites to the most promising techniques.
Nanomaterials	The capacity for verifying, synthesizing and designing materials in the nanometric scale (10^{-9} m) features one of the main directions to progress, which is to use those materials for developing their new applications, scrap and waste reduction as well as for optimizing properties in all basic groups of material engineering.
Smart materials	Smart materials, different from others, are designed in such a way that they react to external stimuli and improve their properties thus adapting themselves to environmental conditions. All work in the discussed area is considered especially vanguard in its character.
Biomimetic materials	Thanks to a better understanding of the development of minerals and composites by live organisms, biomimetic materials have become a fast developing area of material engineering thus enabling to copy biological processes and materials, both organic and inorganic.

Algirdas Vaclovas Valiulis
A History of Materials and Technologies Development

Textbook

The purpose of the book is to provide the students with the text that presents an introductory knowledge about the development of materials and technologies and includes the most commonly available information on human development. The idea of the publication has been generated referring to the materials taken from the organic and non-organic evolution of nature. The suggested texts might be found a purposeful tool for the University students proceeding with studying engineering due to the fact that all subjects in this particular field more or less have to cover the history and development of the studied object. It is expected that studying different materials and technologies will help the students with a better understanding of driving forces, positive and negative consequences of technological development, etc.

Vadovėlyje nagrinėjama inžinerinių medžiagų (medžio, kaulo, keramikos, cemento, betono, juodųjų ir spalvotųjų metalų, polimerų ir kt.) ir technologijų istorinė raida. Nagrinėjama architektūros, statybos inžinerijos, metalurgijos, klasikinių ir naujų medžiagų apdorojimo, pramoninės gamybos, vandens, oro ir sausumos transporto, spaudos, garso, vaizdo, naftos ir dujų gavybos, medicininės diagnostikos, namų ūkio, energetikos, kosmoso, ginkluotės ir kitų technologijų raida, medžiagų ir technologijų kūrimo ir diegimo įtaka visuomenės vystymuisi. Pasaulinė technologinė raida iliustruojama ir Lietuvos pavyzdžiais. Vadovėlis skirtas universitetinių aukštųjų mokyklų technikos specialybių studentams, neuniversitetinių aukštųjų technikos mokyklų (kolegijų) dėstytojams ir studentams, taip pat visiems, kurie domisi medžiagų ir technologijų plėtra.