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Well-posedness, regularity, asymptotic behavior and analyticity for some plate-membrane type transmission problems

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A dissertation submitted to the Doctorado en Ciencias Naturales in partial fulfillment of the requirements for the degree of doctor in sciences

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To my wife and to my son

## Abstract

In this thesis some plate-membrane type transmission problems are studied. See system (1.1)-(1.9). Three dampings are considered on the structure: thermal and structural for the plate, and global viscoelastic of Kelvin-Voigt type on the membrane. Sometimes some damping is removed from the structure. The plate may or may not have an inertial term. In the presence and/or absence of any of the elements mentioned above, we establish existence and uniqueness of solution of the system, which depends continuously on the initial data. We also obtain results of regularity, stability and analyticity. We use the semigroup approach to show the well-posedness our system. Following an idea of proof of regularity developed by Avalos and Lasiecka, we prove that if the inertial term is present or absent then the boundary and transmission conditions hold in the strong sense of the trace when the initial data are smooth enough. Then, using a general criteria of Arendt-Batty, we show the strong stability of our system when the membrane is damped and the plate is with or without rotational inertia. Employing a spectral approach, we indirectly prove exponential stability when the plate has rotational inertia and the structure is totally damped. This asymptotic behavior of the solutions is lost when we remove the viscoelastic component of the membrane. Under this situation, we impose a geometrical condition on the membrane boundary and obtain that the solutions decay polynomially with a rate of $t^{-1 / 25}$ when the plate has rotational inertia and structural damping. Finally, using a wellknown Liu-Zheng criterion we prove by contradiction the analyticity of the system when the membrane has Kelvin-Voigt damping and the thermoelastic plate is considered without inertial term and without structural damping.

Keywords: Analyticity, exponential stability, lack of exponential stability, polynomial stability, regularity, strong stability, transmission problem, wellposedness.

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## Contents

Abstract ..... vii
Acknowledgments ..... ix
1 Introduction ..... 1
1.1 Description of the problem ..... 1
1.2 Physical aspects of thermoelastic plate-membrane model ..... 3
1.3 Literature ..... 9
1.4 Document structure ..... 15
1.5 Generalities ..... 17
2 Preliminaries ..... 19
2.1 Basic spaces ..... 19
2.2 Regular boundaries ..... 23
2.3 Some useful inequalities ..... 28
2.4 Semigroups and groups of bounded linear operators ..... 30
2.5 Interpolation-extrapolation scales ..... 37
2.6 Powers of positive self-adjoint operators ..... 39
2.7 Elliptic and boundary operators ..... 41
3 Well-posedness and regularity of the solutions of a plate- membrane system ..... 45
3.1 Existence and uniqueness of the solutions ..... 45
3.1.1 Basic operators ..... 47
3.1.2 Abstract formulation of the problem ..... 52
3.1.3 The semigroup approach ..... 57
3.2 Regularity ..... 624 Asymptotic behavior of the solutions of some plate-membraneproblems65
4.1 Strong stability ..... 65
4.2 Exponential stability ..... 75
4.3 Lack of exponential stability ..... 77
4.4 Polynomial stability ..... 83
5 Analyticity of the semigroup associated with a transmission problem ..... 93
Bibliography ..... 103
List of symbols ..... 117
Index ..... 119

## Chapter 1

## Introduction

In recent years, many researchers have investigated transmission problems due to their applications in engineering, such as: aircrafts, satellite antennas and road traffic (see [63] and the references therein). The physical phenomenon consisting of the interaction between an acoustic medium and the vibrations of an elastic structure can be described with transmission equations (see [64]). There are also other interactive physical processes modeled through coupled partial differential equations, for example, composite laminates in smart materials and structures, and fluid-structure interactions (see [126]).

In this work, a plate-membrane type system within the $L^{2}$ theory is studied. In addition to proving the existence and uniqueness of the solution to the transmission problem, in certain appropriate spaces, its asymptotic behavior will be analyzed and subsequently the analyticity of a semigroup associated with the problem will be proved. The results depend on the dampings that are placed on the structure. The plate may or may not have thermal or structural damping, and the membrane may or may not have Kelvin-Voigt damping. The geometric description of the structure to be worked on is shown below.

### 1.1 Description of the problem

Let $\Omega$ and $\Omega_{2}$ be bounded domains in $\mathbb{R}^{2}$ with boundaries $\Gamma:=\partial \Omega$ and $I:=\partial \Omega_{2}$, both of class $C^{4}$, such that $\overline{\Omega_{2}} \subset \Omega$. We set $\Omega_{1}:=\Omega \backslash \overline{\Omega_{2}}$. The unit normal vector pointing towards the exterior of $\Omega_{1}$, both on $\Gamma$ and on $I$, is denoted by $\nu$ (see Fig. 1.1). Note that unit outward normal vector to $\Omega_{2}$ along the interface $I$ is $-\nu$. We will assume that $\Omega_{1}$ and $\Omega_{2}$ are occupied by the


Fig. 1.1. The set $\Omega=\Omega_{1} \cup I \cup \Omega_{2}$.
middle surface of the thermoelastic plate and by the membrane, respectively. In this case, $u=u(t, x)$ and $v=v(t, x)$ denote the vertical deflections of the plate and the membrane, respectively. The temperature difference of the plate is denoted by $\theta=\theta(t, x)$. We consider the following problem

$$
\begin{array}{rll}
\rho_{1} u_{t t}-\gamma \Delta u_{t t}+\beta_{1} \Delta^{2} u-m_{1} \Delta u_{t}+\alpha \Delta \theta=0 & \text { in } & \mathbb{R}^{+} \times \Omega_{1}, \\
\rho_{0} \theta_{t}+\sigma \theta-\beta \Delta \theta-\alpha \Delta u_{t}=0 & \text { in } & \mathbb{R}^{+} \times \Omega_{1}, \\
\rho_{2} v_{t t}-\beta_{2} \Delta v-m_{2} \Delta v_{t}=0 & \text { in } & \mathbb{R}^{+} \times \Omega_{2} \tag{1.3}
\end{array}
$$

The parameters $\gamma, m_{1}, \sigma$ and $m_{2}$ are all non-negative, and the rest of the constants $\rho_{1}, \beta_{1}, \alpha, \rho_{0}, \beta, \rho_{2}$ and $\beta_{2}$ are positive. When $\gamma>0$ it is because the rotational inertia of the plate filaments is considered and whose value is proportional to the square of the plate thickness, in this case (1.1) with $m_{1}=\alpha=0$ is known as the Kirchhoff plate equation (see (1) in [81, p. $2]$ ). The case $\gamma=0$ corresponds to a thin plate, and when $m_{1}=\alpha=0$ the equation (1.1) is called Euler-Bernoulli plate equation. The parameters $m_{1}>0$ and $m_{2}>0$ make the plate and the membrane have structural and Kelvin-Voigt damping, respectively. The constant $\sigma$ depends on the thickness of the plate and the ratio of the external thermal conductivity to the thermal conductivity of the plate, in the case of being positive (see [79]). The coupling parameter $\alpha$ is called coefficient of thermal expansion (see [9]), $\beta$ is the thermal conductivity (see [39]) and $\rho_{0}$ has the meaning of heat/thermal capacity (see [35, p. 244]). The constants $\beta_{1}, \beta_{2}, \rho_{1}$ and $\rho_{2}$ are the plate flexural rigidity, the in-plane stress, the plate mass/area and the membrane mass/area, respectively (see [129]). We will assume that the plate is embedded and attached to the membrane, this is interpreted by

$$
\begin{equation*}
u=\partial_{\nu} u=0 \text { on } \mathbb{R}^{+} \times \Gamma \quad \text { and } \quad u=v \text { on } \mathbb{R}^{+} \times I \tag{1.4}
\end{equation*}
$$

We will assume that the temperature satisfies Newton's cooling law (see (2.5.7) in [95, p. 43]) with coefficient $\kappa>0$ along $\partial \Omega_{1}$, this is,

$$
\begin{equation*}
\partial_{\nu} \theta+\kappa \theta=0 \text { on } \mathbb{R}^{+} \times \partial \Omega_{1} . \tag{1.5}
\end{equation*}
$$

Besides the condition on the interface $I$ in (1.4), we consider the following boundary and transmission conditions

$$
\begin{align*}
\beta_{1} \mathscr{B}_{1} u+\alpha \theta & =0
\end{aligned} \begin{aligned}
& \text { on } \mathbb{R}^{+} \times I,  \tag{1.6}\\
\gamma \partial_{\nu} u_{t t}-\beta_{1} \mathscr{B}_{2} u+m_{1} \partial_{\nu} u_{t}-\alpha \partial_{\nu} \theta-\beta_{2} \partial_{\nu} v-m_{2} \partial_{\nu} v_{t} & =0 \tag{1.7}
\end{align*} \text { on } \mathbb{R}^{+} \times I, ~ \$
$$

where

$$
\mathscr{B}_{1} u:=\Delta u+(1-\mu) B_{1} u \quad \text { and } \quad \mathscr{B}_{2} u:=\partial_{\nu} \Delta u+(1-\mu) \partial_{\tau} B_{2} u
$$

are the boundary operators introduced in [79, p. 119], $\tau:=\left(-\nu_{2}, \nu_{1}\right)^{\top}$ is the unit tangent vector along $\partial \Omega_{1}$ with $\nu_{1}$ and $\nu_{2}$ being the first and second components of $\nu$, and $\mu$ is the Poisson ratio whose value belongs to the interval $(0,1 / 2)$, which comes from certain physical considerations on the plate. The operators $B_{1}$ and $B_{2}$ are defined by the relations

$$
\begin{aligned}
& B_{1} u:=2 \nu_{1} \nu_{2} u_{x_{1} x_{2}}-\nu_{1}^{2} u_{x_{2} x_{2}}-\nu_{2}^{2} u_{x_{1} x_{1}}, \\
& B_{2} u:=\nu_{1} \nu_{2}\left(u_{x_{2} x_{2}}-u_{x_{1} x_{1}}\right)+\left(\nu_{1}^{2}-\nu_{2}^{2}\right) u_{x_{1} x_{2}}
\end{aligned}
$$

which appear in [79, p. 2]. The initial conditions for the system are given by

$$
\begin{align*}
u(0)=u_{0}, u_{t}(0)=u_{1}, \theta(0)=\theta_{0} & \text { in } \quad \Omega_{1},  \tag{1.8}\\
v(0)=v_{0}, v_{t}(0)=v_{1} & \text { in } \quad \Omega_{2} . \tag{1.9}
\end{align*}
$$

In this thesis it is guaranteed that the system (1.1)-(1.9) has a unique solution that can be weak or strong, this will depend on the choice of the initial data. Moreover, here there are results concerning regularity, strong stability, exponential stability, non-exponential stability, polynomial stability and analyticity, which depend on the presence or absence of the parameters $\gamma, m_{1}$ and $m_{2}$.

### 1.2 Physical aspects of thermoelastic platemembrane model

Some paragraphs should be written regarding the derivation of the equations that govern the model.

Let us consider a thin plate which in equilibrium occupies the set $\overline{\Omega_{1}} \times$ [ $-h / 2, h / 2]$, where $h>0$ is the plate's thickness. Following the ideas given in [79] and [81], let

$$
\mathbf{U}\left(x_{1}, x_{2}, x_{3}, t\right):=\left(U_{1}\left(x_{1}, x_{2}, x_{3}, t\right), U_{2}\left(x_{1}, x_{2}, x_{3}, t\right), U_{3}\left(x_{1}, x_{2}, x_{3}, t\right)\right)
$$

be the displacement vector at time $t$ of the particle which when the plate is in equilibrium occupies position $\left(x_{1}, x_{2}, x_{3}\right)$. Similarly, let $\mathbf{u}\left(x_{1}, x_{2}, t\right):=$ $\left(u_{1}\left(x_{1}, x_{2}, t\right), u_{2}\left(x_{1}, x_{2}, t\right), u_{3}\left(x_{1}, x_{2}, t\right)\right)$ be the displacement vector at time $t$ of the particle which when the plate is in equilibrium occupies position $\left(x_{1}, x_{2}, 0\right)$. Both $\mathbf{U}$ and $\mathbf{u}$ are supposed smooth enough for the present discussion. The strain energy in the plate is given by

$$
\begin{equation*}
\mathscr{P}^{\mathrm{P}}:=\frac{1}{2} \int_{-h / 2}^{h / 2} \int_{\Omega_{1}} \sum_{i, j=1}^{3} \varepsilon_{i j} \sigma_{i j} d x^{\prime} d x_{3}, \tag{1.10}
\end{equation*}
$$

where $\varepsilon_{i j}$ and $\sigma_{i j}$ are, respectively, the components of the strain and stress tensors and $x^{\prime}:=\left(x_{1}, x_{2}\right)$. The components $\varepsilon_{i j}$ depend on $\mathbf{U}$ and hence on $\left(x_{1}, x_{2}, x_{3}, t\right)$. In the linear theory, the components of the stress tensor are linear combinations of the components of the strain tensor and the thermal strain (when the plate is subject to a temperature distribution). On the other hand, the kinetic energy of the plate is given by

$$
\begin{equation*}
\mathscr{K}^{\mathrm{P}}:=\frac{1}{2} \int_{-h / 2}^{h / 2} \int_{\Omega_{1}} \rho_{1} \sum_{i=1}^{3}\left(\partial_{t} U_{i}\right)^{2} d x^{\prime} d x_{3} \tag{1.11}
\end{equation*}
$$

where $\rho_{1}$ is the density (mass per unit volume) of the plate.
If the plate is subject to a temperature distribution $\tau\left(x_{1}, x_{2}, x_{3}, t\right)$, measured with respect to a reference temperature distribution $\tau_{0}$, at which the thermal stresses and strains in the plate are zero, and assuming that the plate is homogeneous, elastically and thermally isotropic (see [79]), we have that the stress-strain relations are given by

$$
\begin{equation*}
\sigma_{i j}=\frac{E}{1+\mu}\left(\varepsilon_{i j}+\frac{\mu}{1-2 \mu} \delta_{i j} \sum_{k=1}^{3} \varepsilon_{k k}\right)-\frac{E}{1-2 \mu} \varepsilon_{\tau} \delta_{i j}, \quad i, j=1,2,3 \tag{1.12}
\end{equation*}
$$

Here, the constants $E$ and $\mu$ are, respectively, the Young's modulus and the Poisson's ratio $(0<\mu<1 / 2)$, and $\delta_{i j}$ is the well-known Kronecker delta
notation. The symbol $\varepsilon_{\tau}$ denotes the thermal strain which satisfies $\varepsilon_{\tau}=0$ when $\tau=0$.

For the thin plates theory it is assumed that $\sigma_{33}=0$. This implies in (1.12) that

$$
0=\frac{E}{1+\mu}\left(\varepsilon_{33}+\frac{\mu}{1-2 \mu} \sum_{k=1}^{3} \varepsilon_{k k}\right)-\frac{E}{1-2 \mu} \varepsilon_{\tau}
$$

and therefore

$$
\varepsilon_{33}=-\frac{\mu}{1-\mu}\left(\varepsilon_{11}+\varepsilon_{22}\right)+\frac{1+\mu}{1-\mu} \varepsilon_{\tau} .
$$

Then, we have

$$
\begin{align*}
\sigma_{11} & =\frac{E}{1-\mu^{2}}\left[\varepsilon_{11}+\mu \varepsilon_{22}-(1+\mu) \varepsilon_{\tau}\right]  \tag{1.13}\\
\sigma_{22} & =\frac{E}{1-\mu^{2}}\left[\mu \varepsilon_{11}+\varepsilon_{22}-(1+\mu) \varepsilon_{\tau}\right] \tag{1.14}
\end{align*}
$$

and

$$
\begin{equation*}
\sigma_{i j}=\frac{E}{1+\mu} \varepsilon_{i j} \quad(i \neq j, i, j=1,2,3) . \tag{1.15}
\end{equation*}
$$

In the Kirchhoff model for a thin plate it is assumed that the components of the strain tensor are given by

$$
\varepsilon_{i j}=\frac{1}{2}\left(\frac{\partial U_{i}}{\partial x_{j}}+\frac{\partial U_{j}}{\partial x_{i}}\right) \quad(i, j=1,2,3) .
$$

In addition transverse shear effects are neglected. Thus, $\sigma_{i 3}=\sigma_{3 i}=0$ and $\varepsilon_{i 3}=\varepsilon_{3 i}=0$ for $i=1,2$. It follows that

$$
0=\varepsilon_{13}=\varepsilon_{31}=\frac{1}{2}\left(\frac{\partial U_{1}}{\partial x_{3}}+\frac{\partial U_{3}}{\partial x_{1}}\right)
$$

and therefore

$$
\begin{equation*}
\frac{\partial U_{1}}{\partial x_{3}}=-\frac{\partial U_{3}}{\partial x_{1}} . \tag{1.16}
\end{equation*}
$$

Now,

$$
\begin{align*}
U_{3}\left(x_{1}, x_{2}, z, t\right) & =U_{3}\left(x_{1}, x_{2}, 0, t\right)+\int_{0}^{1} \frac{\partial}{\partial s} U_{3}\left(x_{1}, x_{2}, s z, t\right) d s \\
& =u_{3}\left(x_{1}, x_{2}, t\right)+z \int_{0}^{1}\left(\partial_{3} U_{3}\right)\left(x_{1}, x_{2}, s z, t\right) d s \tag{1.17}
\end{align*}
$$

From (1.16) and (1.17), we have

$$
\begin{aligned}
& U_{1}\left(x_{1}, x_{2}, x_{3}, t\right)=U_{1}\left(x_{1}, x_{2}, 0, t\right)-\int_{0}^{x_{3}} \partial_{1} U_{3}\left(x_{1}, x_{2}, z, t\right) d z \\
& =u_{1}\left(x_{1}, x_{2}, t\right)-x_{3} \frac{\partial u_{3}}{\partial x_{1}}\left(x_{1}, x_{2}, t\right)-\int_{0}^{x_{3}} z \int_{0}^{1}\left(\partial_{1} \partial_{3} U_{3}\right)\left(x_{1}, x_{2}, s z, t\right) d s d z
\end{aligned}
$$

After linearization (with respect to $x_{3}$ ), we obtain

$$
\begin{equation*}
U_{1}\left(x_{1}, x_{2}, x_{3}, t\right)=u_{1}\left(x_{1}, x_{2}, t\right)-x_{3} \frac{\partial u_{3}}{\partial x_{1}}\left(x_{1}, x_{2}, t\right) \tag{1.18}
\end{equation*}
$$

In the same way it holds

$$
\begin{equation*}
U_{2}\left(x_{1}, x_{2}, x_{3}, t\right)=u_{2}\left(x_{1}, x_{2}, t\right)-x_{3} \frac{\partial u_{3}}{\partial x_{2}}\left(x_{1}, x_{2}, t\right) \tag{1.19}
\end{equation*}
$$

and

$$
\begin{equation*}
U_{3}\left(x_{1}, x_{2}, x_{3}, t\right)=u_{3}\left(x_{1}, x_{2}, t\right) \tag{1.20}
\end{equation*}
$$

Now we can write the components $\varepsilon_{i j}, i, j=1,2$, in terms of the $u_{i}, i=1,2,3$. In fact:

$$
\begin{gather*}
\varepsilon_{12}=\varepsilon_{21}=\frac{1}{2}\left(\frac{\partial u_{1}}{\partial x_{2}}+\frac{\partial u_{2}}{\partial x_{1}}-2 x_{3} \frac{\partial^{2} u_{3}}{\partial x_{1} \partial x_{2}}\right)  \tag{1.21}\\
\varepsilon_{11}=\frac{\partial u_{1}}{\partial x_{1}}-x_{3} \frac{\partial^{2} u_{3}}{\partial x_{1}^{2}} \tag{1.22}
\end{gather*}
$$

and

$$
\begin{equation*}
\varepsilon_{22}=\frac{\partial u_{2}}{\partial x_{2}}-x_{3} \frac{\partial^{2} u_{3}}{\partial x_{2}^{2}} \tag{1.23}
\end{equation*}
$$

From (1.13)-(1.15) and (1.21)-(1.23), it follows that

$$
\begin{align*}
& \sigma_{11}=\frac{E}{1-\mu^{2}}\left[\frac{\partial u_{1}}{\partial x_{1}}-x_{3} \frac{\partial^{2} u_{3}}{\partial x_{1}^{2}}+\mu \frac{\partial u_{2}}{\partial x_{2}}-\mu x_{3} \frac{\partial^{2} u_{3}}{\partial x_{2}^{2}}-(1+\mu) \varepsilon_{\tau}\right],  \tag{1.24}\\
& \sigma_{22}=\frac{E}{1-\mu^{2}}\left[\mu \frac{\partial u_{1}}{\partial x_{1}}-\mu x_{3} \frac{\partial^{2} u_{3}}{\partial x_{1}^{2}}+\frac{\partial u_{2}}{\partial x_{2}}-x_{3} \frac{\partial^{2} u_{3}}{\partial x_{2}^{2}}-(1+\mu) \varepsilon_{\tau}\right], \tag{1.25}
\end{align*}
$$

and

$$
\begin{equation*}
\sigma_{12}=\sigma_{21}=\frac{E}{2(1+\mu)}\left(\frac{\partial u_{1}}{\partial x_{2}}+\frac{\partial u_{2}}{\partial x_{1}}-2 x_{3} \frac{\partial^{2} u_{3}}{\partial x_{1} \partial x_{2}}\right) . \tag{1.26}
\end{equation*}
$$

We recall (1.10) for the strain energy in the plate. For the case of a thin plate in the Kirchhoff model, it reduces to

$$
\mathscr{P}^{\mathrm{P}}=\frac{1}{2} \int_{-h / 2}^{h / 2} \int_{\Omega_{1}}\left(\varepsilon_{11} \sigma_{11}+2 \varepsilon_{12} \sigma_{12}+\varepsilon_{22} \sigma_{22}\right) d x^{\prime} d x_{3} .
$$

Using (1.21)-(1.26), after integration with respect to $x_{3}$, it can be found that $\mathscr{P}^{\mathrm{P}}$ splits into two terms $\mathscr{P}_{b}^{\mathrm{P}}$ and $\mathscr{P}_{s}^{\mathrm{P}}$, which uncouple the components $u_{1}$, $u_{2}$, representing in-plane stretching, and the component $u_{3}$ related to bending (see [79]). The term $\mathscr{P}_{s}^{\mathrm{P}}$ depends on $u_{1}, u_{2}$ and $\varepsilon_{\tau}$, and it is the part of the strain energy due to the in-plane stretching. On the other side, $\mathscr{P}_{b}^{\mathrm{P}}$ depend only on $u_{3}$ and $\varepsilon_{\tau}$, and it is the part of the strain energy due to bending. In this work we consider only the equations of motion for $u_{3}$. Exactly $\mathscr{P}_{b}^{\mathrm{P}}$ is given by

$$
\begin{aligned}
\mathscr{P}_{b}^{\mathrm{P}}:=\frac{E h^{3}}{24\left(1-\mu^{2}\right)} \int_{\Omega_{1}} & {\left[\left(\frac{\partial^{2} u_{3}}{\partial x_{1}^{2}}\right)^{2}+\left(\frac{\partial^{2} u_{3}}{\partial x_{2}^{2}}\right)^{2}+2(1-\mu)\left(\frac{\partial^{2} u_{3}}{\partial x_{1} \partial x_{2}}\right)^{2}\right.} \\
& \left.+2 \mu \frac{\partial^{2} u_{3}}{\partial x_{1}^{2}} \frac{\partial^{2} u_{3}}{\partial x_{2}^{2}}+\frac{12(1+\mu)}{h^{3}}\left(\Delta u_{3}\right) \int_{-h / 2}^{h / 2} x_{3} \varepsilon_{\tau} d x_{3}\right] d x^{\prime}
\end{aligned}
$$

where $\Delta$ denotes the Laplacian with respect to the two variables $x_{1}$ and $x_{2}$. Using the expression for the modulus of flexural rigidity of the plate $D:=E h^{3} / 12\left(1-\mu^{2}\right)$, setting $u:=u_{3}$ and changing the notations for the derivatives to $\partial_{i}:=\partial / \partial x_{i}$, we can write

$$
\begin{aligned}
\mathscr{P}_{b}^{P}=\frac{D}{2} \int_{\Omega_{1}}\left[\left(\partial_{1}^{2} u\right)^{2}+\left(\partial_{2}^{2} u\right)^{2}\right. & +2(1-\mu)\left(\partial_{1} \partial_{2} u\right)^{2} \\
& \left.+2 \mu\left(\partial_{1}^{2} u\right)\left(\partial_{2}^{2} u\right)+(1+\mu) \theta \Delta u\right] d x^{\prime}
\end{aligned}
$$

where

$$
\theta:=\theta\left(x_{1}, x_{2}, t\right):=\frac{12}{h^{3}} \int_{-h / 2}^{h / 2} x_{3} \varepsilon_{\tau} d x_{3} .
$$

Now, substitution of (1.18), (1.19) and (1.20) in (1.11) and integration with respect to $x_{3}$ gives

$$
\mathscr{K}^{\mathrm{P}}=\frac{h}{2} \int_{\Omega_{1}} \rho_{1}\left[\left(\partial_{t} u_{1}\right)^{2}+\left(\partial_{t} u_{2}\right)^{2}+\left(\partial_{t} u_{3}\right)^{2}+\frac{h^{2}}{12}\left|\nabla\left(\partial_{t} u_{3}\right)\right|^{2}\right] d x^{\prime}
$$

where $\nabla$ is the gradient with respect to the variables $x_{1}$ and $x_{2}$. We see that the components $u_{1}, u_{2}$ and $u_{3}$ are uncoupled in this expression for $\mathscr{K}^{\mathrm{P}}$. Again, we are only interested in the part of the kinetic energy due to bending, which is denoted by $\mathscr{K}_{b}^{\mathrm{P}}$ and it is given by

$$
\mathscr{K}_{b}^{\mathrm{P}}:=\frac{h}{2} \int_{\Omega_{1}} \rho_{1}\left[\left(\partial_{t} u\right)^{2}+\frac{h^{2}}{12}\left|\nabla\left(\partial_{t} u\right)\right|^{2}\right] d x^{\prime}
$$

where $u:=u_{3}$ again.
Suppose now that $\Omega_{2}$ is occupied by an elastic membrane for which its vertical deflection is denoted by $v=v\left(x_{1}, x_{2}\right)$. The strain energy of the membrane is given by

$$
\mathscr{P}^{\mathrm{M}}:=\frac{C}{2} \int_{\Omega_{2}}|\nabla v|^{2} d x^{\prime}
$$

where $C$ depends on the material and the initial tension in the membrane. On the other side, the kinetic energy for the membrane is given by

$$
\mathscr{K}^{\mathrm{M}}:=\frac{1}{2} \int_{\Omega_{2}} \rho_{2}\left(\partial_{t} v\right)^{2} d x^{\prime},
$$

where $\rho_{2}$ is the surface density of the membrane.
We assume that the structure is clamped on $\Gamma$, that is $u=\partial_{\nu} u=0$ on $\Gamma \times[0, T)$ for $0<T \leq \infty$, where $\nu$ is the outward unit normal vector to the boundary of $\Omega_{1}$. Furthermore, we assume that $u=v$ on $I \times[0, T)$. The meaning of this condition is that the structure does not break. We recall that small deflections are being considered.

In order to obtain an initial boundary value problem which models the coupled structure plate-membrane we have to set to zero the first variation in the time interval $[0, T)$ of the Lagrangian

$$
\begin{equation*}
L_{b}(u, v):=\int_{0}^{T}\left[\mathscr{K}_{b}^{\mathrm{P}}(t)+\mathscr{K}^{\mathrm{M}}(t)+\mathscr{W}_{b}(t)-\mathscr{P}_{b}^{\mathrm{P}}(t)-\mathscr{P}^{\mathrm{M}}(t)\right] d t \tag{1.27}
\end{equation*}
$$

with respect to $(u, v)$ in the space of kinematically admissible displacements, that is

$$
\begin{equation*}
\delta L_{b}(u, v):=\left.\frac{\partial}{\partial \lambda} L_{b}(u+\lambda \widetilde{u}, v+\lambda \widetilde{v})\right|_{\lambda=0}=0 \tag{1.28}
\end{equation*}
$$

with $\left.\widetilde{u}\right|_{\Gamma}=\left.\partial_{\nu} \widetilde{u}\right|_{\Gamma}=0$ and $\left.\widetilde{u}\right|_{I}=\left.\widetilde{v}\right|_{I}$. Here $\widetilde{u}$ and $\widetilde{v}$ are chosen, such that $\widetilde{u}(0)=\widetilde{u}^{\prime}(0)=\widetilde{u}(T)=\widetilde{u}^{\prime}(T)=0$ in $\Omega_{1}$ and $\widetilde{v}(0)=\widetilde{v}^{\prime}(0)=\widetilde{v}(T)=\widetilde{v}^{\prime}(T)=0$ in $\Omega_{2}$, where for a function $\varphi: \overline{\Omega_{i}} \times[0, \infty) \rightarrow \mathbb{R}$ we write $\varphi(t)\left(x_{1}, x_{2}\right):=$ $\varphi\left(x_{1}, x_{2}, t\right)$ and ' denotes the derivative with respect to $t$. This approach is equivalent to the use of the Virtual Work Principle from the Classical Mechanic.

In (1.27) $\mathscr{W}_{b}$ is the part of the work done on the structure plate-membrane that contributes to bending due to external forces acting on it.

After renaming constants, equation (1.28) leads to the general equations (1.1)- $(1.3)$ of the model, as well as the boundary and transmission conditions (1.4)-(1.7). Changing the space of kinematically admissible displacements, several models with different boundary and transmission conditions can be obtained. Finally, the damping which can be considered in each model can be incorporated in $\mathscr{W}_{b}$.

### 1.3 Literature

Transmission problems appear very frequently in various fields of physics and technics, e.g., there are applications in the electrodynamics of solid media when working with electromagnetic processes in ferromagnetic media with different dielectric constants. This type of problem also occurs in the solid mechanics of composite materials. The previous exposition can be found in [26]. Some authors also associate these problems with structures consisting of a finite number of interconnected flexible elements such as waves, beams, plates, casings and combinations thereof, which are representative of frames, robot arms, solar panels, antennas and deformable mirrors (see [80]). There are also transmission problems associated with acoustics (see [7, 75]), the automotive industry (see [13, 114]), polymers (see [24, 53, [54, 76, 82]), fluid behaviour in certain structures (see [34, 36, 115]), vibration suppression (see [42]), electrostatics and static magnets (see [88]), and material composition (see [101]).

The transmission problems were initiated by Meuro Picone in 1954 (see [61]), which are framed within the theory of partial differential equations and control theory, (see [35, 87]). Many researchers are interested in the asymptotic behavior of the solutions of the transmission problem in which
they are investigating. In the literature we find this kind of problems with different types of decay, such as: exponential decay (see [14, 15, 22, 30, 31, [37, 42, 52, 63, 64, 65, 90, 102, 122]), polynomial decay (see [4, 14, 15, 19, [46, 52, [63, [64, [83, 90, 105, 122, [125, 127]) and logarithmic decay (see [46, 50, 68, 71, 90]). Of the works mentioned, only in [14, 15, 52, 64, 122, in addition to analyzing the exponential stability and the polynomial stability of their respective systems, the authors studied the absence of exponential stability. All this asymptotic analysis is done in this document for the system (1.1)-(1.9). In [68] and [71], the authors study a transmission problem where the membrane surrounds the plate that has localized Kelvin-Voigt damping. Instead, we consider the plate around the membrane and it is the latter that has the Kelvin-Voigt damping, but uniform.

The Kelvin-Voigt damping is a type of linear damping that suppresses the vibrations of elastic structures and is caused by the internal friction of the material of the vibrating structures and is thus called internal damping. In this case the material is viscoelastic because it has properties of viscosity and elasticity. The term "viscous" implies that it slowly deforms when exposed to an external force. The term "elastic" implies that once a deforming force has been removed, the material will return to its original configuration. For further explanations, see [69, p. 12]. Mathematically, the Kelvin-Voigt damping is an operator of the same order as the principal operator of the equation that describes the vibrating structure (see [71, p. 2242]). The model for a membrane occupying a bounded region $\widetilde{\Omega}$ of $\mathbb{R}^{2}$ is given by

$$
w_{t t}-\Delta w+\widetilde{\beta} L_{w}=0 \quad \text { in } \quad \widetilde{\Omega} \times \mathbb{R}^{+}
$$

where $\widetilde{\beta} \geq 0$ and $L_{w}$ is a dissipative mechanism. In [110, Subsection 2.4.4], the authors determine the equation of motion of the vertically moving particles of the vibrating membrane when $\widetilde{\beta}=0$. If the membrane has the frictional damping ( $\widetilde{\beta}>0$ and $L_{w}=w_{t}$ ), then the semigroup decays exponentially (see [111, p. 16]). In [73, the author proved that if the membrane has a global Kelvin-Voigt damping (i.e., $\widetilde{\beta}>0$ and $L_{w}=-\Delta w_{t}$ ) then the corresponding semigroup is not only exponentially stable, but also is analytic. For this reason, Kelvin-Voigt damping is said to be stronger than frictional damping.

Another type of damping that is considered on some vibrating structures is thermoelastic damping, see [69, p. 13]. This damping originates from the coupling of the elasticity of the structure with a heat source. The thermal
effect contributes to the deformation of the structure since its material expands when the temperature increases and contracts when the temperature decreases. Our attention is directed towards thermoelastic plates.

The first models for thin plates were developed by the German physicist Gustav Robert Kirchhoff (1824-1887). In the first chapter of [81], one finds linear and non-linear models for thin plates subject to different physical situations such as temperature, elasticity and viscosity, depending on different assumptions and conditions according to each model. A detailed study of deflections and plate deformations is in [119, here the authors work with thin plates with small deflections, thin plates with large deflections and thick plates. For plate deformations, see also [96]. On the theory of vibrating plates, we recommend to the reader [99.

The literature on thermoplastic plates is very extensive. The following are some of the works on this subject. The classic linear model for thermoelastic plates is given by

$$
\begin{array}{rll}
\rho u_{t t}-\widetilde{\mu} \Delta u_{t t}+\widetilde{\gamma} \Delta^{2} u+\widetilde{\alpha} \Delta \theta=0 & \text { in } & \widetilde{\Omega} \times \mathbb{R}^{+}, \\
c \theta_{t}-\widetilde{\kappa} \Delta \theta-\widetilde{\alpha} \Delta u_{t}=0 & \text { in } & \widetilde{\Omega} \times \mathbb{R}^{+},
\end{array}
$$

where $\rho, \widetilde{\mu}, \widetilde{\gamma}, \widetilde{\alpha}, c$ and $\widetilde{\kappa}$ are all positive constants and $\widetilde{\Omega} \subset \mathbb{R}^{2}$ is the region occupied by the middle surface of the plate. For the physical model we refer to [79]. In [74], Kim studies the Euler-Bernoulli plate equation $(\widetilde{\mu}=0)$ with thermal effect and homogeneous Dirichlet boundary conditions $u=\partial_{\nu} u=$ $\theta=0$ on $\partial \widetilde{\Omega}$. He proved exponential decay of the energy of his system. In [93], Liu and Renardy worked on the Kim's problem and obtain a much stronger result, showing that the semigroup associated is of analytic type. In [43], the authors do not account for rotational forces of the thermoelastic plate $(\widetilde{\mu}=0)$ and consider homogeneous Dirichlet boundary conditions. They generalize the result of Liu and Renardy, introducing the space $W_{p, D}^{2}(\widetilde{\Omega}):=$ $\left\{u \in W^{2, p}(\widetilde{\Omega}): u=\partial_{\nu} u=0\right.$ on $\left.\partial \widetilde{\Omega}\right\}$ with $\widetilde{\Omega}$ being a bounded domain of $\mathbb{R}^{n}(n \geq 2)$ and showing that if the initial data $\left.\left(u, u_{t}, \theta\right)\right|_{t=0} \in W_{p, D}^{2}(\widetilde{\Omega}) \times$ $L^{p}(\widetilde{\Omega}) \times L^{p}(\widetilde{\Omega})$ then the semigroup is analytic for any $p \in(1, \infty)$ and also exponentially stable. In [86], the authors set $\widetilde{\mu}=0$ and, focusing on the case of free boundary conditions, proved that the associated semigroup is analytic. The addition of the term $\Delta u_{t t}$ not only makes the problem physically more meaningful, it also makes it more mathematically interesting, because we need
more regularity to establish the decay of the solutions. For the case $\widetilde{\mu}>0$ and Dirichlet boundary conditions it was proved in [103] that the solutions decay exponentially.

Going back to analytical results, it is known that the literature pertaining to analytic semigroups is very extensive and has a wide variety of topics. For example, one finds works on: materials with porosity (see [20]), thermoelastic plate with variable coefficients (see [33]), viscoelastic plate equation of Moore-Gibson-Thompson type (see [39]), wave equation with Kelvin-Voigt damping and hyperbolic dynamic boundary conditions (see [60]), system of couple plate equations with different dampings (see [66]), laminated beam (see [67]), thermoelastic plate equations with different boundary conditions (see [84]), thermoelastic plate with hinged mechanical B.C. and Neumann thermal B.C. (see [85]), homogeneous and isotropic prestressed type III thermoelastic thin plate (see [91]), qualitative properties for the semigroup generated by certain matrix operator (see 94]) and thermoelastic plate with dynamical boundary conditions (see [128]). Other work on analyticity can be seen in [106]. Here the authors study a system consisting of a thermoelastic Euler-Bernoulli plate coupled with a membrane that has a global Kelvin-Voigt type damping. They consider appropriate initial conditions and certain boundary conditions, and show that the semigroup associated with the problem is analytic. Regarding transmission problems, it seems that the only work so far with analyticity has been [51]. There, the system studied by the authors is constituted by a pair of thermoelastic plates, none of which have the inertial term and their structure is like the one in Fig. 1.2. In the present manuscript is the analyticity proof of a transmission problem of a thermoelastic thin plate with a membrane that has global viscoelastic damping of Kelvin-Voigt type, this is one of the main results of this work. According to our search, it seems that there is no such result in the entire literature and for this reason it was submitted for a publication according to reference [16].

Many researchers studied transmission problems involving systems of the type: plate-plate, wave-wave and plate-wave. Below we recall some works concerning the three structures mentioned.

In [42], the authors investigate a coupled system of linear plate equations, one non-damped plate and the other structurally-damped plate in two sufficiently smooth and bounded subdomains, and consider a geometric configuration as Fig. 1.1. They showed, independently of the size of the damped
part, that the damping is strong enough to produce exponential stability. In the literature, it is common to find the following configuration: Let $\Omega_{1}^{\prime}, \Omega_{2}^{\prime}$ and $\Omega^{\prime}$ be bounded domains in $\mathbb{R}^{2}$ such that $\Gamma_{0}:=\overline{\Omega_{1}^{\prime}} \cap \overline{\Omega_{2}^{\prime}}, \Gamma_{1}:=\partial \Omega_{1}^{\prime} \backslash \Gamma_{0}$ and $\Gamma_{2}:=\partial \Omega_{2}^{\prime} \backslash \Gamma_{0}$ are smooth curves. Set $\Omega^{\prime}=\Omega_{1}^{\prime} \cup \Omega_{2}^{\prime} \cup \Gamma_{0}$ and $\overline{\Gamma_{1}} \cap \overline{\Gamma_{2}}=\emptyset$ (see Fig. 1.2). In [102], the authors consider a problem of transmission, with


Fig. 1.2. The set $\Omega^{\prime}=\Omega_{1}^{\prime} \cup \Gamma_{0} \cup \Omega_{2}^{\prime}$.
the configuration of Fig. 1.2, of a plate $\Omega^{\prime}$ constituted by two parts $\Omega_{1}^{\prime}$ and $\Omega_{2}^{\prime}$, being $\Omega_{1}^{\prime}$ sensitive to temperature changes. They work on the following model

$$
\begin{array}{rll}
\widehat{\rho}_{1} u_{t t}-\gamma_{1} \Delta u_{t t}+\widehat{\beta}_{1} \Delta^{2} u+\widehat{\mu} \Delta \theta=0 & \text { in } & \Omega_{1}^{\prime} \times \mathbb{R}^{+} \\
\widehat{\rho}_{0} \theta_{t}-\beta_{0} \Delta \theta+\gamma_{0} \theta-\widehat{\mu} \Delta u_{t}=0 & \text { in } & \Omega_{1}^{\prime} \times \mathbb{R}^{+} \\
\widehat{\rho}_{2} v_{t t}-\gamma_{2} \Delta v_{t t}+\widehat{\beta}_{2} \Delta^{2} v=0 & \text { in } & \Omega_{2}^{\prime} \times \mathbb{R}^{+} .
\end{array}
$$

Under certain initial conditions and with several boundary and transmission conditions different from those of problem (1.1)-(1.9), they show that there is a solution and that the local thermal effect is strong enough to produce exponential decay. Afterwards, the author of [112] studies a transmission problem (Fig. 1.2) for a plate consisting of a thermoelastic part $\Omega_{1}^{\prime}$ and an isothermal part $\Omega_{2}^{\prime}$. He demonstrates the existence of a compact global attractor when the non-linearity is of the Berger or scalar type. The model worked was

$$
\begin{array}{rll}
\widehat{\rho}_{1} u_{t t}+\widehat{\beta}_{1} \Delta^{2} u+\widehat{\mu} \Delta \theta+F_{1}(u, v)=0 & \text { in } & \Omega_{1}^{\prime} \times \mathbb{R}^{+} \\
\widehat{\rho}_{0} \theta_{t}-\beta_{0} \Delta \theta-\widehat{\mu} \Delta u_{t}=0 & \text { in } & \Omega_{1}^{\prime} \times \mathbb{R}^{+} \\
\widehat{\rho}_{2} v_{t t}+\widehat{\beta}_{2} \Delta^{2} v+F_{2}(u, v)=0 & \text { in } & \Omega_{2}^{\prime} \times \mathbb{R}^{+}
\end{array}
$$

satisfying boundary conditions and initial conditions, where the functions $F_{1}: H^{2}\left(\Omega_{1}^{\prime}\right) \times H^{2}\left(\Omega_{2}^{\prime}\right) \rightarrow L^{2}\left(\Omega_{1}^{\prime}\right)$ and $F_{2}: H^{2}\left(\Omega_{1}^{\prime}\right) \times H^{2}\left(\Omega_{2}^{\prime}\right) \rightarrow L^{2}\left(\Omega_{2}^{\prime}\right)$
are both non-linear. The author studies three problems where $F_{1}$ and $F_{2}$ are taken differently. Another transmission problem with the configuration of Fig. 1.2 is treated in [121. The authors consider a system constituted by a thermoelastic plate with localized thermal dissipation of memory type coupled with an isothermal plate, both plates have inertial term. They prove that the solutions have exponential decay.

Much study has been made of the wave equation with different dampings. When a vibrating source disturbs the medium, a wave is formed. Several dampings can be added to a system of wave equations to control the vibrations. In [80], Chapter VI, the authors consider the movement of interconnected elastic membranes and determine the equations of this physical phenomenon. In [31], the authors consider the configuration of Fig. 1.1 and study the propagation of waves on a domain consisting of two different materials: one component is elastic where a frictional damping is acting while the other one possesses a viscoelastic component with a memory with past history. They establish exponential stability for the solutions of the problem. In 65], the authors analyze a transmission problem of the wave equation (Fig. 1.2) with linear dynamical feedback control. They prove that the energy of system exponentially decays. In [122], the authors investigate a locally coupled wave equations with only one internal viscoelastic damping of Kelvin-Voigt type. The damping and the coupling coefficients are non smooth. They show that the energy of smooth solutions of the system decays polynomially. In [18], the authors consider a wave-wave system, in one space dimension, with frictional damping. They study the wave propagation in a medium with a component with attrition and another simply elastic, and show that for this type of material, the dissipation produced by the frictional part is strong enough to produce exponential decay of the solution, no matter how small is its size.

There are several works about transmission problems that involve these plate-wave equations, here we mention some. In [70], Hernández works on a semi-linear problem of initial and boundary values, this models a thin elastic plate coupled with an elastic membrane, considering homogeneous boundary conditions. Using semigroup theory, the author obtained existence and uniqueness of weak solutions. In [68], Hassine studies a transmission problem of a membrane coupled with an Euller-Bernoulli plate which has a localized Kelvin-Voigt damping, the membrane surrounds the plate (see Fig. 1.1). He
proves that sufficiently smooth solutions decay logarithmically. Afterwards, in [71], the authors work in the Hassine system but adding the rotational term in the plate. They show that the energy of the transmission system is stable with logarithmic decay. In [63], the authors investigate a transmission wave-plate model with different localized frictional damping, in their structure the plate is next to the wave. They show respectively that the energy of the system decays polynomially under some geometric condition when the frictional damping only acts on the part of the plate, and the energy of the system is exponentially stable when the frictional damping acts only on the other part of the wave. These same authors, in [64], study another plate-wave system but now its structure is as in Fig. 1.2 and obtained the same results as in 63. We further mention two works about plate-membrane transmission problems with transmission conditions different from each other but with exactly the same configuration of Fig. 1.1, see [14] and [15]. The plate in [15] is isothermal and without rotational inertia, while in [14] the plate may or may not have temperature and rotational term. Both works have stability results, but in neither of them is analyticity treated, perhaps because the frictional damping that they place on the membrane is not strong enough to achieve this type of result in the semigroup associated with their problem.

In this document there is an extensive analysis of a transmission problem constituted by a thermoelastic plate-membrane structure with the configuration of Fig. 1.1. In the presence and/or absence of inertial term, structural damping for the plate and Kelvin-Voigt damping on the membrane, we show results of existence and uniqueness of solutions, regularity and asymptotic behavior. These results complement what is already in the literature. Additionally, we establish that the semigroup associated with problem (1.1)-(1.9) is analytic when $\gamma=m_{1}=0$ and $m_{2}>0$. For more details see the next section.

### 1.4 Document structure

The document contains five chapters. In the first of them, the transmission problem is described, some literature on the subject is presented and are mentioned the results obtained in the thesis.

In Chapter 2, the reader will find the fundamental function spaces for the development of this work. We consider Sobolev spaces with some of their
properties, among which we mention those regarding continuous immersion, integration formulas, and important inequalities that satisfy some of their elements. There is a brief mention of some semigroup results regarding generation, stability and analyticity. Subsequently, Stone's theorem for group generation is given. The theory of interpolation-extrapolation scales is presented concisely but providing references where the details of the mentioned results are carried out. The chapter continues with a section devoted to powers of positive self-adjoint operators and follows with some theory about regular elliptic problems to establish a characterization that satisfies the domains of powers of operators that are realizations of some elliptic operator. We end with a result taken from [15] containing the estimate (2.30), that comes from a boundary value problem, which is used in the last two chapters to obtain polynomial stability and analyticity.

Chapter 3 is dedicated to an entire analysis that leads to the existence, uniqueness and regularity of the solutions of the transmission problem. This is achieved by introducing operators defined in appropriate spaces that satisfy properties that allow establishing a Cauchy problem to which semigroup theory can be applied to prove existence and uniqueness of solutions. These results are in Section 3.1 and correspond to the well-posedness of the problem (1.1)-(1.9) when $\gamma>0$ and $m_{1}, m_{2} \geq 0$, and in the case $\gamma=m_{1}=0$ with $m_{2} \geq 0$ (see Theorem 3.15 and Remark 3.16). In the second section of the chapter regularity results appear that will be very useful to obtain asymptotic behavior of the solutions of some transmission problems. One of them can be seen in Remark 3.17, which contains the estimate (3.44) that we will use in the next two chapters in the proofs of strong stability, of polynomial stability and in the one of analyticity. The main result of Section 3.2 is Theorem 3.18, for its demonstration we follow the ideas of the proof of Theorem 1.2 in [10], there we get a shared regularity that $v$ and $v_{t}$ win when $m_{2}>0$ and we establish that for smooth data (1.6)-(1.7) are satisfied in the strong sense of the trace, which will depend on the consideration or not of the parameters $\gamma$, $m_{1}$ and $m_{2}$. Remark 3.19 indicates the regularity that the components of the solution have when the initial data are taken with less smoothness compared to those of Theorem 3.18.

In Chapter 4, the stability of several plate-membrane type transmission problems is analyzed, depending on the types of damping that the structure possesses. In Section 4.1, using a general criteria of Arendt and Batty (see

Theorem 2.39) we show that our problem possesses strong stability when the membrane has the Kelvin-Voigt damping ( $m_{2}>0$ ) and in the structure it is considered the Kirchhoff plate with or without structural damping ( $m_{1} \geq 0$ ), or the Euler-Bernoulli plate without structural damping ( $m_{1}=0$ ), see Proposition 4.2 and Proposition 4.5. The Remark 4.4 and Remark 4.7 highlight the role of the thermal effect on the plate in achieving this stability result. In Section 4.2, employing a well-known characterization (see Theorem 2.41) we establish by contradiction the exponential stability of the solutions of the plate-membrane system when the inertial term and the mechanical dampings are considered, that is, the constants $\gamma, m_{1}$ and $m_{2}$ are all positive, see Theorem 4.8. In Section 4.3, we prove that the system conformed by a undamped membrane coupled to a thermoelastic Kirchhoff plate with or without structural damping lacks exponential stability (see Theorem 4.15). To obtain the absence of this asymptotic behavior, we apply Theorem 4.11 and for this we make use of Stone's Theorem on unitary groups, the interpolation-extrapolation scales theory and compactness arguments. But in Section 4.4, for $\gamma, m_{1}>0$ and $m_{2}=0$ we prove polynomial stability when we consider the geometric condition (4.99). We base the proof on a multiplier method and then we apply a generalization due to Muñoz Rivera and Racke (see Theorem 2.44) of the well-known Borichev and Tomilov characterization on polynomial decay, see Theorem 2.4 in [25].

In Chapter 5, we proof by contradiction the existence of an analytic semigroup associated with a transmission problem of a plate without rotational term that has a thermal effect coupled to a membrane that has a global viscoelastic Kelvin-Voigt damping. In other words, the solutions of the system (1.1)-(1.9) are analytic functions when $\gamma=m_{1}=0$ and $m_{2}>0$. This can be proved thanks to the Liu and Zheng analyticity criterion, see Theorem 2.46. The main result of this chapter is Theorem 5.1 and as a consequence we have a regularizing effect of the solutions and a couple of corollaries (see Corollary 5.3 and Corollary 5.4). We end with Remark 5.5, which indicates the non-analyticity of a semigroup associated with our plate-membrane system.

### 1.5 Generalities

Throughout the thesis, $C$ represents a positive constant which is not necessarily the same every time it appears, it can change from one line to another
line. If $X$ and $Y$ are topological spaces, then the collection of all bounded linear operators from $X$ to $Y$ be denoted by $\mathcal{L}(X, Y)$. We shall write $\mathcal{L}(X)$ instead of $\mathcal{L}(X, X)$. In this work the term domain means open and connected subset of $\mathbb{R}^{n}$, which is different from the concept of the domain of a function.

Let $X$ and $Y$ be two locally convex spaces such that $X \subset Y$ so that the identity operator id : $X \rightarrow Y$ is continuous. Let $A: D(A) \subset Y \rightarrow Y$ be a linear operator. The $X$-realization of $A$, denoted by $A_{X}$, is the linear operator

$$
A_{X}: D\left(A_{X}\right) \subset X \rightarrow X \quad \text { given by } \quad A_{X} x:=A x
$$

where $D\left(A_{X}\right):=\{x \in X \cap D(A): A x \in X\}$. It is easy to see that if $A$ is closed then $A_{X}$ is closed. This definition is taken from [5, p. 7].

Let $X$ be a Banach space and $A: D(A) \subset X \rightarrow X$ be a linear operator. The smallest closed extension of $A$, if it exists, is called the closure of $A$ and is denoted by $\bar{A}$. Operators having a closure are called closable. This definition is taken from [48, Definition A.7].

## Chapter 2

## Preliminaries

In this second chapter, the reader will find notations, definitions and results that will be used later.

### 2.1 Basic spaces

This section has the purpose of presenting the Sobolev spaces in a structured way. In addition, we will establish important properties of these spaces. It is known that these elements are used in the formulation of problems of partial differential equations. Some concepts and properties given here can be found in [1, 55, 72, 123, 131].

Let $\mathbb{N}_{0}:=\mathbb{N} \cup\{0\}$ and $n \in \mathbb{N}$. A multi-index $\alpha$ is an $n$-tuple of non-negative integers, this is, $\alpha=\left(\alpha_{1}, \alpha_{2}, \ldots, \alpha_{n}\right) \in \mathbb{N}_{0}^{n}$. We denote $|\alpha|:=\sum_{j=1}^{n} \alpha_{j}$. As usual $\partial_{j}$ represents the partial derivative with respect to the $j$-th variable $x_{j}$. We set

$$
\partial^{\alpha}:=\prod_{j=1}^{n} \partial_{j}^{\alpha_{j}}:=\frac{\partial^{|\alpha|}}{\partial x_{1}^{\alpha_{1}} \cdots \partial x_{n}^{\alpha_{n}}} .
$$

If $\alpha=0_{\mathbb{N}_{0}^{n}}$, then $\partial^{\alpha}$ is the identity operator. Let $k \in \mathbb{N}_{0}$ and $\Omega \subset \mathbb{R}^{n}$ be an open set. $C^{k}(\Omega)$ denotes the vector space of all functions $f: \Omega \rightarrow \mathbb{C}$ such that $\partial^{\alpha} f: \Omega \rightarrow \mathbb{C}$ are continuous for every $\alpha \in \mathbb{N}_{0}^{n}$ with $|\alpha| \leq k$. Note that $C(\Omega):=C^{0}(\Omega)$ is constituted by continuous complex-valued functions on $\Omega$. We define $C^{k}(\bar{\Omega})$ as the space of all functions $f \in C^{k}(\Omega)$ such that $\partial^{\alpha} f$ extends continuously to the closure $\bar{\Omega}$ for $|\alpha| \leq k$. We fix $C^{\infty}(\Omega):=\bigcap_{k=0}^{\infty} C^{k}(\Omega)$. The
support of $f: \Omega \rightarrow \mathbb{C}$ is given by

$$
\operatorname{supp} f:=\overline{\{x \in \Omega: f(x) \neq 0\}} .
$$

The space formed by the functions of $C^{\infty}(\Omega)$ that have compact support in $\Omega$ is denoted by $C_{c}^{\infty}(\Omega)$ and its elements are called test functions. We say that a sequence $\left(\varphi_{j}\right)_{j \in \mathbb{N}}$ in $C_{c}^{\infty}(\Omega)$ converges to zero if the following two conditions are satisfied:
i) There is a compact subset $K$ of $\Omega$ such that $\operatorname{supp} \varphi_{j} \subset K$ for all $j \in \mathbb{N}$.
ii) $\partial^{\alpha} \varphi_{j} \rightarrow 0$ uniformly in $K$ for every $\alpha \in \mathbb{N}_{0}^{n}$, this is,

$$
\max _{x \in K}\left|\partial^{\alpha} \varphi_{j}(x)\right| \rightarrow 0 \text { as } j \rightarrow \infty \text { for all } \alpha \in \mathbb{N}_{0}^{n}
$$

The space $C_{c}^{\infty}(\Omega)$ equipped with this notion of convergence is symbolized by $\mathscr{D}(\Omega)$. The topological dual $\mathscr{D}^{\prime}(\Omega):=\mathcal{L}(\mathscr{D}(\Omega), \mathbb{C})$ of $\mathscr{D}(\Omega)$ is called the space of distributions on $\Omega$. The complex value of $T \in \mathscr{D}^{\prime}(\Omega)$ on $\varphi \in C_{c}^{\infty}(\Omega)$ is denoted by $\langle T, \varphi\rangle$. Note that the continuity of a linear map $T$ of $\mathscr{D}(\Omega)$ in $\mathbb{C}$ is equivalent to: For every sequence $\left(\varphi_{j}\right)_{j \in \mathbb{N}}$ convergent to zero in $\mathscr{D}(\Omega)$ we have that $\left\langle T, \varphi_{j}\right\rangle \rightarrow 0$ in $\mathbb{C}$. For $\alpha \in \mathbb{N}_{0}^{n}$ and $T \in \mathscr{D}^{\prime}(\Omega)$, we define the distributional derivative $\partial^{\alpha} T$ by

$$
\left\langle\partial^{\alpha} T, \varphi\right\rangle:=(-1)^{|\alpha|}\left\langle T, \partial^{\alpha} \varphi\right\rangle \text { for any } \varphi \in \mathscr{D}(\Omega)
$$

Suppose that $1 \leq p<\infty$. The space of all equivalence classes of Lebesgue measurable functions $u: \Omega \rightarrow \mathbb{C}$ such that $|u|^{p}$ is a integrable function on $\Omega$ will be denoted by $L^{p}(\Omega)$. This space endowed with the norm

$$
\|u\|_{L^{p}(\Omega)}:=\left(\int_{\Omega}|u(x)|^{p} d x\right)^{1 / p}
$$

is a Banach space. The space $L^{2}(\Omega)$ considered with the inner product $(u, v)_{L^{2}(\Omega)}:=\int_{\Omega} u(x) \overline{v(x)} d x$ is a Hilbert space.

The space $L_{l o c}^{1}(\Omega)$ consists of all functions $f: \Omega \rightarrow \mathbb{C}$ such that the integral $\int_{K}|f| d x$ is finite for each compact subset $K$ of $\Omega$. If $f \in L_{l o c}^{1}(\Omega)$, we say that $f$ is locally integrable on $\Omega$. Each $u \in L_{l o c}^{1}(\Omega)$ defines a distribution $T_{u} \in \mathscr{D}^{\prime}(\Omega)$ given by

$$
\left\langle T_{u}, \varphi\right\rangle:=\int_{\Omega} u(x) \varphi(x) d x, \quad \varphi \in \mathscr{D}(\Omega) .
$$

We shall write $u$ instead of $T_{u}$. Due to $L^{p}(\Omega) \subset L_{l o c}^{1}(\Omega)$, we can consider any function of $L^{p}(\Omega)$ as a distribution.

The Sobolev space of non-negative integer order $W^{k, p}(\Omega)$ is composed of all functions $u \in L^{p}(\Omega)$ whose distributional derivatives $\partial^{\alpha} u \in L^{p}(\Omega)$ for all $|\alpha| \leq k$. We introduce the following norm on $W^{k, p}(\Omega)$ :

$$
\|u\|_{W^{k, p}(\Omega)}:=\left(\sum_{|\alpha| \leq k}\left\|\partial^{\alpha} u\right\|_{L^{p}(\Omega)}^{p}\right)^{1 / p}
$$

In particular, $H^{k}(\Omega):=W^{k, 2}(\Omega)$ is a Hilbert space with the scalar product $(u, v)_{H^{k}(\Omega)}:=\sum_{|\alpha| \leq k}\left(\partial^{\alpha} u, \partial^{\alpha} v\right)_{L^{2}(\Omega)}$. Note that $H^{0}(\Omega)=L^{2}(\Omega)$. Let $k_{1}, k_{2} \in$ $\mathbb{N}$ with $k_{1}>k_{2}$, then we have the strict inclusions

$$
H^{k_{1}}(\Omega) \subset H^{k_{2}}(\Omega) \subset L^{2}(\Omega)
$$

see Remark 1.2 in [89, p. 3].
Next we define some non-negative real order Sobolev spaces. Let $s \geq 0$. We shall denote by $[s]$ the integer part of $s$ (i.e., $[s]:=\max \{z \in \mathbb{Z}: z \leq s\}$ ) and by $\sigma$ its fractional part. So, $s=[s]+\sigma$ with $0 \leq \sigma<1$. The Sobolev space $H^{s}(\Omega)$ is defined as $W^{s, 2}(\Omega)$ provided that $s \in \mathbb{N}_{0}$ and if $s \notin \mathbb{N}_{0}$ then $H^{s}(\Omega)$ contains all functions $u \in H^{[s]}(\Omega)$ that have finite norm $\|u\|_{H^{s}(\Omega)}$, which comes from the following inner product:

$$
(u, v)_{H^{s}(\Omega)}:=(u, v)_{H^{[s]}(\Omega)}+(u, v)_{H^{s, \sigma}(\Omega)},
$$

where

$$
(u, v)_{H^{s, \sigma}(\Omega)}:=\sum_{|\alpha|=[s]} \int_{\Omega} \int_{\Omega} \frac{\left(\partial^{\alpha} u(x)-\partial^{\alpha} u(y)\right) \overline{\left(\partial^{\alpha} v(x)-\partial^{\alpha} v(y)\right)}}{|x-y|^{n+2 \sigma}} d x d y
$$

The pair $\left(H^{s}(\Omega),(\cdot, \cdot)_{H^{s}(\Omega)}\right)$ is a Hilbert space. Other definitions of these spaces are in [2, Subsection 5.1], [89, p. 40] and [118, p. 49]. In the literature we also find the $W^{s, p}(\Omega)$ spaces, see for example [45, Section 6]. We put $W_{0}^{s, p}(\Omega)$ to denote the closure of $C_{c}^{\infty}(\Omega)$ in $W^{s, p}(\Omega)$. We set $H_{0}^{k}(\Omega):=$ $W_{0}^{k, 2}(\Omega)$. It can be proved that $\left(H_{0}^{k}(\Omega),(\cdot, \cdot)_{H^{k}(\Omega)}\right)$ is a Hilbert space. The dual of $H_{0}^{k}(\Omega)$ is denoted by $H^{-k}(\Omega)$.

We continue this section with a space which is constituted by abstract functions or vector-valued functions (see, e.g., [35, Subsection 1.1.4], [44, Chapter 3] or [78, Chapter 1]). Let $X$ be a Banach space, $1 \leq p<\infty$ and $(a, b) \subset \mathbb{R}$. The symbol $L^{p}((a, b), X)$ denotes the space of all equivalence classes of strongly Bochner-measurable functions $u:(a, b) \rightarrow X$ such that $t \mapsto\|u(t)\|_{X}$ belongs to $L^{p}(a, b)$, which is a Banach space with the norm

$$
\|u\|_{L^{p}((a, b), X)}:=\left(\int_{a}^{b}\|u(t)\|_{X}^{p} d t\right)^{1 / p}
$$

When $p=2$ and $X$ is a Hilbert space, then $L^{2}((a, b), X)$ is a Hilbert space with the scalar product

$$
(u, v)_{L^{2}((a, b), X)}:=\int_{a}^{b}(u(t), v(t))_{X} d t
$$

$\mathscr{D}^{\prime}((a, b), X):=\mathcal{L}(\mathscr{D}(a, b), X)$ denotes the space of all vector-valued distributions $T: \mathscr{D}(a, b) \rightarrow X$. The vector distributional derivative of order $k$ of $T \in \mathscr{D}^{\prime}((a, b), X)$, denoted by $T^{(k)}$, is defined as

$$
\left\langle T^{(k)}, \varphi\right\rangle:=(-1)^{k}\left\langle T, \varphi^{(k)}\right\rangle \quad \forall \varphi \in \mathscr{D}(a, b),
$$

where $\varphi^{(k)}:=d^{k} \varphi / d t^{k}$. The previous equality holds in $X$. Now, if $u \in$ $L^{p}((a, b), X)$ then $\widetilde{u}: \mathscr{D}(a, b) \rightarrow X$ given by

$$
\langle\widetilde{u}, \varphi\rangle:=\int_{a}^{b} u(t) \varphi(t) d t
$$

is a continuous linear mapping, this is, $\widetilde{u} \in \mathscr{D}^{\prime}((a, b), X)$. The above integral should be understood as a Bochner integral (for details see [1, p. 206] or [47, Appendix C]). The mapping $L^{p}((a, b), X) \ni u \mapsto \widetilde{u} \in \mathscr{D}^{\prime}((a, b), X)$ is injective and thus identifying $\widetilde{u}$ with $u$ we obtain that

$$
L^{p}((a, b), X) \subset \mathscr{D}^{\prime}((a, b), X)
$$

see Section 1.3 in [89, p. 6].
We conclude this section with the Aubin-Lions-Simon lemma. Interesting historical comments on this result can be found in [116].

Definition 2.1. Let $X$ and $Y$ be two normed spaces such that $X \subset Y$. If id : $X \rightarrow Y$ is continuous, we say $X$ is continuously embedded in $Y$ and write $X \hookrightarrow Y$. We use the symbol $X \stackrel{c}{\hookrightarrow} Y$ when id : $X \rightarrow Y$ is a compact operator.

Theorem 2.2 (Aubin-Lions-Simon lemma). Let $B_{0}, B, B_{1}$ be three Banach spaces such that $B_{0} \stackrel{c}{\hookrightarrow} B$ and $B \hookrightarrow B_{1}$. Let $1 \leq p, q<\infty$ and $t>0$. Let us consider the space

$$
\mathcal{W}_{p, q}\left((0, t) ; B_{0}, B_{1}\right):=\left\{u \in L^{p}\left((0, t), B_{0}\right): u_{t} \in L^{q}\left((0, t), B_{1}\right)\right\}
$$

with the norm

$$
\|u\|_{\mathcal{W}_{p, q}\left((0, t) ; B_{0}, B_{1}\right)}:=\|u\|_{L^{p}\left((0, t), B_{0}\right)}+\left\|u_{t}\right\|_{L^{q}\left((0, t), B_{1}\right)}
$$

where $u_{t}:=d u / d t$. Then,

$$
\mathcal{W}_{p, q}\left((0, t) ; B_{0}, B_{1}\right) \stackrel{c}{\hookrightarrow} L^{p}((0, t), B) .
$$

Proof. See Theorem II.5.16 in [27].

### 2.2 Regular boundaries

The boundary of an open set $\Omega \subset \mathbb{R}^{n}$ is denoted by $\partial \Omega:=\bar{\Omega} \cap\left(\mathbb{R}^{n} \backslash \Omega\right)$. Here we will find properties, which depend on the regularity of $\partial \Omega$, that satisfy the Sobolev spaces defined over $\Omega$. One of these results is known as Rellich-Kondrachov theorem. For a vector-valued version, see [6, Theorem 5.1]. Before presenting the theorem we will give Definition 1.2.1.1 of [62]. There a couple of types of functions are mentioned, among others, whose definitions we present below.

Let $\mathscr{O}$ be an open set of $\mathbb{R}^{n-1}$. We recall that a function $\varphi: \mathscr{O} \rightarrow \mathbb{R}$ is said to be Lipschitz continuous function or simply Lipschitz function if it satisfies

$$
\left|\varphi\left(x^{\prime}\right)-\varphi\left(y^{\prime}\right)\right| \leq C\left|x^{\prime}-y^{\prime}\right| \quad \text { for all } \quad x^{\prime}, y^{\prime} \in \mathscr{O}
$$

where the constant $C>0$ does not depend on $x^{\prime}$ and $y^{\prime}$. We say that $\varphi$ belongs to the class $C^{k, 1}$ if it is $k$ times continuously differentiable and its derivatives of order $k$ are Lipschitz functions.

Definition 2.3. Let $\Omega$ be an open subset of $\mathbb{R}^{n}$. We say that $\partial \Omega$ is Lipschitz (resp. of class $C^{k, 1}$ ) if for every $x \in \partial \Omega$ there exists a neighbourhood $V$ of $x$ in $\mathbb{R}^{n}$ and new orthogonal coordinates $\left\{y_{1}, \ldots, y_{n}\right\}$ such that $V$ is a hypercube in the new coordinates given by

$$
V=\left\{\left(y_{1}, \ldots, y_{n}\right):-a_{j}<y_{j}<a_{j}, 1 \leq j \leq n\right\}
$$

and moreover there exists a Lipschitz (resp. of class $C^{k, 1}$ ) function $\varphi$, defined on the set

$$
V^{\prime}:=\left\{y^{\prime}=\left(y_{1}, \ldots, y_{n-1}\right):-a_{j}<y_{j}<a_{j}, 1 \leq j \leq n-1\right\}
$$

and such that

$$
\begin{aligned}
\left|\varphi\left(y^{\prime}\right)\right| & \leq a_{n} / 2 \text { for every } y^{\prime} \in V^{\prime} \\
\Omega \cap V & =\left\{y=\left(y^{\prime}, y_{n}\right) \in V: y_{n}<\varphi\left(y^{\prime}\right)\right\} \\
\partial \Omega \cap V & =\left\{y=\left(y^{\prime}, y_{n}\right) \in V: y_{n}=\varphi\left(y^{\prime}\right)\right\} .
\end{aligned}
$$

Remark 2.4. The above definition means that, in a neighbourhood of $x, \Omega$ is below the graph of $\varphi$ and that the boundary $\partial \Omega$ coincides with the graph of $\varphi$.

Theorem 2.5 (cf. [108, Theorem 1.15]). Let $\Omega$ be a bounded open subset of $\mathbb{R}^{n}$ with a Lipschitz boundary. If $s>t \geq 0$ and $s-n / p>t-n / q$, then

$$
W^{s, p}(\Omega) \stackrel{c}{\hookrightarrow} W^{t, q}(\Omega) .
$$

For $1<p<\infty$ and $s \in \mathbb{R}$, the boundary spaces $W^{s, p}(\partial \Omega)$ are defined in different ways which are endowed with a norm that makes them Banach spaces; see Remark 4 in [40, p. 146], [62, Definition 1.3.3.2] and [107, p. 90]. We fix $H^{s}(\partial \Omega):=W^{s, 2}(\partial \Omega)$, these spaces are rigorously introduced in [56, Section 9.2], [72, Section 4.2], [89, p. 34] and [98, Section 2.7]. We put $L^{p}(\partial \Omega):=W^{0, p}(\partial \Omega)$. Due to (7.17) in [89, p. 35], we can affirm that $H^{s}(\partial \Omega)$ is dense in $L^{2}(\partial \Omega)$ for $s \geq 0$. When $s>0$, the dual space of the antilinear continuous functionals on $H^{s}(\partial \Omega)$ is denoted by $H^{-s}(\partial \Omega)$. For the boundary spaces $H^{s}(\partial \Omega)$ there is a result of compact embedding and consequently also continuous.

Theorem 2.6 (cf. [72, Theorem 4.2.2]). Let $\Omega$ be a bounded open set of $\mathbb{R}^{n}$ with a $C^{k, 1}$ boundary and let $t, s \in \mathbb{R}$ with $|t|,|s| \leq k+1 / 2$. Then,

$$
H^{s}(\partial \Omega) \stackrel{c}{\hookrightarrow} H^{t}(\partial \Omega) \text { for } t<s
$$

There is a relation among the boundary spaces and the Sobolev spaces given by the trace theorem, which is stated below.

Theorem 2.7 (cf. [62, Theorem 1.5.1.2]). Let $\Omega \subset \mathbb{R}^{n}$ be a bounded open set with a $C^{k, 1}$ boundary. Let $1<p<\infty$ and $s \in \mathbb{R}$ such that $s-1 / p \notin \mathbb{Z}$, $s \leq k+1$ and $0 \leq[s-1 / p]=: l$. The mapping

$$
C^{k, 1}(\bar{\Omega}) \ni u \mapsto\left(\left.u\right|_{\partial \Omega},\left.\frac{\partial u}{\partial \nu}\right|_{\partial \Omega}, \ldots,\left.\frac{\partial^{l} u}{\partial \nu^{l}}\right|_{\partial \Omega}\right)
$$

has a unique extension to a surjective continuous linear operator

$$
\gamma: W^{s, p}(\Omega) \rightarrow \prod_{j=0}^{l} W^{s-j-1 / 2, p}(\partial \Omega)
$$

where $\frac{\partial^{j} u}{\partial \nu^{j}}$ is the $j$-order normal derivative of $u$ on $\partial \Omega$.
The operator $\gamma$ is called trace operator and its components are denoted by $\gamma_{0}, \gamma_{1}, \ldots, \gamma_{l}$ and known as trace of order 0 , trace of order $1, \ldots$, trace of order $l$; respectively. So, for $u \in W^{s, p}(\Omega)$ we can write $\gamma u=\left(\gamma_{0} u, \gamma_{1} u, \ldots, \gamma_{l} u\right)$. Due to Theorem 2.7, for certain Sobolev spaces $W_{0}^{s, p}(\Omega)$ with regular boundary $\partial \Omega$ we have a characterization stated in the following result.

Theorem 2.8 (cf. [62, Corollary 1.5.1.6]). Let $\Omega \subset \mathbb{R}^{n}$ be a bounded open set with a $C^{k, 1}$ boundary. Let $1<p<\infty$ and $s \in \mathbb{R}$ such that $s-1 / p \notin \mathbb{Z}$, $s \leq k+1$ and $0 \leq[s-1 / p]=: l$. Then, $u \in W_{0}^{s, p}(\Omega)$ if and only if $u \in W^{s, p}(\Omega)$ and

$$
u=\frac{\partial u}{\partial \nu}=\cdots=\frac{\partial^{l} u}{\partial \nu^{l}}=0 \text { on } \partial \Omega
$$

Green's formulas for integration are given below. The following result is known as integration by parts. A more general version can be found in Section 3.1.2 from 107.

Theorem 2.9 (cf. [62, Theorem 1.5.3.1]). Let $\Omega$ be a bounded open subset of $\mathbb{R}^{n}$ with Lipschitz boundary $\partial \Omega$. For $u \in W^{1, p}(\Omega)$ and $v \in W^{1, q}(\Omega)$ with $1<p<\infty$ and $1 / p+1 / q=1$, we have

$$
\int_{\Omega}\left(\partial_{j} u\right) \bar{v} d x=-\int_{\Omega} u\left(\partial_{j} \bar{v}\right) d x+\int_{\partial \Omega} u \bar{v} \nu_{j} d S
$$

for $j=1, \ldots, n$. Here $\nu_{j}$ denotes the $j$-th component of $\nu$ which is the unit outer normal vector along $\partial \Omega$.

Remark 2.10. The product $u \bar{v}$ in the integral over $\partial \Omega$ should be understood as $\left(\gamma_{0} u\right)\left(\gamma_{0} \bar{v}\right)$. The same holds in similar situations in the boundary integrals that appear later.

Next we present an integration formula that will be widely used in this work which is an easy consequence of Theorem 2.9.
Theorem 2.11 (cf. [62, Theorem 1.5.3.7]). Let $\Omega$ be a bounded open subset of $\mathbb{R}^{n}$ with a Lipschitz boundary. Then for every $u \in H^{2}(\Omega)$ and $v \in H^{1}(\Omega)$, we have

$$
\begin{equation*}
\int_{\Omega}(\Delta u) \bar{v} d x=-\int_{\Omega} \nabla u \cdot \nabla \bar{v} d x+\int_{\partial \Omega}\left(\partial_{\nu} u\right) \bar{v} d S \tag{2.1}
\end{equation*}
$$

where $\Delta u:=\sum_{j=1}^{n} \partial_{j}^{2} u$ is the Laplace operator of $u, \nabla u:=\left(\partial_{1} u, \ldots, \partial_{n} u\right)$ is the gradient of $u$ and $\partial_{\nu} u$ denotes the derivative of $u$ in the direction of $\nu$.
Corollary 2.12. Let $\Omega \subset \mathbb{R}^{n}$ be a bounded open set with a Lipschitz boundary. Then,

$$
\begin{equation*}
\int_{\Omega}(\Delta u) \bar{v} d x=\int_{\Omega} u(\Delta \bar{v}) d x+\int_{\partial \Omega}\left[\left(\partial_{\nu} u\right) \bar{v}-u\left(\partial_{\nu} \bar{v}\right)\right] d S \tag{2.2}
\end{equation*}
$$

for all $u, v \in H^{2}(\Omega)$.
Proof. An application of (2.1) twice allows to get the desired formula.
Identities (2.1) and (2.2) are known as first Green's formula and second Green's formula, respectively. The next theorem is an extension of Theorem 2.11 .

Definition 2.13 (cf. [49, APPENDIX C]). Let $U \subset \mathbb{R}^{n}$ be open and bounded and $k \in \mathbb{N}$. We say that the boundary $\partial U$ is $C^{k}$ if for each point $x \in \partial U$ there exist $r>0$ and a $C^{k}$ function $\phi: \mathbb{R}^{n-1} \rightarrow \mathbb{R}$ such that -upon relabeling and reorienting the coordinates axis if necessary- we have

$$
U \cap B(x, r)=\left\{x \in B(x, r): x_{n}>\phi\left(x_{1}, \ldots, x_{n-1}\right)\right\},
$$

where $B(x, r):=\left\{y \in \mathbb{R}^{n}:|x-y|<r\right\}$. We also say that $\partial U$ is a boundary of class $C^{k}$ or simply $C^{k}$-boundary $\partial U$. The boundary $\partial U$ is of class $C^{\infty}$ if it is $C^{k}$-boundary $\partial U$ for all $k \in \mathbb{N}$.
Theorem 2.14 (cf. [21, Lemma 8.2.4]). Let $\Omega \subset \mathbb{R}^{n}$ be a bounded domain with $C^{2}$-boundary $\partial \Omega$. For $u \in H^{1}(\Omega)$ with $\Delta u \in L^{2}(\Omega)$ and $v \in H^{1}(\Omega)$, we have that

$$
\begin{equation*}
\int_{\Omega}(\Delta u) \bar{v} d x=-\int_{\Omega} \nabla u \cdot \nabla \bar{v} d x+\left\langle\partial_{\nu} u, v\right\rangle_{H^{-1 / 2}(\partial \Omega) \times H^{1 / 2}(\partial \Omega)} . \tag{2.3}
\end{equation*}
$$

Remark 2.15. In Theorem 2.14, $\Delta u$ is a regular distribution generated by a unique function $f \in L^{2}(\Omega)$ through the expression

$$
\langle\Delta u, \varphi\rangle:=\int_{\Omega} f \bar{\varphi} d x, \quad \varphi \in C_{c}^{\infty}(\Omega)
$$

In this case, $f$ is denoted by $\Delta u$.
In [62, p. 62], formula (2.3) appears but now: $u \in H^{1}(\Omega)$ with $\Delta u \in L^{p}(\Omega)$ and $\Omega$ being a bounded open subset of $\mathbb{R}^{n}$ with a $C^{1,1}$ boundary.

The following result is an integration formula known as Rellich identity and it is a immediate consequence of Corollary 2.1 in [100. This formula will allow obtaining a key estimate in the proof of polynomial stability in Section 4.4.

Theorem 2.16. Let $\Omega \subset \mathbb{R}^{n}$ be a bounded domain with boundary of class $C^{2}$. If $u \in H^{2}(\Omega)$ and $h \in C^{1}\left(\bar{\Omega}, \mathbb{R}^{n}\right)$, then

$$
\begin{aligned}
2 \operatorname{Re} \int_{\Omega} \Delta u(h \cdot \nabla \bar{u}) d x= & 2 \operatorname{Re} \int_{\partial \Omega} \partial_{\nu} u(h \cdot \nabla \bar{u}) d S-\int_{\partial \Omega}(h \cdot \nu)|\nabla u|^{2} d S \\
& +\int_{\Omega} \operatorname{div} h|\nabla u|^{2} d x-2 \operatorname{Re} \sum_{i, j=1}^{n} \int_{\Omega} \partial_{i} h_{j} \partial_{i} u \partial_{j} \bar{u} d x
\end{aligned}
$$

where $\operatorname{div} h:=\partial_{1} h_{1}+\cdots+\partial_{n} h_{n}$.
We continue this section with an integration formula involving the bounded open set $\Omega_{1} \subset \mathbb{R}^{2}$, which has a boundary of class $C^{4}$, from Section 1.1.

Proposition 2.17 (cf. [15, Lemma 2.1]). For $u \in H^{4}\left(\Omega_{1}\right)$ and $v \in H^{2}\left(\Omega_{1}\right)$ such that $u=\partial_{\nu} u=v=\partial_{\nu} v=0$ on $\Gamma$, it holds

$$
\begin{align*}
\int_{\Omega_{1}}\left(\Delta^{2} u\right) \bar{v} d x= & \mu \int_{\Omega_{1}} \Delta u \Delta \bar{v} d x+(1-\mu) \int_{\Omega_{1}} \nabla^{2} u: \nabla^{2} \bar{v} d x  \tag{2.4}\\
& -\int_{I}\left(\mathscr{B}_{1} u\right) \partial_{\nu} \bar{v} d S+\int_{I}\left(\mathscr{B}_{2} u\right) \bar{v} d S
\end{align*}
$$

where

$$
\begin{equation*}
\nabla^{2} u: \nabla^{2} \bar{v}:=u_{x_{1} x_{1}} \bar{v}_{x_{1} x_{1}}+u_{x_{2} x_{2}} \bar{v}_{x_{2} x_{2}}+2 u_{x_{1} x_{2}} \bar{v}_{x_{1} x_{2}} \tag{2.5}
\end{equation*}
$$

Generalizations of the previous result can be found in [3], formula (2.13), or in [128], identity (3.1). Next we give the version of the first reference, which we will use in Subsection 3.1.2,

Proposition 2.18. Let $\Omega \subset \mathbb{R}^{2}$ be a bounded open set with boundary of class $C^{4}$. If $u \in H^{2}(\Omega)$ with $\Delta^{2} u \in L^{2}(\Omega)$ and $v \in H^{2}(\Omega)$, then

$$
\begin{aligned}
\int_{\Omega}\left(\Delta^{2} u\right) \bar{v} d x= & \mu(\Delta u, \Delta v)_{L^{2}(\Omega)}+(1-\mu) \int_{\Omega} \nabla^{2} u: \nabla^{2} \bar{v} d x \\
& -\left\langle\mathscr{B}_{1} u, \partial_{\nu} v\right\rangle_{H^{-1 / 2}(\partial \Omega) \times H^{1 / 2}(\partial \Omega)}+\left\langle\mathscr{B}_{2} u, v\right\rangle_{H^{-3 / 2}(\partial \Omega) \times H^{3 / 2}(\partial \Omega)}
\end{aligned}
$$

### 2.3 Some useful inequalities

In [49, p. 706] appears the well-known Young's inequality: Let $p, q>1$ such that $1 / p+1 / q=1$. For $a, b \geq 0$ we have

$$
\begin{equation*}
a b \leq \frac{1}{p} a^{p}+\frac{1}{q} b^{q} . \tag{2.6}
\end{equation*}
$$

Proposition 2.19. Let $\alpha \beta>0$. For $a, b \geq 0$ we have that

$$
\begin{equation*}
a^{\alpha} b^{\beta} \leq \frac{\alpha}{\alpha+\beta} a^{\alpha+\beta}+\frac{\beta}{\alpha+\beta} b^{\alpha+\beta} \tag{2.7}
\end{equation*}
$$

Proof. If $\alpha \beta>0$, then $p:=\frac{\alpha+\beta}{\alpha}>1$ and $q:=\frac{\alpha+\beta}{\beta}>1$ satisfy the equality $1 / p+1 / q=1$. Now, (2.7) follows from the application of (2.6) to the product $a^{\alpha} b^{\beta}$.

Corollary 2.20. Let $a, b \geq 0$ and $\alpha \in(0,2)$. We have

$$
\begin{equation*}
a^{2-\alpha} b^{\alpha} \leq \varepsilon a^{2}+C_{\alpha, \varepsilon} b^{2} \quad \forall \varepsilon>0 \tag{2.8}
\end{equation*}
$$

where $C_{\alpha, \varepsilon}:=\frac{\alpha}{2}\left(\frac{2-\alpha}{2 \varepsilon}\right)^{\frac{2-\alpha}{\alpha}}$.
Proof. As $0<\alpha<2$, then $(2-\alpha) \alpha>0$. Let $\varepsilon>0$. Applying (2.7) to the right-hand side of the expression

$$
a^{2-\alpha} b^{\alpha}=\left(\sqrt{\frac{2 \varepsilon}{2-\alpha}} a\right)^{2-\alpha}\left[\left(\frac{2 \varepsilon}{2-\alpha}\right)^{\frac{\alpha-2}{2 \alpha}} b\right]^{\alpha},
$$

we get (2.8).

Now, we present a generalization of Theorem 1.4.4 in [95], see [17, Proposition 7.3] or [28, Theorem 1.6.6], followed by a simple proof.

Theorem 2.21. Let $1<p<\infty$ and let $\Omega$ be a bounded domain in $\mathbb{R}^{n}$ with $C^{1}$-boundary $\partial \Omega$. For $u \in W^{1, p}(\Omega)$, the following estimate holds:

$$
\begin{equation*}
\|u\|_{L^{p}(\partial \Omega)} \leq C\|u\|_{W^{1, p}(\Omega)}^{1 / p}\|u\|_{L^{p}(\Omega)}^{1-1 / p} \tag{2.9}
\end{equation*}
$$

with $C$ being a positive constant independent of $u$.
Proof. By Theorem 1.5.1.10 in [62], there exists a constant $C>0$ such that

$$
\begin{equation*}
\|u\|_{L^{p}(\partial \Omega)}^{p} \leq C\left[\varepsilon^{1-1 / p}\|u\|_{W^{1, p}(\Omega)}^{p}+\varepsilon^{-1 / p}\|u\|_{L^{p}(\Omega)}^{p}\right] \tag{2.10}
\end{equation*}
$$

for any $u \in W^{1, p}(\Omega)$ and for all $0<\varepsilon<1$. Note that 2.9 is trivially satisfied if $u=0$. Let $u \in W^{1, p}(\Omega)$ with $u \neq 0$. Taking $\varepsilon=\frac{\|u\|_{L^{p}(\Omega)}^{p}}{\|u\|_{W^{1, p}(\Omega)}}$, we obtain (2.9) from (2.10).

Corollary 2.22. Let $\Omega \subset \mathbb{R}^{n}$ be a bounded domain with $C^{1}$-boundary $\partial \Omega$. For any $u \in H^{2}(\Omega)$, the following estimate holds:

$$
\left\|\partial_{\nu} u\right\|_{L^{2}(\partial \Omega)} \leq \widetilde{C}\|u\|_{H^{2}(\Omega)}^{1 / 2}\|\nabla u\|_{L^{2}(\Omega)^{n}}^{1 / 2}
$$

with $\widetilde{C}$ being a positive constant independent of $u$.
Proof. Let $u \in C^{2}(\bar{\Omega})$. Because of Theorem 2.21, we can write

$$
\begin{aligned}
\left\|\partial_{\nu} u\right\|_{L^{2}(\partial \Omega)} & \leq \sum_{j=1}^{n}\left\|\nu_{j} \partial_{j} u\right\|_{L^{2}(\partial \Omega)} \\
& \leq C \max _{\partial \Omega}|\nu| \sum_{j=1}^{n}\left\|\partial_{j} u\right\|_{H^{1}(\Omega)}^{1 / 2}\left\|\partial_{j} u\right\|_{L^{2}(\Omega)}^{1 / 2} \\
& \leq \widetilde{C}\|u\|_{H^{2}(\Omega)}^{1 / 2}\|\nabla u\|_{L^{2}(\Omega)^{n}}^{1 / 2},
\end{aligned}
$$

where $\widetilde{C}:=n C \max \left\{\left|\nu\left(x^{\prime}\right)\right|: x^{\prime} \in \partial \Omega\right\}$. Since $\partial \Omega \in C^{1}$, then it satisfies the segment property ${ }^{1}$ and thus $C^{2}(\bar{\Omega})$ is dense in $H^{2}(\Omega)$, see [1, p. 68]. This density allows us to finish the proof.

[^0]In the next theorem, we present the well-known interpolation inequality for Sobolev spaces. This result is the product of the combination of Theorem 2 of Subsection 4.3.1 in [120, p. 317], Remark 5 of Subsection 4.2.3 in [120, p. 314] and part (f) of Theorem of Subsection 1.9.3 in [120, p. 59].

Theorem 2.23. Let $\Omega$ be a bounded domain of $\mathbb{R}^{n}$ with $C^{1}$-boundary. For $0 \leq s_{2}<s_{1}$ and $0<\theta<1$, there exists a constant $C:=C_{\theta, s_{1}, s_{2}}>0$ such that

$$
\begin{equation*}
\|u\|_{H^{(1-\theta) s_{1}+\theta s_{2}}(\Omega)} \leq C\|u\|_{H^{s_{1}}(\Omega)}^{1-\theta}\|u\|_{H^{s_{2}}(\Omega)}^{\theta} \tag{2.11}
\end{equation*}
$$

for all $u \in H^{s_{1}}(\Omega)$.
Next we present the Friedrichs inequality also known as the generalized Poincaré inequality.

Theorem 2.24 (cf. [107, Theorem 1.9]). Let $\Omega \subset \mathbb{R}^{n}$ be a bounded domain with Lipschitz boundary. Let $\Gamma \subset \partial \Omega$ with $\mu(\Gamma) \neq 0$, where $\mu$ is the $(n-1)$ dimensional Lebesgue surface measure. Then, for $u \in H^{1}(\Omega)$ we have that

$$
\|u\|_{H^{1}(\Omega)} \leq C\left(\|u\|_{L^{2}(\Gamma)}^{2}+\|\nabla u\|_{L^{2}(\Omega)^{n}}^{2}\right)^{1 / 2}
$$

where the positive constant $C$ depends only on $\Omega$.

### 2.4 Semigroups and groups of bounded linear operators

We will give some concepts (taken from the books [47, 48, 58, 95, 109]) and results of our interest corresponding to the theory of semigroups and groups of bounded linear operators, which have to do with generation, stability and analyticity.

Throughout this section $X$ will be a Banach space over $\mathbb{C}$. The space $\mathcal{L}(X)$ endowed with the usual operator norm is a Banach space.

Definition 2.25. A one-parameter family $(T(t))_{t \geq 0} \subset \mathcal{L}(X)$ is a semigroup of bounded linear operators or simply semigroup on $X$ if satisfies
i) $T(0)=I$, where $I$ is the identity operator on $X$.
ii) $T(s+t)=T(s) T(t)$ for every $s, t \geq 0$.

Definition 2.26. An operator $A$ is the infinitesimal generator or simply generator of a semigroup $(T(t))_{t \geq 0}$ when its domain is given by

$$
D(A):=\left\{x \in X: \lim _{t \rightarrow 0^{+}} \frac{T(t) x-x}{t} \text { exists }\right\}
$$

and

$$
A x:=\lim _{t \rightarrow 0^{+}} \frac{T(t) x-x}{t} \quad \text { for } \quad x \in D(A) .
$$

A semigroup $(T(t))_{t \geq 0}$ on $X$ is called strongly continuous semigroup or simply a $C_{0}$-semigroup if

$$
\lim _{t \rightarrow 0^{+}} T(t) x=x \quad \text { for every } \quad x \in X
$$

Proposition 2.27. Let $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup and $A$ be its infinitesimal generator. If $x \in D(A)$, then $T(t) x \in D(A)$ and

$$
\frac{d}{d t} T(t) x=A T(t) x=T(t) A x .
$$

Proof. See part c) of Theorem 2.4 in [109, p. 4].
Proposition 2.28 (cf. [109, Corollary 2.5 on p. 5]). If $A$ is the infinitesimal generator of a $C_{0}$-semigroup $(T(t))_{t \geq 0}$, then $D(A)$ is dense in $X$ and $A$ is a closed linear operator.

Let $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup. Due to Theorem 2.2 in [109, p. 4], there exist constants $\omega \geq 0$ and $M \geq 1$ such that $\|T(t)\|_{\mathcal{L}(X)} \leq M e^{\omega t}$ for all $t \geq 0$. When $\omega=0$ and $M=1,(T(t))_{t \geq 0}$ is called $C_{0}$-semigroup of contractions.

Let $X^{\prime}$ be the dual of $X$. The value of $x^{\prime} \in X^{\prime}$ at $x \in X$ is denoted by $\left\langle x, x^{\prime}\right\rangle_{X \times X^{\prime}}$ or $\left\langle x^{\prime}, x\right\rangle_{X^{\prime} \times X}$. For $x \in X$ the duality set $\mathcal{F}(x)$ is defined as

$$
\mathcal{F}(x):=\left\{x^{\prime} \in X^{\prime}:\left\langle x, x^{\prime}\right\rangle_{X \times X^{\prime}}=\|x\|_{X}^{2}=\left\|x^{\prime}\right\|_{X^{\prime}}^{2}\right\} .
$$

The Hahn-Banach theorem (complex case) allows to obtain the Corollary 1.3 in [29] when the vector space is defined over $\mathbb{C}$. So, we have that for every $x \in X$ there exists $x^{\prime} \in X^{\prime}$ such that $\left\|x^{\prime}\right\|_{X^{\prime}}=\|x\|_{X}$ and $\left\langle x^{\prime}, x\right\rangle_{X^{\prime} \times X}=\|x\|_{X}^{2}$. In consequence, $\mathcal{F}(x) \neq \emptyset$ for every $x \in X$.

Definition 2.29. A linear operator $A: D(A) \subset X \rightarrow X$ is dissipative if for every $x \in D(A)$ there exists $x^{\prime} \in \mathcal{F}(x)$ such that $\operatorname{Re}\left\langle A x, x^{\prime}\right\rangle_{X \times X^{\prime}} \leq 0$.

Theorem 2.30 (Lumer-Phillips). Let $A: D(A) \subset X \rightarrow X$ be a linear operator such that $\overline{D(A)}=X$.
i) If $A$ is dissipative and there is $\lambda_{0}>0$ such that the range, $R\left(\lambda_{0} I-A\right)$, of $\lambda_{0} I-A$ is $X$, then $A$ is the infinitesimal generator of a $C_{0}$-semigroup of contractions on $X$.
ii) If $A$ is the infinitesimal generator of a $C_{0}$-semigroup of contractions on $X$, then $R(\lambda I-A)=X$ for all $\lambda>0$ and $A$ is dissipative.

Proof. See Theorem 4.3 in [109, p. 14].
Theorem 2.31 (cf. [109, Theorem 4.6 on p. 16]). Let $A: D(A) \subset X \rightarrow X$ be dissipative with $R(I-A)=X$. If $X$ is reflexive, then $\overline{D(A)}=X$.

Remark 2.32. Let $V$ and $H$ be two Hilbert spaces over $\mathbb{C}$ such that $V \stackrel{d}{\hookrightarrow} H$, i.e., $V$ is dense in $H$ and is continuously embedded in $H$. Identifying $H$ with its antidual, we have $V \stackrel{d}{\hookrightarrow} H \stackrel{d}{\hookrightarrow} V^{\prime}$ and moreover

$$
\begin{equation*}
\langle f, v\rangle_{V^{\prime} \times V}=(f, v)_{H} \tag{2.12}
\end{equation*}
$$

for all $f \in H$ and $v \in V$ (see Observação 4.11 in [32]). So, if in the definition of dissipativity we put $H$ instead of $X$, we have that $A: D(A) \subset H \rightarrow H$ is dissipative if for any $x \in D(A)$,

$$
\operatorname{Re}(A x, x)_{H} \leq 0
$$

Compare this remark with item (iii) from [48, p. 83].
Now, let $A: D(A) \subset X \rightarrow X$ be a linear operator and $u_{0} \in X$. The abstract Cauchy problem for $A$ with initial data $u_{0}$ is given by

$$
\left\{\begin{align*}
\frac{d u(t)}{d t} & =A u(t) \text { for } t>0  \tag{2.13}\\
u(0) & =u_{0}
\end{align*}\right.
$$

Two types of solutions of the previous system can be considered, which will depend on the choice of the initial data, thanks to the following two theorems.

Definition 2.33. A classic solution or simply solution of (2.13) is a function $u:[0, \infty) \rightarrow X$ such that $u(t)$ is continuous for $t \geq 0$ and continuously differentiable for $t>0$ with $u(t) \in D(A)$ for $t>0$ and satisfying (2.13).

Theorem 2.34. Let $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup of contractions on $X$ and let be $A$ its infinitesimal generator. If $u_{0} \in D(A)$, then (2.13) has an unique solution $u(t)=T(t) u_{0}$, for $t \geq 0$, such that

$$
u \in C([0, \infty), D(A)) \cap C^{1}([0, \infty), X)
$$

where $D(A)$ is equipped with the graph norm $|x|_{D(A)}:=\|x\|_{X}+\|A x\|_{X}$ for $x \in D(A)$.

Proof. See Remark 2.2.1 and Theorem 2.2.2 in 130 .
Definition 2.35. A continuous function $u:[0, \infty) \rightarrow X$ is called a mild solution of (2.13) if $\int_{0}^{t} u(s) d s \in D(A)$ for all $t \geq 0$ and moreover

$$
u(t)=A \int_{0}^{t} u(s) d s+u_{0}
$$

Theorem 2.36. Let $A$ be the generator of a $C_{0}$-semigroup $(T(t))_{t>0}$ on $X$. If $u_{0} \in X$, then $u(t):=T(t) u_{0}$ is the unique mild solution of (2.13).

Proof. See Proposition 6.4 in [47, p. 146].
Next we define the iterations of unbounded operators. Let $n \in \mathbb{N}$. The $n$-th power $A^{n}$ of $A: D(A) \subset X \rightarrow X$ is defined successively as

$$
\begin{aligned}
A^{n} x & :=A\left(A^{n-1} x\right), \\
D\left(A^{n}\right) & :=\left\{x \in D(A): A^{n-1} x \in D(A)\right\},
\end{aligned}
$$

where $A^{0}$ is the identity operator.
Theorem 2.37 (Regularity of solutions). Let $A$ be the infinitesimal generator of a $C_{0}$-semigroup of contractions on $X$. If $u_{0} \in D\left(A^{k}\right), k \in \mathbb{N}$, then the unique solution of problem (2.13) belongs to

$$
\bigcap_{j=0}^{k} C^{k-j}\left([0, \infty), D\left(A^{j}\right)\right)
$$

Proof. See Theorem 7.5 in [29] or Theorem 2.3.1 in [130].
Now we present some notions of stability followed by certain classical characterizations. Subsequently, we give the concept of analytic semigroup and an analyticity result for $C_{0}$-semigroups of contractions.

Definition 2.38. Let $(T(t))_{t \geq 0}$ be a semigroup on $X$. Then, $(T(t))_{t \geq 0}$ is said to be strongly stable if

$$
\lim _{t \rightarrow \infty}\|T(t) x\|_{X}=0, \quad \forall x \in X
$$

The next theorem corresponds to a simpler version of the well-known general result for strong stability established by Arendt and Batty (see [8, Theorem 2.4]). Before presenting the theorem we define a pair of sets. Let $A: D(A) \subset X \rightarrow X$ be a linear operator. The resolvent set $\rho(A)$ of $A$ is the set of all $\lambda \in \mathbb{C}$ for which $\lambda I-A: D(A) \rightarrow X$ is bijective and $(\lambda I-A)^{-1} \in \mathcal{L}(X)$. The set $\sigma(A):=\mathbb{C} \backslash \rho(A)$ is called spectrum of $A$.
Theorem 2.39 (cf. [47, Corollary 2.22 on p. 327]). Let us suppose that $A$ is the infinitesimal generator of a $C_{0}$-semigroup of contractions $(T(t))_{t \geq 0}$ on a Hilbert space $H$. If $A$ has no purely imaginary eigenvalues and $\sigma(A) \cap i \mathbb{R}$ is countable, then $(T(t))_{t \geq 0}$ is strongly stable.
Definition 2.40. A semigroup $(T(t))_{t>0}$ on $X$ is called exponentially stable if there exist constants $\omega>0$ and $M \geq 1$ such that

$$
\begin{equation*}
\|T(t)\|_{\mathcal{L}(X)} \leq M e^{-\omega t} \text { for all } t \geq 0 \tag{2.14}
\end{equation*}
$$

The following characterization for exponential stability is due to Gearhat [57] and Prüss [113], see [95, Theorem 1.3.2]. An elementary proof of this result, but now for bounded $C_{0}$-semigroups, can be seen in [41].
Theorem 2.41. Let $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup of contractions on a Hilbert space $H$ and $A$ be its infinitesimal generator. Then, $(T(t))_{t \geq 0}$ is exponentially stable if and only if $i \mathbb{R} \subset \rho(A)$ and moreover

$$
\begin{equation*}
\limsup _{\lambda \in \mathbb{R},|\lambda| \rightarrow \infty}\left\|(i \lambda I-A)^{-1}\right\|_{\mathcal{L}(H)}<\infty \tag{2.15}
\end{equation*}
$$

In case it is not possible to determine the exponential stability, one tries to look for another type of asymptotic decay. Next we present the notion of polynomial stability, which is slower than exponential stability.
Definition 2.42. Let $(T(t))_{t \geq 0}$ be a semigroup on $X$ and $A$ be its generator. Then, $(T(t))_{t \geq 0}$ is polynomially stable of order $\alpha>0$ if there exists $C>0$ such that

$$
\begin{equation*}
\|T(t) x\|_{X} \leq C t^{-\alpha}|x|_{D(A)} \tag{2.16}
\end{equation*}
$$

for any $t>0$ and for all $x \in D(A)$. In that case, one says that $(T(t))_{t \geq 0}$ decays at a rate $t^{-\alpha}$.

The following observation is based on comments made by Liu and Rao in Introduction from [92].
Remark 2.43. The norm on the right-hand side of (2.16) can not be the norm in $X$ because if $(T(t))_{t>0}$ is bounded, in particular it could be a $C_{0^{-}}$ semigroup of contractions, then (2.16) implies (2.14). Indeed, let us assume

$$
\|T(t) x\|_{X} \leq C t^{-\alpha}\|x\|_{X}, \quad t>0
$$

for all $x \in X$. As $\lim _{t \rightarrow \infty} t^{-\alpha}=0$, then there exists $t_{0} \in \mathbb{R}^{+}$such that $t_{0}^{-\alpha}<1 / C$ and so $\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}<1$. If $(T(t))_{t \geq 0}$ is bounded, there is $\widetilde{C} \geq 1$ such that $\|T(s)\|_{\mathcal{L}(X)} \leq \widetilde{C}$ for all $s \geq 0$. Let $t>0$. By the division algorithm, $t=n_{t} t_{0}+r_{t}$ where $n_{t} \in \mathbb{N}_{0}$ and $0 \leq r_{t}<t_{0}$. In consequence,

$$
\begin{equation*}
\|T(t)\|_{\mathcal{L}(X)}=\left\|\left(T\left(t_{0}\right)\right)^{n_{t}} T\left(r_{t}\right)\right\|_{\mathcal{L}(X)} \leq \widetilde{C}\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{n_{t}} \tag{2.17}
\end{equation*}
$$

If $\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}=0$, then (2.14) holds trivially. Let us assume $\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)} \neq 0$. From 2.17), we get

$$
\|T(t)\|_{\mathcal{L}(X)} \leq \widetilde{C}\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{-r_{t} / t_{0}}\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{t / t_{0}}
$$

Since $0<\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}<1$, we have $\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{-1}>\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{-r_{t} / t_{0}} \geq 1$ (note that $\left.-1<-r_{t} / t_{0} \leq 0\right)$ and there is $\delta>0$ such that $\frac{1}{t_{0}} \ln \left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}=-\delta$. Therefore,

$$
\|T(t)\|_{\mathcal{L}(X)}<M e^{-\delta t}
$$

where $M:=\widetilde{C}\left\|T\left(t_{0}\right)\right\|_{\mathcal{L}(X)}^{-1}>1$.
In [25, Theorem 2.4], Borichev and Tomilov have a characterization for the polynomial stability of bounded $C_{0}$-semigroups. This result is extended by Muñoz Rivera and Racke, now considering $C_{0}$-semigroup of contractions.

Theorem 2.44 (cf. [105, Lemma 5.2]). Let $(T(t))_{t \geq 0}$ be a contraction semigroup on a Hilbert space $H$ with generator $A$ such that $i \mathbb{R} \cap \sigma(A)$ is empty. Then, for $\alpha^{\prime} \in \mathbb{N}_{0}$ and $\beta^{\prime}>0$ fixed the following assertions are equivalent:
i) There exist $C>0$ and $\lambda_{0}>0$ such that for all $\lambda \in \mathbb{R}$ with $|\lambda|>\lambda_{0}$ and all $F \in D\left(A^{\alpha^{\prime}}\right)$ it holds

$$
\begin{equation*}
\left\|(i \lambda \mathcal{I}-A)^{-1} F\right\|_{H} \leq C|\lambda|^{\beta^{\prime}}\left\|A^{\alpha^{\prime}} F\right\|_{H} \tag{2.18}
\end{equation*}
$$

ii) There exists some $C>0$ such that for all $t>0$ it holds

$$
\begin{equation*}
\left\|T(t) A^{-1}\right\|_{\mathcal{L}(H)} \leq C t^{-\frac{1}{\alpha^{\prime}+\beta^{\prime}}} \tag{2.19}
\end{equation*}
$$

Asymptotic behavior is also a characteristic of some analytic semigroups. More precisely, the decay would be of the exponential type (see Corollary 5.3). On the other side, analytic semigroups have the property of smoothing effect, that is, the solution of $(2.13)$ is of class $C^{\infty}$ regardless of the irregularity of the initial data (see [51, Introduction] and [104, Observação 7.1.1]).

Definition 2.45. For $0<\vartheta \leq \pi$, we set $\Sigma_{\vartheta}:=\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda|<\vartheta\}$. A semigroup $(T(t))_{t \geq 0}$ on $X$ is said to be analytic if it admits an extension $T(\lambda) \in \mathcal{L}(X)$ for $\lambda \in \Sigma_{\theta} \cup\{0\}$ for some $0<\theta \leq \pi / 2$ such that
i) $\lambda \mapsto T(\lambda)$ is analytic on the sector $\Sigma_{\theta}$.
ii) $\lim _{\Sigma_{\theta} \ni \ni \lambda \rightarrow 0}\|T(\lambda) x-x\|_{X}=0$ for every $x \in X$ and each $0<\varepsilon<\theta$.
iii) $T(\lambda+\mu)=T(\lambda) T(\mu)$ for all $\lambda, \mu \in \Sigma_{\theta}$.

Theorem 2.46 (cf. [95, Theorem 1.3.3]). Let $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup of contractions on a Hilbert space $H$ and let $A$ be its generator. Suppose that $i \mathbb{R} \subset \rho(A)$. Then, $T(t)$ is analytic if and only if

$$
\begin{equation*}
\limsup _{\lambda \in \mathbb{R},|\lambda| \rightarrow \infty}\left\|\lambda(i \lambda I-A)^{-1}\right\|_{\mathcal{L}(H)}<\infty \tag{2.20}
\end{equation*}
$$

holds.
We continue with some notions about groups of bounded linear operators and close with a classical result of group generation which was initially proved by the American mathematician Marshall Harvey Stone in 1932.

Definition 2.47. A one-parameter family $(T(t))_{t \in \mathbb{R}} \subset \mathcal{L}(X)$ is a $C_{0}$-group or simply group on $X$ if it satisfies
i) $T(0)=I$.
ii) $T(t+s)=T(t) T(s)$ for any $t, s \in \mathbb{R}$.
iii) $\lim _{t \rightarrow 0} T(t) x=x$ for $x \in X$.

Definition 2.48. The infinitesimal generator $A: D(A) \subset X \rightarrow X$ of a group $(T(t))_{t \in \mathbb{R}}$ on $X$ is the operator

$$
A x:=\lim _{t \rightarrow 0} \frac{1}{t}[T(t) x-x]
$$

defined for every $x$ in its domain

$$
D(A):=\left\{x \in X: \lim _{t \rightarrow 0} \frac{1}{t}[T(t) x-x] \text { exists }\right\} .
$$

Before giving Stone's theorem, we will recall some definitions. Let $H$ be a complex Hilbert space and $A: D(A) \subset H \rightarrow H$ be a densely defined linear operator. The set

$$
D\left(A^{*}\right):=\left\{y \in H: \exists y^{*} \in H \text { such that }(A x, y)_{H}=\left(x, y^{*}\right)_{H} \forall x \in D(A)\right\}
$$

is a vector subspace of $H$. Since $\overline{D(A)}=H$, then for each $y \in D\left(A^{*}\right)$ there exists an unique $y^{*} \in H$ such that $(A x, y)_{H}=\left(x, y^{*}\right)_{H}$ for all $x \in D(A)$. The adjoint of $A$ denoted $A^{*}$ is given by $A^{*}: D\left(A^{*}\right) \rightarrow H$ with $y \mapsto A^{*} y:=y^{*}$. It can be easily proved that $A^{*}$ is a linear operator. Note that

$$
(A x, y)_{H}=\left(x, A^{*} y\right)_{H}
$$

for all $x \in D(A)$ and for any $y \in D\left(A^{*}\right)$. For details of the above statements, see for example [32, p. 292-294] and [56, p. 313]. The operator $A$ is said to be self-adjoint if $A=A^{*}$. The operator $A$ is called skew-adjoint if and only if $i A$ is self-adjoint, i.e., $A^{*}=-A$ since $(\lambda A)^{*}=\bar{\lambda} A^{*}$ for all $\lambda \in \mathbb{C}$ (see $(i)$ of Proposição 5.97 in [32]).

Definition 2.49. Let $H$ be a Hilbert space. A bijective linear operator $U: H \rightarrow H$ is unitary if $U^{*}=U^{-1}$.

It is easy to prove that $U$ is unitary if and only if $U$ is an isometry and $R(U)=H$. So, if $U$ is unitary then $U$ is a bounded linear operator.

Theorem 2.50 (Stone). Let $A: D(A) \subset H \rightarrow H$ be a densely defined linear operator on a complex Hilbert space H. Then, A generates a unitary group $(T(t))_{t \in \mathbb{R}}$ on $H$ if and only if $A$ is skew-adjoint.

Proof. See Theorem 3.24 in [48].

### 2.5 Interpolation-extrapolation scales

Let $X$ be a Banach space and $A: D(A) \subset X \rightarrow X$ be a closed linear operator with $0 \in \rho(A)$. Since $A^{-1} \in \mathcal{L}(X)$, then $\xi \mapsto\left\|A^{-1} \xi\right\|_{X}$ defines a norm on $X$ such that $\left\|A^{-1} \xi\right\|_{X} \leq C\|\xi\|_{X}$ for each $\xi \in X$. Equivalence of norms cannot be guaranteed between $\|\cdot\|_{X}$ and $\left\|A^{-1} \cdot\right\|_{X}$ on $X$. By Theorem 2.3-2 in [77], the normed space $\left(X,\left\|A^{-1} \cdot\right\|_{X}\right)$ has a completion

$$
\left(X_{-1},\|\cdot\|_{-1}\right):=\left(X,\left\|A^{-1} \cdot\right\|_{X}\right)^{\sim}
$$

called the extrapolation space of $X$ generated by $A$, which is a Banach space such that $X \stackrel{d}{\hookrightarrow} X_{-1}$ with $\|\xi\|_{-1}=\left\|A^{-1} \xi\right\|_{X}$ for any $\xi \in X$, see [5, p. 262].

Let us suppose that the operator $A$ is the infinitesimal generator of a $C_{0}$-semigroup $(T(t))_{t \geq 0}$ on $X$. For each $t \geq 0$,

$$
T(t): X \subset X_{-1} \rightarrow X_{-1}
$$

is a densely defined bounded linear operator. The extension by continuity (see, e.g., [32, Teorema 2.42]) implies that there exists a unique extension of $T(t)$ to $X_{-1}$, which is denoted by $T_{-1}(t)$. The family $\left(T_{-1}(t)\right)_{t \geq 0}$ is a $C_{0^{-}}$ semigroup on $X_{-1}$, called extrapolated semigroup in $X_{-1}$, whose generator $A_{-1}: D\left(A_{-1}\right) \subset X_{-1} \rightarrow X_{-1}$ is an extension of $A$ with domain $D\left(A_{-1}\right)=X$ such that $0 \in \rho\left(A_{-1}\right)$, see [47, Theorem 5.5] and [117, Proposition 4.2].

Definition 2.51 ([48, Definition A.12]). Let $X^{\prime}$ be the dual space of $X$. As the domain of $A$ is dense in $X$, the adjoint operator $A^{\prime}: D\left(A^{\prime}\right) \subset X^{\prime} \rightarrow X^{\prime}$ exists and is given by

$$
\begin{aligned}
D\left(A^{\prime}\right) & :=\left\{x^{\prime} \in X^{\prime}: \exists y^{\prime} \in X^{\prime} \text { such that }\left\langle A x, x^{\prime}\right\rangle=\left\langle x, y^{\prime}\right\rangle \forall x \in D(A)\right\}, \\
A^{\prime} x^{\prime} & :=y^{\prime} \text { for all } x^{\prime} \in D\left(A^{\prime}\right) .
\end{aligned}
$$

Here $\langle\cdot, \cdot\rangle:=\langle\cdot, \cdot\rangle_{X \times X^{\prime}}$.
If $X$ is reflexive, then $A^{\prime}: D\left(A^{\prime}\right) \subset X^{\prime} \rightarrow X^{\prime}$ is a densely defined close linear operator such that $\rho\left(A^{\prime}\right)=\rho(A)$. We set $X^{\sharp}:=X^{\prime}, A^{\sharp}:=A^{\prime}$ and $X_{1}^{\sharp}:=\left(D\left(A^{\sharp}\right),\left\|A^{\sharp} \cdot\right\|_{X^{\sharp}}\right)$, see Section 1.2 and Section 1.4 in [5, Chapter V]. Thanks to Corollary 1.4.7 in [5, p. 271], the space $X_{-1}$ is the dual space of $D\left(A^{\prime}\right)$, i.e., $X_{-1}=\left[D\left(A^{\prime}\right)\right]^{\prime}$.

For $m \in \mathbb{N}_{0}$ arbitrary but fixed, we have that the extrapolated discrete power scale $\left[\left(X_{k}, A_{k}\right) ; k \in \mathbb{Z} \cap[-m, \infty)\right]$ of order $m$ is well-defined and it is a densely injected Banach scale. For details, see Section 1.3 in [5, Chapter V]. Here,

$$
X_{k}:=\left\{\begin{array}{l}
\left(D\left(A^{k}\right),\|\cdot\|_{k}\right) \quad \text { if } k \geq 0 \\
\left(X,\|\cdot\|_{k}\right)^{\sim} \text { if } k<0
\end{array}\right.
$$

where $\|x\|_{k}:=\left\|A^{k} x\right\|_{X}$ for $x \in D\left(A^{k}\right)$ and $k \in \mathbb{Z} . A_{k}$ is the $X_{k}$-realization of $A$ if $k \geq 0$ and the closure of $A$ in $X_{k}$ if $k<0$. Putting $T_{k}(t):=\left.T(t)\right|_{X_{k}}$ and
denoting $T_{-k}(t)$ as the continuous extension of $T_{-k+1}(t)$ to $X_{-k}$, for $k \geq 0$, the following diagram we obtain


For $0<\theta<1$ consider an admissible interpolation functor $(\cdot, \cdot)_{\theta}$ of exponent $\theta$, e.g., the real interpolation functor, the complex interpolation functor or the continuous interpolation functor, see Section 2 in [5, Chapter I] and the references therein. Let $\alpha:=k+\theta$ with $k \in \mathbb{Z} \cap[-m, \infty)$. Defining

$$
X_{\alpha}:=\left(X_{k}, X_{k+1}\right)_{\theta}
$$

and $A_{\alpha}$ as the $X_{\alpha}$-realization of $A_{k}$, one has that $X_{\alpha}$ is an intermediate space between $X_{k+1}$ and $X_{k}$, and $\left[\left(X_{\alpha}, A_{\alpha}\right) ; \alpha \in[-m, \infty)\right]$ is a densely injected Banach scale, which is called the interpolation-extrapolation scale of order $m$ generated by $(X, A)$ and $(\cdot, \cdot)_{\theta}$. See Theorem 1.5.1 in [5, p. 275]. Therefore,

$$
X_{\alpha} \stackrel{d}{\hookrightarrow} X_{\beta}
$$

provided that $\alpha>\beta$. Additionally, if $A^{-1}: X \rightarrow X$ is a compact operator then the previous interpolation-extrapolation scale is compactly injected, i.e., $X_{\alpha} \stackrel{c}{\hookrightarrow} X_{\beta}$ for $\alpha>\beta$.

### 2.6 Powers of positive self-adjoint operators

Here we indicate when a self-adjoint operator is positive and present some properties that satisfy its fractional powers. Along this section $H$ denotes a Hilbert space.

A self-adjoint operator $A: D(A) \subset H \rightarrow H$ is said to be positive if there exists $\delta>0$ such that

$$
\begin{equation*}
(A x, x)_{H} \geq \delta\|x\|_{H}^{2} \quad \text { for all } x \in D(A) \tag{2.21}
\end{equation*}
$$

Remark 2.52. If $A: D(A) \rightarrow H$ is a positive self-adjoint operator, then Lemma 4.31 in [97] implies that $(-\infty, \delta) \subset \rho(A)$ and moreover

$$
\begin{equation*}
\left\|(\lambda I-A)^{-1}\right\|_{\mathcal{L}(H)} \leq \frac{1}{\delta-\lambda} \text { for } \lambda<\delta \tag{2.22}
\end{equation*}
$$

where $\delta$ is the positive constant of (2.21). Putting $\varphi(\lambda):=\frac{1-\lambda}{\delta-\lambda}$, it follows that $\varphi(\lambda)$ converges to 1 , as $\lambda \rightarrow-\infty$, and so there exists $\lambda_{0}<0$ such that $|\varphi(\lambda)-1|<1$ for any $\lambda<\lambda_{0}$. Note that the function $\lambda \mapsto \varphi(\lambda)$ is continuous on the compact interval $\left[\lambda_{0}, 0\right]$. Therefore, one obtains that there exists $M>1$ such that

$$
\begin{equation*}
\frac{1-\lambda}{\delta-\lambda} \leq M \text { for } \lambda \leq 0 \tag{2.23}
\end{equation*}
$$

Obviously $(-\infty, 0] \subset \rho(A)$, and the inequalities (2.22) and (2.23) imply that

$$
\left\|(\lambda I-A)^{-1}\right\|_{\mathcal{L}(H)} \leq \frac{M}{1+|\lambda|} \text { for } \lambda \leq 0
$$

Suppose $A: D(A) \rightarrow H$ is a positive self-adjoint operator. Then, $0 \in \rho(A)$ and Remark 2.52 implies that $A$ is a positive operator in the sense of Definition 4.1 in [97] or Definition in Subsection 1.14.1 of [120], and is also positive of type $K$ (see [5, p. 147]). The spectral theory allows to define the fractional powers of $A$. The previous affirmation appears, for instance, in [118, p. 57]. It is known that $A^{s}: D\left(A^{s}\right) \subset H \rightarrow H$ is positive self-adjoint operator for $s>0$. The space $D\left(A^{s}\right)$ is endowed with the inner product

$$
\begin{equation*}
(u, v)_{D\left(A^{s}\right)}:=\left(A^{s} u, A^{s} v\right)_{H} \tag{2.24}
\end{equation*}
$$

for all $u, v \in D\left(A^{s}\right)$ and for any $s>0$, which makes it a Hilbert space because $A^{s}$ is closed and since the norm induced by 2.24$)$, denoted $\|\cdot\|_{D\left(A^{s}\right)}$, is equivalent to the graph norm $|\cdot|_{D\left(A^{s}\right)}$. It can be easily proved that the isometric isomorphism

$$
\begin{equation*}
\widetilde{A}: H \rightarrow[D(A)]^{\prime} \quad \text { given by }\langle\widetilde{A} \xi, \zeta\rangle_{[D(A)]^{\prime} \times D(A)}:=(\xi, A \zeta)_{H} \tag{2.25}
\end{equation*}
$$

is a extension of $A$ which is called standard extension of $A$.
The following theorem collects some of the properties that the positive self-adjoint operators satisfy, which can be found in [5, Section 4.6], 97, Chapter 4], [118, p. 57] and [120, Subsection 1.15.2].

Theorem 2.53. Let $A: D(A) \subset H \rightarrow H$ be a positive self-adjoint operator. Their fractional powers satisfy the following properties:
a) If $s_{1}, s_{2}>0$, then $A^{s_{1}} A^{s_{2}} x=A^{s_{1}+s_{2}} x$ for each $x \in D\left(A^{s_{1}+s_{2}}\right)$.
b) If $s_{2}>s_{1}>0$, then $D\left(A^{s_{2}}\right) \stackrel{d}{\hookrightarrow} D\left(A^{s_{1}}\right) \stackrel{d}{\hookrightarrow} H$.
c) If $D\left(A^{-s}\right):=\left[D\left(A^{s}\right)\right]^{\prime}$ for $s>0$, then $A^{s_{1}-s_{2}}: D\left(A^{s_{1}}\right) \rightarrow D\left(A^{s_{2}}\right)$ is an isomorphism $\forall s_{1}, s_{2} \in \mathbb{R}$ with $s_{1}>s_{2}$.

### 2.7 Elliptic and boundary operators

This section recalls fundamental aspects of elliptic boundary value problems, and is largely based on Chapter 2 of [89. Let $\Omega$ be a bounded domain in $\mathbb{R}^{n}$ with boundary of class $C^{\infty}$ and let

$$
\begin{equation*}
A(D) u:=\sum_{|\alpha| \leq 2 m} a_{\alpha} D^{\alpha} u \tag{2.26}
\end{equation*}
$$

be a linear differential operator in $\Omega$ of order $2 m$. Here $u: \Omega \rightarrow \mathbb{C}, \alpha \in \mathbb{N}_{0}^{n}$, $m \in \mathbb{N}, a_{\alpha} \in \mathbb{C}$ and $D^{\alpha}:=(-i)^{|\alpha|} \partial^{\alpha}$. Its principal symbol is the polynomial $A^{0}(\xi):=\sum_{|\alpha|=2 m} a_{\alpha} \xi^{\alpha}$, where $\xi \in \mathbb{R}^{n}$ and $\xi^{\alpha}:=\xi_{1}^{\alpha_{1}} \cdots \xi_{n}^{\alpha_{n}}$.

Definition 2.54. The operator $A$ is said to be elliptic if the following holds

$$
A^{0}(\xi) \neq 0, \quad \forall \xi \in \mathbb{R}^{n} \backslash\{0\}
$$

If moreover for every linearly independent couple of vectors $\xi_{1}$ and $\xi_{2}$ of $\mathbb{R}^{n}$, the polynomial $A^{0}\left(\xi_{1}+\zeta \xi_{2}\right)$ in the complex variable $\zeta$ has $m$ roots with positive imaginary part, then $A$ is called properly elliptic.

Example 2.55. It is well known that the Laplacian $\Delta$ and also the Bi Laplacian $\Delta^{2}$ are properly elliptic (see, e.g., Section 3.2 in [12]).

Let $B_{1}, B_{2}, \ldots, B_{m}$ be boundary operators of order $m_{j}$ with $j=1, \ldots, m$, respectively, defined as follows

$$
\begin{equation*}
B_{j}(D) v:=\sum_{|\beta| \leq m_{j}} b_{j \beta} D^{\beta} v, \tag{2.27}
\end{equation*}
$$

where $v: \bar{\Omega} \rightarrow \mathbb{C}, \beta \in \mathbb{N}_{0}^{n}, b_{j \beta} \in \mathbb{C}$ and $D^{\beta} v$ must be understood as $\gamma_{0}\left(D^{\beta} v\right)$, and whose principal symbols are given by $B_{j}^{0}(\xi):=\sum_{|\beta|=m_{j}} b_{j \beta} \xi^{\beta}$. In this part of the document $\Gamma$ is considered as a subset of $\partial \Omega$.

Definition 2.56. The system of operators $\left\{B_{j}\right\}_{1 \leq j \leq m}$ is a normal system on $\Gamma$ if the system satisfies the following two conditions:
a) For $x \in \Gamma$, it holds $B_{j}^{0}(\xi) \neq 0$ for any $0 \neq \xi \in \mathbb{R}^{n}$ normal to $\Gamma$ at $x$ and for all $1 \leq j \leq m$.
b) $m_{i} \neq m_{j}$ for $i \neq j$ with $1 \leq i, j \leq m$.

Definition 2.57. The system $\left\{B_{j}\right\}_{1 \leq j \leq m}$ covers the operator $A$ on $\Gamma$ if for all $x \in \Gamma$, all $0 \neq \tau \in \mathbb{R}^{n}$ tangent to $\Gamma$ at $x$, and all $0 \neq \eta \in \mathbb{R}^{n}$ normal to $\Gamma$ at $x$, the polynomials $\mathbb{C} \ni \zeta \mapsto \sum_{|\beta|=m_{j}} b_{j \beta}(\tau+\zeta \eta)^{\beta}$ with $j=1, \ldots, m$ are linearly independent modulo the polynomial $\prod_{k=1}^{m}\left[\zeta-\zeta_{k}^{+}(\tau, \eta)\right]$, where $\zeta_{k}^{+}(\tau, \eta)$ are the roots of the polynomial $A^{0}(\tau+\zeta \eta)$ with positive imaginary part.

The operator $A$ of (2.26) and the boundary operators of (2.27) constitute a system called boundary value problem, which is given by

$$
\left\{\begin{array}{l}
A u=f \quad \text { in } \quad \Omega,  \tag{2.28}\\
B_{j} u=g_{j} \quad \text { on } \quad \partial \Omega
\end{array}\right.
$$

and denoted by $\left(A, B_{1}, \ldots, B_{m}\right)$. Here $f$ and $g_{j}$ are given functions belonging to suitable spaces.

Remark 2.58. The general theory of boundary value problems is developed for boundaries of class $C^{\infty}$. This assumption allows to prove general results. However, in some cases we do not need too much regularity on the boundary (see, e.g., Proposition 2.60 and Remark 3.17).

Definition 2.59. The problem (2.28) is a regular elliptic problem if and only if $A$ is properly elliptic in $\bar{\Omega}$, the operators $B_{j}$ have order $m_{j} \leq 2 m-1$, the system $\left\{B_{j}\right\}_{1 \leq j \leq m}$ is normal on $\partial \Omega$ and covers the operator $A$ on $\partial \Omega$.

Suppose the system $\left(A, B_{1}, \ldots, B_{m}\right)$ is a regular elliptic problem and now consider $\mathcal{A}: D(\mathcal{A}) \subset L^{2}(\Omega) \rightarrow L^{2}(\Omega)$ defined as follows:

$$
\begin{aligned}
D(\mathcal{A}) & :=\left\{u \in H^{2 m}(\Omega): B_{j}(D) u=0 \text { on } \partial \Omega \text { for } j=1, \ldots, m\right\} \\
\mathcal{A} u & :=A(D) u .
\end{aligned}
$$

The space $D(\mathcal{A})$ is endowed with the graph norm. The operator $\mathcal{A}$ is called the realization of $A$ in $L^{2}(\Omega)$ under the boundary conditions $\left\{B_{j}\right\}_{1 \leq j \leq m}$. The space $D(\mathcal{A})$ will be denoted by $H_{B}^{2 m}(\Omega)$ when it is equipped with the norm
of $H^{2 m}(\Omega)$. As $|\cdot|_{D(\mathcal{A})} \sim\|\cdot\|_{H^{2 m}}$ on $D(\mathcal{A})$, where $\sim$ means equivalence in norm, then the following characterizations hold:

$$
\begin{align*}
D\left(\mathcal{A}^{\theta}\right) & =\left[D(\mathcal{A}), L^{2}(\Omega)\right]_{1-\theta}=\left[H_{B}^{2 m}(\Omega), H^{0}(\Omega)\right]_{1-\theta} \\
& =\left\{u \in H^{2 m \theta}(\Omega): B_{j}(D) u=0 \text { on } \partial \Omega, m_{j}<2 m \theta-\frac{1}{2}\right\} \tag{2.29}
\end{align*}
$$

for $0<\theta<1$. A summary explanation of how to get the previous equalities can be seen in [23, p. 172] or [87, p. 284]. Here [ $X, Y]_{\tilde{\theta}}$, with $0 \leq \widetilde{\theta} \leq 1$, is the complex interpolation space defined for an interpolation couple $(X, Y)$ consisting of complex Banach spaces $X$ and $Y$. See [97, Chapter 2] and [120, Section 1.9] for details.

Next, we present a proposition that treats a boundary value problem defined in $\Omega_{1}$ with mixed boundary conditions ( $\Omega_{1}$ as in Section 1.1). In addition to guaranteeing the existence and uniqueness of the problem, this result offers an estimate that will be very useful later on.

Proposition 2.60 (cf. [15, Corollary 4.3]). Let $f \in L^{2}\left(\Omega_{1}\right), g_{1} \in H^{7 / 2}(\Gamma)$, $g_{2} \in H^{5 / 2}(\Gamma), h_{1} \in H^{3 / 2}(I)$ and $h_{2} \in H^{1 / 2}(I)$. Then for sufficiently large $\lambda_{0}>0$, we have that the boundary value problem

$$
\begin{aligned}
\left(\lambda_{0}+\Delta^{2}\right) u=f & \text { in } \Omega_{1}, \\
u=g_{1} & \text { on } \Gamma, \\
\partial_{\nu} u=g_{2} & \text { on } \Gamma, \\
\mathscr{B}_{1} u=h_{1} & \text { on } I, \\
\mathscr{B}_{2} u=h_{2} & \text { on } I,
\end{aligned}
$$

has a unique solution $u \in H^{4}\left(\Omega_{1}\right)$. Moreover, the a priori-estimate

$$
\begin{align*}
\|u\|_{H^{4}\left(\Omega_{1}\right)} \leq C\left(\|f\|_{L^{2}\left(\Omega_{1}\right)}+\right. & \left\|g_{1}\right\|_{H^{7 / 2}(\Gamma)}+\left\|g_{2}\right\|_{H^{5 / 2}(\Gamma)}  \tag{2.30}\\
& \left.+\left\|h_{1}\right\|_{H^{3 / 2}(I)}+\left\|h_{2}\right\|_{H^{1 / 2}(I)}\right)
\end{align*}
$$

holds with a constant $C>0$ which depends on $\lambda_{0}$ but not on $u$ or on the data.

## Chapter 3

## Well-posedness and regularity of the solutions of a plate-membrane system

Using semigroup theory of linear operators it will be proved that the platemembrane transmission problem with initial conditions (1.1)-(1.9) has a unique solution, either a classical or a weak solution depending on the choice of the initial datum, the Lumer-Phillips theorem being the key tool for this purpose. In addition, the regularity of the solutions will be established.

### 3.1 Existence and uniqueness of the solutions

In this section the system (1.1)-(1.9) will be written as an abstract Cauchy problem. For that, certain linear operators defined on appropriate function spaces are necessary. The energy of the system (1.1)-(1.7) is what allows these spaces to be determined, which is defined by

$$
\begin{align*}
E_{\gamma}(t):= & \frac{1}{2} \int_{\Omega_{1}} \beta_{1} \mu|\Delta u|^{2}+\beta_{1}(1-\mu)\left|\nabla^{2} u\right|^{2}+\rho_{1}\left|u_{t}\right|^{2}+\gamma\left|\nabla u_{t}\right|^{2} d x \\
& +\frac{1}{2} \int_{\Omega_{2}} \beta_{2}|\nabla v|^{2}+\rho_{2}\left|v_{t}\right|^{2} d x+\frac{1}{2} \int_{\Omega_{1}} \rho_{0}|\theta|^{2} d x \tag{3.1}
\end{align*}
$$

where $\left|\nabla^{2} u\right|^{2}:=\nabla^{2} u: \nabla^{2} \bar{u}$. The above notation was introduced in (2.5).
All the spaces given in this document are taken over the field of complex numbers. Let $k$ be a natural number. We start by considering the following

Chapter 3. Well-posedness and regularity of the solutions of a plate-membrane system
space $H_{\Gamma}^{k}\left(\Omega_{1}\right):=\left\{w \in H^{k}\left(\Omega_{1}\right): \frac{\partial^{j} w}{\partial \nu^{j}}=0\right.$ on $\Gamma$ for $\left.j=0, \ldots, k-1\right\}$. For the particular case $k=2$ one has that the space $H_{\Gamma}^{2}\left(\Omega_{1}\right)$ endowed with the inner product

$$
(w, \widetilde{w})_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}:=\mu(\Delta w, \Delta \widetilde{w})_{L^{2}\left(\Omega_{1}\right)}+(1-\mu)\left(\nabla^{2} w, \nabla^{2} \widetilde{w}\right)_{L^{2}\left(\Omega_{1}\right)^{4}}
$$

where

$$
\begin{equation*}
\left(\nabla^{2} u, \nabla^{2} v\right)_{L^{2}\left(\Omega_{1}\right)^{4}}:=\int_{\Omega_{1}} \nabla^{2} u: \nabla^{2} \bar{v} d x \tag{3.2}
\end{equation*}
$$

is a Hilbert space. Indeed, if $u, v \in H_{\Gamma}^{2}\left(\Omega_{1}\right)$ then

$$
\begin{align*}
(u, v)_{H^{2}\left(\Omega_{1}\right)}= & \sum_{|\alpha| \leq 2} \int_{\Omega_{1}} \partial^{\alpha} u \partial^{\alpha} \bar{v} d x=(u, v)_{L^{2}\left(\Omega_{1}\right)}+(\nabla u, \nabla v)_{L^{2}\left(\Omega_{1}\right)^{2}}  \tag{3.3}\\
& +\int_{\Omega_{1}} u_{x_{1} x_{1}} \bar{v}_{x_{1} x_{1}}+u_{x_{1} x_{2}} \bar{v}_{x_{1} x_{2}}+u_{x_{2} x_{2}} \bar{v}_{x_{2} x_{2}} d x .
\end{align*}
$$

Due to $u=\partial_{\nu} u=0$ on $\Gamma$, it is easy to see that $\nabla u=0$ on $\Gamma$ and therefore $u_{x_{1}}=u_{x_{2}}=0$ on $\Gamma$. By Friedrichs inequality or Corollary 5.5 in [17],

$$
\begin{align*}
\|\nabla u\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2} & =\left\|u_{x_{1}}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\left\|u_{x_{2}}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} \\
& \leq C^{2}\left(\left\|\nabla u_{x_{1}}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\left\|\nabla u_{x_{2}}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}\right)  \tag{3.4}\\
& \leq 2 C^{2}\left\|\nabla^{2} u\right\|_{L^{2}\left(\Omega_{1}\right)^{4}}^{2} .
\end{align*}
$$

By (3.2)-(3.4) and again Friedrichs inequality,

$$
\begin{equation*}
\|u\|_{H^{2}\left(\Omega_{1}\right)}^{2} \leq\|u\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\|\nabla u\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\left\|\nabla^{2} u\right\|_{L^{2}\left(\Omega_{1}\right)^{4}}^{2} \leq \widetilde{C}\|u\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2} . \tag{3.5}
\end{equation*}
$$

Hölder's inequality allows us to write

$$
\begin{aligned}
\|u\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}= & \mu \int_{\Omega_{1}}\left|u_{x_{1} x_{1}}\right|^{2}+2 \operatorname{Re}\left(u_{x_{1} x_{1}} \bar{u}_{x_{2} x_{2}}\right)+\left|u_{x_{2} x_{2}}\right|^{2} d x \\
& +(1-\mu) \int_{\Omega_{1}}\left|u_{x_{1} x_{1}}\right|^{2}+\left|u_{x_{2} x_{2}}\right|^{2}+2\left|u_{x_{1} x_{2}}\right|^{2} d x \leq 4\|u\|_{H^{2}\left(\Omega_{1}\right)}^{2}
\end{aligned}
$$

This shows that $\|\cdot\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}$ and $\|\cdot\|_{H^{2}\left(\Omega_{1}\right)}$ are equivalent norms on $H_{\Gamma}^{2}\left(\Omega_{1}\right)$. Theorem 2.7 guarantees that $H_{\Gamma}^{2}\left(\Omega_{1}\right)$ is a closed subspace of $H^{2}\left(\Omega_{1}\right)$.

Let $\eta:=\frac{\gamma}{\rho_{1}}$. If the inertial term is present on the plate, that is $\gamma>0$, then $H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)$ will be the space $H_{\Gamma}^{1}\left(\Omega_{1}\right)$ and when $\gamma=0$ it will be $L^{2}\left(\Omega_{1}\right)$. For $\eta \geq 0$, the space $H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)$ is endowed with the scalar product

$$
(w, \widetilde{w})_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}:=(w, \widetilde{w})_{L^{2}\left(\Omega_{1}\right)}+\eta(\nabla w, \nabla \widetilde{w})_{L^{2}\left(\Omega_{1}\right)^{2}}
$$

Note that $\|\cdot\|_{H_{\Gamma}^{1}\left(\Omega_{1}\right)}$ and $\|\cdot\|_{H^{1}\left(\Omega_{1}\right)}$ are equivalent norms on the space $H_{\Gamma}^{1}\left(\Omega_{1}\right)$. The continuity of the zero order trace operator $H^{1}\left(\Omega_{1}\right) \ni u \mapsto \gamma_{0} u \in H^{1 / 2}(\Gamma)$ allows us to prove that $H_{\Gamma}^{1}\left(\Omega_{1}\right)$ is a closed subspace in $H^{1}\left(\Omega_{1}\right)$. Therefore, $\left(H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right),(\cdot, \cdot)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}\right)$ is a Hilbert space for any $\eta \geq 0$.

Based on the expression (3.1), the energy space or phase space is defined, which is presented below. For $\eta \geq 0$, we set

$$
\mathscr{H}_{\eta}:=\left\{\varphi=\left(\varphi_{1}, \varphi_{2}, \varphi_{3}, \varphi_{4}, \varphi_{5}\right)^{\top} \in \mathscr{X}_{\eta}: \varphi_{1}=\varphi_{3} \text { on } I\right\},
$$

where $\mathscr{X}_{\eta}:=H_{\Gamma}^{2}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right) \times H^{1}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{1}\right)$. The energy space is equipped with the inner product

$$
\begin{aligned}
(\varphi, \psi)_{\mathscr{H}_{\eta}}:= & \beta_{1}\left(\varphi_{1}, \psi_{1}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}+\rho_{1}\left(\varphi_{2}, \psi_{2}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+\beta_{2}\left(\nabla \varphi_{3}, \nabla \psi_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& +\rho_{2}\left(\varphi_{4}, \psi_{4}\right)_{L^{2}\left(\Omega_{2}\right)}+\rho_{0}\left(\varphi_{5}, \psi_{5}\right)_{L^{2}\left(\Omega_{1}\right)}
\end{aligned}
$$

for all $\varphi, \psi \in \mathscr{H}_{\eta}$. It is easy to see that the operator $T: \mathscr{X}_{\eta} \rightarrow H^{1 / 2}(I)$ given by $T\left(\varphi_{1}, \varphi_{2}, \varphi_{3}, \varphi_{4}, \varphi_{5}\right):=\varphi_{1}-\left.\varphi_{3}\right|_{I}$ is linear, and also continuous thanks to the trace theorem. As $\mathscr{H}_{\eta}=T^{-1}(\{0\})$, then $\mathscr{H}_{\eta}$ is closed in $\mathscr{X}_{\eta}$. By Theorem 2.24 and the trace theorem,

$$
\left\|\varphi_{3}\right\|_{H^{1}\left(\Omega_{2}\right)}^{2} \leq C\left(\left\|\varphi_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+\left\|\nabla \varphi_{3}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}\right)
$$

for $\varphi=\left(\varphi_{1}, \varphi_{2}, \varphi_{3}, \varphi_{4}, \varphi_{5}\right) \in \mathscr{H}_{\eta}$. This allows us to prove that there is $C>0$ such that $\|\varphi\|_{\mathscr{X}_{\eta}} \leq C\|\varphi\|_{\mathscr{H}_{\eta}}$ for any $\varphi \in \mathscr{H}_{\eta}$. From the definition of the space $\mathscr{X}_{\eta}$ follows immediately the inequality $\|\varphi\|_{\mathscr{H}_{\eta}} \leq C\|\varphi\|_{\mathscr{X}_{\eta}}$ for all $\varphi \in \mathscr{H}_{\eta}$. In consequence, $\left(\mathscr{H}_{\eta},(\cdot, \cdot)_{\mathscr{H}_{\eta}}\right)$ is a Hilbert space.

### 3.1.1 Basic operators

At this point in the document, the contents of Section 2.6 and Section 2.7 are necessary to obtain properties of the operators that will be part of the abstract formulation of the problem (1.1)-(1.9). The following definition is taken from Subsection 7.1 of [29].

Definition 3.1. Let $H$ be a Hilbert space. An unbounded linear operator $A: D(A) \subset H \rightarrow H$ is said to be monotone if and only if $(A u, u)_{H} \geq 0$ for all $u \in D(A)$. If also $R(\mathcal{I}+A)=H$, then $A$ is called maximal monotone.

Throughout this document $\mathcal{I}$ denotes the identity operator. The following two results will allows us to show that certain operators that are of interest to us are self-adjoint.

Lemma 3.2. Let $H$ be a Hilbert space and $A: D(A) \subset H \rightarrow H$ a monotone operator. If there exists $\lambda_{0}>0$ such that $R\left(\lambda_{0} \mathcal{I}+A\right)=H$, then $A$ is a maximal monotone operator.

Proof. Let $B:=\frac{1}{\lambda_{0}} A$. Note that $B: D(B) \subset H \rightarrow H$ is an unbounded linear operator with $D(B):=D(A)$. As $(B u, u)_{H} \geq 0$ for all $u \in D(B)$ and $R(\mathcal{I}+B)=H$, we have that $B$ is a maximal monotone operator. The part (c) of Proposition 7.1 from [29] implies $R(\mathcal{I}+\lambda B)=H$ for every $\lambda>0$. In particular, taking $\lambda=\lambda_{0}$ we get $R(\mathcal{I}+A)=H$.

Lemma 3.3 (cf. [29, Proposition 7.6]). Let $A: D(A) \subset H \rightarrow H$ be a maximal monotone symmetric operator. Then $A$ is self-adjoint.

Proposition 3.4. If $\mathcal{A}_{B}: L^{2}\left(\Omega_{1}\right) \supset D\left(\mathcal{A}_{B}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ is the Bi-Laplacian, this is $\mathcal{A}_{B}:=\Delta^{2}$, with domain
$D\left(\mathcal{A}_{B}\right):=\left\{w \in H^{4}\left(\Omega_{1}\right): w=\partial_{\nu} w=0\right.$ on $\Gamma$ and $\mathscr{B}_{1} w=\mathscr{B}_{2} w=0$ on $\left.I\right\}$,
then $\mathcal{A}_{B}$ is a positive self-adjoint operator.
Proof. By Proposition 2.17, we have $\left(\mathcal{A}_{B} w, w\right)_{L^{2}\left(\Omega_{1}\right)}=\|w\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2} \geq 0$ for all $w \in D\left(\mathcal{A}_{B}\right)$. The operator $\mathcal{A}_{B}$ is maximal monotone. In fact: According to Lemma 3.2 it is enough to prove that $R\left(\lambda \mathcal{I}+\mathcal{A}_{B}\right)=L^{2}\left(\Omega_{1}\right)$ for some $\lambda>0$, this is, there is $\lambda>0$ such that for any $f \in L^{2}\left(\Omega_{1}\right)$ there exists $w \in D\left(\mathcal{A}_{B}\right)$ so that $\left(\lambda \mathcal{I}+\mathcal{A}_{B}\right) w=f$ and this in turn is equivalent to the following formulation

$$
\left\{\begin{array}{l}
\text { Given } f \in L^{2}\left(\Omega_{1}\right) \text { there exists } w \in H^{4}\left(\Omega_{1}\right) \text { such that }  \tag{3.6}\\
\lambda w+\Delta^{2} w=f \text { in } \Omega_{1}, \\
w=\partial_{\nu} w=0 \text { on } \Gamma, \\
\mathscr{B}_{1} w=\mathscr{B}_{2} w=0 \text { on } I .
\end{array}\right.
$$

Proposition 2.60 guarantees the existence of $\lambda>0$ so that 3.6 holds. It is immediate from formula (2.4) that $\left(\mathcal{A}_{B} w, \widetilde{w}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(w, \mathcal{A}_{B} \widetilde{w}\right)_{L^{2}\left(\Omega_{1}\right)}$ for all $w, \widetilde{w} \in D\left(\mathcal{A}_{B}\right)$. Thus, $\mathcal{A}_{B}$ is a symmetric operator. Due to Lemma 3.3, $\mathcal{A}_{B}$ is a self-adjoint operator. Estimation (3.5) implies the existence of $C>0$ such that $\left(\mathcal{A}_{B} w, w\right)_{L^{2}\left(\Omega_{1}\right)} \geq C\|w\|_{L^{2}\left(\Omega_{1}\right)}^{2}$ for all $w \in D\left(\mathcal{A}_{B}\right)$.

The operator $\mathcal{A}_{B}^{s}: D\left(\mathcal{A}_{B}^{s}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ is a positive self-adjoint operator for any $s>0$, see Section 2.6 and Proposition 3.4. Recall that $D\left(\mathcal{A}_{B}^{s}\right)$ is endowed with the scalar product $(w, \widetilde{w})_{D\left(\mathcal{A}_{B}^{s}\right)}:=\left(\mathcal{A}_{B}^{s} w, \mathcal{A}_{B}^{s} \widetilde{w}\right)_{L^{2}\left(\Omega_{1}\right)}$ for $w, \widetilde{w} \in D\left(\mathcal{A}_{B}^{s}\right)$, see (2.24), and becomes a Hilbert space. For $s_{1}>s_{2}>0$, we have

$$
\begin{equation*}
D\left(\mathcal{A}_{B}^{s_{1}}\right) \stackrel{d}{\hookrightarrow} D\left(\mathcal{A}_{B}^{s_{2}}\right) \stackrel{d}{\hookrightarrow} L^{2}\left(\Omega_{1}\right), \tag{3.7}
\end{equation*}
$$

see part $b$ ) of Theorem 2.53. Characterization (2.29) allows us to explicitly write the domain of some fractional powers of the operator $\mathcal{A}_{B}$. Hereinafter, the symbol $\cong$ means equality between sets and equivalence in norm. We have,

$$
\begin{align*}
& D\left(\mathcal{A}_{B}^{1 / 4}\right) \cong H_{\Gamma}^{1}\left(\Omega_{1}\right) \\
& D\left(\mathcal{A}_{B}^{1 / 2}\right) \cong H_{\Gamma}^{2}\left(\Omega_{1}\right)  \tag{3.8}\\
& D\left(\mathcal{A}_{B}^{3 / 4}\right) \cong\left\{w \in H^{3}\left(\Omega_{1}\right) \cap H_{\Gamma}^{2}\left(\Omega_{1}\right): \mathscr{B}_{1} w=0 \text { on } I\right\}
\end{align*}
$$

Due to (3.7), we have that $D\left(\mathcal{A}_{B}\right)$ is dense in $D\left(\mathcal{A}_{B}^{1 / 2}\right)$. If $\xi, \zeta \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$, there exist two sequences $\left(\xi_{n}\right)_{n \in \mathbb{N}}$ and $\left(\zeta_{n}\right)_{n \in \mathbb{N}}$ contained in $D\left(\mathcal{A}_{B}\right)$ such that $\xi_{n} \rightarrow \xi$ and $\zeta_{n} \rightarrow \zeta$ in $D\left(\mathcal{A}_{B}^{1 / 2}\right)$. Taking into account (2.12), part a) of Theorem 2.53 and the fact that $\mathcal{A}_{B}^{1 / 2}$ is a self-adjoint operator, we obtain

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \xi_{n}, \zeta_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(\mathcal{A}_{B} \xi_{n}, \zeta_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\mathcal{A}_{B}^{1 / 2} \xi_{n}, \mathcal{A}_{B}^{1 / 2} \zeta_{n}\right)_{L^{2}\left(\Omega_{1}\right)} \tag{3.9}
\end{equation*}
$$

for each $n \in \mathbb{N}$. Note that $\mathcal{A}_{B} \xi \in\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$, see part $\left.c\right)$ of Theorem 2.53. We compute that

$$
\begin{aligned}
& \left|\left\langle\mathcal{A}_{B} \xi_{n}, \zeta_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}-\left\langle\mathcal{A}_{B} \xi, \zeta\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}\right| \\
& =\left|\left\langle\mathcal{A}_{B} \xi_{n}, \zeta_{n}-\zeta\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}+\left\langle\mathcal{A}_{B}\left(\xi_{n}-\xi\right), \zeta\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}\right| \\
& \leq\left\|\mathcal{A}_{B} \xi_{n}\right\|_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}}\left\|\zeta_{n}-\zeta\right\|_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}+\left\|\mathcal{A}_{B}\left(\xi_{n}-\xi\right)\right\|_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}}\|\zeta\|_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}
\end{aligned}
$$

Now the continuity of the operator $\mathcal{A}_{B}: D\left(\mathcal{A}_{B}^{1 / 2}\right) \rightarrow\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$ implies that

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \xi_{n}, \zeta_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} \rightarrow\left\langle\mathcal{A}_{B} \xi, \zeta\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} \tag{3.10}
\end{equation*}
$$

By Proposition 2.17 and the equivalence of norms of the second characterization in (3.8), we get the convergence

$$
\begin{equation*}
\left(\mathcal{A}_{B} \xi_{n}, \zeta_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\xi_{n}, \zeta_{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow(\xi, \zeta)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \tag{3.11}
\end{equation*}
$$

We find that

$$
\left(\mathcal{A}_{B}^{1 / 2} \xi_{n}, \mathcal{A}_{B}^{1 / 2} \zeta_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\xi_{n}, \zeta_{n}\right)_{D\left(\mathcal{A}_{B}^{1 / 2}\right)} \rightarrow(\xi, \zeta)_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(\mathcal{A}_{B}^{1 / 2} \xi, \mathcal{A}_{B}^{1 / 2} \zeta\right)_{L^{2}\left(\Omega_{1}\right)}
$$

By (3.9)-(3.11) and the last limit, we can write the following equalities

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \xi, \zeta\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(\mathcal{A}_{B}^{1 / 2} \xi, \mathcal{A}_{B}^{1 / 2} \zeta\right)_{L^{2}\left(\Omega_{1}\right)}=(\xi, \zeta)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \tag{3.12}
\end{equation*}
$$

for all $\xi, \zeta \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$.
Proposition 3.5. The operator $\mathcal{A}_{L}: L^{2}\left(\Omega_{1}\right) \supset D\left(\mathcal{A}_{L}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ given by $\mathcal{A}_{L}:=-\Delta$ is a positive self-adjoint operator, where

$$
D\left(\mathcal{A}_{L}\right):=\left\{w \in H^{2}\left(\Omega_{1}\right): w=0 \text { on } \Gamma \quad \text { and } \quad \partial_{\nu} w=0 \text { on } I\right\} .
$$

Proof. Using the integration by parts formula (2.1) and Friedrichs inequality, one has that $\left(\mathcal{A}_{L} u, u\right)_{L^{2}\left(\Omega_{1}\right)} \geq C\|u\|_{L^{2}\left(\Omega_{1}\right)}^{2}$ for any $u \in D\left(\mathcal{A}_{L}\right)$. Thus, $\mathcal{A}_{L}$ is monotone. By elliptic regularity theory (see, e.g., Theorem 3.2 and Remark 3.3 in [14]) one has that for sufficiently large $\lambda>0$, the solution $u$ of the boundary value problem

$$
\left\{\begin{array}{l}
\lambda u-\Delta u=f \text { in } \Omega_{1} \\
u=0 \text { on } \Gamma \\
\partial_{\nu} u=0 \text { on } I
\end{array}\right.
$$

belongs to $H^{2}\left(\Omega_{1}\right)$ for any $f \in L^{2}\left(\Omega_{1}\right)$. Thus, $\mathcal{A}_{L}$ is maximal monotone and it is easy to see that it is symmetric.

The square root operator $\mathcal{A}_{L}^{1 / 2}$ of $\mathcal{A}_{L}$ is self-adjoint and positive, and its domain has the following characterization:

$$
\begin{equation*}
D\left(\mathcal{A}_{L}^{1 / 2}\right) \cong H_{\Gamma}^{1}\left(\Omega_{1}\right) \tag{3.13}
\end{equation*}
$$

It is clear that $\left(\mathcal{A}_{L} w, \widetilde{w}\right)_{L^{2}\left(\Omega_{1}\right)}=(\nabla w, \nabla \widetilde{w})_{L^{2}\left(\Omega_{1}\right)^{2}} \forall w, \widetilde{w} \in D\left(\mathcal{A}_{L}\right)$. Using the density of $D\left(\mathcal{A}_{L}\right)$ in $D\left(\mathcal{A}_{L}^{1 / 2}\right)$, the equivalence of norms $\|\cdot\|_{H_{\Gamma}^{1}\left(\Omega_{1}\right)} \sim\|\cdot\|_{H^{1}\left(\Omega_{1}\right)}$ on $H_{\Gamma}^{1}\left(\Omega_{1}\right),(3.13)$ and an argument as in the previous page, we derive

$$
\begin{equation*}
\left\langle\mathcal{A}_{L} \xi, \zeta\right\rangle_{\left[D\left(\mathcal{A}_{L}^{1 / 2}\right)\right]^{\times} \times D\left(\mathcal{A}_{L}^{1 / 2}\right)}=\left(\mathcal{A}_{L}^{1 / 2} \xi, \mathcal{A}_{L}^{1 / 2} \zeta\right)_{L^{2}\left(\Omega_{1}\right)}=(\nabla \xi, \nabla \zeta)_{L^{2}\left(\Omega_{1}\right)^{2}} \tag{3.14}
\end{equation*}
$$

for all $\xi, \zeta \in D\left(\mathcal{A}_{L}^{1 / 2}\right)$.
Now, we define the inertia operator $\mathcal{M}_{\eta}:=\mathcal{I}+\eta \mathcal{A}_{L}$ with $D\left(\mathcal{M}_{\eta}\right):=$ $D\left(\mathcal{A}_{L}\right)$ if $\eta>0$ and $D\left(\mathcal{M}_{0}\right):=L^{2}\left(\Omega_{1}\right)$. For $\xi \in D\left(\mathcal{A}_{L}^{1 / 2}\right)$, we have that the part $c$ ) of Theorem 2.53 implies $\mathcal{A}_{L} \xi \in\left[D\left(\mathcal{A}_{L}^{1 / 2}\right)\right]^{\prime}$ and in consequence $\mathcal{M}_{\eta} \xi=\xi+\eta \mathcal{A}_{L} \xi \in H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right):=\left[H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)\right]^{\prime}$. From 2.12, 3.13) and 3.14) it follows that

$$
\begin{equation*}
\left\langle\mathcal{M}_{\eta} \xi, \zeta\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}=(\xi, \zeta)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \text { for all } \xi, \zeta \in H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right) . \tag{3.15}
\end{equation*}
$$

By the Lax-Milgram theorem, see [40, Theorem 7 on p. 368], we obtain that $\mathcal{M}_{\eta}: H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right) \rightarrow H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right)$ is an isomorphism. In particular, $\mathcal{M}_{\eta}^{-1}$ exists for $\eta \geq 0$, where $\mathcal{M}_{0}^{-1}$ is the identity in $L^{2}\left(\Omega_{1}\right)$.

Proposition 3.6. The operator $\mathcal{A}_{T}: L^{2}\left(\Omega_{1}\right) \supset D\left(\mathcal{A}_{T}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ defined by $\mathcal{A}_{T}:=-\Delta+\frac{\sigma}{\beta} \mathcal{I}$, with domain

$$
D\left(\mathcal{A}_{T}\right):=\left\{\vartheta \in H^{2}\left(\Omega_{1}\right): \partial_{\nu} \vartheta+\kappa \vartheta=0 \text { on } \partial \Omega_{1}\right\},
$$

is self-adjoint and positive.
Proof. Using integration by parts, see Theorem 2.11, it is easy to get that

$$
\begin{equation*}
\left(\mathcal{A}_{T} \vartheta, \vartheta\right)_{L^{2}\left(\Omega_{1}\right)}=\|\nabla \vartheta\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\kappa\|\vartheta\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2}+\frac{\sigma}{\beta}\|\vartheta\|_{L^{2}\left(\Omega_{1}\right)}^{2} \tag{3.16}
\end{equation*}
$$

for all $\vartheta \in D\left(\mathcal{A}_{T}\right)$. Hence, $\left(\mathcal{A}_{T} \widetilde{\vartheta}, \widetilde{\vartheta}\right)_{L^{2}\left(\Omega_{1}\right)} \geq C\|\widetilde{\vartheta}\|_{L^{2}\left(\Omega_{1}\right)}^{2}$ for any $\widetilde{\vartheta} \in D\left(\mathcal{A}_{T}\right)$, which is trivially true if $\sigma>0$ and when $\sigma=0$ is also true thanks to Theorem 2.24 since $\kappa>0$. By Theorem 3.2 and part (b) of Remark 3.3 in [14]: For sufficiently large $\lambda_{0}>0$, the boundary value problem

$$
\begin{aligned}
\lambda_{0} u-\Delta u & =f \text { in } \Omega_{1}, \\
\partial_{\nu} u+\kappa u=0 & \text { on } \partial \Omega_{1},
\end{aligned}
$$

has a unique solution $u \in H^{2}\left(\Omega_{1}\right)$ whenever $f \in L^{2}\left(\Omega_{1}\right)$. This allows to assert that $R\left(\lambda \mathcal{I}+\mathcal{A}_{T}\right)=L^{2}\left(\Omega_{1}\right)$ for some $\lambda>0$. It is very simple to check that $\mathcal{A}_{T}$ is symmetric. An application of Lemma 3.2 and Lemma 3.3 completes the proof.

Chapter 3. Well-posedness and regularity of the solutions of a plate-membrane system

The same arguments used for the operators $\mathcal{A}_{B}$ and $\mathcal{A}_{L}$ are considered for the operator $\mathcal{A}_{T}$ to obtain the following characterization

$$
\begin{equation*}
D\left(\mathcal{A}_{T}^{1 / 2}\right) \cong H^{1}\left(\Omega_{1}\right) \tag{3.17}
\end{equation*}
$$

Therefore, one can establish via density and integration by parts the identity

$$
\begin{equation*}
(\xi, \zeta)_{D\left(\mathcal{A}_{T}^{1 / 2}\right)}=(\nabla \xi, \nabla \zeta)_{L^{2}\left(\Omega_{1}\right)^{2}}+\kappa(\xi, \zeta)_{L^{2}\left(\partial \Omega_{1}\right)}+\frac{\sigma}{\beta}(\xi, \zeta)_{L^{2}\left(\Omega_{1}\right)} \tag{3.18}
\end{equation*}
$$

for all $\xi, \zeta \in D\left(\mathcal{A}_{T}^{1 / 2}\right)$.

### 3.1.2 Abstract formulation of the problem

The operators $\mathcal{A}_{B}, \mathcal{A}_{L}, \mathcal{A}_{T}$ and $\mathcal{M}_{\eta}$ together with others, will be part of a matrix operator which will allow to write system (1.1)-(1.9) as a Cauchy problem on the Hilbert space $\mathscr{H}_{\eta}$. We will make use of two Green maps introduced in [10, p. 158]. Let $\mathcal{G}_{1}$ and $\mathcal{G}_{2}$ be defined by the formulas $\mathcal{G}_{1} x:=\widetilde{u}$ and $\mathcal{G}_{2} y:=\widetilde{v}$, where $\widetilde{u}$ and $\widetilde{v}$ solve the following problems

$$
\left\{\begin{array} { l } 
{ \Delta ^ { 2 } \widetilde { u } = 0 \text { in } \Omega _ { 1 } , }  \tag{3.19}\\
{ \widetilde { u } = \partial _ { \nu } \widetilde { u } = 0 \text { on } \Gamma , } \\
{ \mathscr { B } _ { 1 } \widetilde { u } = x \text { on } I , } \\
{ \mathscr { B } _ { 2 } \widetilde { u } = 0 \text { on } I , }
\end{array} \quad \text { and } \quad \left\{\begin{array}{l}
\Delta^{2} \widetilde{v}=0 \text { in } \Omega_{1}, \\
\widetilde{v}=\partial_{\nu} \widetilde{v}=0 \text { on } \Gamma, \\
\mathscr{B}_{1} \widetilde{v}=0 \text { on } I, \\
\mathscr{B}_{2} \widetilde{v}=y \text { on } I .
\end{array}\right.\right.
$$

We define the Neumann map $\mathcal{N}$ (introduced in [11, p. 405]) given by $\mathcal{N} z:=\widetilde{w}$ where $\widetilde{w}$ is solution of the problem

$$
\left\{\begin{array}{l}
\Delta \widetilde{w}=0 \text { in } \Omega_{1}  \tag{3.20}\\
\widetilde{w}=0 \text { on } \Gamma \\
\partial_{\nu} \widetilde{w}=z \text { on } I
\end{array}\right.
$$

From elliptic theory, see Section 7.3 of Chapter 2 in [89], we obtain that $\mathcal{N}: H^{s}(I) \rightarrow H^{s+\frac{3}{2}}\left(\Omega_{1}\right)$ and $\mathcal{G}_{i}: H^{s}(I) \rightarrow H^{s+\frac{3}{2}+i}\left(\Omega_{1}\right)$, with $i=1,2$, are bounded linear maps for any $s \geq 0$.

Lemma 3.7. For all $f \in D\left(\mathcal{A}_{B}\right)$, we have that a) $\mathcal{G}_{1}^{*} \mathcal{A}_{B} f=\gamma_{1} f$ on $I$ and b) $\mathcal{G}_{2}^{*} \mathcal{A}_{B} f=-\gamma_{0} f$ on $I$, where $\gamma_{0} w:=\left.w\right|_{I}, \gamma_{1} w:=\left.\partial_{\nu} w\right|_{I}$ and $\mathcal{G}_{i}^{*}$ denotes the adjoint of $\mathcal{G}_{i}$ in the sense

$$
\left(\mathcal{G}_{i} \varphi, \psi\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\varphi, \mathcal{G}_{i}^{*} \psi\right)_{L^{2}(I)} \quad \text { for } \varphi \in L^{2}(I) \quad \text { and } \quad \psi \in L^{2}\left(\Omega_{1}\right)
$$

Proof. Let $f \in D\left(\mathcal{A}_{B}\right)$ and $g \in L^{2}(I)$. Note that $\mathcal{G}_{1} g \in H^{5 / 2}\left(\Omega_{1}\right)$. Taking into account 2.12 and the integration formula of Proposition 2.18, we obtain

$$
\begin{aligned}
& \left(\Delta^{2} f, \mathcal{G}_{1} g\right)_{L^{2}\left(\Omega_{1}\right)}-\left(f, \Delta^{2} \mathcal{G}_{1} g\right)_{L^{2}\left(\Omega_{1}\right)}=-\left(\mathscr{B}_{1} f, \partial_{\nu} \mathcal{G}_{1} g\right)_{L^{2}(I)} \\
& +\left(\mathscr{B}_{2} f, \mathcal{G}_{1} g\right)_{L^{2}(I)}+\left(\partial_{\nu} f, \mathscr{B}_{1} \mathcal{G}_{1} g\right)_{L^{2}(I)}-\left(f, \mathscr{B}_{2} \mathcal{G}_{1} g\right)_{L^{2}(I)}
\end{aligned}
$$

On account of $\Delta^{2} \mathcal{G}_{1} g=0$ in $\Omega_{1}$ and $\mathscr{B}_{1} f=\mathscr{B}_{2} f=\mathscr{B}_{2} \mathcal{G}_{1} g=0$ on $I$, we deduce

$$
\left(\mathcal{G}_{1}^{*} \mathcal{A}_{B} f, g\right)_{L^{2}(I)}=\left(\mathcal{A}_{B} f, \mathcal{G}_{1} g\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\Delta^{2} f, \mathcal{G}_{1} g\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\gamma_{1} f, g\right)_{L^{2}(I)}
$$

This last one shows a). The proof of $b$ ) is similar.
Proposition 3.8. For any $w \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and $g \in L^{2}(I)$, we have

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \mathcal{G}_{i} g, w\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(g,(-1)^{i-1} \gamma_{2-i} w\right)_{L^{2}(I)} \tag{3.21}
\end{equation*}
$$

for $i=1,2$.
Proof. Let $w \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and $g \in L^{2}(I)$. As $D\left(\mathcal{A}_{B}\right)$ is dense in $D\left(\mathcal{A}_{B}^{1 / 2}\right)$, then there is $\left(w_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathcal{A}_{B}\right)$ such that $\lim _{n \rightarrow \infty} w_{n}=w$ in $D\left(\mathcal{A}_{B}^{1 / 2}\right)$. For any $n \in \mathbb{N}$, the part a) of Lemma 3.7 implies

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \mathcal{G}_{1} g, w_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(\mathcal{G}_{1} g, \mathcal{A}_{B} w_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(g, \gamma_{1} w_{n}\right)_{L^{2}(I)} \tag{3.22}
\end{equation*}
$$

As $\mathcal{G}_{1} g, w_{n} \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$, then identity (3.12) allows us to write the convergence

$$
\begin{align*}
& \lim _{n \rightarrow \infty}\left\langle\mathcal{A}_{B} \mathcal{G}_{1} g, w_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\lim _{n \rightarrow \infty}\left(\mathcal{G}_{1} g, w_{n}\right)_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}  \tag{3.23}\\
& =\left(\mathcal{G}_{1} g, w\right)_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left\langle\mathcal{A}_{B} \mathcal{G}_{1} g, w\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}
\end{align*}
$$

The continuous embedding $H^{1 / 2}(I) \hookrightarrow L^{2}(I)$, see Theorem 2.6, and the continuity of the trace operator $\gamma_{1}: H^{2}\left(\Omega_{1}\right) \rightarrow H^{1 / 2}(I)$ imply

$$
\left\|\gamma_{1} w_{n}-\gamma_{1} w\right\|_{L^{2}(I)} \leq C\left\|w_{n}-w\right\|_{H^{2}\left(\Omega_{1}\right)} \leq C\left\|w_{n}-w\right\|_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}
$$

because $\|\cdot\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}$ and $\|\cdot\|_{H^{2}\left(\Omega_{1}\right)}$ are equivalent norms on $H_{\Gamma}^{2}\left(\Omega_{1}\right)$. Therefore, $\lim _{n \rightarrow \infty} \gamma_{1} w_{n}=\gamma_{1} w$ in $L^{2}(I)$. Thus,

$$
\begin{equation*}
\lim _{n \rightarrow \infty}\left(g, \gamma_{1} w_{n}\right)_{L^{2}(I)}=\left(g, \gamma_{1} w\right)_{L^{2}(I)} \tag{3.24}
\end{equation*}
$$

From (3.22)-(3.24) it follows (3.21) when $i=1$. Using the same arguments as before, (3.21) is proved when $i=2$.

Lemma 3.9. The $L^{2}\left(\Omega_{1}\right)$-adjoint $\mathcal{N}^{*}$ of the Neumann map $\mathcal{N}$ is given by $\left(\varphi, \mathcal{N}^{*} \psi\right)_{L^{2}(I)}=(\mathcal{N} \varphi, \psi)_{L^{2}\left(\Omega_{1}\right)}$ for $\varphi \in L^{2}(I)$ and $\psi \in L^{2}\left(\Omega_{1}\right)$. Then, for any $f \in D\left(\mathcal{A}_{L}\right)$, we have

$$
\begin{equation*}
\mathcal{N}^{*} \mathcal{A}_{L} f=\gamma_{0} f . \tag{3.25}
\end{equation*}
$$

Proof. Let $g \in L^{2}(I)$ and $f \in D\left(\mathcal{A}_{L}\right)$. Note that $\mathcal{N} g \in H^{3 / 2}\left(\Omega_{1}\right)$ satisfies $\Delta \mathcal{N} g \in L^{2}\left(\Omega_{1}\right)$. Applying the generalization of Green's first formula (see Theorem 2.14 twice, we may write

$$
\begin{aligned}
\left(\mathcal{N}^{*} \mathcal{A}_{L} f, g\right)_{L^{2}(I)} & =\left(\mathcal{A}_{L} f, \mathcal{N} g\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =(\nabla f, \nabla \mathcal{N} g)_{L^{2}\left(\Omega_{1}\right)^{2}}-\left(\partial_{\nu} f, \mathcal{N} g\right)_{L^{2}(\Gamma)}-\left(\partial_{\nu} f, \mathcal{N} g\right)_{L^{2}(I)} \\
& =-(f, \Delta \mathcal{N} g)_{L^{2}\left(\Omega_{1}\right)}+\left(f, \partial_{\nu} \mathcal{N} g\right)_{L^{2}(\Gamma)}+\left(f, \partial_{\nu} \mathcal{N} g\right)_{L^{2}(I)} \\
& =(f, g)_{L^{2}(I)} .
\end{aligned}
$$

Above we have considered the definitions of $D\left(\mathcal{A}_{L}\right)$ and $\mathcal{N}$, see Proposition 3.5 and (3.20). Therefore, (3.25) is true.

Proposition 3.10. For all $w \in D\left(\mathcal{A}_{L}^{1 / 2}\right)$ and $g \in L^{2}(I)$, it holds

$$
\begin{equation*}
\left\langle\mathcal{A}_{L} \mathcal{N} g, w\right\rangle_{\left[D\left(\mathcal{A}_{L}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{L}^{1 / 2}\right)}=\left(g, \gamma_{0} w\right)_{L^{2}(I)} \tag{3.26}
\end{equation*}
$$

Proof. Similar to the proof of Proposition 3.8. Here we use the fact that $D\left(\mathcal{A}_{L}\right)$ is dense in $D\left(\mathcal{A}_{L}^{1 / 2}\right)$ and apply Lemma 3.9, also consider (3.14).

Proposition 3.11. For $g \in L^{2}(I)$, we have

$$
\mathcal{A}_{B} \mathcal{G}_{2} g+\mathcal{A}_{L} \mathcal{N} g=0 \quad \text { in } \quad H_{\Gamma}^{-1}\left(\Omega_{1}\right)
$$

Proof. Let $g \in L^{2}(I)$. Note that $\mathcal{G}_{2} g \in D\left(\mathcal{A}_{B}^{3 / 4}\right)$ and $\mathcal{N} g \in D\left(\mathcal{A}_{L}^{1 / 2}\right)$. So, $\mathcal{A}_{B}^{3 / 4} \mathcal{G}_{2} g, \mathcal{A}_{L}^{1 / 2} \mathcal{N} g \in L^{2}\left(\Omega_{1}\right)$. Parts $\left.a\right)$ and $c$ ) of Theorem 2.53, and the first characterization of (3.8) imply

$$
\mathcal{A}_{B} \mathcal{G}_{2} g=\mathcal{A}_{B}^{1 / 4} \mathcal{A}_{B}^{3 / 4} \mathcal{G}_{2} g \in H_{\Gamma}^{-1}\left(\Omega_{1}\right)
$$

Similar arguments allows to affirm

$$
\mathcal{A}_{L} \mathcal{N} g \in H_{\Gamma}^{-1}\left(\Omega_{1}\right)
$$

Let $f \in D\left(\mathcal{A}_{B}^{1 / 4}\right)$. Now, we will argue analogously as in the proof of Proposition 3.8. The density of $D\left(\mathcal{A}_{B}\right)$ in $D\left(\mathcal{A}_{B}^{1 / 4}\right)$ implies the existence of a sequence
$\left(f_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathcal{A}_{B}\right)$ such that $\lim _{n \rightarrow \infty} f_{n}=f$ in $D\left(\mathcal{A}_{B}^{1 / 4}\right)$. Using the fact that $\mathcal{A}_{B}^{3 / 4}$ is self-adjoint and part b) of Lemma 3.7, we get that

$$
\left\langle\mathcal{A}_{B} \mathcal{G}_{2} g, f_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 4}\right)}=\left(\mathcal{A}_{B}^{3 / 4} \mathcal{G}_{2} g, \mathcal{A}_{B}^{1 / 4} f_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(g,-\gamma_{0} f_{n}\right)_{L^{2}(I)}
$$

for all $n \in \mathbb{N}$. Now, putting $\mathcal{A}_{B}^{1 / 4} \mathcal{A}_{B}^{1 / 2}$ instead of $\mathcal{A}_{B}^{3 / 4}$ below, we obtain that

$$
\left\langle\mathcal{A}_{B} \mathcal{G}_{2} g, f_{n}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 4}\right)}=\left(\mathcal{A}_{B}^{3 / 4} \mathcal{G}_{2} g, \mathcal{A}_{B}^{1 / 4} f_{n}\right)_{L^{2}\left(\Omega_{1}\right)}=\left(\mathcal{A}_{B}^{1 / 2} \mathcal{G}_{2} g, f_{n}\right)_{D\left(\mathcal{A}_{B}^{1 / 4}\right)}
$$

for any $n \in \mathbb{N}$. Taking limit as $n \rightarrow \infty$ in the two equalities above, we have

$$
\begin{equation*}
\left\langle\mathcal{A}_{B} \mathcal{G}_{2} g, f\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 4}\right)}=\left(g,-\gamma_{0} f\right)_{L^{2}(I)} \tag{3.27}
\end{equation*}
$$

To end, let $w \in H_{\Gamma}^{1}\left(\Omega_{1}\right)$. An application of identities (3.26) and (3.27) together with the characterizations (3.8), the first of them, and (3.13) allows us to write the following equalities

$$
\begin{aligned}
& \left\langle\mathcal{A}_{B} \mathcal{G}_{2} g+\mathcal{A}_{L} \mathcal{N} g, w\right\rangle_{H_{\Gamma}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma}^{1}\left(\Omega_{1}\right)} \\
& =\left\langle\mathcal{A}_{B} \mathcal{G}_{2} g, w\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 4}\right)}+\left\langle\mathcal{A}_{L} \mathcal{N} g, w\right\rangle_{\left[D\left(\mathcal{A}_{L}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{L}^{1 / 2}\right)} \\
& =\left(g,-\gamma_{0} w\right)_{L^{2}(I)}+\left(g, \gamma_{0} w\right)_{L^{2}(I)}=0 .
\end{aligned}
$$

This shows what we wanted.
Now, let $u, \theta: \mathbb{R}^{+} \times \overline{\Omega_{1}} \rightarrow \mathbb{C}$ and $v: \mathbb{R}^{+} \times \overline{\Omega_{2}} \rightarrow \mathbb{C}$ be three arbitrary functions with sufficient regularity. Taking into account the definition of the operators $\mathcal{G}_{1}$ and $\mathcal{G}_{2}$ in (3.19), it follows in $\Omega_{1}$ that

$$
\begin{align*}
\beta_{1} \Delta^{2} u= & \beta_{1} \Delta^{2} u+\alpha \Delta^{2} \mathcal{G}_{1} \theta-\gamma \Delta^{2} \mathcal{G}_{2} \partial_{\nu} u_{t t}-m_{1} \Delta^{2} \mathcal{G}_{2} \partial_{\nu} u_{t} \\
& +\alpha \Delta^{2} \mathcal{G}_{2} \partial_{\nu} \theta+\beta_{2} \Delta^{2} \mathcal{G}_{2} \partial_{\nu} v+m_{2} \Delta^{2} \mathcal{G}_{2} \partial_{\nu} v_{t}=\mathcal{A}_{B} \Psi, \tag{3.28}
\end{align*}
$$

where $\Psi:=\beta_{1} u+\alpha \mathcal{G}_{1} \theta-\gamma \mathcal{G}_{2} \partial_{\nu} u_{t t}-m_{1} \mathcal{G}_{2} \partial_{\nu} u_{t}+\alpha \mathcal{G}_{2} \partial_{\nu} \theta+\beta_{2} \mathcal{G}_{2} \partial_{\nu} v+m_{2} \mathcal{G}_{2} \partial_{\nu} v_{t}$. The first conditions of (1.4) are equivalent to $\Psi=\partial_{\nu} \Psi=0$ on $\mathbb{R}^{+} \times \Gamma$ and the transmission conditions on the interface (1.6) and (1.7) can be replaced by $\mathscr{B}_{1} \Psi=\mathscr{B}_{2} \Psi=0$ on $\mathbb{R}^{+} \times I$. As $\mathcal{A}_{B}$ is a positive self-adjoint operator, it follows from 2.25 that it has a standard extension $\widetilde{\mathcal{A}}_{B}: L^{2}\left(\Omega_{1}\right) \rightarrow\left[D\left(\mathcal{A}_{B}\right)\right]^{\prime}$. For simplicity of notation, we shall write $\mathcal{A}_{B}$ instead of $\widetilde{\mathcal{A}}_{B}$, with no fear of confusion. From (3.28) it follows that

$$
\begin{align*}
\beta_{1} \Delta^{2} u= & \beta_{1} \mathcal{A}_{B} u+\alpha \mathcal{A}_{B} \mathcal{G}_{1} \theta-\gamma \mathcal{A}_{B} \mathcal{G}_{2} \partial_{\nu} u_{t t}-m_{1} \mathcal{A}_{B} \mathcal{G}_{2} \partial_{\nu} u_{t}  \tag{3.29}\\
& +\alpha \mathcal{A}_{B} \mathcal{G}_{2} \partial_{\nu} \theta+\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \partial_{\nu} v+m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \partial_{\nu} v_{t}
\end{align*}
$$

Chapter 3. Well-posedness and regularity of the solutions of a plate-membrane system
in $\left[D\left(\mathcal{A}_{B}\right)\right]^{\prime}$. Let $u_{t t} \in H^{2}\left(\Omega_{1}\right)$ with $u_{t t}=0$ on $\Gamma$. Note that $\Delta \mathcal{N} \partial_{\nu} u_{t t}=0$ in $\Omega_{1}, \mathcal{N} \partial_{\nu} u_{t t}=0$ on $\Gamma$ and $\partial_{\nu} \mathcal{N} \partial_{\nu} u_{t t}=\partial_{\nu} u_{t t}$ on $I$. Then, $u_{t t}-\mathcal{N} \partial_{\nu} u_{t t}=0$ on $\Gamma$ and $\partial_{\nu}\left(u_{t t}-\mathcal{N} \partial_{\nu} u_{t t}\right)=0$ on $I$. So, $u_{t t}-\mathcal{N} \partial_{\nu} u_{t t} \in D\left(\mathcal{A}_{L}\right)$. Using the standard extension of $\mathcal{A}_{L}$, writing in the last equality below $\mathcal{A}_{L}$ instead of $\widetilde{\mathcal{A}}_{L}$, we get

$$
\begin{equation*}
\Delta u_{t t}=\Delta u_{t t}-\Delta \mathcal{N} \partial_{\nu} u_{t t}=-\mathcal{A}_{L}\left(u_{t t}-\mathcal{N} \partial_{\nu} u_{t t}\right)=-\mathcal{A}_{L} u_{t t}+\mathcal{A}_{L} \mathcal{N} \partial_{\nu} u_{t t} \tag{3.30}
\end{equation*}
$$

in $\left[D\left(\mathcal{A}_{L}\right)\right]^{\prime}$. Let $u_{t} \in H^{2}\left(\Omega_{1}\right)$ with $u_{t}=0$ on $\Gamma$. Reasoning similarly as above,

$$
\begin{equation*}
\Delta u_{t}=-\mathcal{A}_{L} u_{t}+\mathcal{A}_{L} \mathcal{N} \partial_{\nu} u_{t} \text { in }\left[D\left(\mathcal{A}_{L}\right)\right]^{\prime} \tag{3.31}
\end{equation*}
$$

The equalities (3.29-(3.31) together with Proposition 3.11 allow us to rewrite the problem (1.1)-(1.7) as

$$
\left\{\begin{array}{l}
\rho_{1} \mathcal{M}_{\eta} u_{t t}+\beta_{1} \mathcal{A}_{B} u+\alpha \mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} \theta-\alpha \kappa \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} \theta+\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} v  \tag{3.32}\\
+m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} v_{t}+m_{1} \mathcal{A}_{L} u_{t}-\alpha\left(\mathcal{A}_{T}-\frac{\sigma}{\beta} \mathcal{I}\right) \theta=0 \\
\rho_{0} \theta_{t}+\beta \mathcal{A}_{T} \theta+\alpha \mathcal{A}_{L} u_{t}-\alpha \mathcal{A}_{L} \mathcal{N} \partial_{\nu} u_{t}=0 \\
\rho_{2} v_{t t}-\beta_{2} \Delta v-m_{2} \Delta v_{t}=0
\end{array}\right.
$$

If $w:=\left(u, u_{t}, v, v_{t}, \theta\right)^{\top}$ is in an appropriate space, we can write (3.32) together with the initial conditions (1.8) and (1.9) as the following Cauchy problem

$$
\begin{equation*}
\partial_{t} w(t)=\mathscr{A}_{\eta} w(t), t>0, \quad \text { and } w(0)=w_{0} \tag{3.33}
\end{equation*}
$$

where $w_{0}:=\left(u_{0}, u_{1}, v_{0}, v_{1}, \theta_{0}\right)^{\top}$. For $\eta>0$, the operator $\mathscr{A}_{\eta}$ is given by
$\mathscr{A}_{\eta}:=D\left(\begin{array}{ccccc}0 & \mathcal{I} & 0 & 0 & 0 \\ -\beta_{1} \mathcal{A}_{B} & -m_{1} \mathcal{A}_{L} & -\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} & -m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} & \alpha \mathscr{P} \\ 0 & 0 & 0 & \mathcal{I} & 0 \\ 0 & 0 & \beta_{2} \Delta & m_{2} \Delta & 0 \\ 0 & -\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) & 0 & 0 & -\beta \mathcal{A}_{T}\end{array}\right)$,
where $D$ is the diagonal matrix given by $\operatorname{diag}\left(\mathcal{I}, \frac{1}{\rho_{1}} \mathcal{M}_{\eta}^{-1}, \mathcal{I}, \frac{1}{\rho_{2}} \mathcal{I}, \frac{1}{\rho_{0}} \mathcal{I}\right)$ and $\mathscr{P}:=\mathcal{A}_{T}-\frac{\sigma}{\beta} \mathcal{I}-\mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0}+\kappa \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0}$. Now, introducing the term $\mathcal{W}:=$ $\beta_{1} \mathcal{A}_{B} w_{1}+\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{3}+m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{4}+\alpha \mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} w_{5}-\alpha \kappa \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} w_{5}$ allows us to write

$$
\mathscr{A}_{\eta} w=\left(\begin{array}{c}
w_{2} \\
-\frac{1}{\rho_{1}} \mathcal{M}_{\eta}^{-1} \mathcal{W}-\frac{m_{1}}{\rho_{1}} \mathcal{M}_{\eta}^{-1} \mathcal{A}_{L} w_{2}+\frac{\alpha}{\rho_{1}} \mathcal{M}_{\eta}^{-1} \mathcal{A}_{T} w_{5}-\frac{\alpha \sigma}{\beta \rho_{1}} \mathcal{M}_{\eta}^{-1} w_{5} \\
w_{4} \\
\frac{\beta_{2}}{\rho_{2}} \Delta w_{3}+\frac{m_{2}}{\rho_{2}} \Delta w_{4} \\
-\frac{\alpha}{\rho_{0}} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}-\frac{\beta}{\rho_{0}} \mathcal{A}_{T} w_{5}
\end{array}\right),
$$

where $w:=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top}$. The domain of the operator $\mathscr{A}_{\eta}$ is defined in the following way

$$
\begin{aligned}
D\left(\mathscr{A}_{\eta}\right):= & \left\{w=\left(w_{j}\right)_{1, \ldots, 5}^{\top} \in \mathscr{H}_{\eta}: w_{2} \in H_{\Gamma}^{2}\left(\Omega_{1}\right), w_{4} \in H^{1}\left(\Omega_{2}\right), w_{5} \in D\left(\mathcal{A}_{T}\right)\right. \\
& \left.\mathcal{W} \in H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right), \beta_{2} \Delta w_{3}+m_{2} \Delta w_{4} \in L^{2}\left(\Omega_{2}\right) \text { and } w_{2}=w_{4} \text { on } I\right\} .
\end{aligned}
$$

Remark 3.12. If $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$, then $\mathscr{A}_{\eta} w \in \mathscr{H}_{\eta}$. In effect, since $w_{2} \in H_{\Gamma}^{1}\left(\Omega_{1}\right)$ we get from (3.13) and part $\left.c\right)$ of Theorem 2.53 that $\mathcal{A}_{L} w_{2} \in H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right)$. As $w_{2} \in H_{\Gamma}^{2}\left(\Omega_{1}\right)$ we have that $\left.\partial_{\nu} w_{2}\right|_{\partial \Omega_{1}} \in H^{1 / 2}\left(\partial \Omega_{1}\right)$ and so $\mathcal{N} \gamma_{1} w_{2} \in H^{2}\left(\Omega_{1}\right)$. Therefore, $w_{2}-\mathcal{N} \gamma_{1} w_{2} \in H^{2}\left(\Omega_{1}\right)$. From the definition of the Neumann map $\mathcal{N}$ it follows that $w_{2}-\mathcal{N} \gamma_{1} w_{2}=0$ on $\Gamma$ and $\partial_{\nu}\left(w_{2}-\mathcal{N} \gamma_{1} w_{2}\right)=0$ on $I$. Thus, $\mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2} \in L^{2}\left(\Omega_{1}\right)$.

When $\eta=0$, we not consider structural damping (i.e., $m_{1}=0$ ). Under these conditions, we define the operator $\mathscr{A}_{0}: \mathscr{H}_{0} \supset D\left(\mathscr{A}_{0}\right) \rightarrow \mathscr{H}_{0}$ as follows

$$
\mathscr{A}_{0}\left(\begin{array}{c}
w_{1}  \tag{3.34}\\
w_{2} \\
w_{3} \\
w_{4} \\
w_{5}
\end{array}\right):=\left(\begin{array}{c}
w_{2} \\
-\frac{1}{\rho_{1}} \mathcal{A}_{B} \mathcal{W}_{0}+\frac{\alpha}{\rho_{1}} \mathcal{A}_{T} w_{5}-\frac{\alpha \sigma}{\beta \rho_{1}} w_{5} \\
w_{4} \\
\frac{\beta_{2}}{\rho_{2}} \Delta w_{3}+\frac{m_{2}}{\rho_{2}} \Delta w_{4} \\
-\frac{\alpha}{\rho_{0}} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}-\frac{\beta}{\rho_{0}} \mathcal{A}_{T} w_{5}
\end{array}\right)
$$

where $\mathcal{W}_{0}:=\beta_{1} w_{1}+\beta_{2} \mathcal{G}_{2} \gamma_{1} w_{3}+m_{2} \mathcal{G}_{2} \gamma_{1} w_{4}+\alpha \mathcal{G}_{1} \gamma_{0} w_{5}-\alpha \kappa \mathcal{G}_{2} \gamma_{0} w_{5}$. The domain of $\mathscr{A}_{0}$ is given by

$$
\begin{aligned}
D\left(\mathscr{A}_{0}\right):= & \left\{w=\left(w_{j}\right)_{1, \ldots, 5}^{\top} \in\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{2} \times\left[H^{1}\left(\Omega_{2}\right)\right]^{2} \times D\left(\mathcal{A}_{T}\right): \mathcal{W}_{0} \in D\left(\mathcal{A}_{B}\right),\right. \\
& \left.\beta_{2} \Delta w_{3}+m_{2} \Delta w_{4} \in L^{2}\left(\Omega_{2}\right) \text { and } w_{j}=w_{j+2} \text { on } I \text { for } j=1,2\right\} .
\end{aligned}
$$

### 3.1.3 The semigroup approach

We will prove that the linear operator $\mathscr{A}_{\eta}: D\left(\mathscr{A}_{\eta}\right) \subset \mathscr{H}_{\eta} \rightarrow \mathscr{H}_{\eta}$, for $\eta \geq 0$, is dissipative. This result is associated with the derivative of the energy of our system (3.1), which is given by

$$
\begin{aligned}
\partial_{t} E_{\gamma}(t)= & -m_{1}\left\|\nabla u_{t}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-m_{2}\left\|\nabla v_{t}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2} \\
& -\sigma\|\theta\|_{L^{2}\left(\Omega_{1}\right)}^{2}-\beta\|\nabla \theta\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\beta \kappa\|\theta\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2}
\end{aligned}
$$

Chapter 3. Well-posedness and regularity of the solutions of a plate-membrane system
and for this reason the energy of the system decreases as time passes. Physically, the system is said to be dissipative. The Lumer-Phillips theorem, see Theorem 2.30, will establish that $\mathscr{A}_{\eta}$ is the generator of a $C_{0}$-semigroup of contractions on $\mathscr{H}_{\eta}$ and as a consequence we will have that the Cauchy problem (3.33) has a unique solution.

Proposition 3.13. Let $\eta \geq 0$. For $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$, we have

$$
\begin{align*}
\operatorname{Re}\left(\mathscr{A}_{\eta} w, w\right)_{\mathscr{H}_{\eta}}= & -m_{1}\left\|\nabla w_{2}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-m_{2}\left\|\nabla w_{4}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}  \tag{3.35}\\
& -\sigma\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}-\beta\left\|\nabla w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\beta \kappa\left\|w_{5}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2}
\end{align*}
$$

Therefore, $\mathscr{A}_{\eta}$ is dissipative.
Proof. We first assume that $\eta>0$. If $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$, then it follows from the inner product of $\mathscr{H}_{\eta}$ that

$$
\begin{align*}
& \left(\mathscr{A}_{\eta} w, w\right)_{\mathscr{H}_{\eta}}=\beta_{1}\left(w_{2}, w_{1}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}+\left(\beta_{2} \Delta w_{3}+m_{2} \Delta w_{4}, w_{4}\right)_{L^{2}\left(\Omega_{2}\right)} \\
& +\left(-\mathcal{M}_{\eta}^{-1} \mathcal{W}+\mathcal{M}_{\eta}^{-1}\left(-m_{1} \mathcal{A}_{L} w_{2}+\alpha \mathcal{A}_{T} w_{5}-\frac{\alpha \sigma}{\beta} w_{5}\right), w_{2}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}  \tag{3.36}\\
& +\beta_{2}\left(\nabla w_{4}, \nabla w_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}+\left(-\alpha \mathcal{A}_{L}\left(w_{2}-\mathcal{N} \partial_{\nu} w_{2}\right)-\beta \mathcal{A}_{T} w_{5}, w_{5}\right)_{L^{2}\left(\Omega_{1}\right)}
\end{align*}
$$

Using the first Green's formula (2.3) and the fact that $w_{2}=w_{4}$ on $I$, we get

$$
\begin{align*}
& \left(\beta_{2} \Delta w_{3}+m_{2} \Delta w_{4}, w_{4}\right)_{L^{2}\left(\Omega_{2}\right)} \\
& =-\left(\nabla\left(\beta_{2} w_{3}+m_{2} w_{4}\right), \nabla w_{4}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-\left(\partial_{\nu}\left(\beta_{2} w_{3}+m_{2} w_{4}\right), w_{4}\right)_{L^{2}(I)} \\
& =-\beta_{2}\left(\nabla w_{3}, \nabla w_{4}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-m_{2}\left\|\nabla w_{4}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}-\beta_{2}\left(\partial_{\nu} w_{3}, w_{2}\right)_{L^{2}(I)}  \tag{3.37}\\
& \quad-m_{2}\left(\partial_{\nu} w_{4}, w_{2}\right)_{L^{2}(I)} .
\end{align*}
$$

By (2.12), (3.14), (3.15) and integration by parts, we obtain

$$
\begin{align*}
( & \left.-\mathcal{M}_{\eta}^{-1} \mathcal{W}+\mathcal{M}_{\eta}^{-1}\left(-m_{1} \mathcal{A}_{L} w_{2}+\alpha \mathcal{A}_{T} w_{5}-\frac{\alpha \sigma}{\beta} w_{5}\right), w_{2}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \\
= & -\left\langle\mathcal{W}, w_{2}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}-m_{1}\left\langle\mathcal{A}_{L} w_{2}, w_{2}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \\
& +\alpha\left(\mathcal{A}_{T} w_{5}, w_{2}\right)_{L^{2}\left(\Omega_{1}\right)}-\frac{\alpha \sigma}{\beta}\left(w_{5}, w_{2}\right)_{L^{2}\left(\Omega_{1}\right)} \\
= & -\left\langle\mathcal{W}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}-m_{1}\left\|\nabla w_{2}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}  \tag{3.38}\\
& +\alpha\left(\nabla w_{5}, \nabla w_{2}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}-\alpha\left(\partial_{\nu} w_{5}, w_{2}\right)_{L^{2}(I)} .
\end{align*}
$$

We have used that $\mathcal{W} \in H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \subset\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$ and $-\Delta w_{5}=\left(\mathcal{A}_{T}-\frac{\sigma}{\beta} \mathcal{I}\right) w_{5}$ (see definition of $D\left(\mathscr{A}_{\eta}\right),(3.8)$ and Proposition 3.6). On the other hand,

$$
\begin{aligned}
& \left\langle\mathcal{W}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\beta_{1}\left\langle\mathcal{A}_{B} w_{1}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} \\
& +\beta_{2}\left\langle\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{3}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}+m_{2}\left\langle\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{4}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} \\
& +\alpha\left\langle\mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} w_{5}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}-\alpha \kappa\left\langle\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} w_{5}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} .
\end{aligned}
$$

Due to (3.12) and (3.21), we get

$$
\begin{align*}
& \left\langle\mathcal{W}, w_{2}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\beta_{1}\left(w_{1}, w_{2}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}-\beta_{2}\left(\partial_{\nu} w_{3}, w_{2}\right)_{L^{2}(I)}  \tag{3.39}\\
& -m_{2}\left(\partial_{\nu} w_{4}, w_{2}\right)_{L^{2}(I)}+\alpha\left(w_{5}, \partial_{\nu} w_{2}\right)_{L^{2}(I)}+\alpha \kappa\left(w_{5}, w_{2}\right)_{L^{2}(I)}
\end{align*}
$$

From (3.16), (3.30) and integration by parts it follows that

$$
\begin{align*}
& \left(-\alpha \mathcal{A}_{L}\left(w_{2}-\mathcal{N} \partial_{\nu} w_{2}\right)-\beta \mathcal{A}_{T} w_{5}, w_{5}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =-\alpha\left(\nabla w_{2}, \nabla w_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\alpha\left(\partial_{\nu} w_{2}, w_{5}\right)_{L^{2}(I)}-\beta\left\|\nabla w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}  \tag{3.40}\\
& \quad-\beta \kappa\left\|w_{5}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2}-\sigma\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} .
\end{align*}
$$

A combination of equalities (3.36)-(3.40) implies

$$
\begin{aligned}
& \left(\mathscr{A}_{\eta} w, w\right)_{\mathscr{H}_{\eta}}=i 2 \beta_{1} \operatorname{Im}\left(w_{2}, w_{1}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}+i 2 \alpha \operatorname{Im}\left(\partial_{\nu} w_{2}, w_{5}\right)_{L^{2}(I)} \\
& +i 2 \alpha \operatorname{Im}\left(\nabla w_{5}, \nabla w_{2}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+i 2 \beta_{2} \operatorname{Im}\left(\nabla w_{4}, \nabla w_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-m_{1}\left\|\nabla w_{2}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2} \\
& -m_{2}\left\|\nabla w_{4}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}-\sigma\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}-\beta\left\|\nabla w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\beta \kappa\left\|w_{5}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2} .
\end{aligned}
$$

Taking real part above we see that (3.35) holds. For $\eta=0$ and $m_{1}=0$ the proof is similar.

Proposition 3.14. Let $\mathscr{P}$ be as in the definition of $\mathscr{A}_{\eta}$, see page 56. For $w_{1} \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and $w_{5} \in H^{1}\left(\Omega_{1}\right)$, we have

$$
\begin{equation*}
\left\langle\mathscr{P} w_{5}, w_{1}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left(\nabla w_{5}, \nabla w_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}-\left(w_{5}, \partial_{\nu} w_{1}\right)_{L^{2}(I)} . \tag{3.41}
\end{equation*}
$$

Proof. Let $w_{1} \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and $w_{5} \in H^{1}\left(\Omega_{1}\right)$. Characterization (3.17) implies that $w_{5} \in D\left(\mathcal{A}_{T}^{1 / 2}\right)$ and thus $\mathcal{A}_{T} w_{5} \in\left[D\left(\mathcal{A}_{T}^{1 / 2}\right)\right]^{\prime}$. As $D\left(\mathcal{A}_{B}^{1 / 2}\right) \subset D\left(\mathcal{A}_{T}^{1 / 2}\right)$, we have that $\mathcal{A}_{T} w_{5} \in\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$. Note that $\mathcal{G}_{1} \gamma_{0} w_{5}, \mathcal{G}_{2} \gamma_{0} w_{5} \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and thus
$\mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} w_{5}, \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} w_{5} \in\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$. Hence, $\mathscr{P}_{w_{5}} \in\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime}$. Using (2.12), we obtain

$$
\begin{aligned}
& \left\langle\mathscr{P} w_{5}, w_{1}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}=\left\langle\mathcal{A}_{T} w_{5}, w_{1}\right\rangle_{\left[D\left(\mathcal{A}_{T}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{T}^{1 / 2}\right)}-\frac{\sigma}{\beta}\left(w_{5}, w_{1}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& -\left\langle\mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} w_{5}, w_{1}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}+\kappa\left\langle\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} w_{5}, w_{1}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)} .
\end{aligned}
$$

Now, (3.18) and (3.21) leads to the identity (3.41).
Theorem 3.15. For all $\eta>0$ and $m_{1}, m_{2} \geq 0$, the operator $\mathscr{A}_{\eta}$ generates a $C_{0}$-semigroup $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ of contractions on $\mathscr{H}_{\eta}$. Therefore, for any $w_{0} \in \mathscr{H}_{\eta}$ there exists a unique mild solution $w \in C\left([0, \infty), \mathscr{H}_{\eta}\right)$ of (3.33). Furthermore, for any $w_{0} \in D\left(\mathscr{A}_{\eta}^{k}\right)$ with $k \in \mathbb{N}$, there exists a unique classical solution $w$ to the problem (3.33) that belongs to $\cap_{j=0}^{k} C^{k-j}\left([0, \infty), D\left(\mathscr{A}_{\eta}^{j}\right)\right)$.

Proof. Because of Proposition 3.13. Theorem 2.31 and Theorem 2.30 it is sufficient to prove that $\mathcal{I}-\mathscr{A}_{\eta}: D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathscr{H}_{\eta}$ is a surjective operator. Let $f=\left(f_{1}, f_{2}, f_{3}, f_{4}, f_{5}\right)^{\top} \in \mathscr{H}_{\eta}$. We will find $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$ such that $\left(\mathcal{I}-\mathscr{A}_{\eta}\right) w=f$, this is,

$$
\begin{aligned}
& w_{1}-w_{2}=f_{1} \text { in } D\left(\mathcal{A}_{B}^{1 / 2}\right), \\
& w_{2}+\frac{1}{\rho_{1}} \mathcal{M}_{\eta}^{-1}\left(\mathcal{W}+m_{1} \mathcal{A}_{L} w_{2}-\alpha \mathcal{A}_{T} w_{5}+\frac{\alpha \sigma}{\beta} w_{5}\right)=f_{2} \quad \text { in } H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right), \\
& w_{3}-w_{4}=f_{3} \text { in } H^{1}\left(\Omega_{2}\right), \\
& w_{4}-\frac{\beta_{2}}{\rho_{2}} \Delta w_{3}-\frac{m_{2}}{\rho_{2}} \Delta w_{4}=f_{4} \text { in } L^{2}\left(\Omega_{2}\right), \\
& w_{5}+\frac{\alpha}{\rho_{0}} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}+\frac{\beta}{\rho_{0}} \mathcal{A}_{T} w_{5}=f_{5} \quad \text { in } \quad L^{2}\left(\Omega_{1}\right) .
\end{aligned}
$$

Plugging in $w_{j+1}=w_{j}-f_{j}$ for $j=1,3$, we have the following matrix equation

$$
\begin{aligned}
& \left(\begin{array}{ccc}
\rho_{1} \mathcal{M}_{\eta}+\beta_{1} \mathcal{A}_{B}+m_{1} \mathcal{A}_{L} & \left(\beta_{2}+m_{2}\right) \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} & -\alpha \mathscr{P} \\
0 & \rho_{2} \mathcal{I}-\left(\beta_{2}+m_{2}\right) \Delta & 0 \\
\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) & 0 & \rho_{0} \mathcal{I}+\beta \mathcal{A}_{T}
\end{array}\right)\left(\begin{array}{l}
w_{1} \\
w_{3} \\
w_{5}
\end{array}\right) \\
& +\left(\begin{array}{c}
0 \\
m_{2} \Delta f_{3} \\
0
\end{array}\right)=\left(\begin{array}{c}
\rho_{1} \mathcal{M}_{\eta} f_{1}+m_{1} \mathcal{A}_{L} f_{1}+\rho_{1} \mathcal{M}_{\eta} f_{2}+m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} f_{3} \\
\rho_{2} f_{3}+\rho_{2} f_{4} \\
\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) f_{1}+\rho_{0} f_{5}
\end{array}\right)
\end{aligned}
$$

in $H_{\Gamma}^{-2}\left(\Omega_{1}\right) \times L^{2}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{1}\right)$. We set $\mathcal{Y}:=\mathcal{X} \times H^{1}\left(\Omega_{1}\right)$, where the Hilbert space

$$
\begin{equation*}
\mathcal{X}:=\left\{\left(w_{1}, w_{3}\right)^{\top} \in D\left(\mathcal{A}_{B}^{1 / 2}\right) \times H^{1}\left(\Omega_{2}\right): w_{1}=w_{3} \text { on } I\right\} \tag{3.42}
\end{equation*}
$$

is equipped with the scalar product

$$
\left(\left(w_{1}, w_{3}\right),\left(\widetilde{w}_{1}, \widetilde{w}_{3}\right)\right)_{\mathcal{X}}:=\beta_{1}\left(w_{1}, \widetilde{w}_{1}\right)_{D\left(\mathcal{A}_{B}^{1 / 2}\right)}+\beta_{2}\left(\nabla w_{3}, \nabla \widetilde{w}_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}
$$

The sesquilinear form $b: \mathcal{Y} \times \mathcal{Y} \rightarrow \mathbb{C}$, defined by

$$
\begin{aligned}
& b\left(\left(w_{1}, w_{3}, w_{5}\right),\left(\phi_{1}, \phi_{3}, \phi_{5}\right)\right):=\rho_{1}\left(w_{1}, \phi_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+\beta_{1}\left(w_{1}, \phi_{1}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \\
& +m_{1}\left(\nabla w_{1}, \nabla \phi_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}-\alpha\left(\nabla w_{5}, \nabla \phi_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\alpha\left(w_{5}, \partial_{\nu} \phi_{1}\right)_{L^{2}(I)} \\
& +\rho_{2}\left(w_{3}, \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)}+\left(\beta_{2}+m_{2}\right)\left(\nabla w_{3}, \nabla \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}+\alpha\left(\nabla w_{1}, \nabla \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& -\alpha\left(\partial_{\nu} w_{1}, \phi_{5}\right)_{L^{2}(I)}+\rho_{0}\left(w_{5}, \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)}+\beta\left(\nabla w_{5}, \nabla \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& +\beta \kappa\left(w_{5}, \phi_{5}\right)_{L^{2}\left(\partial \Omega_{1}\right)}+\sigma\left(w_{5}, \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)},
\end{aligned}
$$

is continuous. The coercivity of $b$ is immediate, i.e.,

$$
\operatorname{Re} b\left(\left(w_{1}, w_{3}, w_{5}\right),\left(w_{1}, w_{3}, w_{5}\right)\right) \geq C\left\|\left(w_{1}, w_{3}, w_{5}\right)\right\|_{\mathcal{Y}}^{2}
$$

for all $\left(w_{1}, w_{3}, w_{5}\right) \in \mathcal{Y}$. The mapping $\mathcal{K}: \mathcal{Y} \rightarrow \mathbb{C}$ given by

$$
\begin{aligned}
\mathcal{K}\left(\phi_{1}, \phi_{3}, \phi_{5}\right): & \rho_{1}\left(f_{1}+f_{2}, \phi_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+m_{1}\left(\nabla f_{1}, \nabla \phi_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& +\rho_{2}\left(f_{3}+f_{4}, \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)}+m_{2}\left(\nabla f_{3}, \nabla \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& -\alpha\left(\Delta f_{1}, \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)}+\rho_{0}\left(f_{5}, \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)}
\end{aligned}
$$

is antilinear and continuous. By the Lax-Milgram theorem, there exists a unique $\left(w_{1}, w_{3}, w_{5}\right) \in \mathcal{Y}$ such that

$$
\begin{equation*}
b\left(\left(w_{1}, w_{3}, w_{5}\right),\left(\phi_{1}, \phi_{3}, \phi_{5}\right)\right)=\mathcal{K}\left(\phi_{1}, \phi_{3}, \phi_{5}\right) \tag{3.43}
\end{equation*}
$$

for all $\left(\phi_{1}, \phi_{3}, \phi_{5}\right) \in \mathcal{Y}$. By (3.12), (3.14), (3.15), (3.18), (3.21), Proposition 3.14 , the first Green's formula and (3.43), we obtain the matrix equation and its third row implies

$$
\mathcal{A}_{T} w_{5}=\frac{\alpha}{\beta} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right)\left(f_{1}-w_{1}\right)+\frac{\rho_{0}}{\beta}\left(f_{5}-w_{5}\right) \in L^{2}\left(\Omega_{1}\right)
$$

and as $\mathcal{A}_{T}: D\left(\mathcal{A}_{T}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ is a bijection because it is a positive selfadjoint operator, see Remark 2.52 , then $w_{5} \in D\left(\mathcal{A}_{T}\right)$. We set $w_{2}:=w_{1}-f_{1}$ and $w_{4}:=w_{3}-f_{3}$. From the first and second rows of the matrix equation, it follows that $\mathcal{W} \in H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right)$ and $\beta_{2} \Delta w_{3}+m_{2} \Delta w_{4} \in L^{2}\left(\Omega_{2}\right)$, respectively. Note that $w_{2}=w_{4}$ on $I$. Hence, $w:=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$. Finally, we have proven that the range $R\left(\mathcal{I}-\mathscr{A}_{\eta}\right)$ is equal to $\mathscr{H}_{\eta}$.

The unique solution of the Cauchy problem (3.33), classical or weak, is of the form $w(t)=\mathscr{T}_{\eta}(t) w_{0}(t \geq 0)$ and its regularity depends on the choice of the initial datum (see Theorem 2.34, Theorem 2.36 and Theorem 2.37).

Remark 3.16. The statement of the previous theorem is also true when $\eta=m_{1}=0$ and $m_{2} \geq 0$. The proof is analogous to that presented above.

### 3.2 Regularity

The section starts with a regularity result which contains a useful estimate that we will use in the next two chapters. Said result will allow the proof of Theorem 3.18. This theorem indicates that when the initial data have good regularity, then some components of the solution of the problem (1.1)-(1.9) gain regularity.

Remark 3.17. Let $\mathcal{O}$ be a bounded domain in $\mathbb{R}^{n}$ with a $C^{2,1}$ boundary. We recall that the classic map

$$
H^{2}(\mathcal{O}) \ni u \mapsto\left(-\Delta u,\left.u\right|_{\partial \mathcal{O}}\right) \in L^{2}(\mathcal{O}) \times H^{3 / 2}(\partial \mathcal{O})
$$

is an isomorphism, see [123, p. 253], which implies that for each $f \in L^{2}(\mathcal{O})$ and $g \in H^{3 / 2}(\partial \mathcal{O})$ there exists a unique $u \in H^{2}(\mathcal{O})$ such that $\Delta u=f$ in $\mathcal{O}$ with $\left.u\right|_{\partial \mathcal{O}}=\left.g\right|_{\partial \mathcal{O}}$. Moreover,

$$
\begin{equation*}
\|u\|_{H^{2}(\mathcal{O})} \leq C\left(\|f\|_{L^{2}(\mathcal{O})}+\|g\|_{H^{3 / 2}(\partial \mathcal{O})}\right) \tag{3.44}
\end{equation*}
$$

Theorem 3.18. Let $\eta>0$ and $m_{1}, m_{2} \geq 0$. If $w \in D\left(\mathscr{A}_{\eta}^{2}\right)$, then $w_{1} \in$ $H^{4}\left(\Omega_{1}\right), w_{2} \in H^{3}\left(\Omega_{1}\right), \beta_{2} w_{3}+m_{2} w_{4} \in H^{2}\left(\Omega_{2}\right)$ and $w_{5} \in H^{3}\left(\Omega_{1}\right)$. Therefore, if $w_{0} \in D\left(\mathscr{A}_{\eta}^{2}\right)$ then $w(t):=\mathscr{T}_{\eta}(t) w_{0}(t \geq 0)$ is the unique solution of the problem (1.1)-(1.9), which belongs to $D\left(\mathscr{A}_{\eta}^{2}\right)$ and satisfies the boundary and transmission conditions, (1.6) and (1.7), in the strong sense of traces.
Proof. If $\xi=\left(\xi_{1}, \xi_{2}, \xi_{3}, \xi_{4}, \xi_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$, we have that $\beta_{1} \mathcal{A}_{B} \xi_{1}+\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \xi_{3}+$ $m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \xi_{4}+\alpha \mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} \xi_{5}-\alpha \kappa \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} \xi_{5}=: \mathcal{W}_{\xi} \in\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime}$. It is well known that $\mathcal{A}_{B}: D\left(\mathcal{A}_{B}^{3 / 4}\right) \rightarrow\left[D\left(\mathcal{A}_{B}^{1 / 4}\right)\right]^{\prime}$ is an isomorphism, see part $c$ ) of Theorem 2.53. Thus, $\mathcal{A}_{B}^{-1} \mathcal{W}_{\xi} \in H^{3}\left(\Omega_{1}\right)$ due to the third characterization in (3.8). Due to the trace operators $\gamma_{i-1}$ and the regularity of the Green maps $\mathcal{G}_{i}, i=1,2$, we obtain

$$
\xi_{1}=\frac{1}{\beta_{1}} \mathcal{A}_{B}^{-1} \mathcal{W}_{\xi}-\frac{1}{\beta_{1}} \mathcal{G}_{2} \gamma_{1}\left(\beta_{2} \xi_{3}+m_{2} \xi_{4}\right)-\frac{\alpha}{\beta_{1}}\left(\mathcal{G}_{1} \gamma_{0} \xi_{5}-\kappa \mathcal{G}_{2} \gamma_{0} \xi_{5}\right) \in H^{3}\left(\Omega_{1}\right) .
$$

Let $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}^{2}\right)$. Since $\mathscr{A}_{\eta} w \in D\left(\mathscr{A}_{\eta}\right)$, then the above allows us to affirm that $w_{2} \in H^{3}\left(\Omega_{1}\right)$. We also have

$$
\begin{equation*}
\frac{1}{\rho_{1}} \mathcal{M}_{\eta}^{-1}\left(\mathcal{W}+m_{1} \mathcal{A}_{L} w_{2}-\alpha \mathcal{A}_{T} w_{5}+\frac{\alpha \sigma}{\beta} w_{5}\right)=: \phi \in D\left(\mathcal{A}_{B}^{1 / 2}\right) \tag{3.45}
\end{equation*}
$$

Note that $\mathcal{G}_{2} \gamma_{1} \phi \in D\left(\mathcal{A}_{B}^{3 / 4}\right)$ and so $\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \phi \in H_{\Gamma . \eta}^{-1}\left(\Omega_{1}\right)$. Taking $\widetilde{\phi} \in D\left(\mathcal{A}_{B}^{1 / 2}\right)$ and applying integration by parts together with (2.12), (3.14) and (3.21), we get the following equalities

$$
\begin{align*}
& \left\langle\Delta \phi+\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \phi, \widetilde{\phi}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \\
& =(\Delta \phi, \widetilde{\phi})_{L^{2}\left(\Omega_{1}\right)}+\left\langle\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \phi, \widetilde{\phi}\right\rangle_{\left[D\left(\mathcal{A}_{B}^{1 / 2}\right)\right]^{\prime} \times D\left(\mathcal{A}_{B}^{1 / 2}\right)}  \tag{3.46}\\
& =-\left\langle\mathcal{A}_{L} \phi, \widetilde{\phi}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}
\end{align*}
$$

and thus

$$
\begin{equation*}
\mathcal{M}_{\eta} \phi=\phi-\eta \Delta \phi-\eta \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \phi \text { in } H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \tag{3.47}
\end{equation*}
$$

Doing $w_{j k}:=\beta_{2} w_{j}+m_{2} w_{k}$, we obtain that $\Delta w_{34}=: \tilde{f} \in L^{2}\left(\Omega_{2}\right)$. Note that $w_{34}=w_{12}$ on $I$ with $\gamma_{0} w_{12} \in H^{3 / 2}(I)$. By Remark 3.17, there exists a unique $\widetilde{w}_{34} \in H^{2}\left(\Omega_{2}\right)$ such that $\Delta \widetilde{w}_{34}=\widetilde{f}$ in $\Omega_{2}$ and $\left.\widetilde{w}_{34}\right|_{I}=\left.w_{12}\right|_{I}$. Since $w_{34}-\widetilde{w}_{34} \in H_{0}^{1}\left(\Omega_{2}\right)$ is a weak solution of $\Delta\left(w_{34}-\widetilde{w}_{34}\right)=0$, we immediately obtain $w_{34}=\widetilde{w}_{34}$, which leads to $\beta_{2} w_{3}+m_{2} w_{4} \in H^{2}\left(\Omega_{2}\right)$. From (3.45)- (3.47) it follows that

$$
\mathcal{W}-m_{1} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{2}+\gamma \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} \phi=\rho_{1} \phi-\gamma \Delta \phi+m_{1} \Delta w_{2}-\alpha \Delta w_{5} \in L^{2}\left(\Omega_{1}\right)
$$

In consequence,
$\beta_{1} w_{1}-m_{1} \mathcal{G}_{2} \gamma_{1} w_{2}+\mathcal{G}_{2} \gamma_{1}\left(\beta_{2} w_{3}+m_{2} w_{4}\right)+\alpha\left(\mathcal{G}_{1}-\kappa \mathcal{G}_{2}\right) \gamma_{0} w_{5}+\gamma \mathcal{G}_{2} \gamma_{1} \phi \in D\left(\mathcal{A}_{B}\right)$
because $\mathcal{A}_{B}: D\left(\mathcal{A}_{B}\right) \rightarrow L^{2}\left(\Omega_{1}\right)$ is bijective, see Remark 2.52. Once again, appealing to the regularity of the operators $\gamma_{i-1}$ and maps $\mathcal{G}_{i}$ we obtain that $w_{1} \in H^{4}\left(\Omega_{1}\right)$. Now, the last expression above leads us to that the boundary and transmission conditions (1.6) and 1.7) hold in the strong sense of the traces with $w_{2}$ and $w_{4}$ instead of $u_{t}$ and $v_{t}$, respectively. On the other hand, we have that

$$
\begin{equation*}
\frac{\alpha}{\rho_{0}} \widehat{\varphi}-\frac{\beta}{\rho_{0}} \mathcal{A}_{T} w_{5}=: \widehat{\phi} \in D\left(\mathcal{A}_{T}\right) \tag{3.48}
\end{equation*}
$$

where $\widehat{\varphi}:=-\mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}$. This is also due to the fact that $\mathscr{A}_{\eta} w \in D\left(\mathscr{A}_{\eta}\right)$. As $w_{2} \in H^{3}\left(\Omega_{1}\right)$, then $\widehat{\varphi}=\Delta w_{2} \in H^{1}\left(\Omega_{1}\right)$. Because of $\widehat{\phi} \in H^{2}\left(\Omega_{1}\right)$, we get from (3.17) and (3.48) that

$$
\mathcal{A}_{T} w_{5}=\frac{\alpha}{\beta} \widehat{\varphi}-\frac{\rho_{0}}{\beta} \widehat{\phi} \in D\left(\mathcal{A}_{T}^{1 / 2}\right)
$$

Finally, since the operator $\mathcal{A}_{T}: D\left(\mathcal{A}_{T}^{3 / 2}\right) \rightarrow D\left(\mathcal{A}_{T}^{1 / 2}\right)$ is bijective (see again part c) of Theorem 2.53) we conclude that $w_{5} \in D\left(\mathcal{A}_{T}^{3 / 2}\right) \subset H^{3}\left(\Omega_{1}\right)$.

Chapter 3. Well-posedness and regularity of the solutions of a plate-membrane system

If $w_{0} \in D\left(\mathscr{A}_{\eta}^{2}\right)$, then $w_{0} \in D\left(\mathscr{A}_{\eta}\right)$ and $\mathscr{A}_{\eta} w_{0} \in D\left(\mathscr{A}_{\eta}\right)$. By Proposition 2.27, $\mathscr{T}_{\eta}(t) w_{0} \in D\left(\mathscr{A}_{\eta}\right)$ and $\mathscr{A}_{\eta} \mathscr{T}_{\eta}(t) w_{0}=\mathscr{T}_{\eta}(t) \mathscr{A}_{\eta} w_{0} \in D\left(\mathscr{A}_{\eta}\right)$. Hence, the unique solution of the problem (1.1)-(1.9) satisfies $w(t)=\mathscr{T}_{\eta}(t) w_{0} \in D\left(\mathscr{A}_{\eta}^{2}\right)$ for any $t \geq 0$.

Remark 3.19. Let $\eta \geq 0$. Due to the proof of Theorem 3.18, we have: If $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}\right)$ then $\beta_{2} w_{3}+m_{2} w_{4} \in H^{2}\left(\Omega_{2}\right)$ and moreover $w \in H^{4-\chi}\left(\Omega_{1}\right) \times H^{2}\left(\Omega_{1}\right) \times\left[H^{1}\left(\Omega_{2}\right)\right]^{2} \times H^{2}\left(\Omega_{1}\right)$, where $\chi:=\chi_{(0, \infty)}(\eta)$ stands for the characteristic function of $(0, \infty)$.

## Chapter 4

## Asymptotic behavior of the solutions of some plate-membrane problems

In this chapter the reader will find an extensive study of the asymptotic behavior of the solutions of the plate-membrane transmission problem (1.1)(1.9). We will see that for $\gamma \geq 0$ and $m_{2}>0$ the solutions will have strong stability, see Corollary 4.3 and Corollary 4.6. When the structural damping on the plate is not considered $\left(m_{1}=0\right)$ it will be the thermal effect that will help to obtain the strong stability, see part b) of Remark 4.4 and Remark 4.7. We prove by contradiction the exponential stability of the energy of the system when the plate has rotational inertial $(\gamma>0)$ and the structure is damped $\left(m_{1}>0\right.$ and $\left.m_{2}>0\right)$, see Theorem 4.8. When we remove the Kelvin-Voigt damping $\left(m_{2}=0\right)$ and the inertial term is present, an absence of exponential stability is caused (see Theorem 4.15). In this case, the solutions have polynomial stability when we add a geometric condition (see Theorem 4.16.

### 4.1 Strong stability

First we will prove that the operator $\mathscr{A}_{\eta}$ is continuously invertible, this is, 0 is in the resolvent set $\rho\left(\mathscr{A}_{\eta}\right)$. This result will be key to prove the strong stability of the solutions of our problem for when $\gamma \geq 0$ and $m_{2}>0$.

Proposition 4.1. $0 \in \rho\left(\mathscr{A}_{\eta}\right)$ if a) $\eta>0$ and $m_{1}, m_{2} \geq 0$ or b) $\eta=m_{1}=0$ and $m_{2} \geq 0$.

Proof. Due to the similarity between the proofs of the two cases a) and b), we will only prove part a).

We show that the operator $\mathscr{A}_{\eta}: D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathscr{H}_{\eta}$ is bijective. Let $f=$ $\left(f_{1}, f_{2}, f_{3}, f_{4}, f_{5}\right)^{\top} \in \mathscr{H}_{\eta}$. We will find a unique $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in$ $D\left(\mathscr{A}_{\eta}\right)$ such that $\mathscr{A}_{\eta} w=f$. Conveniently we set $w_{2}:=f_{1}, w_{4}:=f_{3}$ and $w_{5}:=-\mathcal{A}_{T}^{-1}\left[\frac{\alpha}{\beta} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) f_{1}+\frac{\rho_{0}}{\beta} f_{5}\right]$. We recall again that Remark 2.52 implies that $\mathcal{A}_{T}$ is bijective. We now consider the following system

$$
\left\{\begin{array}{lll}
\beta_{1} \mathcal{A}_{B} w_{1}+\beta_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} w_{3}=h & \text { in } & H_{\Gamma}^{-2}\left(\Omega_{1}\right),  \tag{4.1}\\
-\beta_{2} \Delta w_{3}-m_{2} \Delta f_{3}=-\rho_{2} f_{4} & \text { in } & L^{2}\left(\Omega_{2}\right),
\end{array}\right.
$$

where $h:=-m_{1} \mathcal{A}_{L} f_{1}-\rho_{1} \mathcal{M}_{\eta} f_{2}-m_{2} \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1} f_{3}+\alpha \mathscr{P} w_{5}$. With $\mathcal{X}$ as in (3.42), we define $\widetilde{b}: \mathcal{X} \times \mathcal{X} \rightarrow \mathbb{C}$ by $\widetilde{b}\left(\left(w_{1}, w_{3}\right),\left(\phi_{1}, \phi_{3}\right)\right):=\left(\left(w_{1}, w_{3}\right),\left(\phi_{1}, \phi_{3}\right)\right)_{\mathcal{X}}$. The function $\widetilde{\mathcal{K}}: \mathcal{X} \rightarrow \mathbb{C}$ given by

$$
\begin{aligned}
\widetilde{\mathcal{K}}\left(\phi_{1}, \phi_{3}\right):= & -m_{1}\left(\nabla f_{1}, \nabla \phi_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}-\rho_{1}\left(f_{2}, \phi_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+\alpha\left(\nabla w_{5}, \nabla \phi_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& -\alpha\left(w_{5}, \partial_{\nu} \phi_{1}\right)_{L^{2}(I)}-m_{2}\left(\nabla f_{3}, \nabla \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-\rho_{2}\left(f_{4}, \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)}
\end{aligned}
$$

is an element of the antidual space $\mathcal{X}^{\prime}$. By the Riesz's representation theorem there exists a unique $\left(w_{1}, w_{3}\right) \in \mathcal{X}$ such that

$$
\begin{equation*}
\widetilde{b}\left(\left(w_{1}, w_{3}\right),\left(\phi_{1}, \phi_{3}\right)\right)=\widetilde{\mathcal{K}}\left(\phi_{1}, \phi_{3}\right) \tag{4.2}
\end{equation*}
$$

for all $\left(\phi_{1}, \phi_{3}\right) \in \mathcal{X}$. From (3.12), (3.14), (3.15), (3.21), Proposition 3.14, the first Green's formula and (4.2), we obtain that (4.1) is true. Reasoning as in the last part of the proof of Theorem 3.15, we get $w:=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in$ $D\left(\mathscr{A}_{\eta}\right)$. Thus, $\mathscr{A}_{\eta}$ is surjective. As the constructed $w$ above is unique, then $\mathscr{A}_{\eta}$ is injective.

Since $\mathscr{A}_{\eta}: D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathscr{H}_{\eta}$ is a closed linear operator (due to Theorem 3.15 and Proposition 2.28) and bijective, we conclude that $\mathscr{A}_{\eta}^{-1} \in \mathcal{L}\left(\mathscr{H}_{\eta}\right)$.

The following proposition together with Theorem 2.39 allow us to affirm that $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ is strongly stable in $\mathscr{H}_{\eta}$, when there is Kelvin-Voigt damping on the membrane.

Proposition 4.2. Let $\eta>0, m_{1} \geq 0$ and $m_{2}>0$. The imaginary axis is contained in the resolvent set of $\mathscr{A}_{\eta}$, this is, $i \mathbb{R} \subset \rho\left(\mathscr{A}_{\eta}\right)$.

Proof. Let $\eta$ and $m_{2}$ be positive, and $m_{1}$ be non-negative. As $0 \in \rho\left(\mathscr{A}_{\eta}\right)$ and $\rho\left(\mathscr{A}_{\eta}\right)$ is open in $\mathbb{C}$, see Proposition 4.1 and Theorem 1 in [124, p. 211], then there is $R>0$ such that $B(0, R) \subset \rho\left(\mathscr{A}_{\eta}\right)$. For real $\delta$ such that $0<\delta<R$, we have $i[-\delta, \delta] \subset \rho\left(\mathscr{A}_{\eta}\right)$. Thus, $\mathcal{R}:=\left\{\lambda>0: i[-\lambda, \lambda] \subset \rho\left(\mathscr{A}_{\eta}\right)\right\}$ is not empty. Let $\lambda^{*}:=\sup \mathcal{R}$. If $\lambda^{*}=\infty$, we have nothing to prove. Let us suppose $\lambda^{*}<\infty$. If $\lambda^{*} \in \mathcal{R}$, then $i\left[-\lambda^{*}, \lambda^{*}\right]$ is contained in $\rho\left(\mathscr{A}_{\eta}\right)$ and so it is possible to find $r>0$ such that $i\left[-\lambda^{*}-r, \lambda^{*}+r\right] \subset \rho\left(\mathscr{A}_{\eta}\right)$. The above indicates that $\lambda^{*}+r \in \mathcal{R}$ which contradicts the assumption that $\lambda^{*}$ is the supremum of $\mathcal{R}$. Hence, $\lambda^{*} \notin \mathcal{R}$. Then, there exists $\left(\lambda_{n}\right)_{n \in \mathbb{N}} \subset \mathcal{R}$ such that $\lim _{n \rightarrow \infty} \lambda_{n}=\lambda^{*}$ and $\lim _{n \rightarrow \infty}\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right)^{-1}\right\|_{\mathcal{L}\left(\mathscr{H}_{\eta}\right)}=\infty$, see proof of Theorem 4.1 in [105]. Hence, there exists $\left(\widetilde{f}_{n}\right)_{n \in \mathbb{N}} \subset \mathscr{H}_{\eta}$ with

$$
\left\|\widetilde{f}_{n}\right\|_{\mathscr{H}_{n}}=1 \quad \text { and } \quad \lim _{n \rightarrow \infty}\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right)^{-1} \widetilde{f}_{n}\right\|_{\mathscr{H}_{\eta}}=\infty .
$$

Here putting $\widetilde{w}_{n}:=\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right)^{-1} \widetilde{f}_{n}, w_{n}:=\widetilde{w}_{n} /\left\|\widetilde{w}_{n}\right\|_{\mathscr{H}_{n}}$ and $f_{n}:=\widetilde{f}_{n} /\left\|\widetilde{w}_{n}\right\|_{\mathscr{H}_{n}}$, we get that

$$
\begin{equation*}
\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}=f_{n} \tag{4.3}
\end{equation*}
$$

with

$$
\begin{equation*}
\left\|w_{n}\right\|_{\mathscr{H}_{n}}=1 \tag{4.4}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}\right\|_{\mathscr{H}_{n}} \xrightarrow[n \rightarrow \infty]{ } 0 \tag{4.5}
\end{equation*}
$$

It is clear that $\left(w_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathscr{A}_{\eta}\right)$. From (4.3) it follows that

$$
\begin{align*}
i \lambda_{n} w_{1}^{n}-w_{2}^{n} & =f_{1}^{n},  \tag{4.6}\\
i \lambda_{n} \rho_{1} w_{2}^{n}+\mathcal{M}_{\eta}^{-1}\left(\mathcal{W}^{n}+m_{1} \mathcal{A}_{L} w_{2}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n}\right) & =\rho_{1} f_{2}^{n},  \tag{4.7}\\
i \lambda_{n} w_{3}^{n}-w_{4}^{n} & =f_{3}^{n},  \tag{4.8}\\
i \lambda_{n} \rho_{2} w_{4}^{n}-\beta_{2} \Delta w_{3}^{n}-m_{2} \Delta w_{4}^{n} & =\rho_{2} f_{4}^{n},  \tag{4.9}\\
i \lambda_{n} \rho_{0} w_{5}^{n}+\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \mathcal{A}_{T} w_{5}^{n} & =\rho_{0} f_{5}^{n}, \tag{4.10}
\end{align*}
$$

where $\mathcal{W}^{n}:=\beta_{1} \mathcal{A}_{B} w_{1}^{n}+\mathcal{A}_{B} \mathcal{G}_{2} \gamma_{1}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)+\alpha \mathcal{A}_{B} \mathcal{G}_{1} \gamma_{0} w_{5}^{n}-\alpha \kappa \mathcal{A}_{B} \mathcal{G}_{2} \gamma_{0} w_{5}^{n}$. Moreover,

$$
\begin{aligned}
\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}\right\|_{\mathscr{H}_{\eta}}^{2}= & \beta_{1}\left\|f_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+\rho_{1}\left\|f_{2}^{n}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2}+\beta_{2}\left\|\nabla f_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2} \\
& +\rho_{2}\left\|f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\rho_{0}\left\|f_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}
\end{aligned}
$$

The limit (4.5) and (4.6)-(4.10) imply that

$$
\begin{align*}
& i \lambda_{n} w_{1}^{n}-w_{2}^{n} \rightarrow 0 \text { in } H_{\Gamma}^{2}\left(\Omega_{1}\right),  \tag{4.11}\\
& i \lambda_{n} \rho_{1} \mathcal{M}_{\eta} w_{2}^{n}+\mathcal{W}^{n}+m_{1} \mathcal{A}_{L} w_{2}^{n}-\alpha\left(\mathcal{A}_{T}-\frac{\sigma}{\beta}\right) w_{5}^{n} \rightarrow 0 \text { in } H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \text {, }  \tag{4.12}\\
& i \lambda_{n} \nabla w_{3}^{n}-\nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right),  \tag{4.13}\\
& i \lambda_{n} \rho_{2} w_{4}^{n}-\beta_{2} \Delta w_{3}^{n}-m_{2} \Delta w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right),  \tag{4.14}\\
& i \lambda_{n} \rho_{0} w_{5}^{n}+\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \mathcal{A}_{T} w_{5}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) . \tag{4.15}
\end{align*}
$$

Due to the dissipativity (3.35) of the operator $\mathscr{A}_{\eta}$, we have

$$
\begin{align*}
& \operatorname{Re}\left(\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}, w_{n}\right)_{\mathscr{H}_{\eta}}=m_{1}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+m_{2}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}  \tag{4.16}\\
& +\sigma\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta\left\|\nabla w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\beta \kappa\left\|w_{5}^{n}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2} .
\end{align*}
$$

By Theorem 2.24, Cauchy-Schwarz inequality and 4.4, we get the estimate

$$
\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{2} \leq C \operatorname{Re}\left(\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}, w_{n}\right)_{\mathscr{H}_{\eta}} \leq C\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}\right\|_{\mathscr{H}_{n}} .
$$

In consequence,

$$
\begin{equation*}
w_{5}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{1}\right) \tag{4.17}
\end{equation*}
$$

Assuming $m_{1}>0$, from (4.5) and 4.16 we obtain

$$
\begin{equation*}
\nabla w_{2}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) \text { and } \nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) . \tag{4.18}
\end{equation*}
$$

The limits $\lambda_{n} \rightarrow \lambda^{*}$, 4.13) and the right-hand side of (4.18) imply

$$
\begin{equation*}
\nabla w_{3}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.19}
\end{equation*}
$$

On the other hand,

$$
\begin{align*}
& \left(i \lambda_{n} \rho_{2} w_{4}^{n}-\beta_{2} \Delta w_{3}^{n}-m_{2} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)} \\
& =i \lambda_{n} \rho_{2}\left\|w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}-\left(\beta_{2} \Delta w_{3}^{n}+m_{2} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)} \tag{4.20}
\end{align*}
$$

Using integration by parts and the fact that $w_{2}^{n}=w_{4}^{n}$ on $I$, we compute

$$
\begin{align*}
& \left(\beta_{2} \Delta w_{3}^{n}+m_{2} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)}=-\beta_{2}\left(\nabla w_{3}^{n}, \nabla w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& -m_{2}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}-\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}, \tag{4.21}
\end{align*}
$$

by Remark 3.19 we have $\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n} \in H^{2}\left(\Omega_{2}\right)$. Thanks to Cauchy-Schwarz inequality and the trace theorem,

$$
\begin{align*}
& \left|\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}\right| \\
& \quad \leq C\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.22}
\end{align*}
$$

Note that $w_{2}^{n} \in H_{\Gamma}^{1}\left(\Omega_{1}\right)$ and in consequence $\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}$, see Theorem 2.24. By the left-hand side of 4.18),

$$
\begin{equation*}
w_{2}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{1}\right) \tag{4.23}
\end{equation*}
$$

As $\Delta\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)=i \lambda_{n} \rho_{2} w_{4}^{n}-\rho_{2} f_{4}^{n}$ in $\Omega_{2}$ and $\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}=\beta_{2} w_{1}^{n}+m_{2} w_{2}^{n}$ on $I$, see (4.9), Remark 3.17 implies

$$
\begin{align*}
& \left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)} \\
& \leq C\left(\left\|i \lambda_{n} \rho_{2} w_{4}^{n}-\rho_{2} f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|\beta_{2} w_{1}^{n}+m_{2} w_{2}^{n}\right\|_{H^{3 / 2}(I)}\right) \\
& \leq C\left(\left|\lambda_{n}\right|\left\|w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|w_{1}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}+\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}\right) \tag{4.24}
\end{align*}
$$

We know that $f_{4}^{n} \rightarrow 0$ in $L^{2}\left(\Omega_{2}\right)$. By (4.4), the sequences $\left(w_{1}^{n}\right)_{n \in \mathbb{N}}$ and $\left(w_{4}^{n}\right)_{n \in \mathbb{N}}$ are bounded in $H^{2}\left(\Omega_{1}\right)$ and $L^{2}\left(\Omega_{2}\right)$, respectively. The limit (4.11) implies that $\left(w_{2}^{n}\right)_{n \in \mathbb{N}}$ is bounded in $H^{2}\left(\Omega_{1}\right)$. Hence,

$$
\begin{equation*}
\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)} \leq C \tag{4.25}
\end{equation*}
$$

From (4.14), (4.18)-(4.23) and (4.25), it follows that

$$
\begin{equation*}
w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.26}
\end{equation*}
$$

Now, let us consider the following equality

$$
\begin{align*}
& \left\langle i \lambda_{n} \rho_{1} \mathcal{M}_{\eta} w_{2}^{n}+\mathcal{W}^{n}+m_{1} \mathcal{A}_{L} w_{2}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}} \\
& =i \lambda_{n} \rho_{1}\left\langle\mathcal{M}_{\eta} w_{2}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}+\left\langle\mathcal{W}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}  \tag{4.27}\\
& \quad+m_{1}\left\langle\mathcal{A}_{L} w_{2}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}+\left\langle-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}
\end{align*}
$$

where $H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}:=H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)$. By (3.15), we have the following

$$
\begin{equation*}
\left|\left\langle\mathcal{M}_{\eta} w_{2}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}\right|=\left|\left(w_{2}^{n}, w_{1}^{n}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}\right| \leq C\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \tag{4.28}
\end{equation*}
$$

Proceeding as in the proof of Proposition 3.13, we obtain

$$
\begin{align*}
\left\langle\mathcal{W}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}= & \beta_{1}\left\|w_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}-\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{1}^{n}\right)_{L^{2}(I)}  \tag{4.29}\\
& +\alpha\left(w_{5}^{n}, \partial_{\nu} w_{1}^{n}\right)_{L^{2}(I)}+\alpha \kappa\left(w_{5}^{n}, w_{1}^{n}\right)_{L^{2}(I)} .
\end{align*}
$$

Arguing as in (4.22) and considering (4.25),

$$
\begin{equation*}
\left|\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{1}^{n}\right)_{L^{2}(I)}\right| \leq C\left\|w_{1}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.30}
\end{equation*}
$$

The equivalence of norms $\|\cdot\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \sim\|\cdot\|_{H^{2}\left(\Omega_{1}\right)}$ and $\|\cdot\|_{H_{\Gamma}^{1}\left(\Omega_{1}\right)} \sim\|\cdot\|_{H^{1}\left(\Omega_{1}\right)}$ on $H_{\Gamma}^{2}\left(\Omega_{1}\right)$ and $H_{\Gamma}^{1}\left(\Omega_{1}\right)$, respectively, together with Theorem 2.5 allow to obtain immediately the continuous embedding $H_{\Gamma}^{2}\left(\Omega_{1}\right) \hookrightarrow H_{\Gamma}^{1}\left(\Omega_{1}\right)$. Now, the limits (4.11) and 4.23) imply

$$
\begin{equation*}
w_{1}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{1}\right) \tag{4.31}
\end{equation*}
$$

By the trace theorem, we get

$$
\begin{equation*}
\left|\left(w_{5}^{n}, \partial_{\nu} w_{1}^{n}\right)_{L^{2}(I)}\right| \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}\left\|w_{1}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)} \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \tag{4.32}
\end{equation*}
$$

and

$$
\begin{equation*}
\left|\left(w_{5}^{n}, w_{1}^{n}\right)_{L^{2}(I)}\right| \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.33}
\end{equation*}
$$

From (3.14), it follows that

$$
\begin{equation*}
\left|\left\langle\mathcal{A}_{L} w_{2}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}\right|=\left|\left(\nabla w_{2}^{n}, \nabla w_{1}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}\right| \leq C\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} \tag{4.34}
\end{equation*}
$$

Using integration by parts,

$$
\begin{align*}
\left\langle-\mathcal{A}_{T} w_{5}^{n}\right. & \left.+\frac{\sigma}{\beta} w_{5}^{n}, w_{1}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}=\left(\Delta w_{5}^{n}, w_{1}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =-\left(\nabla w_{5}^{n}, \nabla w_{1}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\left(\partial_{\nu} w_{5}^{n}, w_{1}^{n}\right)_{L^{2}\left(\partial \Omega_{1}\right)} . \tag{4.35}
\end{align*}
$$

Applying Cauchy-Schwarz inequality,

$$
\begin{equation*}
\left|\left(\nabla w_{5}^{n}, \nabla w_{1}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}\right| \leq\left\|\nabla w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}\left\|\nabla w_{1}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} . \tag{4.36}
\end{equation*}
$$

As $\partial_{\nu} w_{5}^{n}+\kappa w_{5}^{n}=0$ on $\partial \Omega_{1}$, then

$$
\begin{equation*}
\left|\left(\partial_{\nu} w_{5}^{n}, w_{1}^{n}\right)_{L^{2}\left(\partial \Omega_{1}\right)}\right| \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}\left\|w_{1}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.37}
\end{equation*}
$$

From (4.12), (4.17), (4.23) and (4.27)-(4.37), we can affirm

$$
\begin{equation*}
w_{1}^{n} \rightarrow 0 \text { in } H^{2}\left(\Omega_{1}\right) \tag{4.38}
\end{equation*}
$$

The limits (4.17), (4.19), (4.23), (4.26) and (4.38) imply $w_{n} \rightarrow 0$ in $\mathscr{H}_{\eta}$, this contradicts (4.4). Hence, $\lambda^{*}=\infty$.

Let us now consider $m_{1}=0$. In this case, we initially only have the righthand side of (4.18) and consequently (4.19) holds. Note that 4.20), 4.21) and (4.25) are true here too. The boundedness of the sequence $\left(w_{2}^{n}\right)_{n \in \mathbb{N}}$ in $H^{2}\left(\Omega_{1}\right)$, the trace theorem and Corollary 2.22 allows us to write

$$
\begin{align*}
& \left|\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}\right| \\
& \leq C\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)}^{1 / 2}\left\|\nabla\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \\
& \leq C\left(\left\|\nabla w_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2}+\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2}\right) . \tag{4.39}
\end{align*}
$$

By (4.4), (4.14), right-hand side of (4.18), (4.19)-(4.21) and (4.39) we get

$$
\begin{equation*}
w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.40}
\end{equation*}
$$

From 4.15 and 4.17), it follows that

$$
\alpha \Delta w_{2}^{n}+\beta \Delta w_{5}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right)
$$

The integration by parts formula (2.1) allows the calculation

$$
\begin{aligned}
\left(\alpha \Delta w_{2}^{n}+\beta \Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}= & -\alpha\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\alpha\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \\
& -\beta\left(\nabla w_{5}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\beta\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} .
\end{aligned}
$$

It is easy to see that

$$
\left(\nabla w_{5}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \rightarrow 0 \quad \text { and } \quad\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0
$$

Since $w_{2}^{n}=w_{4}^{n}$ on $I$,

$$
\begin{align*}
\left|\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right| & =\left|\left(\partial_{\nu} w_{2}^{n}, w_{4}^{n}\right)_{L^{2}(I)}\right| \leq\left\|\partial_{\nu} w_{2}^{n}\right\|_{L^{2}(I)}\left\|w_{4}^{n}\right\|_{L^{2}(I)} \\
& \leq C\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}\left\|w_{4}^{n}\right\|_{H^{1 / 2}(I)} \leq C\left\|w_{4}^{n}\right\|_{H^{1}\left(\Omega_{2}\right)} . \tag{4.41}
\end{align*}
$$

The limit of the right part of (4.18), (4.40) and (4.41) imply

$$
\begin{equation*}
\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \tag{4.42}
\end{equation*}
$$

Therefore, $\nabla w_{2}^{n} \rightarrow 0$ in $L^{2}\left(\Omega_{1}\right)$. Reasoning as in the previous case, we obtain again the same contradiction.

Corollary 4.3. Let $\eta>0, m_{1} \geq 0$ and $m_{2}>0$. Then $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ is strongly stable in $\mathscr{H}_{\eta}$.

Remark 4.4. a) In absence of temperature on the plate we also have strong stability when the rotational term is present and the structure is damped.
b) The thermal effect on the plate guarantees the strong stability when we remove the structural damping on the plate and leave the Kelvin-Voigt damping on the membrane and there is presence of the rotational term.
c) With an appropriate geometric condition on $\Omega_{2}$, see (4.99), for the situation $\eta, m_{1}>0$ and $m_{2}=0$ we also have strong stability (see the proof of Theorem 4.16).

Now we will see the second strong stability result of the system (1.1)-1.9) when both the inertial term and the structural damping are not present, but there is Kelvin-Voigt damping on the membrane.

Proposition 4.5. If $\eta=m_{1}=0$ and $m_{2}>0$, then the imaginary axis is contained in the resolvent set of $\mathscr{A}_{0}$, i.e., $i \mathbb{R} \subset \rho\left(\mathscr{A}_{0}\right)$.

Proof. Arguing as in Proposition 4.2, there exists $\left(\lambda_{n}\right)_{n \in \mathbb{N}} \subset \mathcal{R}$ and $\lambda^{*} \in \mathbb{R}^{+}$ such that $\lim _{n \rightarrow \infty} \lambda_{n}=\lambda^{*}$. Moreover, there are sequences $\left(w_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathscr{A}_{0}\right)$ and $\left(f_{n}\right)_{n \in \mathbb{N}} \subset \mathscr{H}_{0}$ such that

$$
\begin{equation*}
\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right) w_{n}=f_{n} \tag{4.43}
\end{equation*}
$$

with

$$
\begin{equation*}
\left\|w_{n}\right\|_{\mathscr{H}_{0}}=1 \quad \text { and } \quad\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right) w_{n}\right\|_{\mathscr{H}_{0}} \xrightarrow[n \rightarrow \infty]{ } 0 \tag{4.44}
\end{equation*}
$$

From (3.34) and (4.43) it follows that

$$
\begin{align*}
i \lambda_{n} w_{1}^{n}-w_{2}^{n} & =f_{1}^{n},  \tag{4.45}\\
i \lambda_{n} \rho_{1} w_{2}^{n}+\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n} & =\rho_{1} f_{2}^{n},  \tag{4.46}\\
i \lambda_{n} w_{3}^{n}-w_{4}^{n} & =f_{3}^{n}  \tag{4.47}\\
i \lambda_{n} \rho_{0} w_{5}^{n}+\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \mathcal{A}_{T} w_{5}^{n} & =\rho_{0} f_{5}^{n}, \tag{4.48}
\end{align*}
$$

where $\mathcal{W}_{0}^{n}:=\beta_{1} w_{1}^{n}+\beta_{2} \mathcal{G}_{2} \gamma_{1} w_{3}^{n}+m_{2} \mathcal{G}_{2} \gamma_{1} w_{4}^{n}+\alpha \mathcal{G}_{1} \gamma_{0} w_{5}^{n}-\alpha \kappa \mathcal{G}_{2} \gamma_{0} w_{5}^{n}$. Notice,

$$
\begin{aligned}
\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right) w_{n}\right\|_{\mathscr{H}_{0}}^{2}= & \beta_{1}\left\|f_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+\rho_{1}\left\|f_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta_{2}\left\|\nabla f_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2} \\
& +\rho_{2}\left\|f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\rho_{0}\left\|f_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} .
\end{aligned}
$$

The limit of (4.44) and (4.45)-(4.48) imply that

$$
\begin{align*}
i \lambda_{n} w_{1}^{n}-w_{2}^{n} \rightarrow 0 & \text { in } H_{\Gamma}^{2}\left(\Omega_{1}\right),  \tag{4.49}\\
i \lambda_{n} \rho_{1} w_{2}^{n}+\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n} \rightarrow 0 & \text { in } L^{2}\left(\Omega_{1}\right),  \tag{4.50}\\
i \lambda_{n} \nabla w_{3}^{n}-\nabla w_{4}^{n} \rightarrow 0 & \text { in } L^{2}\left(\Omega_{2}\right),  \tag{4.51}\\
i \lambda_{n} \rho_{0} w_{5}^{n}+\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \mathcal{A}_{T} w_{5}^{n} \rightarrow 0 & \text { in } L^{2}\left(\Omega_{1}\right) . \tag{4.52}
\end{align*}
$$

Due to the dissipativity (3.35) of the operator $\mathscr{A}_{0}$, we have

$$
\begin{align*}
\operatorname{Re}\left(\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right) w_{n}, w_{n}\right)_{\mathscr{H}_{0}}= & m_{2}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}+\sigma\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} \\
& +\beta\left\|\nabla w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\beta \kappa\left\|w_{5}^{n}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2} . \tag{4.53}
\end{align*}
$$

In consequence,

$$
\begin{equation*}
w_{5}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{1}\right) \tag{4.54}
\end{equation*}
$$

As $m_{2}>0$, from (4.44) and (4.53) we obtain

$$
\begin{equation*}
\nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.55}
\end{equation*}
$$

The limits $\lambda_{n} \rightarrow \lambda^{*}$, 4.51 and 4.55 imply

$$
\begin{equation*}
\nabla w_{3}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.56}
\end{equation*}
$$

Here, we can make use of (4.39), 4.40) and (4.42). Now, we will work on the following equation

$$
\begin{align*}
& \left(i \lambda_{n} \rho_{1} w_{2}^{n}+\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\alpha \sigma \beta^{-1} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =i \lambda_{n} \rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}+\alpha\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \tag{4.57}
\end{align*}
$$

Using integration by parts, Cauchy-Schwarz inequality, $\partial_{\nu} w_{5}^{n}=-\kappa w_{5}^{n}$ on $I$, the trace theorem and (4.54) we get

$$
\begin{equation*}
\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \rightarrow 0 \tag{4.58}
\end{equation*}
$$

Due to Proposition 3.8,

$$
\begin{align*}
\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}= & \beta_{1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}-\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)} \\
& +\alpha\left(w_{5}^{n}, \partial_{\nu} w_{2}^{n}\right)_{L^{2}(I)}+\alpha \kappa\left(w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \tag{4.59}
\end{align*}
$$

Chapter 4. Asymptotic behavior of the solutions of some plate-membrane problems

The trace theorem and (4.54) imply

$$
\begin{equation*}
\left(w_{5}^{n}, \partial_{\nu} w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \quad \text { and } \quad\left(w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \tag{4.60}
\end{equation*}
$$

Joining (4.39), the left-hand side of (4.44), (4.50) and 4.55)-4.60 we obtain the following convergence

$$
\begin{equation*}
i \lambda_{n} \rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta_{1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0 \tag{4.61}
\end{equation*}
$$

From (4.49), it follows that

$$
\begin{equation*}
i \lambda_{n}\left\|w_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}-\left(w_{2}^{n}, w_{1}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0 \tag{4.62}
\end{equation*}
$$

Taking into account (4.54) and adding (4.50) with (4.52), we have

$$
i \lambda_{n} \beta \rho_{1} w_{2}^{n}+\beta \mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha^{2} \Delta w_{2}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right)
$$

and thus

$$
\begin{equation*}
\beta\left[i \lambda_{n} \rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta_{1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}\right]-\alpha^{2}\left(\Delta w_{2}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \rightarrow 0 \tag{4.63}
\end{equation*}
$$

By (4.61) and 4.63,

$$
\left(\Delta w_{2}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \rightarrow 0
$$

and in consequence

$$
\begin{equation*}
\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \tag{4.64}
\end{equation*}
$$

From (4.42) and (4.64), it follows that $\nabla w_{2}^{n} \rightarrow 0$ in $L^{2}\left(\Omega_{1}\right)$. By Friedrichs inequality, we obtain

$$
\begin{equation*}
w_{2}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) \tag{4.65}
\end{equation*}
$$

From (4.61), (4.62) and (4.65), it follows that

$$
\begin{equation*}
w_{1}^{n} \rightarrow 0 \text { in } H^{2}\left(\Omega_{1}\right) \tag{4.66}
\end{equation*}
$$

Because of (4.40), 4.54, 4.56, 4.65) and 4.66 we have $w_{n} \rightarrow 0$ in $\mathscr{H}_{0}$. This contradicts the first assertion in (4.44).
Corollary 4.6. Let $\eta=m_{1}=0$ and $m_{2}>0$. Then $\left(\mathscr{T}_{0}(t)\right)_{t \geq 0}$ is strongly stable in $\mathscr{H}_{0}$.

Remark 4.7. The temperature on the plate played a very important role in obtaining the result of Proposition 4.5.

### 4.2 Exponential stability

The main result of the section is the exponential stability of our problem (1.1)-(1.9) when $\gamma>0$ and we maintain all dampings on the structure. In consequence, for the solutions $w(t)=\mathscr{T}_{\eta}(t) w_{0}$ of this system, we will have that $\|w(t)\|_{\mathscr{H}_{\eta}} \leq C e^{-\delta t}\left\|w_{0}\right\|_{\mathscr{H}_{\eta}}$ for all $t \geq 0$, being $C$ and $\delta$ positive constants. We will use the characterization of Theorem 2.41. The proof will be done by contradiction.

Theorem 4.8. If $\eta>0, m_{1}>0$ and $m_{2}>0$, then the semigroup $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ generated by $\mathscr{A}_{\eta}$ is exponentially stable.

Proof. From Proposition 4.2, we have $i \mathbb{R} \subset \rho\left(\mathscr{A}_{\eta}\right)$. Let us suppose 2.15 is not true. Then, there exists a sequence $\left(\lambda_{n}, w_{n}\right)_{n \in \mathbb{N}} \subset \mathbb{R} \times D\left(\mathscr{A}_{\eta}\right)$ with $\left\|w_{n}\right\|_{\mathscr{H}_{\eta}}=1$ such that

$$
\left\|\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}\right\|_{\mathscr{\mathscr { C } _ { \eta }}} \xrightarrow[n \rightarrow \infty]{ } 0
$$

As the resolvent of $\mathscr{A}_{\eta}$ is holomorphic (see Theorem 1 in [124, p. 211]) and therefore bounded on compact subsets of the imaginary axis, we see that the sequence $\left(\lambda_{n}\right)_{n \in \mathbb{N}}$ is unbounded, thereby we may assume $\lim _{n \rightarrow \infty} \lambda_{n}=\infty$. If $f_{n}:=\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{\eta}\right) w_{n}$, then 4.6)-4.18) hold. Note that in this proof we can make use of (4.20)-(4.24). From (4.13),

$$
i \nabla w_{3}^{n}-\lambda_{n}^{-1} \nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right)
$$

By the right-hand side of (4.18), we obtain

$$
\begin{equation*}
\nabla w_{3}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.67}
\end{equation*}
$$

From (4.11),

$$
i w_{1}^{n}-\lambda_{n}^{-1} w_{2}^{n} \rightarrow 0 \text { in } H^{2}\left(\Omega_{1}\right)
$$

and in consequence $\left(\lambda_{n}^{-1} w_{2}\right)_{n \in \mathbb{N}}$ is bounded in $H^{2}\left(\Omega_{1}\right)$. On the other hand, using (4.24) we get

$$
\begin{equation*}
\lambda_{n}^{-1}\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)} \leq C . \tag{4.68}
\end{equation*}
$$

A combining of (4.14), (4.20)-(4.23) and (4.68) causes that

$$
\begin{equation*}
w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{4.69}
\end{equation*}
$$

From (4.11),

$$
\begin{equation*}
i\left\|w_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}-\lambda_{n}^{-1}\left(w_{2}^{n}, w_{1}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0 . \tag{4.70}
\end{equation*}
$$

Now, let us consider the following equality

$$
\begin{align*}
& \left\langle i \lambda_{n} \rho_{1} \mathcal{M}_{\eta} w_{2}^{n}+\mathcal{W}^{n}+m_{1} \mathcal{A}_{L} w_{2}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\alpha \sigma \beta^{-1} w_{5}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}} \\
& =i \lambda_{n} \rho_{1}\left\langle\mathcal{M}_{\eta} w_{2}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}+\left\langle\mathcal{W}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}  \tag{4.71}\\
& \quad+m_{1}\left\langle\mathcal{A}_{L} w_{2}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}+\left\langle-\alpha \mathcal{A}_{T} w_{5}^{n}+\alpha \sigma \beta^{-1} w_{5}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}} .
\end{align*}
$$

By (3.15),

$$
\begin{equation*}
\left|\left\langle\mathcal{M}_{\eta} w_{2}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}\right|=\left\|w_{2}^{n}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2} \tag{4.72}
\end{equation*}
$$

Proceeding as in the proof of Proposition 3.13,

$$
\begin{align*}
\left\langle\mathcal{W}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1} \times H_{\Gamma, \eta}^{1}}= & \beta_{1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}-\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}  \tag{4.73}\\
& +\alpha\left(w_{5}^{n}, \partial_{\nu} w_{2}^{n}\right)_{L^{2}(I)}+\alpha \kappa\left(w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} .
\end{align*}
$$

Taking into account (4.22) and 4.68),

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}\right| \leq C\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.74}
\end{equation*}
$$

By the trace theorem,

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(w_{5}^{n}, \partial_{\nu} w_{2}^{n}\right)_{L^{2}(I)}\right| \leq C \lambda_{n}^{-1}\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \tag{4.75}
\end{equation*}
$$

and

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right| \leq C \lambda_{n}^{-1}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \tag{4.76}
\end{equation*}
$$

From (3.14),

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left\langle\mathcal{A}_{L} w_{2}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}\right|=\lambda_{n}^{-1}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2} \tag{4.77}
\end{equation*}
$$

Using integration by parts,

$$
\begin{align*}
\left\langle-\mathcal{A}_{T} w_{5}^{n}\right. & \left.+\frac{\sigma}{\beta} w_{5}^{n}, w_{2}^{n}\right\rangle_{H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right) \times H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}=\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =-\left(\nabla w_{5}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\partial \Omega_{1}\right)} . \tag{4.78}
\end{align*}
$$

Applying Cauchy-Schwarz inequality,

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\nabla w_{5}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}\right| \leq \lambda_{n}^{-1}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}\left\|\nabla w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} . \tag{4.79}
\end{equation*}
$$

As $\partial_{\nu} w_{5}^{n}+\kappa w_{5}^{n}=0$ on $\partial \Omega_{1}$, then

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\partial \Omega_{1}\right)}\right| \leq C \lambda_{n}^{-1}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} . \tag{4.80}
\end{equation*}
$$

From (4.12), (4.17), (4.23) and (4.71)-(4.80), we get

$$
\lambda_{n}^{-1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0
$$

and therefore

$$
\begin{equation*}
w_{1}^{n} \rightarrow 0 \text { in } H^{2}\left(\Omega_{1}\right) \tag{4.81}
\end{equation*}
$$

see (4.70). Combining (4.17), (4.23), 4.67), 4.69) and (4.81), we can write $\left\|w_{n}\right\|_{\mathscr{H}_{n}} \rightarrow 0$. This is a contradiction to $\left\|w_{n}\right\|_{\mathscr{H}_{n}}=1$.
Remark 4.9. In the previous proof we took into account the temperature. However, Theorem 4.8 continues to be true when we remove the temperature on the plate. But when we remove the structural damping on the plate and we leave the other conditions of the theorem, we know nothing about exponential stability, even with temperature in the plate.

Remark 4.10. In the next chapter we will see that if $\eta=m_{1}=0$ and $m_{2}>0$, then the semigroup $\left(\mathscr{T}_{0}(t)\right)_{t \geq 0}$ is exponentially stable.

### 4.3 Lack of exponential stability

In this section, we will prove that our system (1.1)-(1.9) is not exponentially stable, if we remove the Kelvin-Voigt damping on the membrane and we leave the rotational term on the plate. Our proof is supported by the following theorem.

Theorem 4.11 (cf. [59, Theorem 3.1]). Let $H_{0}$ be a closed subspace of a Hilbert space $H$. Let $\left(T_{0}(t)\right)_{t \in \mathbb{R}}$ be a unitary group on $H_{0}$ and $(T(t))_{t \geq 0}$ be a $C_{0}$-semigroup over $H$. If the difference $T(t)-T_{0}(t): H_{0} \rightarrow H$ is a compact operator for all $t>0$, then $(T(t))_{t \geq 0}$ is not exponentially stable.

We define $\mathscr{H}:=\{0\} \times\{0\} \times H_{0}^{1}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{2}\right) \times\{0\}$ endowed with the norm $\|\widehat{w}\|_{\mathscr{H}}^{2}:=\beta_{2}\left\|\nabla \widehat{w}_{3}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}+\rho_{2}\left\|\widehat{w}_{4}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}$. Let us consider the system determined by the wave equation with zero Dirichlet boundary condition

$$
\left\{\begin{array}{l}
\rho_{2} \widehat{v}_{t t}-\beta_{2} \Delta \widehat{v}=0 \text { in } \mathbb{R}^{+} \times \Omega_{2},  \tag{4.82}\\
\widehat{v}=0 \text { on } \mathbb{R}^{+} \times I, \\
\widehat{v}(0, \cdot)=\widehat{v}^{0}, \widehat{v}_{t}(0, \cdot)=\widehat{v}^{1} \text { in } \Omega_{2},
\end{array}\right.
$$

where the initial data $\widehat{v}^{0}$ and $\widehat{v}^{1}$ lie in appropiate Hilbert spaces. We introduce the operator

$$
\mathscr{A}_{\mathscr{G}}: D\left(\mathscr{A}_{\mathscr{G}}\right) \subset \mathscr{H} \rightarrow \mathscr{H} \text { given by } \mathscr{A}_{\mathscr{G}}\left(\begin{array}{c}
0 \\
0 \\
\widehat{w}_{3} \\
\widehat{w}_{4} \\
0
\end{array}\right):=\left(\begin{array}{c}
0 \\
0 \\
\widehat{w}_{4} \\
\frac{\beta_{2}}{\rho_{2}} \Delta \widehat{w}_{3} \\
0
\end{array}\right)
$$

with domain $D(\mathscr{A}$ gg $):=\{0\} \times\{0\} \times H^{2}\left(\Omega_{2}\right) \cap H_{0}^{1}\left(\Omega_{2}\right) \times H_{0}^{1}\left(\Omega_{2}\right) \times\{0\}$. Putting $\widehat{w}=\left(\widehat{w}_{j}\right)_{j=1, \ldots, 5}^{\top}:=\left(0,0, \widehat{v}, \widehat{v}_{t}, 0\right)^{\top}$ we have that the system 4.82) can be written as the Cauchy problem

$$
\partial_{t} \widehat{w}(t)=\mathscr{A} \not{\mathscr{G}} \widehat{w}(t)(t>0) \text { with } \widehat{w}(0)=\left(0,0, \widehat{v}^{0}, \widehat{v}^{1}, 0\right)^{\top} .
$$

The operator $i \mathscr{A}_{\mathscr{G}}: D\left(\mathscr{A}_{\mathscr{G}}\right) \rightarrow \mathscr{H}$ is densely defined due to $H^{2}\left(\Omega_{2}\right) \cap H_{0}^{1}\left(\Omega_{2}\right)$ and $H_{0}^{1}\left(\Omega_{2}\right)$ are dense in $H_{0}^{1}\left(\Omega_{2}\right)$ and $L^{2}\left(\Omega_{2}\right)$, respectively, symmetric because $\left(i \mathscr{A}_{\mathscr{G}} \widehat{w}, \widehat{w}\right)_{\mathscr{H}} \in \mathbb{R}$ for any $\widehat{w} \in D\left(\mathscr{A}_{\mathscr{C}}\right)$, see [77, p. 534], and surjective (Remark 3.17 allows to conclude that). Proposição 5.122 in [32] implies that $i \mathscr{A}_{\text {ge }}$ is self-adjoint and therefore $\mathscr{A}_{\mathscr{g}}$ is skew-adjoint. By Theorem 2.50, we obtain that $\mathscr{A}_{\mathscr{G}}$ generates a unitary group $\left(\mathscr{U}_{\mathscr{G}}(t)\right)_{t \in \mathbb{R}}$ on $\mathscr{H}$.

Lemma 4.12. Let $\widehat{v}$ be a sufficiently regular solution of (4.82) with $\widehat{v}^{0} \in$ $H^{1}\left(\Omega_{2}\right)$ and $\widehat{v}^{1} \in L^{2}\left(\Omega_{2}\right)$. For $t>0$ the following estimate holds

$$
\begin{equation*}
\left\|\partial_{\nu} \widehat{v}\right\|_{L^{2}\left((0, t), L^{2}(I)\right)}^{2} \leq C\left(\left\|\widehat{v}^{1}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\left\|\nabla \widehat{v}^{0}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}\right) \tag{4.83}
\end{equation*}
$$

with $C$ being a positive constant that depends on $\rho_{2}, \beta_{2}, \Omega_{2}$ and $t$ but independent of the solution and the initial data.

Proof. Let $t>0$. Multiplying the differential equation in 4.82) by the conjugate of $\widehat{v}_{t}$, integrating over $\Omega_{2}$, employing integration by parts and after taking real part yields

$$
\begin{equation*}
\frac{1}{2} \frac{d}{d s}\left(\rho_{2}\left\|\widehat{v}_{t}(s)\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\beta_{2}\|\nabla \widehat{v}(s)\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}\right)=0 \tag{4.84}
\end{equation*}
$$

for $0 \leq s \leq t$. The initial data in (4.82) and (4.84) imply

$$
\begin{equation*}
\int_{\Omega_{2}} \rho_{2}\left|\widehat{v}_{t}(s)\right|^{2}+\beta_{2}|\nabla \widehat{v}(s)|^{2} d x=\int_{\Omega_{2}} \rho_{2}\left|\widehat{v}^{1}\right|^{2}+\beta_{2}\left|\nabla \widehat{v}^{0}\right|^{2} d x . \tag{4.85}
\end{equation*}
$$

Let $\hbar: \overline{\Omega_{2}} \rightarrow \mathbb{R}^{2}$ be a vector field of class $C^{1}$ in $\overline{\Omega_{2}}$ with $\left.\hbar\right|_{I}=-\nu$. Using the conditions that satisfy $\hbar$ and $\widehat{v}$ on $I$, theorem of divergence and the identity $\hbar \cdot \nabla\left|\widehat{v}_{t}(s)\right|^{2}=\operatorname{div}\left(\hbar\left|\widehat{v}_{t}(s)\right|^{2}\right)-(\operatorname{div} \hbar)\left|\widehat{v}_{t}(s)\right|^{2}$, we obtain

$$
\begin{align*}
\int_{\Omega_{2}} \operatorname{Re}\left(\widehat{v}_{t t}(s) \hbar \cdot \overline{\nabla \widehat{v}(s)}\right) d x= & \frac{1}{2} \int_{\Omega_{2}}(\operatorname{div} \hbar)\left|\widehat{v}_{t}(s)\right|^{2} d x  \tag{4.86}\\
& +\frac{d}{d s} \int_{\Omega_{2}} \operatorname{Re}\left(\widehat{v}_{t}(s) \hbar \cdot \overline{\nabla \widehat{v}(s)}\right) d x
\end{align*}
$$

On the other side, using the convention of summation over repeated indices and integration by parts, we have

$$
\begin{align*}
\int_{\Omega_{2}} \operatorname{Re}(\Delta \widehat{v}(s) \hbar & \cdot \overline{\nabla \widehat{v}(s)}) d x=-\operatorname{Re} \int_{\Omega_{2}}\left(\partial_{j} \widehat{v}(s)\right)\left(\partial_{j} \hbar_{k}\right)\left(\overline{\partial_{k} \widehat{v}(s)}\right) d x \\
& +\frac{1}{2} \int_{\Omega_{2}}(\operatorname{div} \hbar)|\nabla \widehat{v}(s)|^{2} d x+\frac{1}{2} \int_{I}\left|\partial_{\nu} \widehat{v}(s)\right|^{2} d S \tag{4.87}
\end{align*}
$$

Multiplying the differential equation in (4.82) by $\hbar \cdot \bar{\nabla} \widehat{v}$ and using 4.86) together with (4.87), we get

$$
\begin{aligned}
\int_{I}\left|\partial_{\nu} \widehat{v}(s)\right|^{2} d S= & 2 \frac{\rho_{2}}{\beta_{2}} \frac{d}{d s} \int_{\Omega_{2}} \operatorname{Re}\left(\widehat{v}_{t}(s) \hbar \cdot \overline{\nabla \widehat{v}(s)}\right) d x+\frac{\rho_{2}}{\beta_{2}} \int_{\Omega_{2}}(\operatorname{div} \hbar)\left|\widehat{v}_{t}(s)\right|^{2} d x \\
& +2 \operatorname{Re} \int_{\Omega_{2}}\left(\partial_{j} \widehat{v}(s)\right)\left(\partial_{j} \hbar_{k}\right)\left(\overline{\partial_{k} \widehat{v}(s)}\right) d x-\int_{\Omega_{2}}(\operatorname{div} \hbar)|\nabla \widehat{v}(s)|^{2} d x .
\end{aligned}
$$

Note that $|\hbar|,|\operatorname{div} \hbar|$ and $\left|\partial_{j} \hbar_{k}\right|$ are bounded scalar fields on $\overline{\Omega_{2}}$. Integrating above with respect to the variable $s$ from 0 to $t$, and then using the inequalities of Cauchy-Schwarz and Young together with equality (4.85), we deduce (4.83).

We next state and prove a result that will be useful in the proof of the next theorem. Moreover, we take the opportunity to introduce Remark 4.14 which will be important for the next section.

Proposition 4.13. If $\eta>0, m_{1} \geq 0$ and $m_{2}=0$, then the operator $\mathscr{A}_{\eta}$ has compact resolvent.

Proof. Let $\eta>0, m_{1} \geq 0$ and $m_{2}=0$. By Remark 3.19 and Lemma 3.5 in [14], we have the continuous embedding

$$
D\left(\mathscr{A}_{\eta}\right) \hookrightarrow H^{3}\left(\Omega_{1}\right) \times H^{2}\left(\Omega_{1}\right) \times H^{2}\left(\Omega_{2}\right) \times H^{1}\left(\Omega_{2}\right) \times H^{2}\left(\Omega_{1}\right)
$$

Reasoning as in Proposition 3.9 from [14], we get that id : $D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathscr{H}_{\eta}$ is compact. Now, the statement is a consequence of Proposition 5.8 in [48, p. 107].

Remark 4.14. Thanks to the previous proposition and Corollary 1.15 in [48, p. 162], we have that the spectrum $\sigma\left(\mathscr{A}_{\eta}\right)$ is composed only by eigenvalues when $\eta>0, m_{1} \geq 0$ and $m_{2}=0$.

Theorem 4.15. For $\eta>0, m_{1} \geq 0$ and $m_{2}=0$, we have that the system (1.1)-(1.9) does not have exponential decay.

Proof. It is very simple to see that $\mathscr{H}$ is a closed subspace of $\mathscr{H}_{\eta}$. We will show that $\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t): \mathscr{H} \rightarrow \mathscr{H}_{\eta}(t>0)$ is compact. It is enough to prove that $\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t): \mathscr{D} \rightarrow \mathscr{H}_{\eta}$ is compact because $\mathscr{D}:=\{0\}^{2} \times\left[\mathscr{D}\left(\Omega_{2}\right)\right]^{2} \times\{0\}$ is a dense subspace of $\mathscr{H}$. For $w_{0} \in \mathscr{D}$ and $t \geq 0$, we set

$$
\mathscr{E}(t):=\frac{1}{2}\left\|\mathscr{T}_{\eta}(t) w_{0}-\mathscr{U}_{\mathscr{G}}(t) w_{0}\right\|_{\mathscr{H}}^{\eta} .
$$

Note that $D\left(\mathscr{A}_{\eta}\right) \ni w(t):=\mathscr{T}_{\eta}(t) w_{0}$ and $D\left(\mathscr{A}_{\mathscr{G}}\right) \ni \widehat{w}(t):=\mathscr{U}_{\mathscr{G}}(t) w_{0}$ because $w_{0} \in D\left(\mathscr{A}_{\eta}\right) \cap D\left(\mathscr{A}_{\mathscr{G}}\right)$, see Theorem 2.27. By (3.33), we can write

$$
\begin{align*}
\frac{d}{d t} \mathscr{E}(t)= & \operatorname{Re}\left(\mathscr{A}_{\eta} w(t), w(t)\right)_{\mathscr{H}_{\eta}}+\operatorname{Re}\left(\mathscr{A}_{\mathscr{G}} \widehat{w}(t), \widehat{w}(t)\right)_{\mathscr{H}_{\eta}}  \tag{4.88}\\
& -\operatorname{Re}\left(\mathscr{A}_{\eta} w(t), \widehat{w}(t)\right)_{\mathscr{H}_{\eta}}-\operatorname{Re}\left(\mathscr{A}_{\mathscr{G}} \widehat{w}(t), w(t)\right)_{\mathscr{H}_{\eta}} .
\end{align*}
$$

From 3.35 we know $\operatorname{Re}\left(\mathscr{A}_{\eta} w(t), w(t)\right)_{\mathscr{C}_{\eta}} \leq 0$. Using integration by parts and taking into account that $\widehat{w}_{4}(t)=0$ on $I$, we get the following expression

$$
\left(\mathscr{A}_{\mathscr{G}} \widehat{w}(t), \widehat{w}(t)\right)_{\mathscr{H}_{n}}=i 2 \beta_{2} \operatorname{Im}\left(\nabla \widehat{w}_{4}(t), \nabla \widehat{w}_{3}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}}
$$

and thus $\operatorname{Re}\left(\mathscr{A}_{\mathscr{g}} \widehat{w}(t), \widehat{w}(t)\right)_{\mathscr{H}_{\eta}}=0$. From the definition of the operator $\mathscr{A}_{\eta}$, see Subsection 3.1.2, it is immediate that

$$
\left(\mathscr{A}_{\eta} w(t), \widehat{w}(t)\right)_{\mathscr{H}_{\eta}}=\beta_{2}\left(\nabla w_{4}(t), \nabla \widehat{w}_{3}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-\beta_{2}\left(\nabla w_{3}(t), \nabla \widehat{w}_{4}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}} .
$$

Employing integration by parts, we obtain

$$
\begin{aligned}
& \left(\mathscr{A}_{\mathscr{G}} \widehat{w}(t), w(t)\right)_{\mathscr{H}_{n}}=\beta_{2}\left(\nabla \widehat{w}_{4}(t), \nabla w_{3}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& -\beta_{2}\left(\nabla \widehat{w}_{3}(t), \nabla w_{4}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}}-\beta_{2}\left(\partial_{\nu} \widehat{w}_{3}(t), w_{4}(t)\right)_{L^{2}(I)} .
\end{aligned}
$$

The last two equalities produce the following expression

$$
\begin{aligned}
& -\left(\mathscr{A}_{\eta} w(t), \widehat{w}(t)\right)_{\mathscr{H}_{\eta}}-\left(\mathscr{A}_{\mathscr{G}} \widehat{w}(t), w(t)\right)_{\mathscr{H}_{\eta}}=i 2 \beta_{2} \operatorname{Im}\left(\nabla \widehat{w}_{3}(t), \nabla w_{4}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& +i 2 \beta_{2} \operatorname{Im}\left(\nabla w_{3}(t), \nabla \widehat{w}_{4}(t)\right)_{L^{2}\left(\Omega_{2}\right)^{2}}+\beta_{2}\left(\partial_{\nu} \widehat{w}_{3}(t), w_{4}(t)\right)_{L^{2}(I)} .
\end{aligned}
$$

Taking real part in the last equality, inserting this into (4.88) and integrating over $(0, t)$, we obtain the estimate

$$
\begin{equation*}
\mathscr{E}(t) \leq \beta_{2} \operatorname{Re}\left(\partial_{\nu} \widehat{w}_{3}, w_{2}\right)_{L^{2}\left((0, t), L^{2}(I)\right)} \tag{4.89}
\end{equation*}
$$

The equality $\mathscr{E}(0)=0$ was used and the fact that $w_{2}(t)=w_{4}(t)$ on $I$.
Let $\left(w_{0}^{k}\right)_{k \in \mathbb{N}} \subset \mathscr{D}$ be a bounded sequence in $\mathscr{H}$. We must prove that $\left(\mathscr{T}_{\eta}(t) w_{0}^{k}-\mathscr{U}_{\mathscr{G}}(t) w_{0}^{k}\right)_{k \in \mathbb{N}}$ possesses a convergent subsequence in $\mathscr{H}_{\eta}$. This would show the compactness of $\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t): \mathscr{D} \rightarrow \mathscr{H}_{\eta}$. We set $w^{k}(t):=$ $\mathscr{T}_{\eta}(t) w_{0}^{k}$ and $\widehat{w}^{k}(t):=\mathscr{U}_{\mathscr{G}}(t) w_{0}^{k}$. Due to Lemma 4.12,

$$
\begin{equation*}
\left\|\partial_{\nu} \widehat{w}_{3}^{k}\right\|_{L^{2}\left((0, t), L^{2}(I)\right)} \leq C\left\|\widehat{w}^{k}(0)\right\|_{\mathscr{H}}=C\left\|w_{0}^{k}\right\|_{\mathscr{H}} \leq C \quad \text { for all } \quad k \in \mathbb{N} . \tag{4.90}
\end{equation*}
$$

Let $\phi \in \mathscr{Y}:=H_{0}^{4}\left(\Omega_{1}\right) \times H_{0}^{2}\left(\Omega_{1}\right) \times H_{0}^{2}\left(\Omega_{2}\right) \times H_{0}^{1}\left(\Omega_{2}\right) \times H_{0}^{2}\left(\Omega_{1}\right)$. Taking $w \in D\left(\mathscr{A}_{\eta}\right)$ and reasoning as in the proof of Proposition 3.13, we get

$$
\begin{aligned}
\left(\mathscr{A}_{\eta} w, \phi\right)_{\mathscr{H}_{\eta}}= & \beta_{1}\left(w_{2}, \phi_{1}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}-\beta_{1}\left(w_{1}, \phi_{2}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}-m_{1}\left(\nabla w_{2}, \nabla \phi_{2}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& -\alpha\left(w_{5}, \Delta \phi_{2}\right)_{L^{2}\left(\Omega_{1}\right)}-\beta_{2}\left(w_{4}, \Delta \phi_{3}\right)_{L^{2}\left(\Omega_{2}\right)}-\beta_{2}\left(\nabla w_{3}, \nabla \phi_{4}\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& -\alpha\left(\nabla w_{2}, \nabla \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\beta_{1}\left(w_{5}, \Delta \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)}-\sigma\left(w_{5}, \phi_{5}\right)_{L^{2}\left(\Omega_{1}\right)} .
\end{aligned}
$$

Let $\varpi \in H^{2}\left(\Omega_{1}\right)$ with $\varpi=0$ on $\partial \Omega_{1}$. We have that $\partial_{\tau}^{k} \varpi=0$ on $\partial \Omega_{1}$ for any $k \in \mathbb{N}$. Because of Propositions 3C. 7 and 3C. 11 in [87], we obtain

$$
\begin{align*}
& \mathscr{B}_{1} \varpi=\partial_{\nu}^{2} \varpi+\mu(\operatorname{div} \nu) \partial_{\nu} \varpi,  \tag{4.91}\\
& \mathscr{B}_{2} \varpi=\partial_{\nu}^{3} \varpi+(2-\mu) \partial_{\tau}^{2} \partial_{\nu} \varpi+\left[\partial_{\nu}(\operatorname{div} \nu)\right] \partial_{\nu} \varpi+(\operatorname{div} \nu) \partial_{\nu}^{2} \varpi . \tag{4.92}
\end{align*}
$$

Indeed: (4.91) is immediate. Thanks to the part (ii) of Corollary 3C. 10 in [87], we obtain $\partial_{\nu} \partial_{\tau} \varpi=\partial_{\tau} \partial_{\nu} \varpi$ and $\partial_{\nu} \partial_{\tau}^{2} \varpi=\partial_{\tau} \partial_{\nu} \partial_{\tau} \varpi$. Introducing these last equalities in (3C.69) from [87], we get (4.92).

As $\phi_{1} \in H_{0}^{4}\left(\Omega_{1}\right)$, then $\partial_{\nu}^{j} \phi_{1}=0$ on $\partial \Omega_{1}$ for $j=0,1,2,3$ (see Theorem 2.8). From (4.91) and 4.92), it follows that $\mathscr{B}_{1} \phi_{1}=\mathscr{B}_{2} \phi_{1}=0$ on $\partial \Omega_{1}$. So, Proposition 2.17 implies

$$
\left(\phi_{1}, w_{2}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}=\left(\Delta^{2} \phi_{1}, w_{2}\right)_{L^{2}\left(\Omega_{1}\right)} .
$$

Therefore,

$$
\left|\left(\mathscr{A}_{\eta} w, \phi\right)_{\mathscr{H}_{\eta}}\right| \leq C_{\phi}\|w\|_{\mathscr{H}_{\eta}},
$$

where $C_{\phi}:=C\|\phi\|_{\mathscr{y}}$. In consequence, the linear application $\mathbb{A}: D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathbb{C}$ defined by $\mathbb{A} w:=\left(\mathscr{A}_{\eta} w, \phi\right)_{\mathscr{H}_{\eta}}$ is continuous. As $D\left(\mathscr{A}_{\eta}\right)$ is dense in $\mathscr{H}_{\eta}$ there exists a unique extension linear and continuous of $\mathbb{A}$ to $\mathscr{H}_{\eta}$, which is denoted by $\widetilde{\mathbb{A}}: \mathscr{H}_{\eta} \rightarrow \mathbb{C}$. Note that $\phi \in \mathscr{H}_{\eta}^{\prime}$, and for $\widetilde{w} \in D\left(\mathscr{A}_{\eta}\right)$ we have

$$
\langle\widetilde{w}, \widetilde{\mathbb{A}}\rangle_{\mathscr{H}_{\eta} \times \mathscr{H}_{\eta}^{\prime}}=\widetilde{\mathbb{A}} \widetilde{w}=\mathbb{A} \widetilde{w}=\left(\mathscr{A}_{\eta} \widetilde{w}, \phi\right)_{\mathscr{H}_{\eta}}=\left\langle\mathscr{A}_{\eta} \widetilde{w}, \phi\right\rangle_{\mathscr{H}_{\eta} \times \mathscr{H}_{\eta}^{\prime}} .
$$

Thus, $\phi \in D\left(\mathscr{A}_{\eta}^{\prime}\right)$. With this it is shown that $\mathscr{Y} \subset D\left(\mathscr{A}_{\eta}^{\prime}\right)$. We will now use the interpolation-extrapolation scales theory presented in Section 2.5. Let $X_{0}:=\mathscr{H}_{\eta}, X_{-1}:=\left(X_{0},\left\|\mathscr{A}_{\eta}^{-1} \cdot\right\|_{\mathscr{H}_{\eta}}\right)^{\sim}$ and $X_{-\widetilde{\alpha}}:=\left[X_{0}, X_{-1}\right]_{\widetilde{\alpha}}$ for $\widetilde{\alpha} \in(0,1)$, where $[\cdot, \cdot]_{\tilde{\alpha}}$ stands for the complex interpolation functor. We have that

$$
X_{0} \stackrel{c}{\hookrightarrow} X_{-\widetilde{\alpha}} \hookrightarrow X_{-1}
$$

for $\widetilde{\alpha} \in(0,1)$. The embedding above is compact because $\mathscr{A}_{\eta}^{-1}: \mathscr{H}_{\eta} \rightarrow \mathscr{H}_{\eta}$ is a compact operator, see Proposition 4.13. As $X_{-1}=\left[D\left(\mathscr{A}_{\eta}^{\prime}\right)\right]^{\prime}$, we can insure that the embedding

$$
\begin{equation*}
X_{-1} \hookrightarrow H^{-4}\left(\Omega_{1}\right) \times H^{-2}\left(\Omega_{1}\right) \times H^{-2}\left(\Omega_{2}\right) \times H^{-1}\left(\Omega_{2}\right) \times H^{-2}\left(\Omega_{1}\right) \tag{4.93}
\end{equation*}
$$

holds and since from the definition of $\mathscr{H}_{\eta}$ we have

$$
\begin{equation*}
X_{0} \hookrightarrow H^{2}\left(\Omega_{1}\right) \times H^{1}\left(\Omega_{1}\right) \times H^{1}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{2}\right) \times L^{2}\left(\Omega_{1}\right) \tag{4.94}
\end{equation*}
$$

then (4.93), (4.94) and Theorem 2.4 in [89] imply that

$$
\begin{equation*}
X_{-\widetilde{\alpha}} \hookrightarrow H^{2-5 \widetilde{\alpha}}\left(\Omega_{1}\right) \times H^{1-3 \widetilde{\alpha}}\left(\Omega_{1}\right) \times H^{1-3 \widetilde{\alpha}}\left(\Omega_{2}\right) \times H^{-\widetilde{\alpha}}\left(\Omega_{2}\right) \times H^{-2 \widetilde{\alpha}}\left(\Omega_{1}\right) \tag{4.95}
\end{equation*}
$$

Let $t>0$ be fixed. For $s \in[0, t]$ we have that $\left\|\mathscr{T}_{\eta}(s)\right\|_{\mathcal{L}\left(\mathscr{H}_{\eta}\right)} \leq 1$. Therefore,

$$
\left\|w^{k}\right\|_{L^{2}\left((0, t), \mathscr{H}_{\eta}\right)}^{2}=\int_{0}^{t}\left\|w^{k}(s)\right\|_{\mathscr{H}_{\eta}}^{2} d s \leq \int_{0}^{t}\left\|\mathscr{T}_{\eta}(s)\right\|_{\mathcal{L}\left(\mathscr{H}_{\eta}\right)}^{2}\left\|w_{0}^{k}\right\|_{\mathscr{H}_{\eta}}^{2} d s \leq C .
$$

Thus, the sequence $\left(w^{k}\right)_{k \in \mathbb{N}}$ is bounded in the Hilbert space $L^{2}\left((0, t), \mathscr{H}_{\eta}\right)$. Now, we use $\partial_{t} w^{k}=\mathscr{A}_{\eta} w^{k}$ to get the following

$$
\sup _{s \in[0, t]}\left\|\partial_{s} w^{k}(s)\right\|_{X_{-1}}=\sup _{s \in[0, t]}\left\|\mathscr{A}_{\eta} w^{k}(s)\right\|_{X_{-1}}=\sup _{s \in[0, t]}\left\|w^{k}(s)\right\|_{\mathscr{H}_{\eta}} \leq C
$$

So, $\left(\partial_{t} w^{k}\right)_{k \in \mathbb{N}}$ is bounded in $L^{2}\left((0, t), X_{-1}\right)$. We have proven that the sequence $\left(w^{k}\right)_{k \in \mathbb{N}}$ is bounded in the space $\mathcal{W}_{2,2}\left((0, t) ; X_{0}, X_{-1}\right)$. By Aubin-Lions-Simon lemma, see Theorem 2.2, yields that there exists a subsequence $\left(w^{k_{j}}\right)_{j \in \mathbb{N}}$ of $\left(w^{k}\right)_{k \in \mathbb{N}}$, which is convergent in $L^{2}\left((0, t), X_{-\widetilde{\alpha}}\right)$. From (4.95) we see that the second component $\left(w_{2}^{k_{j}}\right)_{j \in \mathbb{N}}$ converges in $L^{2}\left((0, t), H^{1-3 \tilde{\alpha}}\left(\Omega_{1}\right)\right)$. Choosing $\widetilde{\alpha}<\frac{1}{6}$ and taking the trace on $I$ we obtain convergence in $L^{2}\left((0, t), H^{\frac{1}{2}-3 \widetilde{\alpha}}(I)\right)$ and therefore in $L^{2}\left((0, t), L^{2}(I)\right)$ for the subsequence $\left(w_{2}^{k_{j}}\right)_{j \in \mathbb{N}}$ of $\left(\left.w_{2}^{k}\right|_{I}\right)_{k \in \mathbb{N}}$.

Note $\left(\left(\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t)\right) w_{0}^{k_{j}}\right)_{j \in \mathbb{N}}$ is a subsequence of $\left(\mathscr{T}_{\eta}(t) w_{0}^{k}-\mathscr{U}_{\mathscr{G}}(t) w_{0}^{k}\right)_{k \in \mathbb{N}}$. We write $a_{n m}:=a_{n}-a_{m}$ for any sequence $\left(a_{n}\right)_{n \in \mathbb{N}}$. With this notation, we have

$$
\begin{equation*}
w^{k_{i} k_{j}}(t)-\widehat{w}^{k_{i} k_{j}}(t)=\mathscr{T}_{\eta}(t)\left(w_{0}^{k_{i}}-w_{0}^{k_{j}}\right)-\mathscr{U}_{\mathscr{G}}(t)\left(w_{0}^{k_{i}}-w_{0}^{k_{j}}\right) \tag{4.96}
\end{equation*}
$$

For $i, j \in \mathbb{N}$ and $t \geq 0$, we now consider

$$
\begin{equation*}
\mathscr{E}^{i j}(t):=\frac{1}{2}\left\|w^{k_{i} k_{j}}(t)-\widehat{w}^{k_{i} k_{j}}(t)\right\|_{\mathscr{H}_{\eta}}^{2} \tag{4.97}
\end{equation*}
$$

By (4.89) and 4.90), we compute that

$$
\begin{align*}
\mathscr{E}^{i j}(t) & \leq \beta_{2}\left|\left(\partial_{\nu} \widehat{w}_{3}^{k_{i} k_{j}}, w_{2}^{k_{i} k_{j}}\right)_{L^{2}\left((0, t), L^{2}(I)\right)}\right| \\
& \leq \beta_{2}\left\|\partial_{\nu} \widehat{w}_{3}^{k_{k} k_{j}}\right\|_{L^{2}\left((0, t), L^{2}(I)\right)}\left\|w_{2}^{k_{i} k_{j}}\right\|_{L^{2}\left((0, t), L^{2}(I)\right)} \\
& \leq C\left\|w_{2}^{k_{i}}-w_{2}^{k_{j}}\right\|_{L^{2}\left((0, t), L^{2}(I)\right)} \rightarrow 0 \quad(\text { as } i, j \rightarrow \infty) . \tag{4.98}
\end{align*}
$$

Thanks to (4.96)-4.98), we get that $\left(\left(\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t)\right) w_{0}^{k_{j}}\right)_{j \in \mathbb{N}}$ is a Cauchy sequence in $\mathscr{H}_{\eta}$ and therefore converges in this Hilbert space. Accordingly, $\mathscr{T}_{\eta}(t)-\mathscr{U}_{\mathscr{G}}(t)$ is a compact operator from $\mathscr{H}$ to $\mathscr{H}_{\eta}$ for any $t>0$. In consequence, Theorem 4.11 leads to the conclusion of the present theorem.

### 4.4 Polynomial stability

In this section we will prove that our system (1.1)-(1.9) is polynomially stable under certain geometric condition, imposed on the domain $\Omega_{2}$, when we do not have the Kelvin-Voigt damping on the membrane, but the inertial term and structural damping are present. For the proof, we will use the characterization of Theorem 2.44.

Chapter 4. Asymptotic behavior of the solutions of some plate-membrane problems

In order to establish the polynomial stability, we need the following usual geometrical condition: There exists a point $x_{0} \in \mathbb{R}^{2}$ such that

$$
\begin{equation*}
q(x) \cdot \nu(x) \leq 0 \quad \text { for } \quad x \in I \tag{4.99}
\end{equation*}
$$

where $q$ is the vector field defined by $q(x):=x-x_{0}$ for $x \in \overline{\Omega_{2}}$.
Theorem 4.16. Let $\eta>0, m_{1}>0$ and $m_{2}=0$ and assume that 4.99) is satiesfied. Then, the semigroup $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ decays polynomially of order at least $1 / 25$. Furthermore, if $w_{0} \in D\left(\mathscr{A}_{\eta}^{k}\right), k \in \mathbb{N}$, then there exists a constant $C_{k}>0$ such that

$$
\begin{equation*}
\left\|\mathscr{T}_{\eta}(t) w_{0}\right\|_{\mathscr{H}_{\eta}} \leq \frac{C_{k}}{t^{k / 25}}\left|w_{0}\right|_{D\left(\mathscr{A}_{\eta}^{k}\right)} \tag{4.100}
\end{equation*}
$$

for all $t>0$.
Proof. First we will prove that the intersection $i \mathbb{R} \cap \sigma\left(\mathscr{A}_{\eta}\right)$ is empty, then we will look for an estimate of the type (2.18) and this will allow us to obtain an inequality like 2.19). Finally, a standard argument leads to 4.100).

Since $0 \in \rho\left(\mathscr{A}_{\eta}\right)$ there is $\lambda_{0}>0$ such that $\left\{i \zeta:-\lambda_{0}<\zeta<\lambda_{0}\right\} \subset \rho\left(\mathscr{A}_{\eta}\right)$. Let $\lambda \in \mathbb{R}$ with $|\lambda| \geq \lambda_{0}$. To prove that $i \mathbb{R} \cap \sigma\left(\mathscr{A}_{\eta}\right)=\emptyset$ it is sufficient to show that $i \lambda \notin \sigma\left(\mathscr{A}_{\eta}\right)$. Taking $w=\left(w_{1}, w_{2}, w_{3}, w_{4}, w_{5}\right)^{\top} \in D\left(\mathscr{A}_{\eta}^{2}\right)$ and setting

$$
\begin{equation*}
f:=\left(i \lambda \mathcal{I}-\mathscr{A}_{\eta}\right) w \tag{4.101}
\end{equation*}
$$

we derive that $f \in D\left(\mathscr{A}_{\eta}\right)$ and further

$$
\begin{align*}
i \lambda w_{1}-w_{2}=f_{1} & \text { in } H_{\Gamma}^{2}\left(\Omega_{1}\right)  \tag{4.102}\\
i \lambda \rho_{1} \mathcal{M}_{\eta} w_{2}+\mathcal{W}+m_{1} \mathcal{A}_{L} w_{2}+\alpha \Delta w_{5}=\rho_{1} \mathcal{M}_{\eta} f_{2} & \text { in } H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right),  \tag{4.103}\\
i \lambda w_{3}-w_{4}=f_{3} & \text { in } H^{1}\left(\Omega_{2}\right),  \tag{4.104}\\
i \lambda \rho_{2} w_{4}-\beta_{2} \Delta w_{3}=\rho_{2} f_{4} & \text { in } L^{2}\left(\Omega_{2}\right),  \tag{4.105}\\
i \lambda \rho_{0} w_{5}+\alpha \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}+\beta \mathcal{A}_{T} w_{5}=\rho_{0} f_{5} & \text { in } L^{2}\left(\Omega_{1}\right) \tag{4.106}
\end{align*}
$$

Here we set $f:=\left(f_{1}, f_{2}, f_{3}, f_{4}, f_{5}\right)^{\top}$. From 4.101), we get the following

$$
\begin{equation*}
-\operatorname{Re}\left(\mathscr{A}_{\eta} w, w\right)_{\mathscr{H}_{\eta}} \leq\|f\|_{\mathscr{H}_{\eta}}\|w\|_{\mathscr{H}_{\eta}} \tag{4.107}
\end{equation*}
$$

By (3.35) and (4.107),

$$
\begin{equation*}
\left\|w_{2}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \leq C\|f\|_{\mathscr{H}_{\eta}}^{1 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2} \tag{4.108}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|w_{5}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C\|f\|_{\mathscr{H}_{\eta}}^{1 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2} . \tag{4.109}
\end{equation*}
$$

Replacing (4.102) in 4.103) and 4.106), we obtain

$$
\begin{align*}
& -\lambda^{2} \rho_{1} \mathcal{M}_{\eta} w_{1}+\mathcal{W}+i \lambda m_{1} \mathcal{A}_{L} w_{1}-\alpha \mathcal{A}_{T} w_{5}+\alpha \beta^{-1} \sigma w_{5}  \tag{4.110}\\
& =i \lambda \rho_{1} \mathcal{M}_{\eta} f_{1}+m_{1} \mathcal{A}_{L} f_{1}+\rho_{1} \mathcal{M}_{\eta} f_{2}
\end{align*}
$$

and

$$
\begin{equation*}
i \lambda \rho_{0} w_{5}-i \lambda \alpha \Delta w_{1}+\beta \mathcal{A}_{T} w_{5}=-\alpha \Delta f_{1}+\rho_{0} f_{5} \tag{4.111}
\end{equation*}
$$

Inserting (4.104) into (4.105) leads to

$$
\begin{equation*}
-\lambda^{2} \rho_{2} w_{3}-\beta_{2} \Delta w_{3}=i \lambda \rho_{2} f_{3}+\rho_{2} f_{4} . \tag{4.112}
\end{equation*}
$$

Multiplying 4.110) by $w_{1}$, under the duality between $H_{\Gamma, \eta}^{-1}\left(\Omega_{1}\right)$ and $H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)$, considering the identities (3.12)-(3.15) and (3.21), using integration by parts and knowing that $\partial_{\nu} w_{5}+\kappa w_{5}=0$ on $I$, we get

$$
\begin{align*}
& -\lambda^{2} \rho_{1}\left\|w_{1}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2}+\beta_{1}\left\|w_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}-\beta_{2}\left(\partial_{\nu} w_{3}, w_{1}\right)_{L^{2}(I)} \\
& +\alpha\left(w_{5}, \partial_{\nu} w_{1}\right)_{L^{2}(I)}+i \lambda m_{1}\left\|\nabla w_{1}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\alpha\left(\nabla w_{5}, \nabla w_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}  \tag{4.113}\\
& =i \lambda \rho_{1}\left(f_{1}, w_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+m_{1}\left(\nabla f_{1}, \nabla w_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\rho_{1}\left(f_{2}, w_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} .
\end{align*}
$$

Multiplying (4.111) by $w_{5}$, using integration by parts, (3.16) and (3.18) we get the following equality

$$
\begin{align*}
& i \lambda \rho_{0}\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+i \lambda \alpha\left(\nabla w_{1}, \nabla w_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& -i \lambda \alpha\left(\partial_{\nu} w_{1}, w_{5}\right)_{L^{2}(I)}+\beta\left\|w_{5}\right\|_{D\left(\mathcal{A}_{T}^{1 / 2}\right)}^{2}=\alpha\left(\nabla f_{1}, \nabla w_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}  \tag{4.114}\\
& -\alpha\left(\partial_{\nu} f_{1}, w_{5}\right)_{L^{2}(I)}+\rho_{0}\left(f_{5}, w_{5}\right)_{L^{2}\left(\Omega_{1}\right)}
\end{align*}
$$

Multiplying (4.112) by $w_{3}$ and using integration by parts, we obtain

$$
\begin{align*}
& -\lambda^{2} \rho_{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\beta_{2}\left\|\nabla w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}+\beta_{2}\left(\partial_{\nu} w_{3}, w_{3}\right)_{L^{2}(I)}  \tag{4.115}\\
& =i \lambda \rho_{2}\left(f_{3}, w_{3}\right)_{L^{2}\left(\Omega_{2}\right)}+\rho_{2}\left(f_{4}, w_{3}\right)_{L^{2}\left(\Omega_{2}\right)}
\end{align*}
$$

Multiplying (4.114) by $-i \lambda^{-1}$, adding together with (4.113) and (4.115), and taking into account that $w_{1}=w_{3}$ on $I$, we see

$$
-\lambda^{2} \rho_{1}\left\|w_{1}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2}+\beta_{1}\left\|w_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+i 2 \alpha \operatorname{Im}\left(w_{5}, \partial_{\nu} w_{1}\right)_{L^{2}(I)}
$$

$$
\begin{aligned}
& +i \lambda m_{1}\left\|\nabla w_{1}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+i 2 \alpha \operatorname{Im}\left(\nabla w_{1}, \nabla w_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\rho_{0}\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} \\
& -i \lambda^{-1} \beta\left\|w_{5}\right\|_{D\left(\mathcal{A}_{T}^{1 / 2}\right)}^{2}-\lambda^{2} \rho_{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\beta_{2}\left\|\nabla w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2} \\
& =i \lambda \rho_{1}\left(f_{1}, w_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}+m_{1}\left(\nabla f_{1}, \nabla w_{1}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\rho_{1}\left(f_{2}, w_{1}\right)_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)} \\
& -i \lambda^{-1} \alpha\left(\nabla f_{1}, \nabla w_{5}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+i \lambda^{-1} \alpha\left(\partial_{\nu} f_{1}, w_{5}\right)_{L^{2}(I)}-i \lambda^{-1} \rho_{0}\left(f_{5}, w_{5}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& +i \lambda \rho_{2}\left(f_{3}, w_{3}\right)_{L^{2}\left(\Omega_{2}\right)}+\rho_{2}\left(f_{4}, w_{3}\right)_{L^{2}\left(\Omega_{2}\right)}
\end{aligned}
$$

Because of Friedrichs's inequality there exists a constant $C>0$ such that $\left\|\Phi_{3}\right\|_{L^{2}\left(\Omega_{2}\right)} \leq C\|\Phi\|_{\mathscr{H}_{\eta}}$ for any $\Phi=\left(\Phi_{1}, \Phi_{2}, \Phi_{3}, \Phi_{4}, \Phi_{5}\right) \in \mathscr{H}_{\eta}$. Taking real part above, using (4.109) and remembering that $1 \leq \lambda_{0}^{-1}|\lambda|$ we achieve

$$
\begin{align*}
& \beta_{1}\left\|w_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+\beta_{2}\left\|\nabla w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}+\rho_{0}\left\|w_{5}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} \leq \lambda^{2} \rho_{1}\left\|w_{1}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2} \\
& +\lambda^{2} \rho_{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+C\left(|\lambda|\|f\|_{\mathscr{H}_{\eta}}\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}^{3 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2}\right) \tag{4.116}
\end{align*}
$$

By (4.104), we find

$$
\begin{equation*}
\rho_{2}\left\|w_{4}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}=\rho_{2}\left\|i \lambda w_{3}-f_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2} \leq C\left(\lambda^{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\|f\|_{\mathscr{H}_{\eta}}^{2}\right) \tag{4.117}
\end{equation*}
$$

By 4.102 and 4.108, we derive

$$
\begin{equation*}
\lambda^{2}\left\|w_{1}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2}=\left\|f_{1}+w_{2}\right\|_{H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)}^{2} \leq C\left(\|f\|_{\mathscr{H}_{\eta}}^{2}+\|f\|_{\mathscr{H}_{\eta}}\|w\|_{\mathscr{H}_{\eta}}\right) \tag{4.118}
\end{equation*}
$$

Combining (4.108) and (4.116)-4.118), we obtain

$$
\begin{equation*}
\|w\|_{\mathscr{H}_{\eta}}^{2} \leq C\left(\lambda^{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+|\lambda|\|f\|_{\mathscr{H}_{\eta}}\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}^{3 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2}+\|f\|_{\mathscr{H}_{\eta}}^{2}\right) \tag{4.119}
\end{equation*}
$$

Now, we will establish an estimate for $\lambda^{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}$. Thanks to the Rellich identity, see Theorem 2.16, choosing there $h=q$, we can write

$$
\begin{equation*}
\operatorname{Re} \int_{\Omega_{2}} \Delta w_{3}\left(q \cdot \nabla \bar{w}_{3}\right) d x=-\operatorname{Re} \int_{I} \partial_{\nu} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right)-\frac{1}{2}(q \cdot \nu)\left|\nabla w_{3}\right|^{2} d S \tag{4.120}
\end{equation*}
$$

Multiplying $(4.112)$ by the scalar field $q \cdot \nabla \bar{w}_{3}$, integrating, taking real part and using 4.120 we get

$$
\begin{align*}
& -\lambda^{2} \rho_{2} \operatorname{Re} \int_{\Omega_{2}} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right) d x+\beta_{2} \operatorname{Re} \int_{I} \partial_{\nu} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right) d S \\
& -\frac{\beta_{2}}{2} \int_{I}(q \cdot \nu)\left|\nabla w_{3}\right|^{2} d S=\operatorname{Re} \int_{\Omega_{2}}\left(i \lambda \rho_{2} f_{3}+\rho_{2} f_{4}\right)\left(q \cdot \nabla \bar{w}_{3}\right) d x \tag{4.121}
\end{align*}
$$

Making use of identity $q \cdot \nabla w_{3}=\operatorname{div}\left(q w_{3}\right)-2 w_{3}$ and employing integrating by parts, it holds

$$
\begin{equation*}
\operatorname{Re} \int_{\Omega_{2}} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right) d x=-\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}-\frac{1}{2} \int_{I}(q \cdot \nu)\left|w_{3}\right|^{2} d S . \tag{4.122}
\end{equation*}
$$

By (4.121) and 4.122,

$$
\begin{aligned}
\lambda^{2} \rho_{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}= & \operatorname{Re} \int_{\Omega_{2}}\left(i \lambda \rho_{2} f_{3}+\rho_{2} f_{4}\right)\left(q \cdot \nabla \bar{w}_{3}\right) d x-\frac{1}{2} \lambda^{2} \rho_{2} \int_{I}(q \cdot \nu)\left|w_{3}\right|^{2} d S \\
& -\beta_{2} \operatorname{Re} \int_{I} \partial_{\nu} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right) d S+\frac{1}{2} \beta_{2} \int_{I}(q \cdot \nu)\left|\nabla w_{3}\right|^{2} d S
\end{aligned}
$$

Keeping in mind that $q \cdot \nu \leq 0$ on $I, w_{1}=w_{3}$ on $I$ and 4.118), we obtain

$$
\begin{equation*}
\lambda^{2}\left\|w_{3}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2} \leq C\left(|\lambda|\|f\|_{\mathscr{H} \mathscr{H}_{\eta}}\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}^{2}+\int_{I}\left|\beta_{2} \partial_{\nu} w_{3}\left(q \cdot \nabla \bar{w}_{3}\right)\right| d S\right) . \tag{4.123}
\end{equation*}
$$

Next we are going to estimate the last integral denoted this by $\mathscr{\mathscr { I }}$. Let $\widetilde{f}=$ $\left(\widetilde{f}_{1}, \widetilde{f}_{2}, \widetilde{f}_{3}, \widetilde{f}_{4}, \widetilde{f}_{5}\right)$ and $\widetilde{f}:=\mathscr{A}_{\eta} w$. Since $w \in D\left(\mathscr{A}_{\eta}^{2}\right)$, we have that $\widetilde{f}_{2} \in H_{\Gamma}^{2}\left(\Omega_{1}\right)$ and also

$$
\begin{equation*}
\beta_{1} w_{1}-m_{1} \mathcal{G}_{2} \gamma_{1} w_{2}+\beta_{2} \mathcal{G}_{2} \gamma_{1} w_{3}+\alpha\left(\mathcal{G}_{1}-\kappa \mathcal{G}_{2}\right) \gamma_{0} w_{5}-\gamma \mathcal{G}_{2} \gamma_{1} \widetilde{f}_{2} \in D\left(\mathcal{A}_{B}\right) \tag{4.124}
\end{equation*}
$$

see the proof of Theorem 3.18. In consequence,

$$
\beta_{2} \partial_{\nu} w_{3}=\gamma \partial_{\nu} \widetilde{f}_{2}-\beta_{1} \mathscr{B}_{2} w_{1}+m_{1} \partial_{\nu} w_{2}+\alpha \kappa w_{5} \quad \text { on } I .
$$

By (4.101), $i \lambda w-\tilde{f}=f$. Thus, $\tilde{f}_{2}=i \lambda w_{2}-f_{2}$. Therefore,

$$
\begin{aligned}
\mathscr{I} & =\int_{I}\left|\beta_{2} \partial_{\nu} w_{3}\right|\left|q \cdot \nabla \bar{w}_{3}\right| d S \leq C\left\|\beta_{2} \partial_{\nu} w_{3}\right\|_{L^{2}(I)}\left\|\nabla w_{3}\right\|_{L^{2}(I)^{2}} \\
& \leq C\left(|\lambda|\left\|\partial_{\nu} w_{2}\right\|_{L^{2}(I)}+\left\|\partial_{\nu} f_{2}\right\|_{L^{2}(I)}+\left\|\mathscr{B}_{2} w_{1}\right\|_{L^{2}(I)}+\left\|w_{5}\right\|_{L^{2}(I)}\right)\left\|\nabla w_{3}\right\|_{L^{2}(I)^{2}}
\end{aligned}
$$

From (4.105), $\Delta w_{3}=i \lambda \beta_{2}^{-1} \rho_{2} w_{4}-\beta_{2}^{-1} \rho_{2} f_{4}$ in $\Omega_{2}$ and as $w_{3}=w_{1}$ on $I$, then by Remark 3.17 and the trace theorem we get

$$
\begin{align*}
\left\|w_{3}\right\|_{H^{2}\left(\Omega_{2}\right)} & \leq C\left(\left\|i \lambda \beta_{2}^{-1} \rho_{2} w_{4}-\beta_{2}^{-1} \rho_{2} f_{4}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|w_{1}\right\|_{H^{3 / 2}(I)}\right) \\
& \leq C\left(|\lambda|\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}\right) \tag{4.125}
\end{align*}
$$

Applying Theorem 2.21 to the partial derivatives of first order of $w_{3}$, we obtain

$$
\begin{align*}
\left\|\nabla w_{3}\right\|_{L^{2}(I)^{2}} & \leq C\left\|w_{3}\right\|_{H^{2}\left(\Omega_{2}\right)}^{1 / 2}\left\|w_{3}\right\|_{H^{1}\left(\Omega_{2}\right)}^{1 / 2} \\
& \leq C\left(|\lambda|^{1 / 2}\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}^{1 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2}\right) . \tag{4.126}
\end{align*}
$$

By (4.102),

$$
\begin{equation*}
\left\|w_{2}\right\|_{H^{2}\left(\Omega_{1}\right)} \leq C\left\|i \lambda w_{1}-f_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \leq C\left(|\lambda|\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}\right) \tag{4.127}
\end{equation*}
$$

Applying Corollary 2.22 to $w_{2}$ and considering (4.108) together with 4.127), we get

$$
\begin{align*}
|\lambda|\left\|\partial_{\nu} w_{2}\right\|_{L^{2}(I)} & \leq C|\lambda|\left\|w_{2}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2}\left\|w_{2}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2} \\
& \leq C\left(|\lambda|^{3 / 2}\|f\|_{\mathscr{H}_{\eta}}^{1 / 4}\|w\|_{\mathscr{H}_{\eta}}^{3 / 4}+|\lambda|\|f\|_{\mathscr{H}_{\eta}}^{3 / 4}\|w\|_{\mathscr{H}_{\eta}}^{1 / 4}\right) . \tag{4.128}
\end{align*}
$$

Doing $\widehat{f}:=\mathscr{A}_{\eta} f$ with $\widehat{f}=\left(\widehat{f_{1}}, \widehat{f_{2}}, \widehat{f_{3}}, \widehat{f_{4}}, \widehat{f_{5}}\right)$, we have that $\widehat{f}_{1}=f_{2}$ and by the trace theorem

$$
\begin{equation*}
\left\|\partial_{\nu} f_{2}\right\|_{L^{2}(I)} \leq C\left\|f_{2}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}=C\left\|\widehat{f}_{1}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \leq C\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}} \tag{4.129}
\end{equation*}
$$

By (3C.53) in [87], we can write $B_{2} w_{1}=\partial_{\nu} \partial_{\tau} w_{1}$. Applying Corollary 2.22 to $\Delta w_{1}$ and Theorem 2.21 to the partial derivatives of third order of $w_{1}$, and then using Sobolev's interpolation inequality, we deduce that

$$
\begin{align*}
\left\|\mathscr{B}_{2} w_{1}\right\|_{L^{2}(I)} & \leq C\left(\left\|\partial_{\nu} \Delta w_{1}\right\|_{L^{2}(I)}+\left\|\partial_{\tau} \partial_{\nu} \partial_{\tau} w_{1}\right\|_{L^{2}(I)}\right) \\
& \leq C\left(\left\|\Delta w_{1}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2}\left\|\nabla \Delta w_{1}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{1 / 2}+\left\|w_{1}\right\|_{H^{4}\left(\Omega_{1}\right)}^{1 / 2}\left\|w_{1}\right\|_{H^{3}\left(\Omega_{1}\right)}^{1 / 2}\right) \\
& \leq C\left\|w_{1}\right\|_{H^{4}\left(\Omega_{1}\right)}^{5 / 6}\left\|w_{1}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 6} . \tag{4.130}
\end{align*}
$$

Using the definitions of $\mathcal{M}_{\eta}, \mathcal{W}, \mathcal{A}_{T}, \mathcal{A}_{B},(3.30),(3.31)$, Proposition 3.11 and (4.124) we can establish that equation (4.103) is equivalent to the following system

$$
\left\{\begin{array}{l}
\Delta^{2} w_{1}=\beta_{1}^{-1}\left(\rho_{1} f_{2}-i \lambda \rho_{1} w_{2}+\gamma \Delta \widetilde{f}_{2}+m_{1} \Delta w_{2}-\alpha \Delta w_{5}\right)=: \widetilde{f}_{*} \text { in } \Omega_{1}, \\
w_{1}=0 \text { on } \Gamma, \\
\partial_{\nu} w_{1}=0 \text { on } \Gamma, \\
\mathscr{B}_{1} w_{1}=-\alpha \beta_{1}^{-1} w_{5} \text { on } I, \\
\mathscr{B}_{2} w_{1}=\beta_{1}^{-1}\left(\gamma \partial_{\nu} \widetilde{f_{2}}+m_{1} \partial_{\nu} w_{2}-\beta_{2} \partial_{\nu} w_{3}+\alpha \kappa w_{5}\right) \text { on } I .
\end{array}\right.
$$

By Proposition 2.60 there are positive constants $\lambda_{0}^{(2)}$ and $C$ such that

$$
\begin{align*}
\left\|w_{1}\right\|_{H^{4}\left(\Omega_{1}\right)} \leq & C\left(\left\|\lambda_{0}^{(2)} w_{1}+\widetilde{f}_{*}\right\|_{L^{2}\left(\Omega_{1}\right)}+\left\|-\alpha \beta_{1}^{-1} w_{5}\right\|_{H^{3 / 2}(I)}\right. \\
& \left.+\left\|\beta_{1}^{-1}\left(\gamma \partial_{\nu} \widetilde{f}_{2}+m_{1} \partial_{\nu} w_{2}-\beta_{2} \partial_{\nu} w_{3}+\alpha \kappa w_{5}\right)\right\|_{H^{1 / 2}(I)}\right) \tag{4.131}
\end{align*}
$$

From 4.127, 4.129) and the fact that $0 \in \rho\left(\mathscr{A}_{\eta}\right)$ it follows that

$$
\begin{align*}
\left\|\Delta \widetilde{f}_{2}\right\|_{L^{2}\left(\Omega_{1}\right)} & \leq C\left\|\widetilde{f}_{2}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}=C\left\|i \lambda w_{2}-f_{2}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \\
& \leq C\left(\lambda^{2}\|w\|_{\mathscr{H}_{\eta}}+|\lambda|\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}\right) \tag{4.132}
\end{align*}
$$

As $\Delta w_{5}=\beta^{-1}\left(i \lambda \rho_{0} w_{5}-\alpha \Delta w_{2}+\sigma w_{5}-\rho_{0} f_{5}\right)=: \widehat{f}_{*}$ in $\Omega_{1}$ and $\partial_{\nu} w_{5}+\kappa w_{5}=$ 0 on $\partial \Omega_{1}$, this is due to $w_{5}$ satisfies 4.106) and is an element of $D\left(\mathcal{A}_{T}\right)$, by regularity theory there exist positive constants $\lambda_{0}^{(3)}$ and $C$ such that

$$
\begin{equation*}
\left\|w_{5}\right\|_{H^{2}\left(\Omega_{1}\right)} \leq C\left\|\lambda_{0}^{(3)} w_{5}-\widehat{f}_{*}\right\|_{L^{2}\left(\Omega_{1}\right)} \leq C\left(|\lambda|\|w\|_{\mathscr{H}_{\eta}}+\|f\|_{\mathscr{H}_{\eta}}\right) \tag{4.133}
\end{equation*}
$$

In the last inequality it was used (4.127). By (4.125), 4.127), 4.129), 4.131)(4.133) and the trace theorem we see that

$$
\left\|w_{1}\right\|_{H^{4}\left(\Omega_{1}\right)} \leq C\left(\lambda^{2}\|w\|_{\mathscr{H}_{\eta}}+|\lambda|\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}\right)
$$

By 4.118,

$$
\left\|w_{1}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C|\lambda|^{-1}\left(\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}+\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}^{1 / 2}\|w\|_{\mathscr{\mathscr { H } _ { \eta }}}^{1 / 2}\right)
$$

From 4.130),

$$
\begin{align*}
\left\|\mathscr{B}_{2} w_{1}\right\|_{L^{2}(I)} \leq & C\left(|\lambda|^{3 / 2}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}^{1 / 6}\|w\|_{\mathscr{H}_{\eta}}^{5 / 6}+|\lambda|^{3 / 2}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}^{1 / 12}\|w\|_{\mathscr{H}_{\eta}}^{11 / 12}\right.  \tag{4.134}\\
& \left.+|\lambda|^{2 / 3}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}+|\lambda|^{2 / 3}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}^{1 / 12}\|w\|_{\mathscr{H}_{\eta}}^{1 / 12}\right) .
\end{align*}
$$

By the trace theorem and 4.109, we get

$$
\begin{equation*}
\left\|w_{5}\right\|_{L^{2}(I)} \leq C\left\|w_{5}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C\|f\|_{\mathscr{H}_{\eta}}^{1 / 2}\|w\|_{\mathscr{H}_{\eta}}^{1 / 2} \tag{4.135}
\end{equation*}
$$

Due to (4.126), 4.128, (4.129), (4.134), 4.135) and Corollary 2.20, we write

$$
\begin{equation*}
\mathscr{I} \leq \varepsilon C\|w\|_{\mathscr{H}_{\eta}}^{2}+C_{\varepsilon}|\lambda|^{48}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}}^{2} \quad \text { for any } \varepsilon>0 \tag{4.136}
\end{equation*}
$$

Due to 4.119), 4.123 and (4.136) we obtain that

$$
\begin{equation*}
\|w\|_{\mathscr{H}_{\eta}} \leq C|\lambda|^{24}\left\|\mathscr{A}_{\eta} f\right\|_{\mathscr{H}_{\eta}} \tag{4.137}
\end{equation*}
$$

Let $\widetilde{w} \in D\left(\mathscr{A}_{\eta}\right)$ and $\widetilde{f}:=\left(i \lambda \mathcal{I}-\mathscr{A}_{\eta}\right) \widetilde{w}=0$. In this situation, we have that $\mathscr{A}_{\eta} \widetilde{w}=i \lambda \widetilde{w} \in D\left(\mathscr{A}_{\eta}\right)$ and so $\widetilde{w} \in D\left(\mathscr{A}_{\eta}^{2}\right)$. Now, an application of 4.137) allows us to obtain $\widetilde{w}=0$. In consequence, $\operatorname{Ker}\left(i \lambda \mathcal{I}-\mathscr{A}_{\eta}\right)=\{0\}$ and thus $i \lambda$ is not an eigenvalue of $\mathscr{A}_{\eta}$. By Remark 4.14, we get that $i \lambda \notin \sigma\left(\mathscr{A}_{\eta}\right)$.

Let us consider $\mathscr{F} \in D\left(\mathscr{A}_{\eta}\right)$ and $\lambda \in \mathbb{R}$ with $|\lambda|>\lambda_{0}$. As $i \lambda \in \rho\left(\mathscr{A}_{\eta}\right)$, then $i \lambda \mathcal{I}-\mathscr{A}_{\eta}: D\left(\mathscr{A}_{\eta}\right) \rightarrow \mathscr{H}_{\eta}$ is invertible and so there exists $\mathscr{U} \in D\left(\mathscr{A}_{\eta}\right)$ such that $\left(i \lambda \mathcal{I}-\mathscr{A}_{\eta}\right) \mathscr{U}=\mathscr{F}$. Since $\mathscr{A}_{\eta} \mathscr{U}=i \lambda \mathscr{U}-\mathscr{F} \in D\left(\mathscr{A}_{\eta}\right)$, then $\mathscr{U} \in D\left(\mathscr{A}_{\eta}^{2}\right)$ and thus 4.137) holds replacing $w$ and $f$ by $\mathscr{U}$ and $\mathscr{F}$, respectively. Hence,

$$
\left\|\left(i \lambda \mathcal{I}-\mathscr{A}_{\eta}\right)^{-1} \mathscr{F}\right\|_{\mathscr{H}_{\eta}} \leq C|\lambda|^{24}\left\|\mathscr{A}_{\eta} \mathscr{F}\right\|_{\mathscr{H}_{\eta}} .
$$

From Theorem 2.44 with $\alpha^{\prime}=1$ and $\beta^{\prime}=24$ it follows the first statement of our theorem,

$$
\begin{equation*}
\left\|\mathscr{T}_{\eta}(t) \mathscr{A}_{\eta}^{-1}\right\|_{\mathcal{L}\left(\mathscr{H}_{\eta}\right)} \leq C t^{-1 / 25} \quad \text { for all } t>0 \tag{4.138}
\end{equation*}
$$

Let $k \in \mathbb{N}, w_{0} \in D\left(\mathscr{A}_{\eta}^{k}\right)$ and $f_{0}:=\mathscr{A}_{\eta}^{k} w_{0}$. Note that $f_{0} \in \mathscr{H}_{\eta}$. Since $\mathscr{A}_{\eta}^{k}$ is invertible, we have that $w_{0}=\mathscr{A}_{\eta}^{-k} f_{0}$. Using part c) of Theorem 2.4 in [109, p. 5], one can prove that $\mathscr{T}_{\eta}^{k}(t) \mathscr{A}_{\eta}^{-k} \phi=\left[\mathscr{T}_{\eta}(t) \mathscr{A}_{\eta}^{-1}\right]^{k} \phi$ for any $\phi \in \mathscr{H}_{\eta}$ and for all $t \geq 0$. Thus, from (4.138) it follows that

$$
\left\|\mathscr{T}_{\eta}(t) w_{0}\right\|_{\mathscr{H}_{\eta}}=\left\|\left(\mathscr{T}_{\eta}(t / k) \mathscr{A}_{\eta}^{-1}\right)^{k} f_{0}\right\|_{\mathscr{H}_{\eta}} \leq\left(C k^{1 / 25}\right)^{k} t^{-k / 25}\left|w_{0}\right|_{D\left(\mathscr{A}_{\eta}^{k}\right)}
$$

for any $t>0$.
Remark 4.17. a) The estimate (4.100) indicates that the more regular are the initial data, the decay of the energy (3.1) is faster. Indeed: If $w_{0} \in D\left(\mathscr{A}_{\eta}^{k}\right)$, then (4.100) implies that

$$
E_{\gamma}(t)=\frac{1}{2}\left\|\mathscr{T}_{\eta}(t) w_{0}\right\|_{\mathscr{H}_{\eta}}^{2} \leq \frac{\widetilde{C}_{k}}{t^{2 k / 25}}\left|w_{0}\right|_{D\left(\mathscr{A}_{\eta}^{k}\right)}^{2} \quad \text { for all } \quad t>0
$$

where $\widetilde{C}_{k}:=C_{k}^{2} / 2$.
b) The result of the previous theorem continues to be true if the plate is isothermal, i.e., the temperature is not necessary to obtain the decay rate of Theorem 4.16
c) If we remove the structural damping $\left(m_{1}=0\right)$ of our thermoelastic plate-membrane system, we do not know anything about asymptotic polynomial behavior for parameters $\eta>0$ and $m_{2} \geq 0$, and when also $\eta=m_{2}=0$.

Remark 4.18. In the proof of Theorem 4.16, we did not aim at optimal polynomial rate.

## Chapter 5

## Analyticity of the semigroup associated with a transmission problem

The goal of this chapter (Theorem 5.1) is an analytical result when both the inertial term and the structural damping are not present, but there is Kelvin-Voigt damping on the membrane and temperature on the plate. As a consequence we will have that $w(t)=\mathscr{T}_{0}(t) w_{0} \in D\left(\mathscr{A}_{0}^{\infty}\right):=\cap_{k=0}^{\infty} D\left(\mathscr{A}_{0}^{k}\right)$ for $w_{0} \in \mathscr{H}_{0}$, i.e., no matter how irregular the initial data always the corresponding solution is of class $C^{\infty}$ (see [51, p. 2] and [104, p. 179]). Other implications of the analyticity of a semigroup can be found in [94]. We will achieve our analytical purpose by making use of the well known result of Theorem 2.46. The proof will be done by contradiction.

Theorem 5.1. If $\eta=m_{1}=0$ and $m_{2}>0$, then the semigroup $\left(\mathscr{T}_{0}(t)\right)_{t \geq 0}$ generated by $\mathscr{A}_{0}$ is analytic.

Proof. From Proposition 4.5, we have that $i \mathbb{R} \subset \rho\left(\mathscr{A}_{0}\right)$. Let us suppose 2.20) is not true. According to Remark 2.7 in [39], there are sequences $\left(\lambda_{n}\right)_{n \in \mathbb{N}} \subset \mathbb{R}$ with $\left|\lambda_{n}\right| \rightarrow \infty$ and $\left(\widehat{f}_{n}\right)_{n \in \mathbb{N}} \subset \mathscr{H}_{0}$ with $\left\|\widehat{f}_{n}\right\|_{\mathscr{H}_{0}}=1$ such that

$$
\lim _{n \rightarrow \infty}\left\|\lambda_{n}\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right)^{-1} \widehat{f}_{n}\right\|_{\mathscr{H}_{0}}=\infty
$$

Without loss of generality, we will assume that $\lambda_{n}$ is positive for each $n \in \mathbb{N}$. The opposite situation, $\lambda_{n}<0$ for any $n \in \mathbb{N}$, is discussed in Remark 5.2.

Chapter 5. Analyticity of the semigroup associated with a transmission problem
Setting $\widehat{w}_{n}:=\lambda_{n}\left(i \lambda_{n} \mathcal{I}-\mathscr{A}_{0}\right)^{-1} \widehat{f}_{n}, w_{n}:=\widehat{w}_{n} /\left\|\widehat{w}_{n}\right\|_{\mathscr{H}_{0}}$ and $f_{n}:=\widehat{f}_{n} /\left\|\widehat{w}_{n}\right\|_{\mathscr{H}_{0}}$, we obtain the following

$$
\begin{equation*}
\left\|w_{n}\right\|_{\mathscr{H}}^{0}=1 \quad \text { and } \quad\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}=f_{n} \tag{5.1}
\end{equation*}
$$

with

$$
\begin{equation*}
\left\|\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}\right\|_{\mathscr{H}_{0}} \xrightarrow[n \rightarrow \infty]{ } 0 \tag{5.2}
\end{equation*}
$$

Note that $\left(w_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathscr{A}_{0}\right)$. From the second assertion in (5.1),

$$
\begin{align*}
& i w_{1}^{n}-\lambda_{n}^{-1} w_{2}^{n}=f_{1}^{n}  \tag{5.3}\\
& i \rho_{1} w_{2}^{n}+\lambda_{n}^{-1}\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n}\right)=\rho_{1} f_{2}^{n},  \tag{5.4}\\
& i w_{3}^{n}-\lambda_{n}^{-1} w_{4}^{n}=f_{3}^{n},  \tag{5.5}\\
& i \rho_{2} w_{4}^{n}-\beta_{2} \lambda_{n}^{-1} \Delta w_{3}^{n}-m_{2} \lambda_{n}^{-1} \Delta w_{4}^{n}=\rho_{2} f_{4}^{n},  \tag{5.6}\\
& i \rho_{0} w_{5}^{n}+\alpha \lambda_{n}^{-1} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \lambda_{n}^{-1} \mathcal{A}_{T} w_{5}^{n}=\rho_{0} f_{5}^{n}, \tag{5.7}
\end{align*}
$$

where $\mathcal{W}_{0}^{n}:=\beta_{1} w_{1}^{n}+\beta_{2} \mathcal{G}_{2} \gamma_{1} w_{3}^{n}+m_{2} \mathcal{G}_{2} \gamma_{1} w_{4}^{n}+\alpha \mathcal{G}_{1} \gamma_{0} w_{5}^{n}-\alpha \kappa \mathcal{G}_{2} \gamma_{0} w_{5}^{n}$. Note that

$$
\begin{aligned}
\left\|\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}\right\|_{\mathscr{H}_{0}}^{2}= & \beta_{1}\left\|f_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}+\rho_{1}\left\|f_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta_{2}\left\|\nabla f_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2} \\
& +\rho_{2}\left\|f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}+\rho_{0}\left\|f_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} .
\end{aligned}
$$

Then, the limit (5.2) and (5.3)-(5.7) imply that

$$
\begin{align*}
& i w_{1}^{n}-\lambda_{n}^{-1} w_{2}^{n} \rightarrow 0 \text { in } H_{\Gamma}^{2}\left(\Omega_{1}\right),  \tag{5.8}\\
& i \rho_{1} w_{2}^{n}+\lambda_{n}^{-1}\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\frac{\alpha \sigma}{\beta} w_{5}^{n}\right) \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right),  \tag{5.9}\\
& i \nabla w_{3}^{n}-\lambda_{n}^{-1} \nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right),  \tag{5.10}\\
& i \rho_{2} w_{4}^{n}-\beta_{2} \lambda_{n}^{-1} \Delta w_{3}^{n}-m_{2} \lambda_{n}^{-1} \Delta w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right),  \tag{5.11}\\
& i \rho_{0} w_{5}^{n}+\alpha \lambda_{n}^{-1} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \lambda_{n}^{-1} \mathcal{A}_{T} w_{5}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) . \tag{5.12}
\end{align*}
$$

Due to (3.35),

$$
\begin{align*}
\operatorname{Re}\left(\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}, w_{n}\right)_{\mathscr{H}_{0}}= & \frac{m_{2}}{\lambda_{n}}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}+\frac{\sigma}{\lambda_{n}}\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2} \\
& +\frac{\beta}{\lambda_{n}}\left\|\nabla w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}+\frac{\beta \kappa}{\lambda_{n}}\left\|w_{5}^{n}\right\|_{L^{2}\left(\partial \Omega_{1}\right)}^{2} . \tag{5.13}
\end{align*}
$$

By the first assertion in (5.1) and (5.13), we get

$$
\lambda_{n}^{-1}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{2} \leq C \operatorname{Re}\left(\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}, w_{n}\right)_{\mathscr{H}_{0}} \leq C\left\|\left(i \mathcal{I}-\lambda_{n}^{-1} \mathscr{A}_{0}\right) w_{n}\right\|_{\mathscr{H}_{0}}
$$

Now, (5.2) implies

$$
\begin{equation*}
\lambda_{n}^{-1 / 2} w_{5}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{1}\right) \tag{5.14}
\end{equation*}
$$

For being $m_{2}$ positive, it follows from (5.2) and (5.13) that

$$
\begin{equation*}
\lambda_{n}^{-1 / 2} \nabla w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{5.15}
\end{equation*}
$$

The limit (5.10) and (5.15) imply

$$
\begin{equation*}
\nabla w_{3}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{5.16}
\end{equation*}
$$

We will obtain a key convergence in this proof through the following equality.

$$
\begin{align*}
& \left(i \rho_{0} w_{5}^{n}+\alpha \lambda_{n}^{-1} \mathcal{A}_{L}\left(\mathcal{I}-\mathcal{N} \gamma_{1}\right) w_{2}^{n}+\beta \lambda_{n}^{-1} \mathcal{A}_{T} w_{5}^{n}, w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}  \tag{5.17}\\
& =i \rho_{0}\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}-\alpha \lambda_{n}^{-1}\left(\Delta w_{2}^{n}, w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}+\beta \lambda_{n}^{-1}\left(\mathcal{A}_{T} w_{5}^{n}, w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}
\end{align*}
$$

Integration by parts implies

$$
\begin{equation*}
\left(\Delta w_{2}^{n}, w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}=-\left(\nabla w_{2}^{n}, \nabla w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\left(\partial_{\nu} w_{2}^{n}, w_{5}^{n}\right)_{L^{2}(I)} \tag{5.18}
\end{equation*}
$$

As $\left(w_{1}^{n}\right)_{n \in \mathbb{N}}$ is bounded in $H^{2}\left(\Omega_{1}\right)$, the limit (5.8) implies that $\left(\lambda_{n}^{-1} w_{2}^{n}\right)_{n \in \mathbb{N}}$ is also bounded in $H^{2}\left(\Omega_{1}\right)$. By interpolation inequality (2.11), we compute that

$$
\begin{equation*}
\frac{1}{\lambda_{n}^{1 / 2}}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \leq C \frac{\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 2}}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{1 / 2} \leq C \tag{5.19}
\end{equation*}
$$

Using Cauchy-Schwarz inequality and (5.19),

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\nabla w_{2}^{n}, \nabla w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}\right| \leq C \lambda_{n}^{-1 / 2}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} \tag{5.20}
\end{equation*}
$$

Next we will apply Theorem 2.21 to $w_{5}^{n}$ and Corollary 2.22 to $w_{2}^{n}$. Considering again (5.19), we have

$$
\begin{align*}
\left|\lambda_{n}^{-1}\left(\partial_{\nu} w_{2}^{n}, w_{5}^{n}\right)_{L^{2}(I)}\right| & \leq \lambda_{n}^{-1}\left\|\partial_{\nu} w_{2}^{n}\right\|_{L^{2}(I)}\left\|w_{5}^{n}\right\|_{L^{2}(I)} \\
& \leq C \frac{\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 2}} \frac{\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}} \frac{\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}}\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{1 / 2} \\
& \leq C \lambda_{n}^{-1 / 4}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2} . \tag{5.21}
\end{align*}
$$

Thanks to the fact that $\mathcal{A}_{T}^{1 / 2}$ is self-adjoint and to (3.17), we can write

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\mathcal{A}_{T} w_{5}^{n}, w_{5}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}\right|=\lambda_{n}^{-1}\left\|w_{5}^{n}\right\|_{D\left(\mathcal{A}_{T}^{1 / 2}\right)}^{2} \leq C \lambda_{n}^{-1}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{2} \tag{5.22}
\end{equation*}
$$

Due to boundedness of $\left(w_{5}^{n}\right)_{n \in \mathbb{N}}$ in $L^{2}\left(\Omega_{1}\right)$, (5.12), (5.14), (5.17), (5.18) and (5.20)-(5.22), we conclude

$$
\begin{equation*}
w_{5}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) . \tag{5.23}
\end{equation*}
$$

On the other hand,

$$
\begin{align*}
& \left(i \rho_{2} w_{4}^{n}-\beta_{2} \lambda_{n}^{-1} \Delta w_{3}^{n}-m_{2} \lambda_{n}^{-1} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)}  \tag{5.24}\\
& =i \rho_{2}\left\|w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{2}-\lambda_{n}^{-1}\left(\beta_{2} \Delta w_{3}^{n}+m_{2} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)}
\end{align*}
$$

By Remark 3.19, integration by parts and the fact that $w_{2}^{n}=w_{4}^{n}$ on $I$ we derive

$$
\begin{align*}
& \lambda_{n}^{-1}\left(\beta_{2} \Delta w_{3}^{n}+m_{2} \Delta w_{4}^{n}, w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)}=-\beta_{2} \lambda_{n}^{-1}\left(\nabla w_{3}^{n}, \nabla w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)^{2}} \\
& -m_{2} \lambda_{n}^{-1}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{2}-\lambda_{n}^{-1}\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)} . \tag{5.25}
\end{align*}
$$

Obviously,

$$
\begin{equation*}
\left|\lambda_{n}^{-1}\left(\nabla w_{3}^{n}, \nabla w_{4}^{n}\right)_{L^{2}\left(\Omega_{2}\right)^{2}}\right| \leq \lambda_{n}^{-1 / 2}\left\|\nabla w_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}} \lambda_{n}^{-1 / 2}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}} \tag{5.26}
\end{equation*}
$$

As $\Delta\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)=i \rho_{2} \lambda_{n} w_{4}^{n}-\rho_{2} \lambda_{n} f_{4}^{n}$ in $\Omega_{2}$ and $\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}=\beta_{2} w_{1}^{n}+$ $m_{2} w_{2}^{n}$ on $I$, see (5.6), Remark 3.17 implies

$$
\begin{aligned}
& \left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)} \\
& \leq C\left(\left\|i \rho_{2} \lambda_{n} w_{4}^{n}-\rho_{2} \lambda_{n} f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|\beta_{2} w_{1}^{n}+m_{2} w_{2}^{n}\right\|_{H^{3 / 2}(I)}\right) \\
& \leq C\left(\lambda_{n}\left\|w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\lambda_{n}\left\|f_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}+\left\|w_{1}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}+\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}\right)
\end{aligned}
$$

Because of $\left(w_{4}^{n}\right)_{n \in \mathbb{N}}$ and $\left(f_{4}^{n}\right)_{n \in \mathbb{N}}$ are bounded sequences in $L^{2}\left(\Omega_{2}\right)$, then

$$
\begin{equation*}
\lambda_{n}^{-1}\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)} \leq C \tag{5.27}
\end{equation*}
$$

Thanks to Cauchy-Schwarz inequality, Theorem 2.21 and (5.19) we obtain

$$
\begin{align*}
& \left|\lambda_{n}^{-1}\left(\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right), w_{2}^{n}\right)_{L^{2}(I)}\right| \\
& \leq C \lambda_{n}^{-3 / 4}\left\|\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)\right\|_{L^{2}(I)} \lambda_{n}^{-1 / 4}\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{1 / 2} \\
& \leq C \lambda_{n}^{-3 / 4}\left\|\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)\right\|_{L^{2}(I)} . \tag{5.28}
\end{align*}
$$

Taking into account (5.27) and applying Corollary 2.22 to $\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}$,

$$
\lambda_{n}^{-3 / 4}\left\|\partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)\right\|_{L^{2}(I)}
$$

$$
\begin{align*}
& \leq C \frac{1}{\lambda_{n}^{1 / 2}}\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)}^{1 / 2} \frac{1}{\lambda_{n}^{1 / 4}}\left\|\nabla\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2} \\
& \leq C\left(\lambda_{n}^{-1 / 4}\left\|\nabla w_{3}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2}+\lambda_{n}^{-1 / 4}\left\|\nabla w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)^{2}}^{1 / 2}\right) \tag{5.29}
\end{align*}
$$

From (5.11), (5.15), (5.16), (5.24)-(5.26), (5.28) and (5.29), it follows that

$$
\begin{equation*}
w_{4}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{2}\right) \tag{5.30}
\end{equation*}
$$

Now, let us consider the following equality

$$
\begin{align*}
& \left(i \rho_{1} w_{2}^{n}+\lambda_{n}^{-1}\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}-\alpha \mathcal{A}_{T} w_{5}^{n}+\alpha \beta^{-1} \sigma w_{5}^{n}\right), w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \\
& =i \rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\lambda_{n}^{-1}\left(\mathcal{A}_{B} \mathcal{W}_{0}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}+\alpha \lambda_{n}^{-1}\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \tag{5.31}
\end{align*}
$$

Applying integration by parts,

$$
\begin{equation*}
\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}=-\left(\nabla w_{5}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \tag{5.32}
\end{equation*}
$$

Because of $\partial_{\nu} w_{5}^{n}+\kappa w_{5}^{n}=0$ on $I$, the trace theorem and (5.19) we obtain that

$$
\begin{align*}
\left|\lambda_{n}^{-1}\left(\partial_{\nu} w_{5}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right| & \leq C \lambda_{n}^{-1}\left\|w_{5}^{n}\right\|_{L^{2}(I)}\left\|w_{2}^{n}\right\|_{L^{2}(I)} \\
& \leq C \frac{\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}}{\lambda_{n}^{1 / 2}} \frac{\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}}{\lambda_{n}^{1 / 2}}  \tag{5.33}\\
& \leq C \lambda_{n}^{-1 / 2}\left\|w_{5}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)} .
\end{align*}
$$

Gathering (4.59), (5.9), (5.14)-(5.16), (5.20), (5.21), (5.28), (5.29) and (5.31)(5.33) we have that

$$
i \rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}+\beta_{1} \lambda_{n}^{-1}\left(w_{1}^{n}, w_{2}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0
$$

Multiplying (5.8) by $w_{1}^{n}$, we obtain

$$
i\left\|w_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2}-\lambda_{n}^{-1}\left(w_{2}^{n}, w_{1}^{n}\right)_{H_{\Gamma}^{2}\left(\Omega_{1}\right)} \rightarrow 0 .
$$

The last two limits allow to write

$$
\begin{equation*}
\rho_{1}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}-\beta_{1}\left\|w_{1}^{n}\right\|_{H_{\Gamma}^{2}\left(\Omega_{1}\right)}^{2} \rightarrow 0 \tag{5.34}
\end{equation*}
$$

Our goal now is to show that the sequence $\left(w_{2}^{n}\right)_{n \in \mathbb{N}}$ converges to zero in $L^{2}\left(\Omega_{1}\right)$. Developing an analogous argument as in the limit (4.64), taking into account that $\lambda_{n} \rightarrow \infty$, we find that

$$
\begin{equation*}
\lambda_{n}^{-1}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{2}-\lambda_{n}^{-1}\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \tag{5.35}
\end{equation*}
$$

Using the fact that $w_{2}^{n}=w_{4}^{n}$ on $I$, Theorem 2.21, Corollary 2.22 and (5.19) we obtain that

$$
\begin{align*}
\left|\lambda_{n}^{-1}\left(\partial_{\nu} w_{2}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right| & \leq \lambda_{n}^{-1}\left\|\partial_{\nu} w_{2}^{n}\right\|_{L^{2}(I)}\left\|w_{4}^{n}\right\|_{L^{2}(I)} \\
& \leq C \frac{\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 2}} \frac{\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}} \frac{\left\|w_{4}^{n}\right\|_{H^{1}\left(\Omega_{2}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}}\left\|w_{4}^{n}\right\|_{L^{2}\left(\Omega_{2}\right)}^{1 / 2} \\
& \leq C \lambda_{n}^{-1 / 4}\left\|w_{4}^{n}\right\|_{H^{1}\left(\Omega_{2}\right)}^{1 / 2} . \tag{5.36}
\end{align*}
$$

Due to $\left(\lambda_{n}^{-1 / 2} w_{4}^{n}\right)_{n \in \mathbb{N}}$ and $\left(\lambda_{n}^{-1 / 2} \nabla w_{4}^{n}\right)_{n \in \mathbb{N}}$ are sequences that converge to zero in $L^{2}\left(\Omega_{2}\right)$, see the first assertion of (5.1) and (5.15), then

$$
\begin{equation*}
\lambda_{n}^{-1 / 2} w_{4}^{n} \rightarrow 0 \text { in } H^{1}\left(\Omega_{2}\right) \tag{5.37}
\end{equation*}
$$

From (5.35)-(5.37), it follows that

$$
\begin{equation*}
\lambda_{n}^{-1 / 2} \nabla w_{2}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) \tag{5.38}
\end{equation*}
$$

By (5.4),

$$
\mathcal{A}_{B} \mathcal{W}_{0}^{n}=\rho_{1} \lambda_{n} f_{2}^{n}-i \rho_{1} \lambda_{n} w_{2}^{n}-\alpha \Delta w_{5}^{n} .
$$

Since $\left(w_{n}\right)_{n \in \mathbb{N}} \subset D\left(\mathscr{A}_{0}\right)$, it is immediate from the definition of the domain of the operator $\mathscr{A}_{0}$ that $\mathcal{W}_{0}^{n} \in D\left(\mathcal{A}_{B}\right)$. Because of the latter and the definitions of the Green maps $\mathcal{G}_{1}$ and $\mathcal{G}_{2}$ given in (3.19), we have that the last equation is equivalent to the following elliptic boundary value problem

$$
\left\{\begin{array}{l}
\Delta^{2} w_{1}^{n}=\beta_{1}^{-1} \rho_{1} \lambda_{n} f_{2}^{n}-i \beta_{1}^{-1} \rho_{1} \lambda_{n} w_{2}^{n}-\alpha \beta_{1}^{-1} \Delta w_{5}^{n}=: f^{*} \text { in } \Omega_{1},  \tag{5.39}\\
w_{1}^{n}=0 \text { on } \Gamma, \\
\partial_{\nu} w_{1}^{n}=0 \text { on } \Gamma, \\
\mathscr{B}_{1} w_{1}^{n}=-\alpha \beta_{1}^{-1} w_{5}^{n} \text { on } I, \\
\mathscr{B}_{2} w_{1}^{n}=-\beta_{1}^{-1} \partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)+\alpha \beta_{1}^{-1} \kappa w_{5}^{n} \text { on } I .
\end{array}\right.
$$

By Proposition 2.60, there are two positive constants $\lambda_{0}^{(1)}$ and $C$ that satisfy

$$
\begin{align*}
\left\|w_{1}^{n}\right\|_{H^{4}\left(\Omega_{1}\right)} \leq & C\left(\left\|\lambda_{0}^{(1)} w_{1}^{n}+f^{*}\right\|_{L^{2}\left(\Omega_{1}\right)}+\left\|-\alpha \beta_{1}^{-1} w_{5}^{n}\right\|_{H^{3 / 2}(I)}\right. \\
& \left.+\left\|-\beta_{1}^{-1} \partial_{\nu}\left(\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right)+\alpha \beta_{1}^{-1} \kappa w_{5}^{n}\right\|_{H^{1 / 2}(I)}\right) \\
\leq & C\left(\left\|w_{1}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}+\lambda_{n}\left\|f_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}+\lambda_{n}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}\right.  \tag{5.40}\\
& \left.+\left\|w_{5}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}+\left\|\beta_{2} w_{3}^{n}+m_{2} w_{4}^{n}\right\|_{H^{2}\left(\Omega_{2}\right)}\right) .
\end{align*}
$$

From (5.7) and the fact $w_{5}^{n} \in D\left(\mathcal{A}_{T}\right)$, it follows that: $\Delta w_{5}^{n}=i \beta^{-1} \rho_{0} \lambda_{n} w_{5}^{n}-$ $\alpha \beta^{-1} \Delta w_{2}^{n}+\sigma \beta^{-1} w_{5}^{n}-\rho_{0} \beta^{-1} \lambda_{n} f_{5}^{n}=: g^{*}$ in $\Omega_{1}$ and $\partial_{\nu} w_{5}^{n}+\kappa w_{5}^{n}=0$ on $\partial \Omega_{1}$. By regularity theory, see for instance part b) of Remark 3.3 in [14], there are constants $\lambda_{0}^{(2)}>0$ and $C>0$ such that

$$
\begin{aligned}
\left\|w_{5}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)} & \leq C\left\|\lambda_{0}^{(2)} w_{5}^{n}-g^{*}\right\|_{L^{2}\left(\Omega_{1}\right)} \\
& \leq C\left(\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}+\lambda_{n}\left\|w_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}+\left\|w_{2}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}+\lambda_{n}\left\|f_{5}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}\right)
\end{aligned}
$$

Therefore,

$$
\begin{equation*}
\lambda_{n}^{-1}\left\|w_{5}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)} \leq C \tag{5.41}
\end{equation*}
$$

Thanks to (5.27), 5.40) and (5.41) we can affirm that $\left(\lambda_{n}^{-1} w_{1}^{n}\right)_{n \in \mathbb{N}}$ is bounded in $H^{4}\left(\Omega_{1}\right)$. By interpolation inequality,

$$
\begin{equation*}
\frac{1}{\lambda_{n}^{1 / 2}}\left\|w_{1}^{n}\right\|_{H^{3}\left(\Omega_{1}\right)} \leq C \frac{\left\|w_{1}^{n}\right\|_{H^{4}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 2}}\left\|w_{1}^{n}\right\|_{H^{2}\left(\Omega_{1}\right)}^{1 / 2} \leq C \tag{5.42}
\end{equation*}
$$

Due to the first equation in (5.39),

$$
\begin{equation*}
i w_{2}^{n}=-\frac{\beta_{1}}{\rho_{1}} \lambda_{n}^{-1} \Delta^{2} w_{1}^{n}+f_{2}^{n}-\frac{\alpha}{\rho_{1}} \lambda_{n}^{-1} \Delta w_{5}^{n} \text { in } \Omega_{1} . \tag{5.43}
\end{equation*}
$$

Taking the inner product in $L^{2}\left(\Omega_{1}\right)$ of (5.43) with $w_{2}^{n}$ and applying integration by parts, we obtain

$$
\begin{align*}
i\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{2}= & -\frac{\beta_{1}}{\rho_{1}} \lambda_{n}^{-1}\left[-\left(\nabla \Delta w_{1}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}+\left(\partial_{\nu} \Delta w_{1}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right]  \tag{5.44}\\
& +\left(f_{2}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)}-\frac{\alpha}{\rho_{1}} \lambda_{n}^{-1}\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} .
\end{align*}
$$

Due to (5.42), we can write

$$
\begin{align*}
\left|\lambda_{n}^{-1}\left(\nabla \Delta w_{1}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}}\right| & \leq \lambda_{n}^{-1}\left\|\nabla \Delta w_{1}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& \leq C \lambda_{n}^{-1 / 2}\left\|w_{1}^{n}\right\|_{H^{3}\left(\Omega_{1}\right)} \lambda_{n}^{-1 / 2}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} \\
& \leq C \lambda_{n}^{-1 / 2}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}} \tag{5.45}
\end{align*}
$$

The boundedness of the sequences $\left(\lambda_{n}^{-1} w_{1}^{n}\right)_{n \in \mathbb{N}}$ in $H^{4}\left(\Omega_{1}\right)$ and $\left(w_{2}^{n}\right)_{n \in \mathbb{N}}$ in $L^{2}\left(\Omega_{1}\right)$ together with Theorem 2.21. Corollary 2.22, Friedrichs inequality and (5.42) allows us to write

$$
\left|\lambda_{n}^{-1}\left(\partial_{\nu} \Delta w_{1}^{n}, w_{2}^{n}\right)_{L^{2}(I)}\right| \leq \lambda_{n}^{-1}\left\|\partial_{\nu} \Delta w_{1}^{n}\right\|_{L^{2}(I)}\left\|w_{2}^{n}\right\|_{L^{2}(I)}
$$

Chapter 5. Analyticity of the semigroup associated with a transmission problem

$$
\begin{align*}
& \leq C \frac{\left\|w_{1}^{n}\right\|_{H^{4}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 2}} \frac{\left\|w_{1}^{n}\right\|_{H^{3}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}} \frac{\left\|w_{2}^{n}\right\|_{H^{1}\left(\Omega_{1}\right)}^{1 / 2}}{\lambda_{n}^{1 / 4}}\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)}^{1 / 2} \\
& \leq C \lambda_{n}^{-1 / 4}\left\|\nabla w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)^{2}}^{1 / 2} . \tag{5.46}
\end{align*}
$$

The limit (5.38) together with (5.45) and (5.46) imply

$$
\begin{equation*}
\lambda_{n}^{-1}\left(\nabla \Delta w_{1}^{n}, \nabla w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)^{2}} \rightarrow 0 \quad \text { and } \quad \lambda_{n}^{-1}\left(\partial_{\nu} \Delta w_{1}^{n}, w_{2}^{n}\right)_{L^{2}(I)} \rightarrow 0 \tag{5.47}
\end{equation*}
$$

Because $f_{2}^{n} \rightarrow 0$ in $L^{2}\left(\Omega_{1}\right)$ and $\left\|w_{2}^{n}\right\|_{L^{2}\left(\Omega_{1}\right)} \leq C$ for all $n \in \mathbb{N}$,

$$
\begin{equation*}
\left(f_{2}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \rightarrow 0 . \tag{5.48}
\end{equation*}
$$

By (5.14), (5.20), (5.32) and (5.33),

$$
\begin{equation*}
\lambda_{n}^{-1}\left(\Delta w_{5}^{n}, w_{2}^{n}\right)_{L^{2}\left(\Omega_{1}\right)} \rightarrow 0 . \tag{5.49}
\end{equation*}
$$

Because of (5.34), (5.44) and (5.47)-(5.49) we obtain the convergences

$$
\begin{equation*}
w_{1}^{n} \rightarrow 0 \text { in } H^{2}\left(\Omega_{1}\right) \text { and } w_{2}^{n} \rightarrow 0 \text { in } L^{2}\left(\Omega_{1}\right) . \tag{5.50}
\end{equation*}
$$

Joining the limits (5.16), (5.23), (5.30) and (5.50) we get $w_{n} \rightarrow 0$ in $\mathscr{H}_{0}$, this contradicts the first assertion in (5.1).

Remark 5.2. If $\lambda_{n}<0$ for all $n \in \mathbb{N}$, we multiply (5.13) by -1 . From this point on, the proof is similar to the one presented above with the difference that some $\lambda_{n}$ would now be $-\lambda_{n}$.

Corollary 5.3. If $\eta=m_{1}=0$ and $m_{2}>0$, then the analytic semigroup $\left(\mathscr{T}_{0}(t)\right)_{t \geq 0}$ generated by $\mathscr{A}_{0}$ is exponentially stable. Thus, there exist constants $M>0$ and $\omega<0$ such that $\mathcal{E}(t) \leq M e^{\omega t} \mathcal{E}(0)$ for all $t \geq 0$, where $\mathcal{E}:=E_{0}$ and $E_{0}$ is as in (3.1).

Proof. By Remark 3.16, $\mathscr{A}_{0}$ is the infinitesimal generator of the $C_{0}$-semigroup of contractions $\left(\mathscr{T}_{0}(t)\right)_{t>0}$. Due to Corollary 3.6 in [109, p. 11], we have that $\rho\left(\mathscr{A}_{0}\right) \supset\{\lambda \in \mathbb{C}: \operatorname{Re} \lambda>0\}$ and as $i \mathbb{R} \subset \rho\left(\mathscr{A}_{0}\right)$, see Proposition 4.5, then

$$
\sigma\left(\mathscr{A}_{0}\right) \subset \mathbb{C} \backslash\{\lambda \in \mathbb{C}: \operatorname{Re} \lambda \geq 0\} .
$$

Now, Theorem 5.1] and Theorem 5.2 in [38] imply the exponential stability of the semigroup $\left(\mathscr{T}_{0}(t)\right)_{t \geq 0}$.

Another consequence of the analyticity of the solutions is the impossibility of localization. That means that the only solution that can be identically zero after a finite period of time is the null solution. This can be found in [20, p. 162].

Corollary 5.4. Let $\eta=m_{1}=0$ and $m_{2}>0$. Let us assume that $w=$ $\left(u, u_{t}, v, v_{t}, \theta\right)$ is a solution of the system (1.1)-(1.3) that satisfies (1.4)-(1.7) with initial conditions (1.8) and (1.9) such that $u=v=\theta=0$ after a finite time $t_{0}>0$. Then, $u=v=\theta=0$ for every $t \geq 0$.

Remark 5.5. In the proof of Theorem 4.16 it was shown that if we assume condition (4.99) with $\eta>0, m_{1}>0$ and $m_{2}=0$, then $i \mathbb{R} \subset \rho\left(\mathscr{A}_{\eta}\right)$. On the other hand, Theorem 4.15 implies that the $C_{0}$-semigroup of contractions $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ generated by $\mathscr{A}_{\eta}$ does not have exponential stability. Consequently, under the given conditions, we have that $\left(\mathscr{T}_{\eta}(t)\right)_{t \geq 0}$ is not analytic.

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## List of symbols

$\mathscr{B}_{1}, \mathscr{B}_{2}$
$B_{1}, B_{2}$
$\mathcal{L}(X, Y)$
$\mathcal{L}(X)$
id
$\mathbb{N}_{0}$
$|\alpha|$
$\partial^{\alpha}$
$C^{k}(\Omega)$
$C^{k}(\bar{\Omega})$
$C^{\infty}(\Omega)$
$\operatorname{supp} f$
$C_{c}^{\infty}(\Omega)$
$\mathscr{D}(\Omega)$
$\mathscr{D}^{\prime}(\Omega)$
$L^{p}(\Omega)$
$L_{l o c}^{1}(\Omega)$
$W^{k, p}(\Omega)$
$H^{k}(\Omega)$
$[s]$
$H^{s}(\Omega)$
$W^{s, p}(\Omega)$
$W_{0}^{s, p}(\Omega)$
$H_{0}^{k}(\Omega)$
$L^{p}((a, b), X)$
$\mathscr{D}^{\prime}((a, b), X)$
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| $\hookrightarrow, \stackrel{c}{\hookrightarrow}$ | 23 |
| :--- | :--- |
| $\mathcal{W}_{p, q}\left((0, t) ; B_{0}, B_{1}\right)$ | 23 |
| $\partial \Omega$ | 23 |
| $C^{k, 1}$ | 24 |
| $W^{s, p}(\partial \Omega)$ | 24 |
| $H^{s}(\partial \Omega)$ | 24 |
| $L^{p}(\partial \Omega)$ | 24 |
| $H^{-s}(\partial \Omega)$ | 24 |
| $\frac{\partial^{j} u}{\partial \nu^{j}}$ | 25 |
| $\gamma_{0}, \gamma_{1}$ | 25 |
| $\Delta u$ | 26 |
| $\nabla u$ | 26 |
| $C^{k}$ | 26 |
| $B(x, r)$ | 26 |
| $C^{\infty}$ | 26 |
| $\operatorname{div} h$ | 27 |
| $\nabla^{2} u: \nabla^{2} \bar{v}$ | 27 |
| $\hookrightarrow^{d}$ | 32 |
| $A^{n}$ | 33 |
| $\rho(A)$ | 34 |
| $\sigma(A)$ | 34 |
| $\|\cdot\|_{D(A)}$ | 33 |
| $X$ | 37 |
| $D\left(A^{\prime}\right)$ | 38 |
| $X_{\alpha}$ | 39 |
| $A^{s}$ | 40 |
| $(\cdot, \cdot)_{D\left(A^{s}\right)}$ | 40 |

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$\begin{array}{ll}\operatorname{div} h & 27 \\ \nabla^{2} u: \nabla^{2} \bar{v} & 27\end{array}$
$\nabla_{d}^{2} u: \nabla^{2} \bar{v} \quad 27$
32
$A^{n} \quad 33$
$\rho(A) \quad 34$
$\sigma(A) \quad 34$
$|\cdot|_{D(A)} \quad 33$
$X_{-1} \quad 37$
$D\left(A^{\prime}\right) \quad 38$
$X_{\alpha} \quad 39$
$A^{s} \quad 40$
$(\cdot, \cdot)_{D\left(A^{s}\right)} \quad 40$

List of symbols

| $E_{\gamma}(t)$ | 45 | $\mathcal{A}_{T}$ | 51 |
| :--- | :--- | :--- | :--- |
| $H_{\Gamma}^{k}\left(\Omega_{1}\right)$ | 46 | $\mathcal{G}_{1}, \mathcal{G}_{2}$ | 52 |
| $\eta$ | 47 | $\mathcal{N}$ | 52 |
| $H_{\Gamma, \eta}^{1}\left(\Omega_{1}\right)$ | 47 | $\mathscr{A}_{\eta}$ | 56 |
| $\mathscr{H}_{\eta}$ | 47 | $\mathscr{P}$ | 56 |
| $\mathcal{A}_{B}$ | 48 | $\mathcal{W}$ | 56 |
| $\cong$ | 49 |  |  |
| $\mathcal{A}_{L}$ | 50 | $\mathscr{A}_{0}$ | 57 |
| $\mathcal{M}_{\eta}$ | 51 | $\mathcal{W}_{0}$ | 57 |

## Index

## A

abstract Cauchy problem, 32
adjoint operator $A^{\prime}, 38$
adjoint operator $A^{*}, 37$
analytic operator, 36
Aubin-Lions-Simon lemma, 22

## B

boundary of an open set, 23
boundary spaces, 24
boundary value problem, 42

## C

class $C^{\infty}, 26$
class $C^{k}, 26$
class $C^{k, 1}, 24$
classic solution, 32
closure of an operator, 18
continuously embedded, 23

## D

dissipative operator, 32
distributional derivative, 20
domain, 18

## E

elliptic operator, 41
energy space, 47
Euler-Bernoulli plate equation, 2
exponentially stable semigroup, 34
extrapolated semigroup, 38
extrapolation space, 38

## F

first Green's formula, 26
Friedrichs inequality, 30

G
generalized Poincaré inequality, 30
geometrical condition, 84
gradient vector, 26
graph norm, 33
Green maps, 52
Green's formulas, 25
group, 36

## I

inertia operator, 51
infinitesimal generator, 31
integration by parts, 25
interface, 1
internal damping, 10
interpolation inequality, 30
interpolation-extrapolation scale, 39

## K

Kelvin-Voigt damping, 2
Kirchhoff plate equation, 2

## L

lack of exponential stability, 77
Laplace operator, 26
Lipschitz boundary, 24
Lipschitz function, 23
locally integrable function, 20
Lumer-Phillips theorem, 32
M
maximal monotone operator, 48
mild solution, 33
monotone operator, 48
multi-index, 19

## N

Neumann map, 52
normal derivative, 26
O
operator norm, 30

## P

phase space, 47
polynomially stable semigroup, 34
positive self-adjoint operator, 39

## R

realization of an operator, 18
regularity, 62
Rellich identity, 27
Rellich-Kondrachov theorem, 23
resolvent set, 34
rotational inertia, 2

## S

second Green's formula, 26
self-adjoint operator, 37
semigroup, 30
skew-adjoint operator, 37
Sobolev space, 21
space of distributions, 20
spectrum of an operator, 34
standard extension, 40
Stone theorem, 37
strongly continuous semigroup, 31
strongly stable semigroup, 34
structural damping, 2
support, 20
T
test functions, 20
trace operator, 25
trace theorem, 25

## U

unitary operator, 37

## V

vector distributional derivative, 22
vector-valued distributions, 22

## W

well-posedness, 45

## Y

Young's inequality, 28


[^0]:    ${ }^{1}$ cf. this concept in [123, Definition 2.1].

