# Counting and Sampling Problems on Eulerian Graphs 

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## Abstract

In this thesis we consider two sets of combinatorial structures defined on an Eulerian graph: the Eulerian orientations and Euler tours. We are interested in the computational problems of counting (computing the number of elements in the set) and sampling (generating a random element of the set). Specifically, we are interested in the question of when there exists an efficient algorithm for counting or sampling the elements of either set

The Eulerian orientations of a number of classes of planar lattices are of practical significance as they correspond to configurations of certain models studied in statistical physics. In 1992 Mihail and Winkler showed that counting Eulerian orientations of a general Eulerian graph is \#P-complete and demonstrated that the problem of sampling an Eulerian orientation can be reduced to the tractable problem of sampling a perfect matching of a bipartite graph. We present a proof that this problem remains \#Pcomplete when the input is restricted to being a planar graph, and analyse a natural algorithm for generating random Eulerian orientations of one of the afore-mentioned planar lattices. Moreover, we make some progress towards classifying the range of planar graphs on which this algorithm is rapidly mixing by exhibiting an infinite class of planar graphs for which the algorithm will always take an exponential amount of time to converge.

The problem of counting the Euler tours of undirected graphs has proven to be less amenable to analysis than that of Eulerian orientations. Although it has been known for many years that the number of Euler tours of any directed graph can be computed in polynomial time, until recently very little was known about the complexity of counting Euler tours of an undirected graph. Brightwell and Winkler showed that this problem is \#P-complete in 2005 and, apart from a few very simple examples, e.g., series-parellel graphs, there are no known tractable cases, nor are there any good reasons to believe the problem to be intractable. Moreover, despite several unsuccessful attempts, there has been no progress made on the question of approximability. Indeed, this problem was considered to be one of the more difficult open problems in approximate counting since long before the complexity of exact counting was resolved. By considering a randomised input model, we are able to show that a very simple algorithm can sample or approximately count the Euler tours of almost every $d$-in/ $d$-out directed graph in expected polynomial time. Then, we present some partial results towards showing that this algorithm can be used to sample or approximately count the Euler tours of almost every $2 d$-regular graph in expected polynomial time. We also provide some empirical evidence to support the unproven conjecture required to obtain this result. As a sideresult of this work, we obtain an asymptotic characterisation of the distribution of the number of Eulerian orientations of a random $2 d$-regular graph.

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## Declaration

I declare that this thesis was composed by myself, that the work contained herein is my own except where explicitly stated otherwise in the text, and that this work has not been submitted for any other degree or professional qualification except as specified. Most of the material in Chapter 2 and some of Section 3.3 of this thesis have been published in:

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## Chapter 1

## Introduction

Counting problems are a fundamental class of problems in the fields of combinatorics and algorithms and complexity. Moreover, solutions to such problems are of great practical value across the sciences because many real problems can be modelled as counting problems, particularly those related to the estimation of probabilities. For example, evaluating thermodynamic properties of physical systems [6], calculating the output probabilities of Bayesian networks [24], and calculating the probability of correct reconstruction of DNA from fragments using sequencing by hybridisation [4] are all counting problems.

Many counting problems, including those studied in this thesis, are defined on graphs. Before proceeding, we summarise the definitions of graphs and the graphtheoretic properties we will use later in the thesis. For a thorough introduction to the theory of graphs see, for example, [90].

Definition 1.1. An undirected graph $G=(V, E)$ is a set of vertices $V$ and a set of edges $E$; each $e \in E$ is an unordered pair of vertices, $e=\{u, v\}$, where $u, v \in V$. The degree of a vertex $v \in V$ is the number of edges $e \in E$ containing $v$. We denote the degree of a vertex by $\operatorname{deg}(v)$.

Definition 1.2. A directed $\operatorname{graph} \vec{G}=(V, A)$ is a set of vertices $V$ and a set of arcs $A$; each $a \in A$ is an ordered pair of vertices, $a=(u, v)$, where $u, v \in V$. We say an arc $a=(u, v)$ is directed towards $v$ and away from $u$, and call $u$ and $v$ the source and target of $a$, respectively. The out-degree (resp. in-degree) of a vertex $v \in V$ is the number of arcs directed away from (resp. towards) $v$ in $A$. We denote the in-degree and out-degree of $v$ by $\operatorname{indeg}(v)$ and $\operatorname{outdeg}(v)$, respectively.

Definition 1.3. We say a directed graph $\vec{G}=(V, A)$ is an orientation of a graph
$G=(V, E)$ if there exists a bijection from $E$ to $A$ in which each $\{u, v\} \in E$ is mapped to $(u, v)$ or $(v, u)$.

Typically, we do not allow loops (edges or arcs where $u=v$ ) or multiple copies of the same edge or arc. However, in some cases we wish to allow duplicates, in which case we speak of multigraphs and directed multigraphs.

Definition 1.4. An undirected multigraph $G=(V, E)$ is a set of vertices $V$ and a multiset of edges $E$.

Definition 1.5. A directed multigraph $G=(V, A)$ is a set of vertices $V$ and a multiset of $\operatorname{arcs} A$.

Remark 1.6. A graph or directed graph which does not contain any loops or multiple edges is often said to be simple. A simple directed graph is not necessarily an orientation of a simple graph, since $(u, v)$ and $(v, u)$ do not count as the same arc.

Often, we are interested in some particular substructure of a graph or directed graph $G$.

Definition 1.7. Let $G=(V, E)$ be a graph or directed graph. A subgraph of $G$ is a graph $H=\left(V^{\prime}, E^{\prime}\right)$, where $V^{\prime} \subseteq V$ and $E^{\prime} \subseteq E$. If $V^{\prime}=V$, we say $H$ is a spanning subgraph of $G$. If $H \neq G$ we say $G$ is a proper subgraph of $G$.

Definition 1.8. A graph $G=(V, E)$ with vertex set $V=\left\{v_{0}, v_{1}, \ldots, v_{k-1}\right\}$ and edge set

$$
E=\left\{\left\{v_{i}, v_{i+1} \bmod k\right\}: i=0,1, \ldots, k-1\right\},
$$

is called a cycle. The subgraphs of a graph $G$ which are cycles are called the cycles of G.

Definition 1.9. A directed graph $G=(V, A)$ with vertex set $V=\left\{v_{0}, v_{1}, \ldots, v_{k-1}\right\}$ and arc set

$$
A=\left\{\left(v_{i}, v_{i+1} \bmod k\right): i=0,1, \ldots, k-1\right\},
$$

is called a directed cycle. The subgraphs of a directed graph $\vec{G}$ which are directed cycles are called the directed cycles of $\vec{G}$, or sometimes the cycles of $\vec{G}$.

Definition 1.10. A connected graph is called a tree if it contains no cycles. The spanning subgraphs of $G$ which are trees are called the spanning trees of $G$.

Definition 1.11. A rooted tree is a tree $T=(V, E)$ with a single distinguished root vertex $r$. For each $v \in V$, the first node on the unique path from $v$ to $r$ is called the parent of $v$; all other neighbours of $v$ are called the children of $v$.

Definition 1.12. An arborescence with root $v$ is a rooted tree in which every edge has been directed towards $v$.

Let $\vec{G}=(V, A)$ be a directed graph and let $v \in V$. The set of arborescences of $\vec{G}$ rooted at $v$, denoted $\operatorname{ARB}(\vec{G}, v)$, is the set of spanning subgraphs of $\vec{G}$ which are arborescences rooted at $v$. We denote by $\operatorname{ARB}(\vec{G})$ the set of all arborescences of $G$.

There are several classes of graphs in which some additional conditions are placed on $E$ or $A$.

Definition 1.13. A graph $G=(V, E)$ is said to be bipartite if there exists a partition of $V$ into $A \cup B$ such that for every $e=\{u, v\} \in E$, either $u \in A$ and $v \in B$ or $v \in A$ and $u \in B$.

Definition 1.14. Let $G=(V, E)$ be an undirected graph. We say $G$ is $d$-regular if we have $\operatorname{deg}(v)=d$ for every vertex $v \in V$.

Definition 1.15. Let $\vec{G}=(V, A)$ be a directed graph. We say $\vec{G}$ is $d$-in/d-out if we have $\operatorname{indeg}(v)=\operatorname{outdeg}(v)=d$ for every vertex $v \in V$.

Definition 1.16. A graph $G$ is called planar if it can be drawn in the plane such that

1. each vertex is mapped to a unique point in $\mathbb{R}^{2}$;
2. each edge $\{u, v\}$ is a line whose endpoints are the points for $u$ and $v$;
3. the interior of an edge $\{u, v\}$ contains no vertex and no point of any other edge. This drawing is called an embedding of $G$. An embedding of $G$ can be viewed as a decomposition of the plane $\mathbb{R}^{2}$ into a number of regions, exactly one of which is unbounded. We refer to the sets of edges bounding these regions as the faces of $G$. The unbounded region is referred to as the unbounded face of $G$.

Generally, when we speak of a graph $G$ being planar we are considering the graph $G$ and a fixed embedding of $G$. Assuming we have a fixed embedding of $G$, we have the notion of dual graph of $G$.

Definition 1.17. Let $G$ be a planar graph with some fixed embedding in the plane. The dual graph of $G$, denoted $G^{\star}$, is the graph whose vertices are the faces of $G$ and which has an edge joining each pair of faces $\eta$ and $\sigma$ which share an edge in $G$.

The classes of graphs and directed graphs which are the focus of this thesis are those which satisfy the Eulerian condition.

Definition 1.18. An undirected graph $G=(V, E)$ is said to be Eulerian if all vertices have even degree. A directed graph $\vec{G}=(V, A)$ is said to be Eulerian if we have $\operatorname{indeg}(v)=\operatorname{outdeg}(v)$ for every vertex $v \in V$.

Note that for all $d \geq 1$, every $2 d$-regular undirected graph, and every $d$-in $/ d$-out directed graph, satisfies the Eulerian condition. We are interested in two closely related structures which can only be defined on Eulerian graphs: Eulerian orientations and Euler tours.

Definition 1.19. Let $G=(V, E)$ be an undirected graph and let $\alpha: V \rightarrow \mathbb{Z}$ satisfy

$$
\sum_{v \in V} \alpha(v)=0
$$

We say an orientation of $G$ is an $\alpha$-orientation ${ }^{1}$ of $G$ if for every $v \in V$,

$$
\operatorname{outdeg}(v)-\operatorname{indeg}(v)=\alpha(v) .
$$

If $\alpha(v)=0$ for all $v \in V$, then each $\alpha$-orientation of $G$ is an Eulerian directed graph. In this case, we refer to the $\alpha$-orientations as Eulerian orientations, and denote this set by $\mathrm{EO}(G)$.

Remark 1.20. The set $\operatorname{EO}(G)$ is non-empty if and only if $G$ satisfies the Eulerian condition (Definition 1.18).

Definition 1.21. Let $G=(V, E)$ be a graph. We define a tour of $G$ to be a sequence of vertices $v_{0} v_{1} v_{2} \ldots v_{k-1}$ in which $\left\{v_{i}, v_{i+1} \bmod k\right\} \in E$ for $i=0,1, \ldots, k-1$. An Euler tour of $G$ is a tour that uses every edge exactly once. Two Euler tours are equivalent if one is a cyclic permutation of the other. We denote the set of all Euler tours of a graph $G$ by $\mathrm{ET}(G)$.

Definition 1.22. Let $\vec{G}=(V, A)$ be a directed graph. We define a tour of $G$ to be a sequence of vertices $v_{0} v_{1} v_{2}, \ldots, v_{k-1}$ in which $\left(v_{i}, v_{i+1} \bmod k\right) \in A$ for $i=0,1, \ldots, m-1$. An Euler tour is a tour that uses every arc exactly once. Two Euler tours are equivalent if one is a cyclic permutation of the other. We denote the set of all Euler tours of a directed graph $\vec{G}$ by $\mathrm{ET}(\vec{G})$.

[^0]Remark 1.23. We can also define Euler tours of multigraphs. In this case, we treat the duplicate edges as distinct so a sequence $v_{0} v_{1} \ldots v_{m-1}$ can correspond to multiple Euler tours, as we assign each occurrence of $u v$ or $v u$ to a distinct instance of the duplicated edge $\{u, v\}$.

Remark 1.24. The set $\mathrm{ET}(G)$ is non-empty if and only if $G$ connected and satisfies the Eulerian condition (Definition 1.18).

The remainder of this chapter is organised as follows. In §1.1.1 we summarise the theory of the complexity of counting problems. In §1.1.2 we describe the relationship between counting and sampling that is the key tool for finding approximate solutions to counting problems. In $\S 1.2$ we summarise the basic techniques used to construct sampling algorithms. Finally, in $\S 1.3$ we survey previous work on the problems of counting/sampling Eulerian orientations and Euler tours and put the work of this thesis into its correct context.

### 1.1 Counting and Sampling

### 1.1.1 Computational Complexity of Counting

In theoretical computer science, computational problems are often viewed as questions about relations between problem instances and sets of feasible solutions. Formally, if we consider instances and solutions to be encoded in the same finite alphabet $\Sigma$ then a computational problem can be viewed as a relation $R \subseteq \Sigma^{\star} \times \Sigma^{\star}$ which maps each instance $x \in \Sigma^{\star}$ to a (finite) set of solutions

$$
\Omega(x)=\left\{y \in \Sigma^{\star}:(x, y) \in R\right\} .
$$

The most widely studied computational problems of this form are decision problems, where we are asking if a particular input has a solution.

Definition 1.25. Let $\Sigma$ be a finite alphabet and $R \subset \Sigma^{\star} \times \Sigma^{\star}$. For a particular instance $x \in \Sigma^{\star}$, the decision problem R asks whether there exists some $y \in \Sigma^{\star}$ such that $(x, y) \in R$. We say a decision problem R is in the class NP (non-deterministic polynomial time) if

1. there exists a polynomial $p(\cdot)$ such that for any instance $x \in \Sigma^{\star}$ we have

$$
|y| \leq p(|x|) \quad \forall y \in \Sigma^{\star} \text { such that }(x, y) \in R ;
$$

2. there exists a polynomial time algorithm for testing the predicate $(x, y) \in R$.

The subset of NP consisting of problems for which there exist polynomial time algorithms is known as P . Although it is not known whether or not there exists a polynomial time algorithm for all problems in NP, it is well known [20, 60] (see [89] for an English translation of the second article and [3, Chapter 2] for a general reference) that there exists another subset of NP, known as the NP-complete problems, which are the "hardest" problems in NP; that is, if we had a polynomial time algorithm for any NP-complete problem then we could construct a polynomial time algorithm for any problem in NP. The canonical problem in this class is the Boolean satisfiability problem, defined below.

Example 1.26. The Boolean satisfiability problem (SAT) is a decision problem whose instances are Boolean expressions written using only $\wedge, \vee$, and $\neg$. Given a Boolean expression $\psi$, we want to determine whether there exists an assignment to the variables in $\psi$ that will make $\psi$ true. Given any particular assignment to the variables, we can quickly check whether or not the expression evaluates to true, so the problem is in NP. It was shown by Cook [20] (and, simultaneously, by Levin [60]) that for any instance $x$ of an NP relation R we can construct an instance of SAT, $\psi_{\mathrm{R}}$, such that there exists some $(x, y) \in \mathrm{R}$ if and only if $\psi_{\mathrm{R}}$ has a satisfying assignment. Hence, SAT is NP-complete.

There are two possible scenarios: either $\mathrm{P}=\mathrm{NP}$, in which case every algorithm in NP has a polynomial time algorithm; or $\mathrm{P} \subset \mathrm{NP}$, in which case there does not exist a polynomial time algorithm for any NP-complete problem.

Counting problems are a natural generalisation of decision problems. A decision problem asks whether some set of objects is non-empty, whereas a counting problem asks for something stronger: the exact number of objects in the set. Valiant formalised the computational complexity of counting problems in his seminal 1979 paper [92], in which he defined the complexity class \#P.

Definition 1.27. Let $\Sigma$ be a finite alphabet. Given a relation $R \subset \Sigma^{\star} \times \Sigma^{\star}$ we can define the counting problem for $R$ as the problem of computing a function $N_{R}: \Sigma^{\star} \rightarrow \mathbb{N}$ with values

$$
N_{R}(x)=\left|\left\{y \in \Sigma^{\star}:(x, y) \in R\right\}\right| .
$$

We say the counting problem $N_{R}$ is in the class \# P if

1. there exists a polynomial $p(\cdot)$ such that for any instance $x \in \Sigma^{\star}$ we have

$$
|y| \leq p(|x|) \quad \forall y \in \Sigma^{\star} \text { such that }(x, y) \in R ;
$$

2. there exists a polynomial time algorithm for testing the predicate $(x, y) \in R$.

In general, for any reasonable function $f: \Sigma^{\star} \rightarrow \mathbb{N}$ there exists some relation $R \subseteq \Sigma^{\star} \times \Sigma^{\star}$ such that $f=N_{R}$. Thus, we will usually refer to counting problems as functions from $\Sigma^{\star}$ to $\mathbb{N}$, without referring to their defining relation $R$.

An algorithm that counts the number of solutions to a problem can also determine whether or not there exists a solution, so we cannot expect every counting problem to be efficiently solvable unless $\mathrm{P}=\mathrm{NP}$. We denote by FP the subset of $\# \mathrm{P}$ which consists of all functions $f: \Sigma^{\star} \rightarrow \mathbb{N}$ which can be computed in polynomial time. As in the decision case, the set of problems which are "hardest" for \#P are called \#P-complete.

Definition 1.28. Given two counting problems $f$ and $g$ we say $f$ is polynomial-time Turing reducible to $g$, denoted $f \leq_{P T} g$, if there exists a Turing machine with an oracle for $f$ that computes $g(x)$ in time polynomial in $|x|$. We say a problem $f$ is \#P-complete if every problem in \#P is polynomial-time Turing reducible to $f$.

Remark 1.29. \#P-completeness is accepted to be strong evidence of computational intractability: exhibiting a polynomial time algorithm for all problems in \#P would imply the collapse of the polynomial-time hierarchy [88] (for a general reference, see [3, Chapter 17]).

The construction used in [20] to show that SAT is NP-complete can be used to construct an instance of \#SAT with the same number of solutions as any \#P problem [92, Lemma 2]. Hence, every \#P problem is reducible to \#SAT; that is to say, \#SAT is \#P-complete. This is the observation used by Valiant in [92], in which \#SAT is taken to be the defining problem for the class of \#P-complete problems. The interesting result in [92] was that there exist \#P-complete problems for which the decision problem is in P. Specifically, Valiant showed that every problem in \#P can be reduced to counting perfect matchings in bipartite graphs; that is, counting perfect matchings of a bipartite graph is \# P -complete. But the question of whether or not a bipartite graph has a perfect matching can be solved by computing a maximum flow in a network, something which can be achieved in polynomial time [98, §3.8].

There are some non-trivial counting problems for which polynomial-time algorithms exist. For example, counting perfect matchings in a bipartite planar graph $G$ can
be achieved by evaluating the determinant of a special matrix, known as the Pfaffian [54]. Similarly, the number of spanning trees of a graph can be computed by evaluating the determinant of another special matrix, known as the Laplacian [55] (see §1.3.2 for a description of a related result). It is well-known that the determinant of any matrix can be computed in polynomial time [85]; hence, both these problems are in FP. See [47, Chapter 1] for a nice presentation of both these results.

However, the majority of interesting counting problems tend to be \#P-complete. For example, counting perfect matchings [92], counting Eulerian orientations [69], counting solutions to propositional formulae in DNF [93], and estimating the partition function of several important models in statistical physics [97] are all examples of \#Pcomplete problems. In all these cases, the best known algorithms for exact counting have exponential running time, despite the corresponding decision problems lying in P . This has shifted the focus towards designing polynomial algorithms for approximate counting. Many of these algorithms use randomisation to produce a close estimate of the true answer with high probability, but run in polynomial time. That is, at some points during the execution of the algorithm a "fair coin" is tossed and the next step of the algorithm depends on the outcome. This leads to algorithms which can produce different outputs on different executions with the same input and sometimes produce an output that does not match the requirements of the problem. The goal in analysing this type of algorithm is not only to prove that they run in polynomial time but also that the probability of the answer being correct is sufficiently high. Formally, we are interested in designing algorithms with the following specifications:

Definition 1.30. Let $\Sigma$ be a finite alphabet and let $f: \Sigma^{*} \rightarrow \mathbb{N}$ be a counting problem on $\Sigma$. A randomised approximation scheme for $f$ with confidence parameter $\delta$ is a randomised algorithm that takes as input an instance $x \in \Sigma^{*}$ and an error parameter $\varepsilon$, and outputs a number $N \in \mathbb{N}$ (this is a random variable of the "coin tosses" made by the algorithm),

$$
\begin{equation*}
\mathbb{P}[(1-\varepsilon) f(x) \leq N \leq(1+\varepsilon) f(x)] \geq 1-\delta . \tag{1.1}
\end{equation*}
$$

We call this a fully polynomial randomised approximation scheme, or an fpras, if the algorithm runs in time bounded by a polynomial in $|x|, \varepsilon^{-1}$ and $\log \left(\delta^{-1}\right)$ for every instance $x$.

Remark 1.31. We cannot expect an fpras to exist for every problem in \#P. For example, suppose we had an fpras for \#SAT. That is, we have an algorithm $\mathcal{A}$ that takes as input
a propositional formula $\psi$ and returns a number $N$ satisfying

$$
(1-\varepsilon) \# \operatorname{SAT}(\psi) \leq N \leq(1+\varepsilon) \# \operatorname{SAT}(\psi)
$$

with probability $3 / 4$. Then, if $\psi$ has no satisfying assignments $\mathcal{A}$ returns 0 with probability $3 / 4$. But if $\psi$ has one or more satisfying assignments $\mathcal{A}$ returns some number $N>0$ with probability $3 / 4$. Hence, $\mathcal{A}$ can determine whether or not $\psi$ is satisfiable with $3 / 4$ accuracy. It is widely believed [3, Chapter 7] that such an algorithm cannot exist for any NP-complete problem; hence, it is unlikely that an fpras for \#SAT exists.

In light of Remark 1.31 we only search for an fpras for a problem $f \in \# \mathrm{P}$ if we know that the corresponding decision problem is in P .

### 1.1.2 The relationship between counting and sampling

In almost all known approximate counting algorithms, the use of randomness referred to in Definition 1.30 is necessary. That is to say, apart from a few notable cases, e.g., [96, 8], the only known approximate counting algorithms are randomised algorithms. One of the reasons for this fact is that most approximation algorithms for \#P-complete problems take advantage of a close relationship between approximate counting and random generation discovered in 1986 by Jerrum, Valiant, and Vazirani [52]. Before we describe this relationship, we define the sampling problem for a relation R .

Definition 1.32. Let $\Sigma$ be a finite alphabet and let $\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}$. The sampling problem for R asks for a randomised algorithm that, given an instance $x \in \Sigma^{\star}$, generates uniformly at random $y \in \Sigma^{\star}$ such that $(x, y) \in \mathrm{R}$.

Often, it is too much to ask for an algorithm that samples exactly from the uniform distribution. However, when this is the case we are sometimes able to come up with an algorithm that generates samples from a distribution that is close enough to uniform for practical purposes. In order to be able to speak about the accuracy of a particular sampling algorithm we need a notion of distance between probability distributions; the variation distance (defined below) is a standard measure of distance between two probability distributions [41, Chapter 2].

Definition 1.33. The total variation distance $\left\|\mu_{1}-\mu_{2}\right\|_{T V}$ between two distributions $\mu_{1}, \mu_{2}$ on a set $\Omega$ is defined as

$$
\left\|\mu_{1}-\mu_{2}\right\|_{T V}=\frac{1}{2} \sum_{\omega \in \Omega}\left|\mu_{1}(\omega)-\mu_{2}(\omega)\right|=\max _{A \subseteq \Omega}\left|\mu_{1}(A)-\mu_{2}(A)\right| .
$$



Figure 1.1: Top three levels of a decomposition tree for the perfect matching problem

Now that we have a notion of distance between probability distributions, we are able to define what it means for an algorithm to be an almost uniform sampler.

Definition 1.34. Let $\Sigma$ be a finite alphabet and $\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}$. An almost uniform sampler for R is a randomised algorithm which, for instance $x \in \Sigma^{*}$ and sampling tolerance $\delta>0$, outputs a random variable $W \in\left\{y \in \Sigma^{\star}:(x, y) \in \mathrm{R}\right\}$ such that the variation distance between the distribution of $W$ and the uniform distribution is no more than $\delta$. We call this a fully polynomial almost uniform sampler, or an fpaus, if the algorithm runs in time polynomial in $|x|$ and $\log \delta^{-1}$ for all instances $x$.

There is a class of relations for which the existence of an fpras for the counting problem is equivalent to the existence of a fpaus for the sampling problem [52]. The relations for which this is the case are those which satisfy a property known as selfreducibility.

Definition 1.35. A relation $\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}$ is said to be self-reducible if and only if

1. there exists a polynomial time computable function $g: \Sigma^{\star} \rightarrow \mathbb{N}$ such that

$$
(x, y) \in R \Rightarrow|y|=g(|x|) ;
$$

2. there exist polynomial time computable functions $\psi: \Sigma^{\star} \times \Sigma^{\star} \rightarrow \Sigma^{\star}$ and $\sigma: \Sigma^{\star} \rightarrow \mathbb{N}$ satisfying

$$
\begin{align*}
& \sigma(x)=O(\log (|x|)),  \tag{1.2}\\
& g(x)>0 \rightarrow \sigma(x)>0 \quad \forall x \in \Sigma^{\star},  \tag{1.3}\\
& |\psi(x, w)|<|x| \quad \forall x, w \in \Sigma^{\star} \tag{1.4}
\end{align*}
$$

and such that, for all $x \in \Sigma^{\star}$ and all $y=y_{1} \cdots y_{g(|x|)} \in \Sigma^{\star}$,

$$
\begin{equation*}
\left(x, y_{1} y_{2} \cdots y_{g(|x|)}\right) \in \mathbf{R} \Leftrightarrow\left(\psi\left(x, y_{1} \cdots y_{\sigma(x)}\right), y_{\sigma(x)+1} \cdots y_{g(|x|)}\right) \in \mathbf{R} . \tag{1.5}
\end{equation*}
$$

The above definition requires that if we choose the first $\sigma(x)$ bits of a solution to $x$ we can construct another smaller instance of the same problem $x^{\prime}=\psi\left(x, y_{1} \cdots y_{\sigma(x)}\right)$ such that the solutions of $x$ starting with $y_{1} \cdots y_{\sigma(x)}$ are obtained by concatenating $y_{1} \cdots y_{\sigma(x)}$ with a solution $y^{\prime}$ of $x^{\prime}$, and that every solution of $x^{\prime}$ gives a solution to $x$ in this way.

Intuitively, this says that, given any instance $x$, we can construct a number of smaller instances of the same problem such that the solutions of $x$ are obtained by extending solutions to the smaller instances. Thus, we can think of the function $\psi$ as decomposing the set of solutions to $x$ into a tree structure such that

1. each vertex is a subset of the solutions to $x$, with all solutions at the root, and the sets at the children of a node partitioning the set at that node ;
2. each edge is labelled with a string $w \in \Sigma^{\star}$;
3. each leaf is of the form $S=\{y\}$, with $y$ equal to the concatenation of the labels on the path from the root to $S$.

We will now demonstrate the concept of self-reducibility with a concrete example.
Example 1.36. Let $G=(V, E)$ be a graph. A matching of $G$ is a set $M \subset E$ such that each vertex is contained in no more than one $e \in M$. We say a matching $M$ is perfect if every vertex in $V$ is contained in some edge $e \in M$. We denote by $\operatorname{PM}(G)$ the set of perfect matchings of $G$. Now, suppose we have a relation R where $(x, y) \in \mathrm{R}$ if and only if $x$ encodes a graph $G=(V, E)$ and $y$ encodes a perfect matching of $G$. Each solution is a string $y_{1} y_{2} \cdots y_{m}$ where $y_{i}=1$ if $e_{i} \in M$ and $y_{i}=0$ if $e_{i} \notin M$ (we are assuming some order on the edges $E=\left\{e_{1}, e_{2}, \ldots, e_{m}\right\}$ ).

Given any edge $e=\{u, v\}$ of $G$ we can construct two graphs $G_{e}$, in which we remove $u$ and $v$ and all edges incident with them from $G$, and $G_{\bar{e}}$, in which we only remove $e$ from the set of edges. Every perfect matching of $G_{\bar{e}}$ is a perfect matching of $G$; in fact, the perfect matchings of $G_{\bar{e}}$ are exactly the set of perfect matchings of $G$ that do not contain $e$. The perfect matchings of $G_{e}$ are not perfect matchings of $G$, since neither $u$ nor $v$ is contained in any perfect matching of $G_{e}$. But we can add $e$ to every perfect matching of $G_{e}$ to obtain a perfect matching of $G$; hence, the perfect
matchings of $G_{e}$ correspond exactly to the perfect matchings of $G$ that contain $e$. By repeating this process, we obtain a tree decomposition of the set of perfect matchings of $G$ in the form of Figure 1.1

Suppose we have a self-reducible relation R and an fpras, COUNT, for the counting problem of R. The fpaus implied by the result of [52] is given (roughly) in Algorithm 1. If we had an algorithm for computing $N_{\mathrm{R}}$ exactly, then Algorithm 1 would be an exact sampler. However, since we only have an fpras we must make do with approximations to each of the $N_{\mathrm{R}}\left(x_{i}\right)$ when we compute the probabilities at Step 4. However, we are able to obtain close enough approximations to the values $N_{\mathrm{R}}\left(x_{i}\right)$ to ensure that the algorithm is an fpaus. See [52] for details.

```
Algorithm 1: GEN
    Input: \(x \in \Sigma^{\star}\) and self-reducible \(\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}\)
    Output: An almost uniformly random \(y \in \Sigma^{\star}\) such that \((x, y) \in \mathrm{R}\)
    begin
        Obtain \(x_{1}, x_{2}, \ldots, x_{n}\) from \(x\) by choosing the first \(\sigma(|x|)\) bits of the solution;
        for \(i \leftarrow 1\) to \(n\) do
            \(\rho_{i} \leftarrow \operatorname{COUNT}\left(x_{i}\right) ;\)
        end
        Set \(x^{\prime}=x_{i}\) with probability proportional to \(\rho_{i}\);
        \(y_{i} \leftarrow \operatorname{GEN}\left(x^{\prime}\right) ;\)
        return the \(y \in \Sigma^{\star}\) with \((x, y) \in R\) corresponding to \(y_{i}\);
    end
```

To go in the other direction, turning an fpaus into an fpras we use Algorithm 2. Assuming we have an fpaus for a self-reducible R, GEN, Algorithm 2 is an fpras for $N_{\mathrm{R}}$. Again, see [52] for the details. In §2.7, we describe the application of Algorithm 2 to a concrete example.

Fortunately, for those of us interested in algorithms for counting and sampling, a great deal of \#P-complete problems are self-reducible. Moreover, in the cases where they are not, e.g., transposition tables [22, 23], the problems often satisfy a slightly weaker self-reducible property which means we can still use Algorithm 2 to construct approximate counting algorithms.

```
Algorithm 2: COUNT
    Input: \(x \in \Sigma^{\star}\) and self-reducible \(\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}\) and a constant \(\varepsilon>0\)
    Output: A number \(N\) such that \((1-\varepsilon) N \leq N_{\mathrm{R}}(x) \leq(1+\varepsilon) N\)
    begin
        \(\Pi:=1, m:=g(x), t:=180|x|^{3 c} ;\)
        while \(g(x)>0\) do
            Let \(S=\left\{y_{1}, \ldots, y_{t}\right\}\) be the set of results obtained by making \(t\) calls to
            GEN with input ( \(x, \varepsilon / 11 m\) );
            Let \(w\) be the most common prefix of elements of \(S\);
            \(\rho:=\mid\{y \in S: w\) is an initial segment of \(y\}|/|S| ;\)
            \(x:=\psi(x, w)\);
            \(\Pi:=\Pi / \rho ;\)
        end
        return \(\Pi\)
    end
```


### 1.1.3 Applications of counting/sampling algorithms

An important class of counting/sampling problems come from the field of statistical mechanics (also known as statistical physics) [7]. Statistical mechanics provides a framework for relating the microscopic (local) properties of individual atoms and molecules in some physical object to the macroscopic (global) properties that can be observed in the real world; in particular, it provides an interpretation of the thermodynamic properties of an object, such as free energy and entropy, in terms of the microscopic properties of the configuration of the particles within that object. These systems will, in general, contain a large number of particles, thus rendering exact computation impractical due to an unfeasibly large number of possible configurations. However, if we have an fpaus for generating random configurations, we can generate a large number of samples and then estimate the thermodynamic property of interest by averaging over all the samples generated. Moreover, an important quantity known as the partition function is no more than the solution to the counting problem which enumerates the number of configurations (possibly with some weight associated with each configuration). In many cases, evaluating the partition function is \#Pcomplete [45, 97] and many algorithms for approximating the partition function use Markov chain Monte Carlo [50, 62]. One of the problems studied in this thesis,
counting and sampling Eulerian orientations of planar graphs, is equivalent to counting and sampling configurations in so-called ice-type models [7, Chapter 8].

### 1.2 Markov chain Monte Carlo

In this section we define Markov chains and explain how they are used in the construction of algorithms for almost uniform sampling. General references for the material of this section are, e.g., [41, Chapter 6], [70, Chapters 7, 10 and 11] and [47].

Definition 1.37. A discrete-time Markov chain $\mathcal{M}$ with finite state space $\Omega$ is a stochastic process $\left(X_{t}\right)_{t \geq 0}$, with $X_{t} \in \Omega$ for all $t=1,2, \ldots$, that satisfies the Markov property:

$$
\begin{equation*}
\mathbb{P}\left[X_{t+1}=y \mid X_{t}=x_{t}, X_{t-1}=x_{t-1}, \ldots, X_{0}=x_{0}\right]=\mathbb{P}\left[X_{t+1}=y \mid X_{t}=x_{t}\right] ; \tag{1.6}
\end{equation*}
$$

that is, the probability of being in a particular state at the $(t+1)$-th step depends only on the state at the $t$-th step.

In the applications of Markov chains to sampling problems we are only ever interested in time-homogeneous Markov chains, where the probability in (1.6) depends only on the state and not on the time $t$. In this situation, we can define a Markov chain by its transition probability matrix $P$ :

$$
\begin{equation*}
P(x, y)=\mathbb{P}\left[X_{t+1}=y \mid X_{t}=x\right] \quad \forall x, y \in \Omega . \tag{1.7}
\end{equation*}
$$

For $t \geq 0$, we then define the $t$-step probability distribution of $\mathcal{M}$ by the following inductive formula:

$$
P^{t}(x, y)= \begin{cases}I(x, y), & \text { if } t=0 \\ \sum_{z \in \Omega} P^{t-1}(x, z) P(z, y), & \text { otherwise }\end{cases}
$$

where $I(x, y)=1$ if $x=y$ and 0 otherwise. Every Markov chain has a number of stationary distributions which are the distributions $P^{t}(x, \cdot)$ can converge to as $t \rightarrow \infty$.

Definition 1.38. Let $\mathcal{M}$ be a time-homogeneous Markov chain with state space $\Omega$ and transition probability matrix $P$. A distribution $\pi: \Omega \rightarrow[0,1]$ is called a stationary distribution of $\mathcal{M}$ if

$$
\begin{equation*}
\sum_{x \in \Omega} \pi(x) P(x, y)=\pi(y) \quad \forall y \in \Omega \tag{1.8}
\end{equation*}
$$

An arbitrary Markov chain can have any number of stationary distributions, but in the application of Markov chains to designing sampling algorithms we will want chains that always converge to the distribution we are interested in. We now define the ergodic property, which ensures that a chain has a unique stationary distribution.

Definition 1.39. Let $\mathcal{M}$ be a time-homogeneous Markov chain with state space $\Omega$ and transition probability matrix $P . \mathcal{M}$ is said to be ergodic if $\mathcal{M}$ is aperiodic and irreducible:

$$
\begin{align*}
& \mathcal{M} \text { is aperiodic if } \forall x \in \Omega, \operatorname{gcd}\left\{t: P^{t}(x, x)>0\right\}=1  \tag{1.9}\\
& \mathcal{M} \text { is irreducible if } \forall x, y \in \Omega, \exists t \text { such that } P^{t}(x, y)>0 . \tag{1.10}
\end{align*}
$$

Remark 1.40. A sufficient (and easily testable) condition for aperiodicity is:

$$
\begin{equation*}
\forall x \in \Omega, P(x, x)>0 . \tag{1.11}
\end{equation*}
$$

The following well-known [41] theorem shows that ergodic Markov chains always converge to a unique stationary distribution.

Theorem 1.41. Every ergodic Markov chain $\mathfrak{M}$ has a unique stationary distribution $\pi$; moreover, $\mathcal{M}$ converges to $\pi$ in the sense that $P^{t}(x, y) \rightarrow \pi(y)$ for all $x, y \in \Omega$.

We say a Markov chain is time-reversible if it satisfies the detailed balance condition

$$
\begin{equation*}
\forall x, y \in \Omega \quad \pi(x) P(x, y)=\pi(y) P(x, y), . \tag{1.12}
\end{equation*}
$$

If $P$ is symmetric then (1.12) is true if and only if $\pi$ is uniform over $\Omega$.
A very simple example of a Markov chain is a random walk on a graph or directed graph.

Example 1.42. Let $G=(V, E)$ be a graph. The random walk on $G$ is a Markov chain with state space $V$ and probability transition matrix

$$
P(u, v)= \begin{cases}\frac{1}{\operatorname{deg}(u)} & \text { if }\{u, v\} \in E  \tag{1.13}\\ 0 & \text { otherwise }\end{cases}
$$

This is just the random process on $V$ which, at each step, moves to a random neighbour of the current vertex. We observe that this Markov chain is ergodic if and only if $G$ is connected and non-bipartite. Obviously, the graph must be connected for the random walk to be irreducible. On the other hand, if $G$ is bipartite then the random walk
is periodic with period 2: after $t$ steps we are in the same part of the bipartition we started in if and only if $t$ is even. The presence of a single odd cycle ensures that the random walk is aperiodic, since this gives an odd length path by which the chain can return to any vertex on the cycle from itself. The stationary distribution of this chain is not uniform; if we take $\pi$ to denote the stationary distribution of the random walk, we have

$$
\pi(v)=\frac{\operatorname{deg}(v)}{2 m}
$$

where $2 m=\sum_{v \in V} \operatorname{deg}(v)$. It is easy to check that this distribution satisfies (1.8) since, for all $v \in V$,

$$
\sum_{u \in V} \pi(u) P(u, v)=\frac{1}{2 m}|\{u \in V: P(u, v)>0\}|=\frac{\operatorname{deg}(v)}{2 m}=\pi(v) .
$$

We can also define random walks on directed graphs. For a directed graph $G=(V, A)$ we replace (1.13) by

$$
P(u, v)=\left\{\begin{array}{ll}
\frac{1}{\operatorname{outdeg}(u)} & \text { if }(u, v) \in A \\
0 & \text { otherwise }
\end{array} .\right.
$$

In this case, at each step we follow a random outgoing arc of current vertex to reach the next state. However, in this case we do not always have a nice stationary distribution as we do for random walks on undirected graphs.

Given a relation $\mathrm{R} \subset \Sigma^{\star} \times \Sigma^{\star}$ we can construct an almost uniform sampler for R by defining a class of Markov chains such that for each instance we have an ergodic chain $\mathcal{M}$ that converges to the uniform distribution on the set of feasible solutions.

If we run a Markov chain $\mathcal{M}=(\Omega, P)$ for an infinite number of steps, and then take the current state as our output, we know that this will be be sampled from the stationary distribution of $\mathcal{M}$. In practice, we can only run the chain for a finite number steps. A Markov chain Monte Carlo (MCMC) algorithm generates random samples from a set $\Omega$ by starting at some state $x_{0} \in \Omega$ and then running a Markov chain for sufficiently many steps that the probability of any particular $y \in \Omega$ being the current state is (approximately) $\pi(y)$. The quality of the sample improves as a function of the number of steps. However, given an arbitrary Markov chain it is far from clear how long we need to run it for before the distribution is "close enough" to the stationary distribution (i.e., within $\varepsilon$ in total variation distance), and this can vary hugely even across Markov chains defined on the same state space. The time it takes for this convergence to occur is known as the mixing time of the Markov chain, defined below.

In the following, we assume we are dealing with an infinite family of Markov chains defined by the instances of some sampling problem. The properties described are asymptotic properties for all chains in such a class.

Definition 1.43. Let $\mathcal{M}$ be a finite, discrete time Markov chain with state space $\Omega$. For $\varepsilon>0$, the mixing time of $\mathcal{M}$ from initial state $x \in \Omega, \tau_{x}(\varepsilon)$, is defined as

$$
\tau_{x}(\varepsilon)=\min \left\{t:\left\|P^{t}(x, \cdot)-\pi\right\|_{T V} \leq \varepsilon\right\} .
$$

We define the mixing time of $\mathcal{M}, \tau(\varepsilon)$, to be the maximum over the mixing times from each state:

$$
\tau(\varepsilon)=\sup _{x \in \Omega} \tau_{x}(\varepsilon) .
$$

Definition 1.44. A Markov chain $\mathcal{M}$ on state space $\Omega$ is said to be rapidly mixing if, for all for all initial states $x \in \Omega$ and all $\varepsilon>0, \tau_{x}(\varepsilon)$ is bounded above by some function which is polynomial in $|x|$ and $\varepsilon^{-1}$. A Markov chain is said to be torpidly mixing if there exist some $\varepsilon>0$ and $x \in \Omega$ for which $\tau_{x}(\varepsilon)$ is bounded below by some function exponential in $|x|$.

For example, if the state space of the Markov chain is the set of perfect matchings (Example 1.36), then the mixing time must be bounded in terms of the number of vertices and edges of $G$.

Suppose R is a relation defining a self-reducible counting problem. If we have a Markov chain $\mathcal{M}$ which is rapidly mixing on the set of solutions,

$$
\Omega(x)=\left\{y \in \Sigma^{\star}:(x, y) \in R\right\},
$$

for all $x \in \Sigma^{\star}$, then $\mathcal{M}$ is an fpaus for R , and we can use the Algorithm COUNT (Algorithm 2) to construct an fpras for the counting problem $N_{\mathrm{R}}$. So, the problem of finding an fpras can be reduced to the problem of finding a rapidly mixing Markov chain.

There are two principal techniques used in the analysis of the mixing times of Markov chains: coupling and conductance. Coupling is used to show that chains are rapidly mixing, whereas conductance can be used to show that chains are rapidly mixing or torpidly mixing.

### 1.2.1 Coupling

Coupling is a standard technique for proving upper bounds on the mixing time of Markov chains [1, 16].

Definition 1.45. A coupling $\mathcal{C}$ of a Markov chain $\mathscr{M}$ is a stochastic process $\left(X_{t}, Y_{t}\right)_{t \geq 0}$ on $\Omega \times \Omega$ such that each of the marginal distributions, $\left(X_{t}\right)_{t \geq 0}$ and $\left(Y_{t}\right)_{t \geq 0}$, is a faithful copy of the original Markov chain.

One approach to bounding the mixing time of a Markov chain is to use the coupling inequality of Aldous [1].

Lemma 1.46 ([1]). Let $\mathcal{M}=(\Omega, P)$ be a time-homogeneous, finite state Markov chain. For every state $x \in \Omega$, the variation distance between $\pi$ and the $t$-step distribution at time $t, P^{t}(x, \cdot)$, is bounded above by the probability of any coupling coalescing by time $t$, i.e.,

$$
\begin{equation*}
\max _{x \in \Omega}\left\|P^{t}-\pi\right\|_{T V} \leq \sup _{X_{0}, Y_{0}} \mathbb{P}\left[X_{t} \neq Y_{t}\right] . \tag{1.14}
\end{equation*}
$$

Thus, in order to obtain a polynomial bound on the mixing time it suffices to construct a coupling which will have coalesced (with high probability) after a polynomial number of steps. In general, couplings can be difficult to describe and analyse, so it is common to restrict attention to Markovian couplings.

Definition 1.47. A coupling $\mathcal{C}$ of a Markov chain $\mathcal{M}=(\Omega, P)$ is said to be Markovian if $\mathcal{C}$ is itself a Markov chain on $\Omega \times \Omega$. That is, the distribution of the pair of states $\left(X_{t+1}, Y_{t+1}\right)$ depends only on the values of $\left(X_{t}, Y_{t}\right)$.

The advantage of Markovian couplings over general couplings is that they allow us to analyse the behaviour of the coupling by considering only single steps at a time. We now give an illustrative example of a Markovian coupling.

Example 1.48. Let $G=(V, E)$ be a graph. A $k$-colouring of $G$ is a mapping

$$
\sigma: V \rightarrow\{1,2, \ldots, k\} .
$$

We say $\sigma$ is a valid $k$-colouring if $\sigma(u) \neq \sigma(v)$ for every $\{u, v\} \in E$. Let $\Omega$ be the set of all valid $k$-colourings of $G$. There is a very simple Markov chain $\mathcal{M}$ on $\Omega$. Let $X_{t} \in \Omega$ denote the current state of the chain. We now define how one step of the Markov chain obtains $X_{t+1}$ from $X_{t}$.

## One step of the chain $\mathcal{M}$

1. Choose $v \in V$ and $c \in\{1,2, \ldots, k\}$ uniformly at random;
2. Let $X^{\prime}$ denote the colouring in which $X^{\prime}(u)=X_{t}(u)$ for all $u \neq v$ and $X^{\prime}(v)=c$;
3. If $X^{\prime} \in \Omega$ then set $X_{t+1}=X^{\prime}$. Otherwise, set $X_{t+1}=X_{t}$.

This chain is ergodic when the number of colours $k$ is greater than the minimum degree of $G$ plus 2, see, e.g., [70, §11.5], and converges to the uniform distribution on $\Omega$. The simplest coupling of $\mathcal{M}$ is the one which chooses the same vertex $v$ and colour $c$ for both chains. If we can re-colour $v$ to $c$ in both chains, then we have increased the number of vertices that the two copies of the chain agree on. However, if we can recolour $v$ in one copy of the chain but not the other, and $v$ has the same colour in the current state of both chains, then the number of disagreements will increase,

It has been shown that the simple coupling described above can be used to prove rapid mixing of $\mathscr{M}$ whenever we have $k \geq 2 \Delta+1$, where $\Delta$ is the maximum degree of $G$ [46]. It is known that the Markov chain $\mathcal{M}$ is rapidly mixing for graphs with fewer colours but these results require more sophisticated couplings, some of which are non-Markovian. See [34] for a survey of these results.

We will give more details of coupling and a technique called path coupling [16, 13] when we analyse the mixing time of a particular Markov chain in Chapter 2.

### 1.2.2 Conductance

The coupling technique can only be used to show that a Markov chain is rapidly mixing and, in the case of Markovian couplings at least, cannot always be applied [58]. An alternative approach is to estimate the conductance of the chain.

Definition 1.49. Let $\mathcal{M}=(\Omega, P)$ be a discrete state, time-homogeneous, Markov chain with stationary distribution $\pi$. For any non-empty set $S \subset \Omega$ the conductance $\Phi(S)$ of $S$ is defined to be

$$
\Phi(S)=\frac{\sum_{x \in S, y \in \Omega \backslash S} \pi(x) P(x, y)}{\pi(S)} .
$$

We define the conductance of $\mathscr{M}$ as being the minimum over all sets:

$$
\Phi(\mathcal{M})=\min _{\substack{S \subset \Omega \\ 0<\pi(S) \leq 1 / 2}} \Phi(S) .
$$

In 1989, Sinclair and Jerrum [79] showed how conductance could be used to analyse the mixing times of Markov chains by exhibiting the following close relationship between the conductance and mixing time of a Markov chain.

Theorem 1.50 ([79]). Let $\mathcal{M}=(\Omega, P)$ be a time-homogeneous, reversible Markov chain and suppose $P(x, x) \geq 1 / 2$ for all $x \in \Omega$. Then, the mixing time, $\tau(\varepsilon)$, and conductance, $\Phi(\mathcal{M})$, satisfy

1. For all $x \in \Omega, \tau_{x}(\varepsilon) \leq \frac{2}{\Phi(\mathcal{M})^{2}}\left(\log \left(\varepsilon^{-1}\right)+\log \left(\pi(x)^{-1}\right)\right)$;
2. $\max _{x \in \Omega} \tau_{x}(\varepsilon) \geq \frac{1-2 \Phi(\mathcal{M})}{2 \Phi(\mathcal{M})} \log \left(\varepsilon^{-1}\right)$.

Thus, if one can show that the conductance of $\mathcal{M}$ is "large", i.e., bounded below by an inverse polynomial in the size of the elements of $\Omega$, then one can deduce that $\mathcal{M}$ is rapidly mixing. On the other hand, if the conductance of $\mathcal{M}$ is exponentially small then $\mathcal{M}$ is torpidly mixing. Moreover, to show torpid mixing we only need to show a bound on the mixing time for some $\varepsilon>0$. Defining the boundary of a set $S$ as

$$
\partial S=\{x \in S: P(x, y)>0 \text { for some } y \in \Omega \backslash S\}
$$

we get

$$
\begin{equation*}
\Phi(S) \leq \pi(\partial S) / \pi(S) \tag{1.15}
\end{equation*}
$$

Hence, to show that a chain is torpidly mixing it suffices to find a set for which the right-hand side of (1.15) is bounded above by some exponentially small function. This is encapsulated in the following theorem, taken from [63]:

Theorem 1.51. If, for some $S \subset \Omega$ satisfying $0<|S| \leq|\Omega| / 2$, the ratio $\pi(\partial S) / \pi(S)$ is exponentially small in the size of the elements of $\Omega$, then the Markov chain is torpidly mixing.

Conductance can also be used to find upper bounds on the mixing time, particularly using a technique known as canonical paths, due to Jerrum and Sinclair [49]. This technique has been successfully applied to analysing many Markov chains, e.g., [22, 51], but will not be used in this thesis. However, in Chapter 2 we will use a technique known as comparison, which is also based on conductance, to analyse the mixing time of a Markov chain (see §2.6.2 for details).

### 1.2.3 Dart throwing

An alternative approach that is sometimes applicable is dart throwing. Suppose we have a set $S$ whose elements we want to count or sample, but do not know of a polynomial time algorithm for doing so. Now suppose that we have a polynomial time algorithm for testing $x \in S$ and that there exists a set $S^{\prime} \supseteq S$ for which we have an fpaus and which satisfies

$$
\begin{equation*}
\frac{|S|}{\left|S^{\prime}\right|} \geq \frac{1}{p(n)} \tag{1.16}
\end{equation*}
$$

where $n$ is the size of the elements of $S$ and $p(\cdot)$ is a polynomial. Then, using our fpaus for $S^{\prime}$, we generate elements of $S^{\prime}$ until we obtain an element of $S$. Each element of $S^{\prime}$ is generated in polynomial time and (1.16) means we will only have to generate a polynomial number of elements of $S^{\prime}$ before we get an element of $S$. Moreover, we can approximate

$$
\begin{equation*}
\frac{|S|}{\left|S^{\prime}\right|} \tag{1.17}
\end{equation*}
$$

by generating a polynomial number of samples from $S^{\prime}$ and then taking the proportion of them which lie in $S$ as an estimation of (1.17). Then, if we can also calculate or approximate $\left|S^{\prime}\right|$, we can use this to obtain an approximation of $|S|$. The fpaus for $S^{\prime}$ could be a rapidly mixing Markov chain, e.g., [36], or $S^{\prime}$ could be some simple set that we can sample from without recourse to MCMC, e.g., [27].

In Chapter 3, we describe how this approach can be used to sample Euler tours of a graph or directed graph. The algorithms resulting from this are shown to run in expected polynomial time for almost every $d$-in/d-out graph and we conjecture that they run in expected polynomial time for almost every $2 d$-regular graph. These results are proven using random graphs.

### 1.3 Eulerian orientations and Euler tours

Recall that we denote the set of Eulerian orientations of a graph $G$ by $\operatorname{EO}(G)$ and the set of Euler tours by $\mathrm{ET}(G)$. It is a classical graph-theoretic result that the Eulerian graphs are the only graphs that admit Eulerian orientations and that Eulerian graphs and directed graphs are the only graphs and directed graphs to admit Euler tours. In this thesis, we are interested in the complexity of the following two problems:

| Name. | \#EO. |
| :--- | :--- |
| Instance. | An undirected graph $G$. |
| Output. | The number of Eulerian orientations of $G$. |

and

| Name. | \#ET. |
| :--- | :--- |
| Instance. | An undirected graph $G$. |
| Output. | The number of Euler tours of $G$. |

We are also interested in these problems under various restrictions, specifically:

| Name. | \#PlanarEO. |
| :--- | :--- |
| Instance. | A planar graph $G$. |
| Output. | The number of Eulerian orientations of $G$. |

and

| Name. | \#DirectedET. |
| :--- | :--- |
| Instance. | A directed graph $\vec{G}$. |
| Output. | The number of Euler tours of $\vec{G}$. |

Recall the conditions for $\mathrm{EO}(G) \mathrm{ET}(G)$ to be non-empty (Remarks 1.20 and 1.24). Hence, to solve the decision problem $\operatorname{EO}(G) \neq \emptyset$ all we need to do is check that each vertex satisfies the Eulerian condition (Definition 1.18). To solve the decision problem $\operatorname{ET}(G) \neq \emptyset$ we need to check that each vertex satisfies the Eulerian condition and that $G$ is connected. Checking that each vertex has even degree and that a graph is connected can both be achieved in polynomial time, so both decision problems are in P. However, both \#EO and \#ET are \#P-complete. Exact counting of Eulerian orientations was shown to be \#P-complete by Mihail and Winkler in 1992 [68, 69]. In [68, 69], Mihail and Winkler also described a method to construct an fpras for \#EO that works for every undirected graph. In the case of Euler tours, the complexity of counting Euler tours remained open until 2005, when it was shown to be \#P-complete by Brightwell and Winkler [15]. More recently, Ge and Stefankovic [37] have shown that counting Euler tours is \#P-complete for 4-regular planar graphs. The status of approximately counting or sampling the Euler tours of an undirected graph is still unresolved, even for simple classes, e.g., 4-regular planar graphs. Indeed, the only positive result to date is the recent proof of Chebolu et al. [17] that we can exactly count the Euler tours of series-parallel graphs in polynomial time. In the following two sections, we present a summary of what is known about these two problems.

### 1.3.1 Eulerian orientations

The number of Eulerian orientations of an undirected graph is a subject that has received attention in statistical physics [61, 6, 29], combinatorics [78, 95, 65, 33], and theoretical computer science [68, 69, 62, 38]. The problem of sampling/counting Eulerian orientations in a general Eulerian graph is almost entirely understood: the complexity of exact counting resolved by Mihail and Winkler in 1992 [68, 69]; Mihail and Winkler (also in $[68,69]$ ) provided a reduction for constructing an fpaus or an fpras for any Eulerian graph, which we will describe in a moment.

The Eulerian orientations of certain planar lattices have special significance in statistical physics as the configurations of so-called "ice models" [61, 6, 7, 29], motivating studying the complexity of counting/sampling Eulerian orientations restricted to planar graphs or planar lattices. Lieb [61] and Baxter [6] have derived asymptotic expressions for the number of Eulerian orientations of the square and triangular lattice.

Definition 1.52. The infinite square lattice is the 4-regular infinite graph with vertex set $\{(i, j): i, j \in \mathbb{Z}\}$ and such that for every $i, j \in \mathbb{Z}$ there is an edge joining $(i, j)$ to $(i, j \pm 1)$ and $(i \pm 1, j)$.

Definition 1.53. The infinite triangular lattice is the 6 -regular infinite graph with vertex set $\{(i, j): i, j \in \mathbb{Z}\}$ and such that for every $i, j \in \mathbb{Z}$ there is an edge joining $(i, j)$ to $(i \pm 1, j),(i, j \pm 1),(i-1, j+1)$, and $(i+1, j-1)$.

The ice models defined on two further Eulerian lattices are studied in [29] but we will not be considering them in this thesis. The lattices defined in Definition 1.52 and 1.53 are both infinite, so finite sub-lattices are used in order to be able to perform computational analysis; boundary conditions are used to simulate the influence of the infinite lattice.

Definition 1.54. A solid subgraph of an infinite lattice is subgraph which can be obtained by taking some finite cycle and everything lying on its interior.

Suppose, for example, we are taking the subgraph of the square lattice induced by the set

$$
\{(i, j): 0 \leq i, j<n\},
$$

which we denote by $G(n, n)$.
Each of the $(n-2)^{2}$ internal vertices of $G(n, n)$ (those in which $i, j \notin\{0, n-1\}$ ) has degree 4 and we enforce the Eulerian condition on each of these. The boundary vertices (those with at least one of $i$ and $j$ equal to 0 or $n-1$ ) have one of the following boundary conditions applied to them.

Fixed Suppose we have fixed an orientation for the rest of the infinite lattice, which extends to an Eulerian orientation of $G(n, n)$. Then, sampling an Eulerian orientation consistent with the fixed boundary condition is equivalent to sampling an $\alpha$-orientation of $G(n, n)$ for a specific $\alpha: V \rightarrow \mathbb{Z}$, which we will now define. For each $(i, j)$ with $0<i, j<n-1$ we set $\alpha\left(v_{i, j}\right)=0$. Each of the


Figure 1.2: Square and triangular ice
non-corner boundary vertices has $\alpha(v)= \pm 1$ and each of the corner vertices has $\alpha(v) \in\{-2,0,2\}$ such that

$$
\sum_{i=0}^{n-1} \alpha\left(v_{0, i}\right)+\alpha\left(v_{n-1, i}\right)+\alpha\left(v_{i, 0}\right)+\alpha\left(v_{i, n-1}\right)=0 .
$$

Clearly, any $\alpha$-orientation of $G(n, n)$ can be extended to an Eulerian orientation of the infinite lattice. Typically, a fixed boundary condition will be of a regular form similar to that given by the dotted arrows in Figure 1.2;

Free Enforce the Eulerian condition on all even degree vertices and insist that

$$
|\operatorname{indeg}(v)-\operatorname{outdeg}(v)|=1
$$

for all vertices $v$ of odd degree. The Eulerian orientations with free boundary conditions could be viewed as a generalisation of $\alpha$-orientations. Here we are giving a set, $\alpha(v)=\{-1,+1\}$, of allowed values for indeg $(v)-\operatorname{outdeg}(v)$;

Torus Assume that the graph is a torus, so the edge leaving a boundary vertex wraps around and attaches to another boundary vertex. In the case of the grid we have edges connecting $(0, j)$ to $(n-1, j)$ and connecting $(j, 0)$ to $(n-1,0)$.

The ice models defined on the square lattice and the triangular lattice are considered solved $[61,6,7]$, in the sense that an asymptotic expression is known for the number of Eulerian orientations of the subgraphs induced by the set $\{(i, j): 0 \leq i, j<n\}$ for a toroidal boundary condition. Specifically, in each case the authors obtain the exponential growth rate of the number of Eulerian orientations of an $n \times n$ section of the lattice with toroidal boundary condition. This is achieved using the transfer matrix method [61, 6, 7, 39].

We now sketch the transfer matrix method, as applied to counting Eulerian orientations of the square grid; this work is originally due to Lieb [61]. Let $\mathcal{C}$ denote a configuration of the arcs in the $k$-th column of an Eulerian orientation of the $n \times n$ grid; that is, an assignment of $u p$ or down to each arc. Each configuration can be defined by choosing the set of down arcs, so there are $2^{n}$ possible configurations. For each configuration $\mathcal{C}^{\prime}$ of arcs for the $(k+1)$-th column, let $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right)$ denote the number of ways to configure the horizontal arcs joining both columns so that the Eulerian condition can be satisfied at each vertex. Note that this is dependent only on $\mathcal{C}$ and $\mathcal{C}^{\prime}$ so $A\left(C, \mathcal{C}^{\prime}\right)$ depends only on $n$ and not $k$. Then, the number of Eulerian orientations of the $n \times n$ grid with toroidal boundary conditions can be written as

$$
\sum_{C_{0}} \sum_{C_{1}} \cdots \sum_{C_{n-1}} \prod_{i=0}^{n} A\left(\mathcal{C}_{i}, \mathcal{C}_{i+1 \bmod n}\right)=\operatorname{Trace}\left(A^{n}\right) .
$$

In [61] Lieb analyses the structure of $A$ to find the maximum eigenvalue of $A$, which can be used to estimate the asymptotic growth rate of $\operatorname{Trace}\left(A^{n}\right)$. In [6] Baxter performs a similar, though more complicated, analysis to obtain a corresponding result for the triangular lattice.

Remark 1.55. The transfer matrix method is a general schema for counting structures defined on grids and similar recursively definable graphs, see [39] for a general description. When applicable, it yields an algorithm for exact counting. However, these algorithms are often inefficient as they require the multiplication of large matrices. This approach only becomes practical when some useful structure exists that allows us to either (a) improve the running time of the algorithm or (b) obtain an asymptotic estimate for the quantity in question. In §3.6, we discuss how the transfer matrix method can be applied to counting Euler tours of the grid.

The research of Lieb and Baxter was carried out in the 1960s. Despite their analyses of the asymptotic growth rate of the EOs of these lattices, the question of whether there existed efficient algorithms for counting (exactly or approximately) was still of interest. The estimates of [61] and [6] only capture the exponential growth rate and so are not useful for the construction of an fpaus for sampling Eulerian orientations either, so the question of whether one could sample Eulerian orientations of these lattices in polynomial time remained open until the result of Mihail and Winkler [68, 69] in 1992. Being able to generate random configurations of models is important in statistical physics as physicists often want to compute average properties of configurations; hence, this was an important open problem in the area of sampling algorithms.

In a related line of research, McKay [65] has obtained asymptotic estimates for the total number of Eulerian directed graphs and Eulerian orientations on $n$ vertices, as well as the number of Eulerian directed graphs and Eulerian orientations on $n$ vertices with exactly $m$ edges ${ }^{2}$. The results of $\S 3.4$ of this thesis can be viewed as complementing those of [65] by providing an asymptotic estimate of the number of Eulerian orientations of random $2 d$-regular graphs, when $d$ is a constant. In fact, we provide a stronger result: an asymptotic expression for the distribution of the number of Eulerian orientations of a random $2 d$-regular graph. We are also able to show that almost every $2 d$-regular graph has few Eulerian orientations, where few is taken to mean within a polynomial factor of Schrijver's lower bound [78] (see Theorem 1.56 below).

The flavour of the calculations in [65] is similar to that of §3.4, in the sense that both obtain the asymptotic estimate by the so-called saddle-point method [25]. However, the actual approaches are quire different. McKay defines a generating function whose constant term is equal to the required quantity and then applies Cauchy's theorem to obtain an integral expression for this value. In §3.4 we take a more direct combinatorial approach, using the configuration model for graphs with fixed degree sequence. The approach of McKay seems to be best suited to the situation where we are enumerating all objects in a set, without regard to the vertex degrees. Although a similar approach has been applied to estimating the number of graphs with fixed degree sequence [67], we do not know of any result applying this approach to counting Eulerian orientations or Eulerian directed graphs with fixed degree sequences. Moreover, McKay's method is not suited to the extra analysis required to obtain the concentration result mentioned above. However, one distinct advantage of the approach used in [65] and [67] is that the results hold for larger degrees. Results using the configuration model only hold when if the degrees of the vertices are very small compared to $n$. Although beyond the scope of this thesis, it would be interesting to see if McKay's approach could be extended to obtain similar results (to those of §3.4) for graphs with larger degrees.

Several authors have derived upper and lower bounds on the number of Eulerian orientations of certain classes of Eulerian graphs. The number of Eulerian orientations of a regular graph has been studied by Schrijver [78] and Las Vergnas [95]. Schrijver proved the following theorem.

Theorem 1.56 (Schrijver [78]). Let d be some fixed integer and let $G$ be a $2 d$-regular

[^1]graph on $n$ vertices. Then,
$$
\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} \leq|\mathrm{EO}(G)| \leq\binom{ 2 d}{d}^{n / 2}
$$

Las Vergnas [95] was able to generalise the bound of Theorem 1.56 by considering also the girth and the number of edge-disjoint cycles in the graphs.

For the planar case, Zickfeld and Felsner [33] have, in the more general context of $\alpha$-orientations, obtained upper and lower bounds on the number of Eulerian orientations of several classes of planar graphs. In particular, they showed that the number of Eulerian orientations of any planar Eulerian graph $G$ is never less than the number of $\alpha$-orientations of $G$, for any $\alpha$.

The number of Eulerian orientations of a graph $G$ is closely related to the number of perfect matchings of a specially constructed bipartite graph. For any Eulerian graph $G=(V, E)$ we can construct a bipartite graph $H_{G}$ such that

$$
\begin{equation*}
\prod_{v \in V}(\operatorname{deg}(v) / 2)!|\mathrm{EO}(G)|=\left|P M\left(H_{G}\right)\right|, \tag{1.18}
\end{equation*}
$$

where $\mathrm{EO}(G)$ denotes the set of Eulerian orientations of $G$ and $P M\left(H_{G}\right)$ denotes the set of perfect matchings of $H_{G}$. This reduction was first made explicit by Mihail and Winkler $[68,69]$ but is implicit in the earlier work of Schrijver [78].

We now describe the reduction and explain how it can be used to construct an fpras or an fpaus for $\mathrm{EO}(G)$, despite the fact that Eulerian orientations are not self-reducible. Let $G=(V, E)$ be an Eulerian graph. Let $H_{G}=(A, B, F)$ be the bipartite graph with bipartition

$$
A=\bigcup_{v \in V} X_{v}, \quad X_{v}=\left\{x_{v, e}: e=\{u, v\} \in E\right\}
$$

and

$$
B=\left\{w_{e}: e \in E\right\} \cup \bigcup_{v \in V} Y_{v}, \quad Y_{v}=\left\{y_{v, i}: 1 \leq i \leq \operatorname{deg}(v) / 2\right\} .
$$

and edge set

$$
F=\left\{\left(x_{u, e}, w_{e}\right),\left(x_{v, e}, w_{e}\right): e=\{u, v\} \in E\right\} \cup \bigcup_{v \in V} X_{v} \times Y_{v} .
$$

Suppose we have an Eulerian orientation $\mathcal{E} \in \operatorname{EO}(G)$. We construct a perfect matching $M \in P M\left(H_{G}\right)$ as follows: for each edge $e=\{u, v\} \in E$ we include $\left(x_{v, e}, w_{e}\right)$ in $M$ if $e$ is oriented $(u, v)$ in $\mathcal{E}$ and include $\left(x_{u, e}, w_{e}\right)$ in $M$ if $e$ is oriented $(v, u)$ in $\mathcal{E}$; complete $M$ by, for each $v \in V$, choosing any perfect matching on $Y_{v}$ and the unmatched vertices
in $X_{v}$. Every edge in $G$ receives an orientation in $\mathcal{E}$, so every vertex in $\left\{w_{e}: e \in E\right\}$ is matched with a vertex in $\bigcup_{v \in V} X_{v}$ in the first step. This uses up half of the vertices in $\bigcup_{v \in V} X_{v}$. The remaining vertices are all matched with $Y_{v}$ in the second step, so the construction always yields a perfect matching. Moreover, each Eulerian orientation gives rise to exactly

$$
\prod_{v \in V}(\operatorname{deg}(v) / 2)!
$$

perfect matchings; this is the number of ways to choose the edges in the second step.
Now, suppose we have a perfect matching $M \in P M\left(H_{G}\right)$. We construct an Eulerian orientation $\mathcal{E}$ of $G$ as follows: for each edge $e=\{u, v\} \in E$ we orient $e$ towards $u$ if $\left(x_{u, e}, w_{e}\right) \in M$ and orient $e$ towards $v$ if $\left(x_{v, e}, w_{e}\right) \in M$. Exactly half of the vertices in $X_{v}$ must be matched to vertices in $Y_{v}$ in any perfect matching. Thus, exactly half of the edges incident with $v$ are oriented towards $v$ in this construction. Moreover, the perfect matching $M$ exactly determines $\mathcal{E} \in \operatorname{EO}(G)$, so there is exactly one Eulerian orientation of $G$ corresponding to each perfect matching of $H_{G}$. The ratio (1.18) follows from the fact that the procedure just described is exactly the reversal of the mapping from Eulerian orientation to perfect matching described in the previous paragraph.

The relationship between $\operatorname{EO}(G)$ and $P M(G)$ has two important consequences: the existence of an fpaus and an fpras for $\mathrm{EO}(G)$. Given a uniformly random perfect matching $M \in P M\left(H_{G}\right)$, applying the reduction described above gives a uniformly random Eulerian orientation $\mathcal{E} \in \mathrm{EO}(G)$, since each $\mathcal{E} \in \mathrm{EO}(G)$ arises as a result of an equal number of perfect matchings. Thus, any fpaus for $P M\left(H_{G}\right)$ can be turned into an fpaus for $\mathrm{EO}(G)$. The celebrated Jerrum-Sinclair chain [49] (see [51, 10] for improvements) can be used to sample a perfect matching of any bipartite graph in time $O\left(n^{7}(\log n)^{4}\right)$.

Similarly, an fpras for counting perfect matchings of bipartite graphs can be used to construct an fpras for the number Eulerian orientations of any graph $G$ : multiplying an $\varepsilon$-approximation for $P M\left(H_{G}\right)$ by

$$
\prod_{v \in V}(\operatorname{deg}(v) / 2)!
$$

gives an $\varepsilon$-approximation for $\operatorname{EO}(G)$. In [51], Jerrum et al. describe how to use the Jerrum-Sinclair chain to construct an fpras for the number of perfect matchings of any bipartite graph that runs in polynomial-time. Thus, by using the construction described in [69], we can use the results of [51] (with an improvement to the running time from
[10]) to show the existence of an fpaus and fpras for the Eulerian orientations of any Eulerian graph.

Several authors [62, 38, 29] have considered other sampling schemes for the Eulerian orientations of the Eulerian planar lattices, due to their connection to the ice models of statistical physics. Eloranta [29] has studied dynamics for generating random configurations of four different ice models, including the two described above (see Figure 1.2). Although called probabilistic cellular automata in [29], these dynamics are nothing more than the face reversing Markov chain studied in [62, 38] and in Chapter 2 of this thesis.

Luby et al. [62] and Goldberg et al. [38] have shown that the face reversing Markov chain mixes rapidly on the Eulerian orientations of the square lattice with fixed and free boundary conditions, respectively. In this thesis, we extend current knowledge by showing that counting Eulerian orientations remains \#P-complete for planar graphs and show that the Markov chain studied in [62] and [38] can be used to construct an fpras or fpaus for the Eulerian orientations of the triangular lattice. Specifically, in Chapter 2 of this thesis we will show that this chain is rapidly mixing on the triangular lattice under any fixed boundary condition. However, we also show that there exists an infinite family of planar graphs for which this chain is torpidly mixing.

### 1.3.2 Euler tours

Euler tours are a classical graph theoretical structure, studied in every undergraduate graph theory class. Indeed, the result believed to be the first theorem in graph theory is the one that introduced Euler tours: Euler's solution to the Königsberg bridge problem [30]. It was in order to solve this problem that Euler defined the concept of a graph and stated what is now known as the Eulerian condition (Definition 1.18). The complexity of counting/sampling Euler tours is quite different depending on whether the input is a directed graph or an undirected graph. For any Eulerian directed graph $\vec{G}$ it is possible to exactly count the number of Euler tours of $\vec{G}$ in polynomial time [94, 47] and there exist many polynomial time sampling algorithms, e.g., [19, 72]. In the undirected case the counting problem is \#P-complete in general and, apart from the special case of series-parallel graphs [17] there are no known polynomial cases. Moreover, apart from the cases of the complete graph [66] and series-parallel graphs [17] there are no known polynomial time sampling algorithms.

### 1.3.2.1 Euler tours of directed graphs

In the case of Eulerian directed graphs, it is possible to compute \#ET $(\vec{G})$ in polynomial time using a relationship between the Euler tours of $\vec{G}$ and the arborescences of $\vec{G}$ (see Definition 1.12).

For any Eulerian directed graph $\vec{G}=(V, A)$, the so-called BEST theorem (due to de Bruijn and van Aardenne-Ehrenfest [94], extending a result of Smith and Tutte [80]) reduces the problem of computing $|\operatorname{ET}(\vec{G})|$ to the problem of computing $|\operatorname{ARB}(\vec{G}, v)|$ for any vertex $v \in V$.

Theorem 1.57 ([94]). Let $\vec{G}=(V, A)$ be an Eulerian directed graph and let $d_{u}=$ outdeg $(u)$ for each $u \in V$. For any $v \in V$, we have

$$
\begin{equation*}
|E T(\vec{G})|=\left[\prod_{u \in V}\left(d_{u}-1\right)!\right]|\operatorname{ARB}(\vec{G}, v)| . \tag{1.19}
\end{equation*}
$$

We now sketch the proof of Theorem 1.57. Let $\vec{G}=(V, A)$ be an Eulerian directed graph and let $v \in V$. Suppose we have an Euler tour $\mathcal{T}=\left(a_{0}, a_{1}, \ldots, a_{m-1}\right)$ of $\vec{G}$ and further suppose that $v$ is the source of $a_{0}$. Given $T$ we can construct a unique $\mathcal{A} \in \operatorname{ARB}(\vec{G}, v):$ for each vertex $u \neq v$ we take the last arc leaving $u$ on $\mathcal{T}$ to be the unique arc leaving $u$ in $\mathcal{A}$. To see that this construction always gives an arborescence, suppose we have a cycle $C=\left(v_{0}, v_{1}, \ldots, v_{k-1}\right)$ such that $\left(v_{i}, v_{(i+1) \bmod k}\right)$ is the last arc leaving $v_{i}$ on $\mathcal{T}$ for each $i=0,1, \ldots, k-1$. Consider how $\mathcal{T}$ behaves on $\vec{G}$. Each time $\mathcal{T}$ reaches a vertex $v_{i}$ on $C$ the next arc on $\mathcal{T}$ is from $\vec{G}-C$, unless we have reached the last occurrence of $v_{i}$ on $\mathcal{T}$. But this cannot occur until $\mathcal{T}$ has passed all paths from $v_{i}$ to $v$ in $\vec{G}-C$, in which case $\mathcal{T}$ would define an Euler tour on $\vec{G}-C$, contradicting the assumption that $\mathcal{T}$ is an Euler tour of $G$. Thus, choosing $a_{0}$ from the $d_{v}$ arcs leaving $v$ in $\vec{G}$ is sufficient to induce a unique arborescence $\mathcal{A} \in \operatorname{ARB}(\vec{G}, v)$.

Next, suppose we have an arborescence $\mathcal{A}$ of $\vec{G}$ with root $v$. Moreover, suppose we have an ordering on the arcs leaving each vertex of $\vec{G}$ such that for each $u \neq v$ the unique arc leaving $u$ in $\mathscr{A}$ is the last arc in the ordering. For each such ordering we can construct a unique Euler tour of $\vec{G}$ as follows. Starting at $v$ we choose the least-most arc in the ordering of the out-arcs of $v$. Then, at each subsequent vertex we choose the least arc which has not yet been used. This process terminates when we reach $v$ and there are no more unused arcs left to choose, and at this point we have constructed an Euler tour of $\vec{G}$.

Thus, there are $d_{v}$ arborescences with root $v$ associated to each Euler tour, one for
each arc leaving $v$, and

$$
\left(d_{v}\right)!\prod_{u \neq v}\left(d_{u}-1\right)!
$$

Euler tours associated to each arborescence with root $v$. Taken together, these two facts imply (1.19).

One immediate corollary to Theorem 1.57 is that there are an equal number of arborescences rooted at each vertex of any Eulerian directed graph. However, more importantly, Tutte [90] showed that we can calculate the number of arborescences rooted at a vertex of any directed graph by computing the determinant of minor of a specially constructed matrix.

Definition 1.58. Let $M$ be a $n \times n$ matrix. For any row/column index $i$, we denote by $M_{i i}$ the matrix obtained by removing the row and column indexed by $i$. This is known as the $i$-th minor of $M$.

Theorem 1.59 (Tutte [90]). Let $\vec{G}=(V, A)$ be a directed graph. We define the Laplacian of $\vec{G}$ to be the matrix $L$ with entries given by

$$
L_{u, v}= \begin{cases}\text { outdeg }(u) & \text { if } u=v \\ -d_{u, v} & \text { if } u \neq v \text { and there are } d_{u, v} \text { arcs from } u \text { to } v \text { in } \vec{G} .\end{cases}
$$

For any vertex $v \in V$, the number of arborescences rooted at $v$ is equal to the determinant of the $v$-th minor of $L$.

We can calculate the determinant of any $n \times n$ matrix in time polynomial in $n$ by, e.g., Gaussian elimination [84]; hence, the above theorem tells us that we can count the Euler tours of any Eulerian directed graph in time polynomial in the number of vertices.

Remark 1.60. For the special case of 2-in/2-out directed graphs, Macris \& Pulé [64] claimed that one could obtain the number of Euler tours by computing the determinant of an alternative matrix; Lauri [59] claimed to have a combinatorial proof of this result. However, the result in [64] is incorrect. Indeed, we can construct an example with 5 vertices for which this method gives the wrong value. This counter-example is contained in Appendix A.

Remark 1.61. The importance of the BEST theorem goes beyond the fact that it allows us to count Euler tours of directed graphs in polynomial time. Indeed, almost every result about counting or sampling Euler tours of any graph (directed or undirected) depends on the relationship implied by this theorem, e.g., [19, 66, 72, 15].

There exist several schemes for generating random Euler tours of an Eulerian directed graph $\vec{G}$, all of them based on the relationship between tours and arborescences described above. These can be grouped into two types: those based on the fact that we can count the number of arborescences in any directed graph in polynomial time [42, 18, 57, 19]; and those based on a random walk in $\vec{G}[53,99,72]$. In the next two paragraphs we will briefly describe algorithms of both types.

In the algorithms based on Theorem 1.59, the arborescence is constructed one arc at a time. At each step the algorithm considers a vertex $u$ which has not yet been assigned its unique out-arc in $\mathcal{A}$. Let $\operatorname{Out}(u)$ denote the set of arcs leaving $u$ in $\vec{G}$ and suppose we want to compute the probability $a=(u, w) \in \operatorname{Out}(u)$ is contained in a uniformly random arborescence of $\vec{G}$. Let $\vec{G}_{a}$ denote the directed graph in which we have replaced $u$ and $w$ by a single vertex $u w$ which has the same outgoing arcs as $w$ in $\vec{G}$ and has the incoming arcs of both $u$ and $w$ in $\vec{G}$. The arborescences in $\operatorname{ARB}(\vec{G}, v)$ containing $(u, w)$ are in one-to-one correspondence with the arborescences in $\operatorname{ARB}\left(\vec{G}_{a}, v\right)$. Hence, the probability $(u, w)$ is contained in a uniformly random element of $\operatorname{ARB}(\vec{G}, v)$ is

$$
\begin{equation*}
\frac{\left|\operatorname{ARB}\left(\vec{G}_{a}, v\right)\right|}{|\operatorname{ARB}(\vec{G}, v)|} . \tag{1.20}
\end{equation*}
$$

By Theorem 1.59, we can compute (1.20) in polynomial time. We need to choose $n-1$ arcs, so the total computational cost of choosing a random arborescence is a polynomial in $n$. Implemented naively, this algorithm will have running time $O\left(n^{3}\right)$ but some clever tricks used in [19] reduce this to the amount of time taken to multiply $n \times n$ matrices (currently $\left.O\left(n^{2.376}\right)[21]^{3}\right)$.

Remark 1.62. The astute reader will notice that the algorithm just described is no more than an instantiation of the general scheme described in Algorithm 1.

The second, and perhaps better-known, class of algorithms are based on random walks [72]. The most efficient of these algorithms is believed to be the Propp-Wilson algorithm [72], which is based on the concept of loop-erased random walks. The running times of these algorithms are in terms of properties of the random walk, e.g., cover time or hitting time (defined below), so a direct comparison of running times is difficult. For some cases, the running time of a random walk based algorithm could be exponential, in which case algorithms based on Theorem 1.57 could be more

[^2]efficient. However, it is believed that for most graphs, the Propp-Wilson algorithm will outperform all known algorithms based on Theorem 1.57 [72].

Definition 1.63. Let $\vec{G}=(V, A)$ be a directed graph and let $u, v \in V$. A random walk $\gamma$ from $u$ to $v$ is obtained by setting $\gamma_{1}=u$ and then setting $\gamma_{i+1}$ to be the target of a randomly chosen arc leaving $\gamma_{i}$ until we reach $\gamma_{t}=v$. A loop erased random walk, $\gamma^{\prime}$, is constructed from $\gamma$ by removing all loops in the order in which they appear on $\gamma$. We define the indices of the positions in $\gamma$ that are not on loops, i.e., the positions that do not occur between two occurrences of some vertex on $\gamma$, inductively by

$$
\begin{aligned}
& i_{1}=1 \\
& i_{j}=\max \left\{i: \gamma_{i}=\gamma_{i_{j-1}}\right\}+1
\end{aligned}
$$

Then $\gamma^{\prime}$, the loop erased version of $\gamma$, is defined by $\gamma_{j}^{\prime}=\gamma_{i_{j}}$.
We define a random variable $\operatorname{LERW}(u, v)$ whose value is a loop erased random walk from $u$ to $v$. Similarly, $\operatorname{LERW}(u, U)$ is a loop erased random walk from $u$ to any vertex in $U$; that is, the random walk $\gamma$ terminates as soon as we reach a vertex from $U$.

Example 1.64. Suppose we have

$$
\gamma=1,2,3,1,4,2,5 .
$$

Then $i_{1}=1, i_{2}=5, i_{3}=6$ and $i_{4}=7$, so

$$
\gamma^{\prime}=1,4,2,5
$$

```
Algorithm 3: Propp-Wilson
    Input: A directed graph \(\vec{G}=(V, A)\) and a vertex \(v \in V\)
    Output: A uniformly random arborescence \(\mathcal{A}\) of \(\vec{G}\) rooted at \(v\)
    begin
        \(U \longleftarrow\{v\}, \mathcal{A} \longleftarrow \mathfrak{0} ;\)
        while \(U \neq V\) do
        Choose \(u \in V-U\);
        \(\gamma \longleftarrow \operatorname{LERW}(u, U) ;\)
        Add the edges of \(\gamma\) to \(T\) and the vertices to \(U\);
        end
    end
```

The Propp-Wilson algorithm is presented as Algorithm 3. Propp and Wilson proved that this algorithm generates a random arborescence from the uniform distribution on $\operatorname{ARB}(\vec{G}, v)$ for any directed graph, and that the running time is bounded by the hitting time.

Definition 1.65. Let $\vec{G}=(V, A)$ be a directed graph. The hitting time of $u, v \in V$ is the random variable recording the amount of time for the random walk on $\vec{G},\left(X_{t}\right)_{t \geq 0}$, to reach $v$ given that it started at $u$ :

$$
H_{u, v}=\inf \left\{t: X_{t}=v: X_{0}=u\right\} .
$$

The hitting time of $\vec{G}$ is the maximum expected hitting time over all pairs of vertices:

$$
h(\vec{G})=\max _{u, v \in V} \mathbb{E}\left(H_{u, v}\right) .
$$

Remark 1.66. The Propp-Wilson algorithm can be regarded as a Markov chain, $\mathcal{M}_{P W}$, on $\operatorname{ARB}(\vec{G})$, which operates as follows. Let $X_{t}$ denote the $t$-th state of the chain and let $r_{t}$ be the root of $X_{t}$. We use $X_{t}(v)$ to denote the unique arc leaving $v$ in $X_{t}$, for each $v \neq r_{t}$. We now describe how $\mathcal{M}_{P W}$ generates the next state of the chain, $X_{t+1}$, given $X_{t}$.

## One step of the Markov chain $\mathcal{M}_{P W}$

1. Choose a random $\operatorname{arc}\left(r_{t}, r_{t+1}\right) \in A$.
2. Obtain $X_{t+1}$ from $X_{t}$ by removing $X_{t}\left(r_{t+1}\right)$ and adding $\left(r_{t}, r_{t+1}\right)$.

Consider what happens when the sequence of root vertices of states, $X_{t}, \ldots, X_{t+k}$, forms a cycle $\left(v_{0}, v_{1}, \ldots, v_{k-1}\right)$. On the transition $\left(X_{t}, X_{t+1}\right)$ we remove ( $v_{0}, v_{1}$ ) from the arborescence and add $\left(v_{1}, v_{2}\right)$. Then, on the transition $\left(X_{t+1}, X_{t+2}\right)$ we remove $\left(v_{1}, v_{2}\right)$ and add $\left(v_{2}, v_{3}\right)$, and so on. Eventually, the transition $\left(X_{t+k-1}, X_{t+k}\right)$ removes $\left(v_{k-1}, v_{0}\right)$ and adds $\left(v_{0}, v_{1}\right)$. Hence, $X_{t+k}=X_{t}$. Thus, at any particular step $t$ of the chain, the arcs which have been used by the random walk but are not on a loop are contained in $X_{t}$, so this chain is equivalent to the Propp-Wilson algorithm [72]. The result of Propp-Wilson [72] was to show that this chain has reached its stationary distribution once all vertices have been visited by the random walk.

### 1.3.2.2 Euler tours of undirected graphs

Next, we turn to the problem of generating and counting Euler tours of an undirected graph $G=(V, E)$. Two alternative representations have been used to reason about the set of Euler tours of an undirected graph: transition systems and orbs.

The first, and more direct representation, is to define a tour $\mathcal{T}$ according to its component 2-paths: the pairs of edges which are visited consecutively on $\mathcal{T}$.

Definition 1.67. Let $G=(V, E)$ be an Eulerian graph and let $v \in V(G)$. We define a transition system at $v$ to be a decomposition of the set of edges incident with $v$ into pairs. A transition system of $G$ is a function $T$ that maps each $v$ to a transition system on $v$. We denote by $\operatorname{TS}(G)$ the set of transition systems of $G$.

We can also define a transition system of an Eulerian directed graph by adding the restriction that an in-arc of $v$ must be paired with an out-arc of $v$.

Suppose $G=(V, E)$ is an Eulerian graph on $n$ vertices such that $\operatorname{deg}(v)=d_{v}$, for each $v \in V$. The number of ways to choose a pairing of the edges at $v$ is

$$
\frac{\binom{2 d_{v}}{d_{v}} d_{v}!}{2^{d_{v}}} ;
$$

hence, the number of transition systems of any graph satisfying $\operatorname{deg}(v)=d_{v}$ is

$$
\prod_{v \in V} \frac{\binom{2 d_{v}}{d_{v}} d_{v}!}{2^{d_{v}}}
$$

In particular, every $2 d$-regular graph has

$$
\left(\frac{\binom{2 d}{d} d!}{2^{d}}\right)^{n}
$$

transition systems. Similarly, a directed graph in which $\operatorname{outdeg}(v)=d_{v}$, for every $v \in V$, has
transition systems, and every $d$-in/ $d$-out directed graph has

$$
d!^{n}
$$

transition systems.
Each transition system $T$ of $G$ corresponds to a decomposition of the edges of $G$ into a set of edge-disjoint cycles, denoted by $\mathcal{C}(T)$, where the edges that are paired together by $T$ lie on the same cycle. Hence, the Euler tours of $G$ are equivalent to the set of transition systems for which $\mathcal{C}(T)$ contains exactly one component. It seems the first method to exactly count the number of Euler tours of any Eulerian graph was developed by Tarry [86] in 1887. This approach chooses a transition system vertex
by vertex, ensuring that the current transition system can always be extended to the transition system of an Euler tour. Suppose $G$ is an Eulerian graph and $v$ is a vertex of $G$. Let $\mathcal{P}(v)$ denote the set of pairings of the edges incident with $v$. For each $\rho \in \mathcal{P}(v)$, let $G_{\rho}$ denote the graph obtained by removing $v$ and adding the edges $u, w$ for each $\{\{u, v\},\{v, w\}\} \in \rho$. Then, for any $v \in V$, we have

$$
\begin{equation*}
\# \mathrm{ET}(G)=\sum_{\substack{\rho \in \mathcal{P}(v) \\ G_{\mathrm{\rho}} \text { is connected }}} \# \mathrm{ET}\left(G_{\rho}\right) . \tag{1.21}
\end{equation*}
$$

The equation in (1.21) can be expanded recursively to compute each of the \# $\mathrm{ET}\left(G_{\rho}\right)$, stopping the recursion when we reach a graph for which we can easily calculate the number of Euler tours, e.g., a graph with 2 vertices. This procedure is not practical for large graphs as the running time is proportional to the number of Euler tours of $G$, which is almost certainly exponential in the number of vertices. Therefore, Tarry's approach is not useful for developing polynomial-time algorithms for counting Euler tours.

Kötzig [56] showed that any Eulerian transition system can be transformed into another by performing a sequence of local changes, each one altering the transition system at exactly one vertex.

Definition 1.68. Let $G=(V, E)$ be an Eulerian graph, let $\mathcal{T}$ be an Euler tour of $G$ and let $T$ denote the transition system corresponding to $\mathcal{T}$. A $\kappa$-transformation at $v$ takes two pairs of edges $\{e, f\}$ and $\left\{e^{\prime}, f^{\prime}\right\}$ in $T(v)$ and obtains $T^{\prime}$ by replacing them by $\left\{e, f^{\prime}\right\}$ and $\left\{e^{\prime}, f\right\}$. If $T^{\prime}$ induces an Euler tour $\mathcal{T}^{\prime}$ we call this an allowed transition and denote this by $\mathcal{T} \Rightarrow_{\nu} \mathcal{T}^{\prime}$.

An allowed $\kappa$-transformation corresponds to choosing two distinct occurrences of $v$ on $\mathcal{T}$ and reversing the segment of $\mathcal{T}$ between them. In [87] Tetali and Vempala used this result to define an ergodic Markov chain that converges to the uniform distribution on the set of Euler tours of any Eulerian graph. The transitions in this chain are very simple. Let $X_{t}$ denote the $t$-th state of the chain. We now describe how a step of the Kötzig move chain obtains $X_{t+1}$, given $X_{t}$.

## One step of the Kötzig chain

1. Choose $v \in V$ uniformly at random and choose a $\kappa$-transformation uniformly at random from the set of potential transformations;
2. If the chosen transformation is allowed in $X_{t}$ then apply it to obtain $X_{t+1}$ as the next state;
3. Otherwise, set $X_{t+1}=X_{t}$.

In [87], Tetali and Vempala attempted to use the multi-commodity flow technique [49] to prove this chain was rapidly mixing, in the case where the given graph is either 4-regular or 6-regular. To apply this technique one needs to define certain canonical paths between states in the Markov chain. The construction for these paths in [87] does not give paths with the claimed properties; hence, the proof does not work. Even now, several years later, no one has been able to fix the argument of [87]. The reason for this is that the structure of the state space of the chain, upon which proofs of rapid mixing depend, is very complicated. Although the change we are making in an individual transition is local, we need to look at the whole tour to determine whether or not a particular change is legal. Generally speaking, the only Markov chains whose mixing times are amenable to analysis by current proof techniques are those in which the moves are very local; that is, we can determine which changes we can apply to a part of the structure by considering the parts of the structure close to the part we want to change. So, even in the simple cases of 4-regular and 6-regular graphs, the problem of sampling Euler tours of undirected graphs in polynomial time remains open.

McKay and Wormald [67] have obtained an asymptotic expression for the number of Euler tours of $K_{n}$ for $n$ odd. One corollary of this result is that a simple dart throwing algorithm can be used to generate random Euler tours of $K_{n}$ in polynomial time. This algorithm, described in Chapter 3 of this thesis, does not run in polynomial time for every Eulerian graph: the graph $B_{n}$ (see Example 1.70 below) is one example of a graph for which it will take exponential time; see $\S 3.6$ for more interesting examples. Thus, it is an interesting open problem to determine on which classes of graphs this algorithm is effective. In Chapter 3 of this thesis we present some partial results towards showing that the simple algorithm studied by McKay and Wormald will run in expected polynomial time for almost all $2 d$-regular graphs. Whether or not this is the case depends on an unproven conjecture, a conjecture which is supported by empirical evidence.

For many years it was not even known whether or not there might exist a polynomial-time algorithm for exactly counting the Euler tours of an undirected graph. This question was finally resolved by Brightwell and Winkler [15] in 2005, when they showed that the problem was indeed \#P-complete ${ }^{4}$. Brightwell and Winkler

[^3]considered a second representation for the Euler tours of a graph $G$.
Definition 1.69. Let $G=(V, E)$ be an Eulerian graph. We define an orb of $G$ rooted at $v$ to be a pair $(\mathcal{E}, \mathcal{A})$, where $\mathcal{E} \in \operatorname{EO}(G)$ and $\mathcal{A} \in \operatorname{ARB}(\mathcal{E}, v)$. We use $\operatorname{ORB}(G, v)$ to denote the set of orbs of $G$ rooted at a particular vertex $v$, and $\operatorname{ORB}(G)$ to denote the set of all orbs of $G$.

From Theorem 1.57, we know that the number of Euler tours of a particular Eulerian orientation $\mathcal{E} \in \operatorname{EO}(G)$ is

$$
\left[\prod_{u \in V(G)}(\operatorname{deg}(u) / 2-1)!\right]|\operatorname{ARB}(\mathcal{E}, v)|,
$$

for an arbitrary vertex $v \in V$. Hence, the number of Euler tours of $G$ satisfies

$$
|\operatorname{ET}(G)|=\frac{1}{n}\left[\prod_{u \in V(G)}(\operatorname{deg}(u) / 2-1)!\right]|\operatorname{ORB}(G)| .
$$

Brightwell and Winkler [15] showed that counting orbs is \#P-complete via a reduction from \#EO; hence, \#ET is also \#P-complete.

Although the reason Brightwell and Winkler came up with the concept of an orb was to prove \#P-completeness of \#ET, the connections between orbs and Euler tours also suggests ideas that might allow us to come up with polynomial-time algorithms for sampling Euler tours. Let $(\mathcal{E}, \mathcal{A}) \in \operatorname{ORB}(G, v)$. We can generate a random Euler tour associated with $(\mathcal{E}, \mathcal{A})$ in polynomial time: all we need to do is choose a random ordering on the non-arborescence out-arcs of each vertex and then follow the deterministic procedure described in the proof of Theorem 1.57. Hence, if we could sample from the uniform distribution on the orbs of $G$, we could also sample from the uniform distribution on the Euler tours of $G$. Unfortunately, this is no easy task. The algorithms for sampling arborescences do not generalise to sampling orbs as most known algorithms for sampling arborescences are heavily dependent on the fact that the input graph is directed.

To illustrate why the problem of sampling orbs of an undirected graph is so different from that of sampling arborescences of a directed graph, we consider the following naive idea for a sampling algorithm. Suppose we try to choose the arborescence part of the orb step by step, in a similar fashion to the algorithms based on Theorem 1.59, e.g., [19]. At each step we want to calculate the probability that an edge incident with a vertex $u$ is oriented away from $u$ and contained in the arborescence part of a random orb. But this would require being able to count the
number of orbs containing a particular set of arcs in their arborescence, and this is clearly \#P-complete! Attempts to generalise the random walk algorithms to the undirected case suffer from similar problems, since we need an algorithm that changes both the orientation part and the arborescence part simultaneously.

We know that we can generate random Eulerian orientations of any graph in polynomial time, and that we can generate Euler tours of any Eulerian directed graph in polynomial time. However, simply choosing a uniformly random Eulerian orientation $\mathcal{E} \in \operatorname{EO}(G)$ and then choosing a random Euler tour $\mathcal{T} \in \mathrm{ET}(\mathcal{E})$ will not give a uniformly random Euler tour, since different Eulerian orientations of a graph can have very different numbers of Euler tours. For example, consider the graph obtained by duplicating each edge of the $n$-cycle.

Example 1.70. Let $B_{n}$ be the multigraph obtained by duplicating each edge in the $n$ cycle $C_{n}=(0,1,2, \ldots, n-1)$. The Eulerian orientations of $B_{n}$ can be grouped into two equivalence classes. For each $i$ we label the edges joining $i$ to $(i+1) \bmod n$ as $e_{i}$ and $f_{i}$. We say two Eulerian orientations $\mathcal{E}, \mathcal{E}^{\prime} \in \mathrm{EO}\left(B_{n}\right)$ are equivalent if either $\mathcal{E}^{\prime}$ is the reverse of $\mathcal{E}$ or $\mathcal{E}^{\prime}$ can be obtained from $\mathcal{E}$ by switching the orientations of $e_{i}$ and $f_{i}$ for each $i$ in some set $S \subset\{1,2, \ldots, n\}$. This relation partitions $\operatorname{EO}(G)$ into two classes: the orientations in which $e_{i}$ and $f_{i}$ are both oriented in the same direction for every $i$ and the orientations in which $e_{i}$ and $f_{i}$ are oriented in opposite directions for every $i$. There are exactly two Eulerian orientations in the first class and $2^{n}$ in the second.

Both of the orientations in the first class have $2^{n-1}$ arborescences rooted at each vertex $v$, and each of the orientations in the second class have $n$ arborescences rooted at each vertex $v$. Hence, if we choose a random orb as suggested above, the probability of obtaining a particular $(\mathcal{E}, \mathcal{A})$ with $\mathcal{E}$ in the the first class is

$$
\frac{1}{2^{n}+2} \frac{1}{n 2^{n-1}}
$$

but the probability of obtaining a particular $(\mathcal{E}, \mathcal{A})$ with $\mathcal{E}$ in the second class is

$$
\frac{1}{2^{n}+2} \frac{1}{n^{2}} .
$$

### 1.4 Contents of the thesis

The problems \#EO and \#ET have been shown to \#P-complete for general graphs by Mihail and Winkler [69] and Brightwell and Winkler [15], respectively. We extend current knowledge by showing that the complexity of counting Eulerian orientations
remains \#P-complete for planar graphs. This result is contained in the first section of Chapter 2.

The Eulerian orientations of any graph can be sampled in polynomial time, by combining the reduction of Mihail and Winkler [69] with the results of Jerrum et al. [49, 51] (improvements in [10]). A more natural chain on the Eulerian orientations of planar graphs has been studied in [62,38], in which the mixing time was analysed on the Eulerian orientations of the square lattice. In Chapter 2 we show that this chain is rapidly mixing on the triangular lattice, a planar lattice that which is of interest in the statistical physics community $[6,29]$. However, we also show that there exist planar graphs for which it will always take an exponential time to converge.

In Chapter 3 we describe a simple algorithm that can be used to generate uniform Euler tours or approximately count the number of Euler tours of any Eulerian graph or Eulerian directed graph. It is known that this algorithm runs in expected polynomial time for the complete graph [66], but also that it can be expected to take an exponential amount of time for other graphs. We use the theory of random graphs to show that the algorithm runs in expected polynomial time for almost every Eulerian directed graph.

Although we are not able to show that this algorithm runs in expected polynomial time on almost every $2 d$-regular graph, we obtain some partial results towards showing that this is the case. This includes an analysis of the number of Eulerian orientations of a random $2 d$-regular graph that is of independent interest. Whether or not the algorithm runs in expected polynomial time for almost every $2 d$-regular graph depends on an unproven conjecture regarding an estimate for the second moment of the number of orbs of a random $2 d$-regular graph. In Chapter 3, we provide some empirical evidence to support our conjecture.

## Chapter 2

## Eulerian orientations of planar graphs

### 2.1 Introduction

Recall that the Eulerian orientations of certain planar graphs are of particular practical interest as they correspond to configurations of the ice-type models studied in the statistical physics community, e.g., the square lattice [61], the triangular lattice [6], the Kagomé lattice [29], and the 3.4.6.4 lattice [29]. While the \#P-completeness of counting the Eulerian orientations of a general Eulerian graph (\#EO) was established in [68, 69], it is not known if the problem remains \#P-complete under the restriction that the input be a planar graph (\#PlanarEO). The first result of this chapter answers this question by showing that \#EO is polynomial-time Turing reducible to \#PlanarEO.

Given the computational hardness of exact counting, we turn our attention to approximation. Recall from §1.1.2 that one approach to approximate counting is via random sampling. It is known that we can generate (almost) uniformly random Eulerian orientations of any Eulerian graph in polynomial time (by combining the result of Mihail and Winkler [68,69] described in §1.3.1 with the rapidly mixing Markov chain of $[49,51,10])$. In this chapter, we are interested in a different approach, which uses a natural Markov chain on the $\alpha$-orientations of a planar graph. The reason for using a chain on the set of $\alpha$-orientations is that sampling Eulerian orientations of a solid subgraph of an infinite lattice with some fixed boundary condition requires one to sample $\alpha$-orientations of a non-Eulerian graph.

The Markov chain we study is in the style of the "Glauber dynamics" often used to generated random configurations of physical models. Although the approach of Mihail and Winkler detailed in §1.3.1 gives an fpaus for the set of Eulerian orientations of any graph, this chain is still of interest as it is the method most commonly used in
practice [28, 2, 29]. The main contribution of this chapter is to show that this Markov chain is rapidly mixing for any solid subgraph of the triangular lattice. However, we are also able to find an infinite family of planar graphs for which this Markov chain will take an exponential number of steps to converge.

The contents of this chapter is as follows. In $\S 2.2$ we describe the reduction that shows \#PlanarEO is \#P-complete. In §2.3, we define the face-reversal Markov chain and show that the stationary distribution is the uniform distribution on the set of $\alpha$-orientations of any planar graph. In this section, we also describe a result of Felsner [32] that is useful in the analysis of this Markov chain. In §2.4, we describe path coupling, a technique for constructing couplings that we will use to analyse the mixing time of the Markov chain. Path coupling is not applicable directly to the facereversal chain, so, in $\S 2.5$, we first extend the chain with additional moves, in the style of $[62,38]$. Then, in $\S 2.6$, we show that the face-reversal chain is rapidly mixing when the input is a solid section of the triangular lattice with a fixed boundary condition. This analysis is broken into two parts. In $\S 2.6 .1$ we use path coupling to analyse the mixing time of the extended chain. Then, in §2.6.2, we use a technique called comparison to infer that the face-reversal chain is rapidly mixing on this class of graphs. In §2.7 we show how the face-reversal chain can be used to construct an fpras for counting Eulerian orientations of the triangular lattice. Finally, in §2.8 we show that there exist planar graphs for which the face-reversal chain is torpidly mixing, using a distributive lattice structure (due to Felsner [32]) on the set of $\alpha$-orientations of a planar graph.

### 2.2 Computational Complexity of \#PlanarEO

Before discussing algorithms for sampling Eulerian orientations of planar graphs, we provide a polynomial-time reduction from \#EO to \#PlanarEO. This shows that \#PlanarEO is \#P-complete since the \#P-completeness of \#EO is already known [68, 69]. Our reduction uses a recursive gadget (suggested by Mark Jerrum [48]) and can be seen as an application of the so-called Fibonacci method of Vadhan [91].

The proof is as follows. Given a non-planar Eulerian graph $G$ with an embedding in the plane, we construct a sequence of planar Eulerian graphs $G_{k}, k=0,1, \ldots, \ell+1$, where $\ell$ is the number of crossing edges in the embedding of $G$. We then show that the values $\# \mathrm{EO}\left(G_{k}\right)$ give evaluations of a degree $\ell$ polynomial $p(x)$ at a sequence of distinct points. This polynomial, $p(x)$, has the value $\# \mathrm{EO}(G)$ as one of its coefficients. Hence, we can obtain the value \# $\mathrm{EO}(G)$ from the values $\# \mathrm{EO}\left(G_{k}\right)$, using the Lagrange


Figure 2.1: The crossover box

## Interpolation Formula:

Theorem 2.1 (Lagrange). Given $d+1$ distinct real values $\beta_{0}, \beta_{1}, \ldots, \beta_{d}, d \geq 2$, and $d+1$ points $\left(\alpha_{0}, \beta_{0}\right),\left(\alpha_{1}, \beta_{1}\right), \ldots,\left(\alpha_{d}, \beta_{d}\right)$, there is a unique polynomial of the form

$$
f(x)=\sum_{i=0}^{d-1} a_{i} x^{i}
$$

passing through these points, specified by

$$
\begin{equation*}
f(x)=\sum_{i=0}^{d} \beta_{i} \prod_{j \neq i} \frac{x-\alpha_{j}}{\alpha_{i}-\alpha_{j}} . \tag{2.1}
\end{equation*}
$$

Theorem 2.2. \#PlanarEO is \#P-complete.
Proof. We prove the theorem by exhibiting a polynomial time reduction from \#EO to \#PlanarEO. That, is given an oracle for computing \#PlanarEO we can construct an algorithm that can compute \#EO for any non-planar graph.

Let $G$ be any non-planar graph and suppose we have an embedding in the plane with $\ell$ crossings. We turn $G$ into a planar graph $G_{0}$ by removing each pair of crossing edges $\{x, y\}$ and $\{u, v\}$ and replacing them by a pair of paths $(x, s, y)$ and $(u, s, v)$, where $s$ is a vertex not contained in $G$, as in Figure 2.1(a). Every Eulerian orientation of $G$ induces an Eulerian orientation of $G_{0}$; however, not all Eulerian orientations of $G_{0}$ correspond to Eulerian orientations of $G$, e.g., $\{(u, s),(v, s)\}$ may be present in an Eulerian orientation of $G_{0}$. We say $\mathcal{E} \in \operatorname{EO}\left(G_{0}\right)$ is valid at $s$ if the orientation of the edges incident with $s$ is consistent with an orientation of $G$, i.e., exactly one of $\{u, s\}$ and $\{v, s\}$, and of $\{x, s\}$ and $\{y, s\}$ is directed towards $s$ in $\mathcal{E}$.

We now define a sequence of graphs $H_{k}$, each of which has 4 vertices of degree 1 , labelled $u_{k}, v_{k}, x_{k}$, and $y_{k}$. Let $H_{0}$ be the graph Figure 2.1(a). For $k \geq 1$, we define $H_{k}$ recursively in terms of $H_{k-1}$ and a 4 -cycle. Each vertex of the cycle connects to a pair of vertices from $\left\{\left(u_{k}, u_{k-1}\right),\left(v_{k}, v_{k-1}\right),\left(x_{k}, x_{k-1}\right),\left(y_{k}, y_{k-1}\right)\right\}$, as illustrated in Figure 2.1(b). We then define $G_{k}$ to be the graph obtained from $G$ by replacing each pair of crossing edges $\{x, y\},\{u, v\}$ with $H_{k}$. To do this we delete the edges $\{x, y\}$ and $\{u, v\}$ and identify $x, y, u$ and $v$ with $x_{k}, y_{k}, u_{k}$, and $v_{k}$. We call the copy of $H_{k}$ used to replace $\{x, y\}$ and $\{u, v\}$ the crossover box for the crossing $\{x, y\}$ and $\{u, v\}$. We say an orientation of $G_{k}$ is valid at particular crossover box if the orientation of the edges (of $H_{k}$ ) incident with $u_{k}, v_{k}, x_{k}$, and $y_{k}$ are consistent with an orientation of $G$.

Let $a_{k}$ (resp. $b_{k}$ ) denote the number of orientations of the edges of $H_{k}$ satisfying the Eulerian condition which correspond to valid (resp. invalid) orientations. These values satisfy

$$
\begin{align*}
a_{k} & =4 a_{k-1}+2 b_{k-1}  \tag{2.2}\\
b_{k} & =4 a_{k-1}+3 b_{k-1} \tag{2.3}
\end{align*}
$$

with $a_{0}=b_{0}=1$. Now, let $N_{i}$ denote the number of Eulerian orientations of $G_{0}$ which are valid at exactly $i$ crossover boxes, so $N_{\ell}=\# \mathrm{EO}(G)$. Each Eulerian orientation of $G_{0}$ counted by $N_{i}$ corresponds to exactly $a_{k}^{i} b_{k}^{\ell-i}$ Eulerian orientations of $G_{k}$, so we can write

$$
\# \operatorname{PlanarEO}\left(G_{k}\right)=\sum_{i=0}^{\ell} N_{i} a_{k}^{i} b_{k}^{\ell-i} .
$$

Hence, computing \#PlanarEO $\left(G_{k}\right) / b_{k}^{\ell}$ corresponds to evaluating the polynomial

$$
p(z)=\sum_{i=0}^{\ell} N_{i} z^{i}
$$

at the point $a_{k} / b_{k}$. It is straightforward to check that $a_{k} / b_{k}$ is a non-repeating sequence (either by basic calculus or an application of [91, Lemma 6.2]). Hence, by evaluating \# $\operatorname{PlanarEO}\left(G_{k}\right) / b_{k}^{\ell}$ at the value $a_{k} / b_{k}$, for $k=0,1, \ldots, \ell+1$, we obtain enough information to recover the values $N_{i}, i=0,1, \ldots, \ell$, by the Lagrange interpolation formula (Theorem 2.1).

Letting $n$ and $m$ denote the number of vertices and edges in $G$ we have

$$
\left|V\left(G_{k}\right)\right|=n+(4 k+1) \ell
$$

and

$$
\left|E\left(G_{k}\right)\right|=m+(8 k+2) \ell .
$$

Hence, we can compute $\# \operatorname{EO}\left(G_{k}\right)$ in time polynomial in $n$, $m$, and $\ell$, for each $k$. We can embed a graph with $m$ edges in the plane with at most $\binom{m}{2}$ crossings: if we lay the vertices out on a circle each pair of edges can cross at most once. Thus, the number of steps required for the whole reduction is bounded by a polynomial in $m$ and $n$.

### 2.3 A natural Markov chain for sampling $\alpha$-orientations

Before we define the natural Markov chain on the set of $\alpha$-orientations of a planar graph, we introduce some structural results of Felsner [32] which will be useful in the definition and subsequent analysis of the chain. Let $G=(V, E)$ be a planar graph and let $\mathcal{F}(G)$ denote the set of bounded faces in some planar embedding of $G$. We will use $f$ to denote the number of elements in $\mathcal{F}(G)$. A face is said to be directed in an orientation of $G$ if its boundary edges form a directed cycle. We say a directed face $\gamma$ is clockwise oriented (resp. counter-clockwise oriented) if the orientation of the edges of $\gamma$ is clockwise (resp. counter-clockwise) on $\gamma$. We recall the definition of $\alpha$-orientation.

Definition 2.3. Let $G=(V, E)$ be a graph and let $\alpha: V \rightarrow \mathbb{Z}$. An $\alpha$-orientation of $G$ is an orientation in which outdeg $(v)-\operatorname{indeg}(v)=\alpha(v)$ for all $v \in V$. Let $\alpha-\mathrm{O}(G)$ denote the set of $\alpha$-orientations of $G$.

Note that the Eulerian orientations of a graph are the $\alpha$-orientations where $\alpha(v)=0$ for all $v \in V$.

Definition 2.4. If an edge $e \in E$ has the same orientation in every $\alpha$-orientation of $G$, we say $e$ is $\alpha$-rigid.

Felsner defined a special set of cycles in a planar graph, with respect to $\alpha$, which he calls the essential cycles.

Definition 2.5 ( [32]). A cycle $C$ of $G$ is essential (with respect to $\alpha$ ) if

1. $C$ is simple and chord free;
2. all edges on the interior of $C$ are $\alpha$-rigid;
3. there exists an $\alpha$-orientation of $G$ in which $C$ is a directed cycle.

We denote by $\mathcal{F}_{\alpha}(G)$ the set of essential cycles of $G$ with respect to $\alpha$.

Remark 2.6. The set of essential cycles are exactly the faces of the planar graph obtained by removing all rigid edges. In particular, for the case of Eulerian orientations we have that $\mathcal{F}(G)$ is the set of essential cycles.

From here on we will assume that all rigid edges have been removed, with $\alpha$ updated accordingly, so the essential cycles of $G$, with respect to $\alpha$, will always be $\mathcal{F}(G)$.

In [32], Felsner has shown that it is possible to convert any $\alpha$-orientation of a planar graph $G$ into another by performing a sequence of reversals of the edges of directed faces. Furthermore, Felsner defined a partial order on $\alpha-\mathrm{O}(G)$.

Definition 2.7. Let $G=(V, E)$ be a planar graph and suppose we have some $\alpha: V \rightarrow \mathbb{N}$. We define a partially ordered set $\operatorname{Fels}(G)=(\alpha-\mathrm{O}(G), \prec)$, where $\mathcal{E} \prec \mathcal{E}^{\prime}$ if $\mathcal{E}^{\prime}$ can be obtained from $\mathcal{E}$ by performing a sequence of reversals of clockwise oriented directed faces.

This order has a unique maximum element and minimum element: the unique $\alpha$-orientation with no clockwise cycles, $\mathcal{E}_{\max }$, and the unique $\alpha$-orientation with no counter-clockwise cycles, $\mathcal{E}_{\text {min }}$. In [32], Felsner proved this order forms a finite distributive lattice.

Definition 2.8. A distributive lattice is a partially ordered set ( $L, \prec$ ) which also satisfies:

1. every pair of elements $x, y \in L$ has a unique least upper bound, $x \vee y$, called the join;
2. every pair of elements $x, y \in L$ has a unique greatest lower bound, $x \wedge y$, called the meet;
3. the join and meet operations are distributive

$$
x \wedge(y \vee z)=(x \wedge y) \vee(x \wedge z) .
$$

That $\operatorname{Fels}(G)$ is a distributive lattice is established by giving a bijection between the $\alpha$-orientations of a planar graph and a set of functions called $\alpha$-potentials: for each $\mathcal{E} \in \alpha-\mathrm{O}(G)$ the $\alpha$-potential associated with $\mathcal{E}$ is a function

$$
\wp_{\mathcal{E}}: \mathcal{F}(G) \rightarrow \mathbb{N} .
$$

To define these functions Felsner used a partial order on $\mathcal{F}(G): \sigma \prec_{\mathcal{F}} \rho$ if $\sigma$ and $\rho$ share an edge and that edge is counter-clockwise on $\sigma$ in the minimum orientation. The $\alpha$-potentials are then defined as the set of functions of the form $\wp: \mathcal{F}(G) \rightarrow \mathbb{N}$ satisfying

$$
\begin{align*}
\sigma \text { and } \rho \text { share an edge } & \Rightarrow|\wp(\sigma)-\wp(\rho)| \leq 1  \tag{2.4}\\
\sigma \text { is on the boundary } & \Rightarrow \wp(\sigma) \leq 1 \tag{2.5}
\end{align*}
$$

$$
\begin{equation*}
\sigma \prec_{\mathcal{F}} \rho \Rightarrow \wp(\sigma) \leq \wp(\rho) \tag{2.6}
\end{equation*}
$$

The bijection is given as follows: for each $\mathcal{E} \in \alpha-\mathrm{O}(G)$, we define $\wp_{\mathcal{E}}$ to be function $\mathcal{F}(G) \rightarrow \mathbb{N}$ where $\wp_{\mathcal{E}}(\sigma)$ is equal to the number of times $\sigma$ is reversed on any shortest path from $\mathcal{E}_{\min }$ to $\mathcal{E}$. See Figure 2.2 for an example. In the figure, each face is labelled with its potential value.


Figure 2.2: The Felsner lattice for a small graph

Any partial order $\prec$ of a set $L$ induces a graph with vertex set $L$ called the cover graph of $L$.

Definition 2.9. The cover graph of a partially ordered set $(L, \prec)$ is the graph with vertex set $L$ and an edge joining $x, y \in L$ if there is no $z \in L$ such that $x \prec z \prec y$.

In the case of the Felsner lattice, two orientations are adjacent in the cover graph if they disagree on the orientation of a single directed face. Felsner's result suggests a natural Markov chain $\mathcal{M}_{\mathcal{F}}$ on the set of $\alpha$-orientations of any planar graph $G$ : a random walk on the cover graph of $\operatorname{Fels}(G)$. Each transition in this chain reverses the orientation of a random directed face, so we call this the face-reversal Markov chain.

We now formally define our face-reversal chain, by specifying how $X_{t+1}$ is generated from the current state $X_{t}$. We use $c w$ and $c c w$ to denote clockwise and counter-clockwise, respectively.

## One step of the chain $\mathcal{M}_{\mathcal{F}}$

1. Choose $\kappa \in \mathcal{F}(G)$ and dir $\in\{c w, c c w\}$ uniformly at random.
2. If $\kappa$ is $d i r$-directed then obtain $X_{t+1}$ from $X_{t}$ by reversing the orientation of all the edges in $\kappa$.
3. Otherwise, set $X_{t+1}=X_{t}$.

Proposition 2.10. For any planar graph $G=(V, E)$ and any $\alpha: V \rightarrow \mathbb{Z}$, the chain $\mathcal{M}_{\mathcal{F}}$ is ergodic and converges to the uniform distribution on $\alpha-\mathrm{O}(G)$.

Proof. By choosing the direction (cw or ccw) in which to perform the reversal as well as choosing a face, we automatically ensure that the probability of $X_{t+1}$ being equal to $X_{t}$ is at least $1 / 2$. Hence, $\mathcal{M}_{\mathcal{F}}$ is aperiodic.

Felsner's result [32] (described above) ensures the chain is connected, and so ergodic. The transition probability matrix of $\mathscr{M}_{\mathcal{F}}$ is

$$
P\left(\mathcal{E}, \mathcal{E}^{\prime}\right)=\left\{\begin{array}{ll}
\frac{1}{2|\mathcal{F}(G)|} & \text { if } \mathcal{E} \oplus \mathcal{E}^{\prime}=\gamma, \text { for } \gamma \in \mathcal{F}(G) \\
1-\sum_{\mathcal{E}^{\prime \prime} \neq \mathcal{E}} P\left(\mathcal{E}, \mathcal{E}^{\prime \prime}\right) & \text { if } \mathcal{E}=\mathcal{E}^{\prime} \\
0 & \text { otherwise }
\end{array} .\right.
$$

Thus, the chain is ergodic and symmetric, and so converges to the uniform distribution on $\alpha-\mathrm{O}(G)$.

Felsner's result has more use than proving ergodicity. We use the distributive lattice structure as part of our proofs of rapid mixing (§2.6), and torpid mixing (§2.8), of $\mathcal{M}_{\mathcal{F}}$ on different classes of graphs.

### 2.4 Path Coupling

Suppose $\mathcal{M}$ is an ergodic time-homogeneous Markov chain on finite state space $\Omega$ with probability transition matrix $P$ and stationary distribution $\pi$. Recall the definition of the mixing time of a Markov chain from §1.2: a function $\tau(\varepsilon)$ such that for any $\varepsilon>0$ and $t>\tau(\varepsilon)$,

$$
\max _{x \in \Omega}\left\|P^{t}(x, \cdot)-\pi(\cdot)\right\|_{T V} \leq \varepsilon .
$$

We say a chain is rapidly mixing if there is a polynomial upper bound on $\tau(\varepsilon)$ and torpidly mixing if there exists some $\varepsilon>0$ such that $\tau(\varepsilon)$ is exponentially bounded from below.

## Path Coupling

Recall from $\S 1.2$ that coupling is a common method used to show a Markov chain is rapidly mixing. The path coupling method, due to Bubley and Dyer [16], simplifies the process of defining and analysing a coupling. Bubley and Dyer showed that, if we define and analyse a Markovian coupling on the adjacent states in the chain then we can also construct a coupling for nonadjacent $X$ and $Y$ by composing the couplings of the pairs of states along a path $X=Z_{0}, Z_{1}, \ldots, Z_{r}=Y$, where $\left(Z_{i}, Z_{i+1}\right)$ is a pair of states adjacent in the chain, for $i=0,1, \ldots, r$. This is encapsulated in the following theorem:

Theorem 2.11 (Bubley and Dyer [16]). Let $\mathcal{M}$ be an ergodic Markov chain with state space $\Omega$ and let $\delta$ be an integer valued metric defined on $\Omega \times \Omega$ which takes values in $\{0,1, \ldots, D\}$. Let $S$ be a subset of $\Omega \times \Omega$ such that for all $(X, Y) \in \Omega \times \Omega$ there exists a path

$$
X=Z_{0}, Z_{1}, \ldots, Z_{r}=Y
$$

between $X$ and $Y$ such that $\left(Z_{i}, Z_{i+1}\right) \in S$ for $0 \leq i<r$ and

$$
\sum_{i=0}^{r-1} \delta\left(Z_{i}, Z_{i+1}\right)=\delta(X, Y)
$$

Now suppose $\left(X_{t}, Y_{t}\right)$ is a coupling of $\mathcal{M}$ defined on $S$. If there exists $\beta \leq 1$ such that for all $(X, Y) \in S$

$$
\mathbb{E}\left[\delta\left(X_{t+1}, Y_{t+1}\right) \mid\left(X_{t}, Y_{t}\right)=(X, Y)\right] \leq \beta \delta\left(X_{t}, Y_{t}\right)
$$

then this coupling can be extended to a coupling $\left(X_{t}, Y_{t}\right)$ defined on the whole of $\Omega \times \Omega$ such that

$$
\mathbb{E}\left[\boldsymbol{\delta}\left(X_{t+1}, Y_{t+1}\right)\right] \leq \beta \boldsymbol{\delta}\left(X_{t}, Y_{t}\right) .
$$

Moreover, if $\beta<1$ then $\tau(\varepsilon) \leq \frac{\log \left(D \varepsilon^{-1}\right)}{1-\beta}$.
Remark 2.12. Although Theorem 2.11 does not require it, in most applications we have that $S$ is the set of pairs of states adjacent in the chain:

$$
S=\{(X, Y) \in \Omega: P(X, Y)>0\} .
$$

If $\beta=1$ in Theorem 2.11 then in order to use the Bubley-Dyer path coupling technique of Theorem 2.11 it must be shown that the variance of the distance between any two states (not just the pairs in $S$ ) after one step of the coupling can be bounded away from 0 , see, e.g., $[16,38]$. This situation is undesirable, as the power of the path coupling technique is that it allows one to avoid considering all pairs of states. A recent result has relaxed this condition:

Theorem 2.13 (Bordewich and Dyer [13]). Suppose we have a path coupling $\mathcal{C}=\left(X_{t}, Y_{t}\right)$ for an ergodic Markov chain $\mathcal{M}$ with distance metric $\delta: \Omega \times \Omega \rightarrow$ $\{0, \ldots, D\}$, where $S \subset \Omega \times \Omega$ is the set of pairs of states at distance 1 . Let $p$ denote the minimum transition probability between pairs of states in $S$. We define a new chain, the lazy chain $\mathcal{M}^{\star}$, with probability transition matrix

$$
P^{\star}(X, Y)=\left\{\begin{array}{ll}
\frac{P(X, Y)+p}{1+p} & X=Y \\
\frac{P(X, Y)}{1+p} & \text { otherwise }
\end{array} .\right.
$$

If $\beta \leq 1$ for the coupling $\mathcal{C}$ then the mixing time of the chain $\mathcal{M}^{\star}$ satisfies

$$
\tau^{\star}(\varepsilon) \leq\left\lceil p^{-1} e D^{2}\right\rceil\left\lceil\log \varepsilon^{-1}\right\rceil .
$$

Moreover, if $\tau^{\prime}(\varepsilon)$ denotes the random time $\operatorname{Bin}\left(\tau^{\star}(\varepsilon),(1+p)^{-1}\right)$, where $\operatorname{Bin}$ denotes the binomial distribution, then the distribution of $X_{\tau^{\prime}(\varepsilon)}$ is within $\varepsilon$ of the stationary distribution of the original chain $\mathcal{M}$.

Observation 2.14. For any Markov chain $\mathcal{M}$ which has a uniform stationary distribution, the lazy chain $\mathcal{M}^{\star}$ will also have a uniform stationary distribution.

The effect of the lazy chain $\mathcal{M}^{\star}$ is to adjust the transition probabilities slightly, so as to introduce a small additional probability that the distance between the two chains will change at any particular step. In many circumstances where $\beta=1$, this is sufficient to ensure that the lazy chain is rapidly mixing. The second part of the above theorem means that, for all practical purposes, we can consider the mixing time of the chain $\mathcal{M}$ to be the same as the mixing time of $\mathcal{M}^{\star}$.

## Applying Path Coupling to $\mathcal{M}_{\mathcal{F}}$

To apply path coupling to bounding the mixing time of the face-reversal chain $\mathcal{M}_{\mathcal{F}}$, we need to define a metric $\delta: \Omega^{2} \rightarrow \mathbb{N}$ on the set of $\alpha$-orientations of a planar graph $G$. The most natural choice is to define $\delta(\mathcal{E}, \hat{\mathcal{E}})$, for each $\mathcal{E}, \hat{\mathcal{E}} \in \alpha-\mathrm{O}(G)$, to be the length of the shortest path between $\mathcal{E}$ and $\hat{\mathcal{E}}$ in the cover graph of $\operatorname{Fels}(G)$. In particular, $\delta(\mathcal{E}, \hat{\mathcal{E}})=1$ if $\mathcal{E}$ and $\hat{\mathcal{E}}$ differ on a single face. Let $\mathcal{E}_{\min }$ and $\mathcal{E}_{\text {max }}$ denote the unique minimum and maximum elements of $\operatorname{Fels}(G)$. The fact that $\operatorname{Fels}(G)$ is a distributive lattice implies that the maximum value the metric $\delta$ can take on pairs from $\alpha-\mathrm{O}(G)$ is $D=\delta\left(\mathcal{E}_{\min }, \mathcal{E}_{\max }\right)$, since we can always find a path of length $\leq D$ between two states $\mathcal{E}$ and $\hat{\mathcal{E}}$ either by going down from $\mathcal{E}$ to $\mathcal{E}_{\text {min }}$ and then up to $\hat{\mathcal{E}}$, or by taking a similar path through $\mathcal{E}_{\text {max }}$.

Let $S$ denote the set of pairs of Eulerian orientations which disagree on the orientation of a single directed face:

$$
S=\{(\mathcal{E}, \hat{\mathcal{E}}) \in \alpha-\mathrm{O}(G) \times \alpha-\mathrm{O}(G): \mathcal{E} \oplus \hat{\mathcal{E}}=\gamma, \text { for some } \gamma \in \mathcal{F}(G)\} .
$$

To apply path coupling (Theorem 2.11 or Theorem 2.13), we need to define a one-step Markovian coupling $\mathcal{C}$ on $S$ and show that, in expectation, the distance between any pair of states in $S$ does not increase after one step of $\mathcal{C}$. The natural choice is the coupling that chooses the same face and direction in both chains. This is a maximal coupling for our chosen distance metric. That is, if this coupling does not contract (for our chosen metric $\delta$ ), we cannot expect any other one-step Markovian coupling to contract.

Proposition 2.15. There is no one-step Markovian coupling for $\mathcal{M}_{\mathcal{F}}$ that contracts under the metric $\delta$.

Proof. Consider the situation in Figure 2.3. This is the subgraph of some graph $G$ incident with a face $\gamma$, which consists of the only edges on which a pair of $\alpha$ orientations, $\mathcal{E}$ and $\hat{\mathcal{E}}$, disagree. Any face in $G$ not contained in the part of the graph shown in Figure 2.3 will either be directed in both $\mathcal{E}$ and $\hat{\mathcal{E}}$ or will be directed in neither $\mathcal{E}$ nor $\hat{\mathcal{E}}$. Hence, the move available, if any, resulting from choosing any face not included in the subgraph shown will be the same for $\mathcal{E}$ and $\hat{\mathcal{E}}$. By choosing one of these faces and the same direction in both chains, we can ensure that the distance between the two copies of the chain does not increase over one step of the coupling.

By choosing $\gamma$ and a different direction in both copies of the chain, we can ensure coalescence with probability $1 / f$. However, each neighbour of $\gamma$ will be reversed with


Figure 2.3: A configuration that gives rise to a non-contracting coupling for $\mathcal{M}_{\mathcal{F}}$
probability $1 / 2 f$ in $\mathcal{E}$ but cannot be reversed in $\hat{\mathcal{E}}$. Hence, the distance will increase by 1 with probability $3 / 2 f$, giving a total expected increase of $1 / 2 f$.

### 2.5 Extending the chain

In §2.4, we saw that no one-step Markovian coupling is sufficient to show rapid mixing of $\mathscr{M}_{\mathcal{F}}$ using the metric $\delta$. However, we will see that if we extend the chain with extra moves in the style of [62,38], we will then be able to design an appropriate path coupling for the extended chain. It is these additional "tower moves" that will allow us to couple with $\beta \leq 1$.

Definition 2.16. Let $G$ be a planar graph, $\mathcal{E}$ an $\alpha$-orientation of $G$, and $\gamma$ a face of G. We say $\gamma$ is almost-directed in $\mathcal{E}$ if all but one of the edges of $\gamma$ have a common direction on $\gamma$. We call the edge oriented in the opposite direction the blocking edge of $\gamma$. For faces $\eta$ and $\sigma$ which are, respectively, almost-directed and directed in the orientation $\mathcal{E}$, we say there is a tower starting at $\eta$ and ending at $\sigma$ if there is a sequence of faces $\eta=\gamma_{1}, \ldots, \gamma_{h}=\sigma$ such that

- $\gamma_{i}$ is almost directed in $\mathcal{E}$, and the blocking edge of $\gamma_{i}$ is shared with $\gamma_{i+1}$ for $1 \leq i \leq h-1$.
- $\gamma_{h}$ is directed in $\mathcal{E}$

We say $h$ is the height of the tower.
Observe that the definition of a tower implies that $C=\bigoplus_{1 \leq i \leq h} \gamma_{i}$ is a directed cycle in $\mathcal{E}$. We say that a tower is clockwise (resp. counter-clockwise) in an orientation if


Figure 2.4: Reversing a tower
this cycle is clockwise (resp. counter-clockwise) in the orientation. We call $\gamma_{h}$ and $\gamma_{1}$ the top and bottom of the tower, respectively, and refer to the right and left sides of the towers in terms of a walk from the bottom to the top. It follows that in a clockwise tower the internal edges are all directed from the right to the left, and vice-versa for counter-clockwise towers. Moreover, if $T=\left(\gamma_{1}, \ldots, \gamma_{h}\right)$ is a clockwise tower in $\mathcal{E}$ and $\mathcal{E}^{\prime}$ is obtained by reversing $C$, then $T^{\prime}=\left(\gamma_{h}, \gamma_{h-1}, \ldots, \gamma_{1}\right)$ is a counter-clockwise tower in $\mathcal{E}^{\prime}$, see, e.g., Figure 2.4.

Let $\mathcal{E} \in \alpha-\mathrm{O}(G)$ and suppose $\eta \in \mathcal{F}(G)$ is almost directed in $\mathcal{E}$. If there is a tower in $\mathcal{E}$ starting at $\eta$ we can find it by walking along the faces of $G$, starting at $\eta$ and choosing the face sharing the blocking edge with the current face at each step. If at any point we reach a directed face then we have found a tower. On the other hand, we can be certain there is no tower starting at $\eta$ if we reach a face $\gamma_{i}$ satisfying one of the following:

- $\gamma_{i}$ is not directed but not almost directed ${ }^{1}$;
- the blocking edge of $\gamma_{i}$ lies on the boundary of the graph;
- the blocking edge of $\gamma_{i}$ is the same edge as the blocking edge of $\gamma_{i-1}$.

To see that the process of forming a tower terminates (i.e., does not wrap around on itself), note that reaching some already explored face implies the existence of a cycle $C$ such that every edge joining $C$ to a vertex on its interior is directed towards $C$. But then none of these vertices can be involved in a directed cycle in $\mathcal{E}$, or, indeed, in any other $\alpha$-orientation of $G$. Hence, these edges must have the same orientation in every $\alpha$-orientation of $G$; that is, all edges joining $C$ to its interior are rigid. We

[^4]have assumed that all rigid edges have already been removed from $G$, and $\alpha$ has been adjusted accordingly, so this situation cannot occur. Furthermore, we observe that when $\alpha(v)=0, v$ cannot be incident with more than $\operatorname{deg}(v) / 2$ faces in any tower.

We now define the tower-moves chain, $\mathcal{M}_{\mathrm{T}}$, by describing how to obtain $X_{t+1}$ from $X_{t}$. The definition includes an undetermined probability $p_{T}$ which will be fixed later.

## One step of the Markov chain $\mathcal{M}_{\mathrm{T}}$

1. Choose $\operatorname{dir} \in\{c w, c c w\}$ and $\kappa \in \mathcal{F}(G)$ u.a.r.
2. If $\kappa$ is a dir-cycle then obtain $X_{t+1}$ from $X_{t}$ by reversing the orientation of all the edges in $\kappa$.
3. If there is a dir-tower $T=\left(\gamma_{i}\right)_{1 \leq i \leq h}$ with $\gamma_{1}=\kappa$ then let $C=\bigoplus_{1 \leq i \leq h} \gamma_{i}$. With probability $p_{T}$ obtain $X_{t+1}$ from $X_{t}$ by reversing all the edges of $C$.
4. Otherwise, set $X_{t+1}=X_{t}$

This type of chain has been used to extend the face reversal chain in the past, see [62, 38]. The irreducibility of this chain is inherited from the irreducibility of $\mathcal{M}_{\mathcal{F}}$, since every transition in $\mathcal{M}_{\mathcal{F}}$ is also a transition in $\mathcal{M}_{\mathrm{T}}$. Again, the fact they we choose the direction of reversal for each transition ensures that there is a selfloop probability of $1 / 2$. Hence, $\mathscr{M}_{\mathrm{T}}$ is aperiodic and so ergodic. As long as $p_{T}$ is chosen to be independent of whether $T$ is a clockwise or a counter-clockwise tower, $\mathcal{M}_{\mathrm{T}}$ converges to the uniform distribution. To see this suppose $\mathcal{E}$ can be obtained from $\hat{\mathcal{E}}$ by reversing a clockwise tower $T$. Then we can obtain $\hat{\mathcal{E}}$ from $\mathcal{E}$ by reversing a counter-clockwise tower containing the same faces as $T$. Hence, $P(\mathcal{E}, \hat{\mathcal{E}})=P(\hat{\mathcal{E}}, \mathcal{E})$ for any $\mathcal{E}, \hat{\mathcal{E}} \in \Omega$, and the stationary distribution of $\mathcal{M}_{\mathrm{T}}$ is uniform.

### 2.6 Rapid mixing on the triangular lattice

Up to now, we have not placed any restriction on $G$ or $\alpha$, beyond the fact that $G$ is a planar graph and $\alpha$ is a function assigning an integer to the vertices of $G$, such that $G$ has no $\alpha$-rigid edges. That is to say, we have shown that both $\mathcal{M}_{\mathcal{F}}$ and $\mathcal{M}_{\mathrm{T}}$ converge to the uniform distribution on the set of $\alpha$-orientations of any planar graph $G=(V, E)$, for any $\alpha: V \rightarrow \mathbb{N}$. In this section, we will prove that $\mathcal{M}_{\mathcal{F}}$ is rapidly mixing on the set of Eulerian orientations, possibly with some fixed boundary condition, of any solid subgraph $G$ of the triangular lattice. These are the $\alpha$-orientations of solid subgraph of the triangular lattice $G$, for some $\alpha$ with $\alpha(v)=0$ for every internal vertex $v$; the boundary condition determines the values for $\alpha(v)$ on boundary vertices. Hereafter,
we are assuming we have some fixed boundary condition, and use $\operatorname{EO}(G)$ to refer to the set of Eulerian orientations of $G$ with this fixed boundary condition.

First, in §2.6.1, we show that $\mathcal{M}_{\mathrm{T}}$ (or, more precisely, the lazy version of $\mathcal{M}_{\mathrm{T}}$ ) is rapidly mixing on the Eulerian orientations of any solid subgraph $G$ of the triangular lattice, possibly with some fixed boundary condition. Then, in §2.6.2, we use the comparison technique of Diaconis and Saloff-Coste [26] to obtain rapid mixing of $\mathcal{M}_{\mathcal{F}}$.

### 2.6.1 Rapid mixing of $\mathscr{M}_{\mathrm{T}}^{\star}$

In this section we use the path coupling technique of Bubley and Dyer [16] to analyse the mixing time of the Markov chain $\mathcal{M}_{\mathrm{T}}$ on $\mathrm{EO}(G)$ when $G$ is a solid subgraph of the triangular lattice with some fixed boundary condition, and $p_{T}$ is chosen appropriately; more precisely, we use the result of Bordewich and Dyer [13] to show that the lazy chain $\mathcal{M}_{\mathrm{T}}^{\star}$ (see Theorem 2.13) mixes rapidly.

This is not the first result regarding the mixing time of this type of chain. $\mathcal{M}_{\mathrm{T}}$ has been shown to be rapidly mixing on the square lattice (using different $p_{T}$ values to ones we shall use) for the case of a particular fixed boundary condition [62] and for the case of a free boundary [38]. Fehrenbach and Rüschendorf [31] attempted to give a proof of rapid mixing for a related Markov chain (on the triangular lattice) in which only towers of height 2 are used. However, the path coupling defined for the chain in [31] does not contract as claimed. In fact, it is possible to show that no one-step path coupling, using the natural distance metric, can prove rapid mixing for the chain in [31]. It could be that a one-step coupling will work with a more complicated metric [14], but this has not been attempted.

Rapid mixing proofs for $\mathscr{M}_{\mathrm{T}}$ are dependent on the correct choice of the probabilities $p_{T}$. For example, the proof of [38] sets $p_{T}$ to $1 / 4 h$ if the tower runs along the boundary, and $1 / 2 h$ otherwise, where $h$ is the height of the tower $T$. We now state the main result of this section but defer the proof until we have presented some useful lemmas.

Theorem 2.17. Let $G$ be a solid section of the triangular lattice with any fixed boundary condition. Then, $\mathcal{M}_{\mathrm{T}}^{\star}$, the lazy version of the Markov chain $\mathcal{M}_{\mathrm{T}}$, is rapidly mixing with mixing time $\tau_{M_{\mathrm{T}}^{\star}}(\varepsilon)$ satisfying

$$
\tau_{\mathscr{M}_{\mathrm{T}}^{*}}(\varepsilon) \in O\left(f^{4} \log \varepsilon^{-1}\right),
$$

when $p_{T}=1 / 3$ for all towers $T$ of height $h$.

Consequentially, the chain $\mathcal{M}_{\mathrm{T}}$ can be considered to mix in this time, for all practical purposes (see [13]). Note that, with $p_{T}=1 / 3 h$, the transition probability matrix of the chain, $P$, is given by

$$
P\left(\mathcal{E}, \mathcal{E}^{\prime}\right)= \begin{cases}\frac{1}{2 f} & \text { if } \mathcal{E} \oplus \mathcal{E}^{\prime} \text { is a single face } \\ \frac{1}{6 f h} & \text { if } \mathcal{E} \oplus \mathcal{E}^{\prime} \text { is a single tower of height } h \\ 1-\sum_{\mathcal{E}^{\prime \prime} \neq \mathcal{E}} P\left(\mathcal{E}, \mathcal{E}^{\prime \prime}\right) & \text { if } \mathcal{E}=\mathcal{E}^{\prime} \\ 0 & \text { otherwise }\end{cases}
$$

We first bound the maximum distance between any pair of $\alpha$-orientations of the triangular lattice.

Lemma 2.18. Let $G$ be any solid subgraph of the triangular lattice with $f$ bounded faces. The maximum distance between any pair of $\alpha$-orientations of $G$ in the Felsner lattice is $O\left(f^{3 / 2}\right)$.

Proof. The maximum distance between any two vertices in the cover graph of the Felsner lattice is equal to the distance between the maximum and minimum elements. Thus, the problem reduces to bounding $\delta\left(\mathcal{E}_{\text {max }}, \mathcal{E}_{\text {min }}\right)$.

We write $\wp_{\text {max }}$ for $\wp_{\mathcal{E}_{\text {max }}}$, the potential function corresponding to the maximum element of $\operatorname{Fels}(G)$. Then, from the definition of the bijection between $\alpha$-orientations and $\alpha$-potentials given in $\S 2.3$ we can conclude that the distance between the maximum and minimum orientations is $\sum_{\gamma \in \mathcal{F}(G)} \wp_{\max }(\gamma)$.

Conditions (2.4) and (2.5) of the definition of $\alpha$-potentials imply that $\wp_{\max }(\gamma)$ is exactly the minimum number of edges in a shortest path in the dual graph of $G$ from $\gamma$ to the unbounded face. Let $G_{k}$ be the smallest graph which contains a face $\gamma$ with $\wp_{\max }(\gamma)=k$. We can construct $G_{k}$ inductively, starting with $G_{1}=K_{3}$. To extend $G_{k}$ to $G_{k+1}$ we add a triangular face onto each edge of the boundary of $G_{k}$; see Figure 2.5 for an example of what these look like. Let $\gamma$ be the single face which is contained in $G_{1}$. Every path in the dual graph of $G_{k}$ from $\gamma$ to the unbounded face has length $k$, so $\wp_{\max }(\gamma)=k$ in the maximum $\alpha$-potential of $G_{k}$. Moreover, removing any boundary face will introduce a path (in the dual graph) from the $\gamma$ to the unbounded face of length $k-1$, so any graph smaller than $G_{k}$ will not have any face which has value $k$ in the maximum $\alpha$-potential.

A simple inductive argument shows that the number of faces added at each step is $3 k$. This implies $\left|\mathcal{F}\left(G_{k}\right)\right| \in \Theta\left(k^{2}\right)$, so $\wp_{\max }(\gamma) \in O(\sqrt{f})$ for any face $\gamma$.


Figure 2.5: The graph $G_{3}$ from Lemma 2.18. The edges from $G_{1}$ and $G_{2} \backslash G_{1}$ are marked by dashed and dotted lines respectively

Let $S$ be the set of pairs of orientations at distance 1 in the Felsner lattice. We want to define a coupling $\mathcal{C}$ for every pair $(\mathcal{E}, \hat{\mathcal{E}}) \in S$. Suppose $(\mathcal{E}, \hat{\mathcal{E}}) \in S$, let $\gamma$ be the face on which they disagree, and let $N(\gamma)$ denote the set of faces which share an edge with $\gamma$. Suppose $\left(X_{t}, Y_{t}\right)=(\mathcal{E}, \hat{\mathcal{E}})$. Our path coupling $\mathcal{C}$ for $\mathcal{M}_{\mathrm{T}}$ chooses the same face $\kappa$ in both chains and couples the transitions as follows:

- if $\kappa=\gamma$ then reverse $\kappa$ in $\mathcal{E}$ but not $\hat{\mathcal{E}}$ with probability $1 / 2$, and reverse $\kappa$ in $\hat{\mathcal{E}}$ but not in $\mathcal{E}$ with probability $1 / 2$;
- if the same transition is available for the choice of $\kappa$ in both $\mathcal{E}$ and $\hat{\mathcal{E}}$ then apply this change to both orientations with the relevant probability and leave both unchanged otherwise;
- if different transitions are available for $\kappa$ in both orientations (with the transition in $\mathcal{E}$ happens with higher probability) then apply both transitions with the lower probability, apply the transition in $\mathcal{E}$ alone with its remaining probability, and leave both orientations unchanged otherwise.

In order to apply Bordewich and Dyer's path coupling theorem (Theorem 2.13) we need to show that the expected distance between the two copies of the chain does not increase after a single step of the coupling. To do this we need to consider which choice of faces will cause the distance to increase, which choices will leave the distance unchanged, and which choices will cause the distance to decrease.

Definition 2.19. Suppose $(\mathcal{E}, \hat{\mathcal{E}})$ is the current state, $\left(X_{t}, Y_{t}\right)$, of the coupling $\mathcal{C}$. We say the move resulting from choosing $\kappa$ involves $\sigma \in N(\gamma)$ if $\kappa=\sigma$ or there is a tower (in either $\mathcal{E}$ or $\hat{\mathcal{E}}$ ) starting at $\kappa$ that contains $\sigma$.

The distance between the coupled chains can change if and only if $\kappa=\gamma$ or the move at $\kappa$ involves some $\sigma \in N(\gamma)$. Note that no move can involve more than one element of $N(\gamma)$, so, in the analysis of our path coupling, we can treat each $\sigma \in N(\gamma)$ separately. In the following analysis, for each $\sigma \in N(\gamma)$, we use $\delta_{\sigma}$ to denote the contribution to the change to the distance between the two Eulerian orientations after a single step of the coupling resulting from moves involving the face $\sigma$. We will now analyse the expected value of $\delta_{\sigma}$, conditioned on whether $\sigma$ is a directed face in $\mathcal{E}$ or $\hat{\mathcal{E}}$, or is directed in neither.

Lemma 2.20. Suppose $\sigma \in N(\gamma)$ is a directed face in $\mathcal{E}$ or $\hat{\mathcal{E}}$. Then $\mathbb{E}\left[\delta_{\sigma}\right]=\frac{1}{3 f}$.
Proof. We assume that $\sigma$ is directed in $\mathcal{E}$; the case when $\sigma$ is directed in $\hat{\mathcal{E}}$ is symmetric. Let $\kappa \in \mathcal{F}(G)$ such that selecting $\kappa$ gives a move which involves $\sigma$ in at least one of the coupled chains. We have two cases to consider:

Case $\kappa=\sigma$ : Since $\sigma$ is a neighbour of $\gamma$, and $\sigma$ is directed in $\mathcal{E}$, it follows that the blocking edge of $\sigma$ in $\hat{\mathcal{E}}$ is the edge shared with $\gamma$. Then $T=\{\sigma, \gamma\}$ is a tower of height 2 in $\hat{\mathcal{E}}$ with a reversal probability of $\frac{1}{12 f}$. The coupling reverses $\sigma$ in $\mathcal{E}$ and $T$ in $\hat{\mathcal{E}}$ with probability $\frac{1}{12 f}$, and reverses $\sigma$ in $\mathcal{E}$ but leaves $\hat{\mathcal{E}}$ unchanged with probability $\frac{5}{12 f}$. The former results in coalescence, whereas the latter yields a pair of orientations which are distance 2 apart. Hence, the contribution to $\mathbb{E}\left[\delta_{\sigma}\right]$ from this case is

$$
\frac{5}{12 f}-\frac{1}{12 f}=\frac{1}{3 f} .
$$

Case $\kappa \neq \sigma$ : Since $\gamma$ and $\sigma$ are both directed in $\mathcal{E}$ it follows that there must be a tower $T_{1}$ starting at $\kappa$ and ending at $\sigma$ in $\mathcal{E}$ that does not contain $\gamma$, and a tower $T_{2}=T_{1} \cup\{\gamma\}$ starting at $\kappa$ in $\hat{\mathcal{E}}$. Let $h$ be the height of $T_{1}$, so $T_{1}$ is reversed in $\mathcal{E}$ with probability $\frac{1}{6 f h}$ and $T_{2}$ is reversed in $\hat{\mathcal{E}}$ with probability $\frac{1}{6 f(h+1)}$. Observe that if we reverse $T_{1}$ in $\mathcal{E}$ we obtain an orientation which is distance $h+1$ from $\hat{\mathcal{E}}$, but if we also reverse $T_{2}$ in $\hat{\mathcal{E}}$ then we have the same orientation in both chains. Therefore, the contribution to $\mathbb{E}\left[\delta_{\sigma}\right]$ from this case is

$$
h\left(\frac{1}{6 f h}-\frac{1}{6 f(h+1)}\right)-\frac{1}{6 f(h+1)}=0 .
$$

Thus, the only face whose selection will result in a move which has non-zero contribution to $\mathbb{E}\left[\delta_{\sigma}\right]$ is $\sigma$ itself, so $\mathbb{E}\left[\delta_{\sigma}\right]=1 / 3 f$.

Lemma 2.21. If $\sigma \in N(\gamma)$ is not directed in $\mathcal{E}$ or $\hat{\mathcal{E}}$ then $\mathbb{E}\left[\delta_{\sigma}\right]$ is no more than $\frac{1}{3 f}$.

Proof. Observe that, since $\sigma$ is not directed in $\mathcal{E}$ or $\hat{\mathcal{E}}$, the blocking edge of $\sigma$ will not be shared with $\gamma$ in either orientation and the blocking edge of $\sigma$ is different in both orientations. Hence, no tower can include both $\sigma$ and $\gamma$. Assuming there is a tower containing $\sigma$ in at least one of the two orientations we have two disjoint cases to consider. Suppose $\mathcal{E}$ has a tower starting at some $\eta \in N(\sigma) \backslash\{\gamma\}$. Then there will be no tower containing $\sigma$ in $\hat{\mathcal{E}}$ as the procedure for constructing towers described in §2.3 is guaranteed to reach a pair of consecutive faces sharing the same blocking edge ( $\sigma$ and $\eta$ ). Thus, when we are in this situation we can assume that one of the orientations will be unchanged after one step of the coupling. In the second case, when there is no tower starting at any $\eta \in N(\sigma) \backslash\{\gamma\}$ in $\mathcal{E}$ or $\hat{\mathcal{E}}$, there may be a tower starting at $\sigma$ in either orientation.


Figure 2.6: Example from Lemma 2.21
$\exists$ a tower starting at $\eta \in N(\sigma) \backslash \gamma$ in $\mathcal{E}$ or $\hat{\mathcal{E}}$ : We can assume, without loss of generality, that $\mathcal{E}$ is the orientation with a tower starting at some $\eta \in N(\sigma) \backslash \gamma$. Since no move involving $\sigma$ is possible in $\hat{\mathcal{E}}$ we only need to bound the expected distance between $\mathcal{E}$ and all $\mathcal{E}^{\prime}$ which can be obtained by making a move involving $\sigma$ in $\mathcal{E}$.

We begin by showing that any tower containing $\sigma$ in $\mathcal{E}$ must start at $\sigma$ or a neighbour of $\sigma$. To see this suppose we have a tower in $\mathcal{E}$ containing $\rho, \eta$, and $\sigma$ where $\eta$ is a neighbour of $\sigma$ and $\rho \in N(\eta) \backslash\{\sigma\}$. Let $u, v, w$ be the vertices of $\sigma$, and suppose that the edges of $\sigma$ are oriented $(u, v),(u, w)$, and $(w, v)$ and that $\{u, w\}$ is the edge shared between $\gamma$ and $\sigma$. Then $\sigma$ must share $\{v, w\}$ with $\eta$. Recall that no vertex can be incident with more then 3 faces in any tower. Therefore, $v$ cannot belong to the edge shared between $\rho$ and $\eta$, so $\rho$ must contain $w$ (as illustrated in Figure 2.6). To satisfy the definition of a tower, both edges of $\rho$ incident with $w$ must be oriented away from $w$. But this implies that there are 4 edges oriented away from $w$ in $\mathcal{E}$, a contradiction (e.g., see Figure 2.6). An identical argument holds if the edges of $\sigma$ are
oriented $(v, u),(u, w)$, and $(v, w)$. Hence, any tower containing $\sigma$ must start at $\eta$ or $\sigma$.
Let $h$ be the height of the tower starting at $\sigma$. Then, the move in which $\sigma$ is the chosen face is made with probability $\frac{1}{6 \mathrm{fh}}$, and the move in which $\eta$ is the chosen face is made with probability $\frac{1}{6 f(h+1)}$ (if it exists). Since these moves increase the distance by $h$ and $h+1$, respectively, we have

$$
\mathbb{E}\left[\delta_{\sigma}\right] \leq \frac{h}{6 f h}+\frac{h+1}{6 f(h+1)}=\frac{1}{3 f} .
$$

$\nexists$ a tower starting at $\eta \in N(\sigma) \backslash \gamma$ in $\mathcal{E}$ or $\hat{\mathcal{E}}$ : In the worst case, we could have a tower starting at $\sigma$ in both orientations. If this occurs, then the two towers will have opposite orientations in $\mathcal{E}$ and $\hat{\mathcal{E}}$, and each will be the largest tower containing $\sigma$ in the orientation. Let $T_{1}$ and $T_{2}$ denote the towers in $\mathcal{E}$ and $\hat{\mathcal{E}}$, respectively, and let $h_{1}$ and $h_{2}$ be the heights of each tower. If $h_{1} \leq h_{2}$ then the coupling reverses the towers in $\mathcal{E}$ and $\hat{\mathcal{E}}$ with probability $\frac{1}{6 f h_{2}}$, and reverses the tower in $\mathcal{E}$ alone with probability $\frac{1}{6 f h_{1}}-\frac{1}{6 f h_{2}}$, conditional on $\sigma$ being the chosen face. The first situation yields a pair orientations which are distance $h_{1}+h_{2}+1$ apart, and the second a pair orientations which are $h_{1}+1$ apart. Hence,

$$
\mathbb{E}\left[\delta_{\sigma}\right] \leq\left(h_{1}+h_{2}\right) \frac{1}{6 f h_{2}}+h_{1}\left(\frac{1}{6 f h_{1}}-\frac{1}{6 f h_{2}}\right)=\frac{1}{3 f} .
$$

The analysis is identical if $h_{2} \leq h_{1}$.
Since two cases are disjoint, we see that $\mathbb{E}\left[\delta_{\sigma}\right]$ is no more than $\frac{1}{3 f}$.
We now apply Theorem 2.13 to obtain a bound on the mixing time.
Proof of Theorem 2.17. We use the coupling defined earlier and analysed in Lemma 2.20 and Lemma 2.21. With probability $1 / f, \gamma$ is reversed in one or the other of the two orientations, causing the two chains to coalesce. Combining this fact with the results of Lemmas 2.20 and 2.21 we find that for all $\mathcal{E}$ and $\hat{\mathcal{E}}$ differing on the orientation of a single face

$$
\mathbb{E}\left[\delta\left(X_{t+1}, Y_{t+1}\right)-\delta\left(X_{t}, Y_{t}\right) \mid\left(X_{t}, Y_{t}\right)=(\mathcal{E}, \hat{\mathbb{E}})\right] \leq 3 \frac{1}{3 f}-\frac{1}{f}=0,
$$

where $\left(X_{t}, Y_{t}\right)_{t \geq 1}$ is the coupling $\mathcal{C}$ defined on $S$, the set of pairs of Eulerian orientations which are adjacent in the cover graph of the Felsner lattice. The transition probability for any pair of states in $S$ is $1 / 2 f$. Moreover, since we have just shown that $\beta=1$ for the path coupling $\mathcal{C}$, we can apply Theorem 2.13 with $p=1 / 2 f$ and $D \in O\left(f^{\frac{3}{2}}\right)$ (from Lemma 2.18) to obtain that the lazy version of the chain, $\mathcal{M}_{\mathrm{T}}^{\star}$, mixes in time $\tau(\varepsilon) \in O\left(f^{4} \log \varepsilon^{-1}\right)$.

### 2.6.2 Rapid mixing of $\mathscr{M}_{\mathcal{F}}$

In the previous section, we applied Theorem 2.13 to show that the lazy version of the tower chain, $\mathcal{M}_{\mathrm{T}}^{\star}$, is rapidly mixing. Since the smallest transition probability of any pair in the set $S$ used to define the path coupling is $1 / 2 f$, it follows that the transition probability matrix of this chain, $P^{\star}$, is given by

$$
P^{\star}\left(\mathcal{E}, \mathcal{E}^{\prime}\right)= \begin{cases}\frac{1}{2 f+1} & \text { if } \mathcal{E} \oplus \mathcal{E}^{\prime}=\text { is a single face } \\ \frac{1}{3(2 f+1) h} & \text { if } \mathcal{E} \oplus \mathcal{E}^{\prime} \text { is a single tower of height } h \\ 1-\sum_{\mathcal{E}^{\prime \prime} \neq \mathcal{E}} P^{\star}\left(\mathcal{E}, \mathcal{E}^{\prime \prime}\right) & \text { if } \mathcal{E}=\mathcal{E}^{\prime} \\ 0 & \text { otherwise }\end{cases}
$$

Although this chain has different transition probabilities, it also converges to the uniform distribution on $\Omega=\mathrm{EO}(G)$. Hence, we can apply the comparison method of Diaconis and Saloff-Coste [26] to obtain a bound on the mixing time of $\mathcal{M}_{\mathcal{F}}$.

## The Comparison Method

If we cannot find a bound on the mixing time of a Markov chain $\mathcal{M}$ by direct analysis of $\mathcal{M}$, but can find a bound on the mixing time of another chain $\widetilde{\mathcal{M}}$, which has the same state space as $\mathfrak{M}$, we can often use the so-called comparison theorem of Diaconis and Saloff-Coste [26] to obtain a bound on the mixing time of $\mathcal{M}$. We will use the formulation of the Diaconis and Saloff-Coste result from [73], restated here for convenience. Note that we are using $E(P)$ to denote the set of edges corresponding to moves between adjacent states in the Markov chain with transition matrix $P$.

Theorem 2.22. ( [73, Proposition 4]) Suppose $P$ and $\widetilde{P}$ are the transition matrices of two reversible Markov chains, $\mathcal{M}$ and $\widetilde{\mathcal{M}}$, both with the state space $\Omega$ and stationary distribution $\pi$, and let $\pi_{\star}=\min _{x \in \Omega} \pi(x)$. For each pair $(u, v) \in E(\widetilde{P})$, define a path $\gamma_{u v}$ which is a sequence of states $u=u_{0}, u_{1},, \ldots, u_{k}=v$ with $\left(u_{i}, u_{i+1}\right) \in E(P)$ for all $i$. For $(x, y) \in E(P)$, let $\Gamma(x, y)=\left\{(u, v) \in E(\widetilde{P}):(x, y) \in \gamma_{u v}\right\}$. Let

$$
A=\max _{(x, y) \in E(P)}\left\{\frac{1}{\pi(x) P(x, y)} \sum_{(u, v) \in \Gamma(x, y)}\left|\gamma_{u v}\right| \pi(u) \widetilde{P}(u, v)\right\} .
$$

Suppose that the second largest eigenvalue, $\lambda_{1}$, of $\widetilde{P}$ satisfies $\lambda_{1} \geq 1 / 2$. Then for any $0<\varepsilon<1$

$$
\tau_{\mathscr{M}}(\varepsilon) \in O\left(A \tau_{\widetilde{\mathcal{M}}}(\varepsilon) \log 1 / \pi_{\star}\right)
$$

To obtain a polynomial bound on the mixing time of $\mathcal{M}$, we want to show that the simulation of the transitions of $\mathcal{M}$ by transitions of $\mathcal{M}$ does not overload any edge in
$E(P)$. That is, we want the value of $A$ in Theorem 2.22 to be bounded by a polynomial in the size of the elements of $\Omega$.

## Comparing $\mathcal{M}_{\mathcal{F}}$ and $\mathscr{M}_{\mathrm{T}}^{\star}$

We first show that every move of $\mathcal{M}_{\mathrm{T}}^{\star}$ can be simulated by moves of the chain $\mathcal{M}_{\mathcal{F}}$. Suppose $T=\left\{\gamma_{1}, \gamma_{2}, \ldots, \gamma_{h}\right\}$ is a tower in $\mathcal{E}$ and that $\hat{\mathcal{E}}$ is the orientation obtained by reversing $T$. Observe that by the definition of a tower $\gamma_{h}$ is a directed cycle in $\mathcal{E}$. Then we can apply the transition in $\mathcal{M}_{\mathcal{F}}$ that reverses $\gamma_{h}$ to $\mathcal{E}$ and obtain a new Eulerian orientation $\mathcal{E}^{\prime \prime}$. But there is now a tower $T^{\prime}=\left\{\gamma_{1}, \gamma_{2}, \ldots, \gamma_{h-1}\right\}$ in $\mathcal{E}^{\prime}$ which can be reversed to obtain $\hat{\mathcal{E}}$. Repeating this process until we reach $\hat{\mathcal{E}}$ gives a decomposition of the tower move reversing $T$ in $\mathcal{E}$ into moves of the chain $\mathcal{M}_{\mathcal{F}}$. Hence, we are able to apply the comparison method (Theorem 2.22) to obtain a bound on the mixing time of $\mathcal{M}_{\mathcal{F}}$.

We are now ready to prove our rapid mixing result for $\mathcal{M}_{\mathcal{F}}$. We bound the mixing time in terms of $f$, the number of unbounded faces, and $h$, the height of the largest tower. Then, we derive mixing times for several different cases by considering the values $h$ can take.

Theorem 2.23. Suppose $G$ is a solid subgraph of the triangular lattice and let $h$ be the maximum height of any tower in an Eulerian orientation of $G$ (possibly with respect to some fixed boundary condition). Then the mixing time of the face-reversal Markov chain $\mathcal{M}_{\mathcal{F}}$ satisfies

$$
\tau_{\mathcal{M}_{\mathcal{F}}}(\varepsilon) \in O\left(h^{2} f^{5} \log \varepsilon^{-1}\right) .
$$

Proof. Note that $\mathcal{M}_{\mathcal{F}}$ and $\mathcal{M}_{\mathrm{T}}^{\star}$ both have a self-loop probability of $1 / 2$, ensuring that the second-largest eigenvalue of either chain will be at least $1 / 2$, so Theorem 2.22 can be applied.

Let $P$ and $P^{\star}$ denote the transition matrices of $\mathcal{M}_{\mathcal{F}}$ and $\mathcal{M}_{\mathrm{T}}^{\star}$. For each pair of states $(x, y)$ that differ on the orientation of exactly one face (i.e., each $(x, y) \in E(P)$ ), we define $\Gamma(x, y)$ to be the set of all transitions in $\mathcal{M}_{\mathrm{T}}^{\star}$ containing the transition $t=(x, y)$ as a sub-move. Each pair $(u, v) \in \Gamma(x, y)$ corresponds to a pair of orientations which
differ on exactly one tower; let $\gamma_{u v}$ denote this tower. For each $(x, y) \in E(P)$ we have

$$
\begin{align*}
A_{x, y} & =\frac{1}{\pi(x) P(x, y)} \sum_{(u, v) \in \Gamma(x, y)}\left|\gamma_{u v}\right| \pi(u) P^{\star}(u, v) \\
& =2 f \sum_{(u, v) \in \Gamma(x, y)}\left|\gamma_{u v}\right| P^{\star}(u, v),  \tag{2.7}\\
& \leq 1+\frac{1}{3}(|\Gamma(x, y)|-1), \tag{2.8}
\end{align*}
$$

where (2.7) is due to the fact that all transition probabilities in $P$ are $\frac{1}{2 f}$ and that $\pi$ is uniform, and (2.8) is due to the fact that $P^{\star}(u, v)=\frac{1}{2 f+1}$ if $(u, v)=(x, y)$, and $P^{\star}(u, v)=$ $\frac{1}{3(2 f+1)\left|\gamma_{u v}\right|}$ if $(u, v)$ is the reversal of a tower.

We will now find an upper bound for $|\Gamma(x, y)|$. Let $\gamma$ be the face that is reversed in the transition $t=(x, y)$. We need to consider the different cases in which $t$ can feature as part of the decomposition of a tower move $(u, v) \in E\left(P^{\star}\right)$. Observe that there are three different directions in which a tower can pass through $\gamma$ and contain $t$ as a submove (one for each pair of edges of $\gamma$ ). Let $h_{i}$ be the height of the maximum length tower passing through $\gamma$ in direction $i$ (over all orientations $\mathcal{E} \in \Omega$ ), let $\left(\gamma_{j}\right)_{1 \leq j \leq h_{j}}$ denote the sequence of faces that make up this tower, and suppose $\gamma_{k}=\gamma$. Any other tower which passes through $\gamma$ in this direction and whose encoding uses $t$ must be subtower of the maximum length tower. Moreover, the bottom of any such tower must be contained in $\left\{\gamma_{j}: 1 \leq j<k\right\}$ and the top of any such tower must be contained in $\left\{\gamma_{j}: k \leq j \leq h_{i}\right\}$. To see this, observe that $x$ is the orientation obtained by reversing the tower $\left(\gamma_{k+1}, \gamma_{k+2}, \ldots, \gamma_{h_{i}}\right)$ (in the orientation containing the maximum length tower). Hence, reversing the tower $\left(\gamma_{j}, \gamma_{j+1}, \ldots, \gamma_{h_{i}}\right)$, for any $j$ satisfying $k<j \leq h_{i}$ will yield an orientation containing a tower whose encoding uses $t$. Moreover, each of these orientations has a tower whose encoding uses $t$ for each element of $\left\{\gamma_{1}, \ldots, \gamma_{k-1}\right\}$. Thus, the number of towers using $t$ that pass through $\gamma$ in direction $i$ is $(k-1)\left(h_{i}-k+\right.$ 1 ), which could be as large as $h_{i}^{2} / 4$. Hence, we have

$$
|\Gamma(x, y)| \leq O\left(h_{0}^{2}\right)+O\left(h_{1}^{2}\right)+O\left(h_{2}^{2}\right)=O\left(h^{2}\right) .
$$

Thus, we have $A_{x, y} \in O\left(h^{2}\right)$ for all $(x, y) \in E(P)$. Finally, we need to find an upper bound on the value $1 / \pi_{\star}$, where $\pi_{\star}$ is the minimum value of the stationary distribution of $\mathscr{M}_{\mathcal{F}}$. Since this is the uniform distribution, $1 / \pi$ has the same value on all $x \in \Omega$ : $|\Omega|$. The number of edges in $G$ is no more than $3 f$, so $2^{3 f}$ provides an upper bound on the number of orientations of $G$, and so also on $1 / \pi_{\star}$. Combining all this with Theorems 2.22 and 2.17 we get

$$
\tau_{\mathcal{M}_{\mathcal{F}}}(\varepsilon) \in O\left(h^{2} f^{5} \log \varepsilon^{-1}\right) .
$$

As a corollary to Theorem 2.23, we can obtain a bound on the mixing time of $\mathcal{M}_{\mathcal{F}}$ for any solid subgraph of the triangular lattice and arbitrary boundary condition (Corollary 2.24). However, this bound is overly pessimistic and we can improve it by a factor of $f$ for two important special cases (Corollary 2.25 and Corollary 2.26).

Corollary 2.24. Let $G$ be a solid subgraph of the triangular lattice and let $\alpha$ be an arbitrary boundary condition. Then the mixing time of $\mathcal{M}_{\mathcal{F}}$ satisfies

$$
\tau_{\mathcal{M}_{\mathcal{F}}}(\varepsilon) \in O\left(f^{7} \log \varepsilon^{-1}\right)
$$

Proof. In general, when we consider an arbitrary solid subgraph of the triangular lattice with an arbitrary fixed boundary condition, we can have towers of height $O(f)$ and so, in this case, Theorem 2.23 gives a bound on the mixing time of

$$
\tau_{\mathcal{M}_{\mathcal{F}}}(\varepsilon) \in O\left(f^{7} \log \varepsilon^{-1}\right) .
$$

Corollary 2.25. Suppose $G$ is an Eulerian solid subgraph of the triangular lattice. Then, the mixing time of $\mathcal{M}_{\mathcal{F}}$ satisfies

$$
\tau_{\mathcal{M}_{\mathcal{F}}}(\varepsilon) \in O\left(f^{6} \log \varepsilon^{-1}\right)
$$

Proof. We need to bound the height of any tower in an Eulerian orientation of $G$, where $G$ is a solid subgraph of the triangular lattice.

Let $T=\left(\gamma_{1}, \gamma_{2}, \ldots, \gamma_{h}\right)$ denote a tower in an Eulerian orientation of $G$. We assume that $T$ is a clockwise tower; the argument is identical for counter-clockwise towers. Recall that all the internal edges of a clockwise tower are directed towards the vertices on the left, so any Eulerian orientation of $G$ must contain a set of $h-1$ edge-disjoint directed paths linking the left-side vertices to the right-side vertices. Each of these paths must go around the top or the bottom of the tower. But each path that goes around the bottom (resp. top) contributes 1 to the distance from $\gamma_{1}$ (resp. $\gamma_{h}$ ) to the boundary. Hence, by (2.4), (2.5) and (2.6), $\max \left(\wp_{\max }\left(\gamma_{1}\right), \wp_{\max }\left(\gamma_{h}\right)\right) \geq h / 2$. But $\wp_{\max }(\gamma) \in O(\sqrt{f})$ for any $\gamma$ (see proof of Lemma 2.18), whence $h \in O(\sqrt{f})$.

Corollary 2.26. Let $G$ be the subgraph of triangular lattice induced by the set of vertices

$$
\{(i, j): 0 \leq i<n, 0 \leq j<n\} .
$$

Then, the mixing time of $\mathfrak{M}_{\mathcal{F}}$ satisfies

$$
\tau_{\mathscr{M}_{\mathcal{F}}}(\varepsilon) \in O\left(f^{6} \log \varepsilon^{-1}\right)
$$

Proof. Recall that no vertex can be incident with more than 3 faces in a tower in an orientation of $G$. That is, towers can only extend in straight lines in the natural embedding of $G$ in the plane. Thus, the longest sequence of any faces which can form a tower is $n$, which is equal to $\sqrt{f}$.

### 2.7 Approximating \#EO $(G)$

Suppose $G$ is a solid subgraph of the triangular lattice and $\alpha$ some fixed boundary condition. If we choose an edge $e=\{u, v\}$ from the unbounded face for the above reduction, then the resulting graph will be a solid subgraph of the triangular lattice and the pair of functions $\alpha_{(u, v)}$ and $\alpha_{(v, u)}$ will both correspond to boundary conditions on $G-e$. In other words, the problem of counting Eulerian orientations with fixed boundary conditions is self-reducible (Definition 1.35) for solid subgraphs of the triangular lattice. Thus, we can use our rapidly mixing Markov chain with Algorithm 2 to approximate the number of Eulerian orientations with fixed boundary condition of any solid subgraph of the triangular lattice.

We now sketch how the algorithm operates. We construct a sequence of graphs $G_{1}, \ldots, G_{m}$, where $m$ is the number of edges, defined inductively as

$$
\begin{aligned}
G_{1} & =G \\
G_{i} & =G_{i-1}-e_{i-1} \text { for } i \geq 2,
\end{aligned}
$$

where $e_{i}$ is an edge from the boundary face of $G_{i}$ for each $i=1,2, \ldots, m-1$, and a sequence of boundary conditions

$$
\begin{aligned}
& \alpha_{1}=\alpha \\
& \alpha_{i}=\left(\alpha_{i-1}\right)_{a_{i}} \text { for } i \geq 2,
\end{aligned}
$$

where $a_{i}$ is an orientation of the edge $e_{i}$. For $i=1,2, \ldots, m-1$, we estimate

$$
\rho_{i}=\frac{\# \alpha_{i}-\mathrm{O}\left(G_{i+1}\right)}{\# \alpha_{i}-\mathrm{O}\left(G_{i}\right)},
$$

by generating random $\alpha_{i}$-orientations of $G_{i}$ and counting how many can be extended to $\alpha_{i+1}$-orientations of $G_{i+1}$. The value returned by Algorithm 2 is an approximation to

$$
\# \alpha-\mathrm{O}(G)=\left(\prod_{i=1}^{m-1} \rho_{i}\right)^{-1} .
$$

### 2.8 Torpid mixing

Given the small collection of positive results regarding the mixing time of the facereversal chain $\mathcal{M}_{\mathcal{F}}$ (Theorem 2.23 and [62, 38]) and given that the reduction of [69] allows us to sample from $\operatorname{EO}(G)$ in polynomial time for any graph, one might hope that $\mathcal{M}_{\mathcal{F}}$ would be rapidly mixing on the set of Eulerian orientations of any planar graph. In fact, this is not true and in this section we exhibit a family of planar graphs for which $\mathcal{M}_{\mathcal{F}}$ is torpidly mixing. Consider the family of graphs $H_{N}$, of which $H_{2}$ is shown in Figure 2.7. Formally, $H_{N}$ is a graph with vertex set

$$
V=\left\{v_{i}: 1 \leq i \leq 12 N\right\} \cup\left\{u_{i}: 1 \leq i \leq 12 N+6\right\} \cup\left\{w_{i}: 1 \leq i \leq 6 N\right\} .
$$

The edges of $H_{N}$ consists of the disjoint union of three large cycles:

$$
\begin{aligned}
& E_{1}=\left(v_{1}, \ldots, v_{12 N}, v_{1}\right) \\
& E_{2}=\left(u_{1}, \ldots, u_{12 N+6}, u_{1}\right) \\
& E_{3}=\left(v_{1}, u_{2}, w_{1}, u_{3}, v_{3}, \ldots, u_{12 N+6}, v_{1}\right) .
\end{aligned}
$$

It is the large face in the centre of each of these graphs that creates the bottleneck in the Markov chain we will use to show torpid mixing. We label this face $C$ and its neighbours $\eta_{i}$ (for $1 \leq i \leq 6 N$ ). The face that is adjacent to both $\eta_{i}$ and $\eta_{i+1}$ is labelled $\sigma_{i}$, and the face that is only adjacent to $\eta_{i}$ is labelled $\rho_{i}$. Note that there are $6 N \sigma_{i}$ faces and $6 N \rho_{i}$ faces.

The reason $C$ causes torpid mixing is that we are able to partition the state space of the chain into two parts $S$ and $\bar{S}$, such that the only transitions crossing from $S$ into $\bar{S}$ are transitions reversing $C$. Because $C$ has so many edges (linear in the number of vertices) we are able to show that the set of orientations in which $C$ is directed is exponentially small. Hence, by Theorem 1.51, the face-reversal chain will be torpidly mixing.

Theorem 2.27. The face-reversal chain $\mathcal{M}_{\mathcal{F}}$ is torpidly mixing on $\mathrm{EO}\left(H_{N}\right)$ for $N \geq 3$.
Proof. From Theorem 1.51 we know that $\mathcal{M}_{\mathcal{F}}$ is torpidly mixing on a set of Eulerian orientations $\Omega$ if there exists some $S \subset \Omega$, with $0<|S| \leq|\Omega| / 2$, such that $|\partial S| /|S|$ is exponentially small in $f$.

Recall the definition of $\alpha$-potentials and the Felsner lattice from §2.3; in particular, recall that for any $\gamma \in \mathcal{F}(G)$ the value $\wp_{\mathcal{E}}(\gamma)$ is equal to the number of times $\gamma$ is reversed on any path from $\mathcal{E}_{\text {min }}$, the unique minimum element of the Felsner lattice,


Figure 2.7: The graph $\mathrm{H}_{2}$
to $\mathcal{E}$. Let $\wp_{\max }=\wp_{\mathcal{E}_{\max }}$, where $\mathcal{E}_{\max }$ is the unique maximum element of the Felsner lattice.

Let $S$ be the set of all Eulerian orientations $\mathcal{E}$ on $H_{N}$ satisfying $\wp_{\mathcal{E}}(C) \leq 1$. From the conditions (2.4), (2.5), and (2.6), we can deduce that the value of $\wp_{\max }(C)$ is 3 . Hence, we can define a bijection between $S$ and $\Omega \backslash S$ by mapping $\wp_{\mathcal{E}}$ to $\wp_{\max }-\wp_{\mathcal{E}}$ for each $\mathcal{E} \in S$ (this corresponds to mapping $\mathcal{E}$ to the orientation with all edges reversed), so $|S|=|\Omega| / 2$.

An Eulerian orientation $\mathcal{E}$ is an element of $\partial S$ if and only if $C$ is a counter-clockwise directed cycle in $\mathcal{E}$ and $\wp_{\mathcal{E}}(C)=1$. For this to occur we must have $\wp_{\mathcal{E}}\left(\eta_{i}\right)=1$ for each $\eta_{i}$. Hence, the number of Eulerian orientations satisfying this condition is exactly $2^{2 k}$ since each of the $\sigma_{i}$ and $\rho_{i}$ can take potential value 0 or 1 , where $k=6 \mathrm{~N}$.

We note that $C$ is the only directed cycle in $\mathcal{E}_{\text {min }}$ so $|S|=\left|S^{\prime}\right|+1$, where

$$
S^{\prime}=\left\{\mathcal{E} \in \Omega: \wp_{\mathcal{E}}(C)=1\right\}
$$

We can partition $S^{\prime}$ as $\bigcup_{I \subset[k]} S_{I}$, where

$$
S_{I}=\left\{\mathcal{E} \in \Omega: \wp_{\mathcal{E}}(C)=1 \wedge \wp_{\mathcal{E}}\left(\eta_{i}\right)=1 \Leftrightarrow i \in I\right\} .
$$

We can find the size of each of the $S_{I}$ by counting the number of potential functions which correspond to members of $S_{I}$. If $\mathcal{E} \in S_{I}$ then there are two possible values for $\wp_{\mathfrak{E}}\left(\sigma_{i}\right)$ for each $i$ with $\wp_{\mathscr{E}}\left(\eta_{i}\right)=1$ and $\wp_{\mathfrak{E}}\left(\eta_{i+1}\right)=1$, and two possible values for $\wp_{\mathfrak{E}}\left(\rho_{i}\right)$ for each $i$ with $\wp_{\mathcal{E}}\left(\eta_{i}\right)=1$. All of the other $\sigma_{i}$ and $\rho_{i}$ must have potential value 0 . Hence,

$$
\left|S_{I}\right|=2^{|I|+c(I)},
$$

where $c(I)$ counts the number of circular successions in $I$. The number of $j$-subsets of $\{1, \ldots, k\}$ containing $m$ circular successions is given by the following expression ${ }^{2}$ :

$$
c(k, j, m)=\left\{\begin{array}{ll}
0 & \text { if } j=0, j>k, \text { or } m<2 j-k  \tag{2.9}\\
\frac{k}{j}\binom{j}{m}\binom{k-j-1}{j-m-1} & \text { otherwise }
\end{array} .\right.
$$

Then,

$$
\begin{align*}
|S| & =1+\sum_{j=0}^{k} 2^{j} \sum_{m=0}^{j} 2^{m} c(k, j, m)  \tag{2.10}\\
& =\sum_{j=1}^{k-1} \sum_{m=\max (0,2 j-k)}^{j-1} \frac{k}{j}\binom{j}{m}\binom{k-j-1}{j-m-1} 2^{j+m}+1+2^{2 k}  \tag{2.11}\\
& >\sum_{j=1}^{k-1}\binom{j}{2 j-k} 2^{3 j-k}  \tag{2.12}\\
& >\binom{\left\lfloor\frac{16}{17} k\right\rfloor}{\left\lceil\frac{1}{17} k\right\rceil} 2^{\frac{31}{17} k-3} \quad \text { if } k \geq 17  \tag{2.13}\\
& \geq 2^{\left(2+\frac{1}{17}\right) k-3} \tag{2.14}
\end{align*}
$$

The last line of this follows from the fact that

$$
\binom{\left\lfloor\frac{16}{17} k\right\rfloor}{\left\lceil\frac{1}{17} k\right\rceil} \geq 2^{\frac{4 k}{17}}
$$

when $k \geq 17$. Hence,

$$
|\partial S| /|S|<8 \cdot 2^{-\frac{1}{17} k} \in O\left(2^{-\frac{1}{51} f}\right) .
$$

Remark 2.28. Observe that we can obtain $H_{N}$ as a subgraph of the infinite triangular lattice (see the embedding of $H_{2}$ in Figure 2.7, for an example). Hence, the requirement that the input is a solid subgraph is necessary for the rapid mixing result of $\$ 2.6$ to hold.

[^5]
## Chapter 3

## Euler tours of Random Graphs

### 3.1 Introduction

For situations where we do not have a rapidly mixing Markov chain (or some other fpaus) for every instance of a sampling problem, we relax the problem and search for a sampling algorithm that works for almost all inputs. Formally, we want to show, for some particular sampling problem on an alphabet $\Sigma$, that there exists an algorithm $\mathcal{A}$ such that for $x$ chosen uniformly at random from $\Sigma^{\star}$, the probability that $\mathcal{A}$ behaves like an a fpaus for $\Omega(x)$ tends to 1 as $|x| \rightarrow \infty$.

When we first started looking at Euler tours, our goal was to find a rapidly mixing Markov chain on the set of Euler tours of an undirected Eulerian graph. This task proved too difficult, even restricted to special classes of graphs such as planar graphs or 4-regular graphs. Inspired by results showing that, with high probability, one could sample or approximately count Hamiltonian cycles of a random $d$-regular graph in polynomial time [36], we turned our attention in this direction. Still, the problem proved too difficult. In this chapter we describe results that we believe to be milestones on the road to showing we can, with high probability, sample or approximately count the Euler tours of a random $2 d$-regular graph, using very simple algorithms.

When studying the behaviour of algorithms on random inputs, it is typical for the algorithm itself to be very simple; the hard work goes into the analysis. A natural and simple algorithm for generating a random Euler tour is given in Algorithm 4. In Algorithm 4, we use $E(v)$ to denote the set of edges incident with $v$ and $\mathcal{T}$ for the partial tour, or set of edges which have already been used. We use $\mathcal{T} \cdot e$ to represent the concatenation of $e$ with the sequence of edges $\mathcal{T}$. Recall the definition of a transition system of an Eulerian graph $G$ from $\S 1.3 .2$ (Definition 1.67), and that we use $\mathcal{C}(T)$ to

```
Algorithm 4: Generate
    Input: An Eulerian graph or directed graph \(G\)
    Output: A uniformly random Euler tour \(\mathcal{T}\) of \(G\)
    repeat
        \(u \in V(G)\);
        \(e=(u, v) \in E(u) ;\)
        \(\mathcal{T}=(e) ;\)
        while \(E(v) \backslash \mathcal{T} \neq \emptyset\) do
            \(e=(v, w) \in_{\text {u.a.r. }} E(v) \backslash \mathcal{T} ;\)
            \(v \leftarrow w ;\)
            \(\mathcal{T} \leftarrow \mathcal{T} \cdot e ;\)
        end
    until \(\mathcal{T}=E(G)\);
    return \(\mathcal{T}\);
```

denote the decomposition of $G$ into a set of edge-disjoint cycles induced by a particular transition system $T$ of $G$. One way of looking at Algorithm 4 is that it generates a random transition system of $G$ by choosing one random edge pairing at a time. If this closes a cycle, i.e., forms a component in $\mathcal{C}(\mathcal{T})$, before it has visited every edge, then the algorithm restarts; eventually the algorithm will generate the transition system of an Euler tour $\mathcal{T}$. The algorithm starts at a particular vertex $u$ and chooses an edge incident with it, setting the "current vertex" $v$ to be the opposite end-point of this edge. Then, at each subsequent step we randomly choose an edge incident with $v$ that has not been used before and pair it with the edge chosen at the previous step, updating the "current vertex" $v$ to be the opposite endpoint of the chosen edge. Eventually, we will return to the initial vertex with no more edges available. At this point we have either constructed a uniformly random Euler tour, or generated a partial transition system that does not extend to a transition system of a tour; i.e., we have generated the transition system of an Euler tour of an Eulerian proper subgraph of $G$.

We will now show that Algorithm 4 generates Euler tours with equal probability. Suppose $G$ is a graph with degree sequence $2 \mathbf{d}=\left(2 d_{1}, 2 d_{2}, \ldots, 2 d_{n}\right)$, let $\mathcal{T}$ be some Euler tour of $G$, and let $u$ be the initial vertex in Algorithm 4. Since Euler tours are equivalent up to circular rotation, we can choose any edge incident with $u$ to be the first edge of the tour. The probability that this edge is directed away from $u$, and so valid as the first edge of the tour, is $1 / 2$. The second time we reach $u$ the probability
of choosing the edge-pairing prescribed by $\mathcal{T}$ is $\frac{1}{2 d_{u}-2}$, then $\frac{1}{2 d_{u}-4}$ for the third time, and so on. At all other vertices $v$, the probability of choosing the correct edge-pairing the first time we pass $v$ is $\frac{1}{2 d_{v}-1}$, then $\frac{1}{2 d_{v}-3}$, and so on. Hence, the probability that this procedure constructs the particular Euler tour $\mathcal{T}$ is

$$
\frac{1}{2^{2 d_{u}+1}} d_{u}\binom{2 d_{u}}{d_{u}}\left(\prod_{v=1}^{n} \frac{\binom{2 d_{v}}{d_{v}} d_{v}!}{2^{d_{v}}}\right)^{-1}=\frac{d_{u}\binom{2 d_{u}}{d_{u}}}{2^{2 d_{u}+1}} \frac{1}{|\operatorname{TS}(G)|},
$$

where $\operatorname{TS}(G)$ is the set of transition systems of $G$. Thus, the probability of Algorithm 4 terminating, i.e., generating an Euler tour, on any iteration is

$$
\begin{equation*}
\frac{d_{u}\binom{2 d_{u}}{d_{u}}}{2^{2 d_{u}+1}} \frac{|\operatorname{ET}(G)|}{|\operatorname{TS}(G)|} \tag{3.1}
\end{equation*}
$$

We can also apply Algorithm 4 to sampling Euler tours of Eulerian directed graphs. Suppose $G$ is an Eulerian directed graph on vertex set $V=\{1,2, \ldots, n\}$ with out-degree sequence $\mathbf{d}=\left(d_{1}, d_{2}, \ldots, d_{n}\right)$. If we now take $E(v)$ to be the set of arcs directed away from $v$ in $G$, for each $v \in V$, then Algorithm 4 returns a uniformly random Euler tour of $G$. When $G$ is directed, the probabilities of choosing the correct arc at each pass of $u$ are $\frac{1}{d_{u}-1}, \frac{1}{d_{u}-2}$, etc., and the probabilities of choosing the correct arc at each pass of $v \neq u$ are $\frac{1}{d_{v}}, \frac{1}{d_{v}-1}$, etc. Hence, the probability of Algorithm 4 terminating on any iteration is

$$
\begin{equation*}
d_{u} \frac{|\mathrm{ET}(G)|}{|\operatorname{TS}(G)|} . \tag{3.2}
\end{equation*}
$$

We can use the same idea to construct an algorithm for approximating the number of Euler tours of any Eulerian graph or directed graph. Algorithm 5 iterates the process of Algorithm 4 a number of times and takes the proportion of these iterations which yield Euler tours as an estimation of the probability (3.1) or (3.2), depending on whether $G$ is a graph or directed graph. Since we have an exact expression for the number of transition systems of a graph or directed graph, we can use this to obtain an approximation of the number of Euler tours of $G$.

The time taken to generate a random Euler tour, and the time required to get a good quality approximation of the number of Euler tours, is strongly dependent on the ratio of the number of transition systems to the number of Euler tours. In particular, if $|\mathrm{ET}(G)| /|\mathrm{TS}(G)|$ is bounded from below by an inverse polynomial in the number of vertices of $G$, then Algorithm 4 and Algorithm 5 will run in expected polynomial time, i.e., the algorithms will behave like an fpaus and an fpras, respectively. One consequence of the results in [66] is that the ratio $|\mathrm{ET}(G)| /|\mathrm{TS}(G)|$ is bounded below

```
Algorithm 5: Count
    Input: An Eulerian graph or directed graph \(G\), an integer \(N\)
    Output: An approximation to (3.1) or (3.2)
    count \(\leftarrow 0\);
    \(u \in V(G)\);
    \(d \leftarrow \operatorname{deg}(u) ;\)
    for \(t \leftarrow 1\) to \(N\) do
        \(e=(v, u) \in E(v) ;\)
        \(\mathcal{T} \leftarrow(e) ;\)
        while \(E(v) \backslash \mathcal{T} \neq \emptyset\) do
            \(e=(v, w) \in_{\text {u.a.r. }} E(v) \backslash \mathcal{T} ;\)
            \(v \leftarrow w ;\)
            \(\mathcal{T} \leftarrow \mathcal{T} \cdot e ;\)
        end
        if \(\mathcal{T}=E(G)\) then
            count \(\leftarrow\) count +1 ;
        end
    end
    return count \(/ N\);
```

by an inverse polynomial when $G$ is the complete graph on $K_{2 n+1}$ or a tournament on $2 n+1$ vertices (an Eulerian orientation of $K_{2 n+1}$ ). However, there exist graphs for which this will not be the case. For example, consider the graph from Example 1.70. This graph has $3^{n}$ transition systems, but only $(n+1) 2^{n}$ Euler tours, so both algorithms will require exponential time. That example, however, is of a particularly contrived multigraph, so we might hope the algorithms will perform better on a typical, i.e., random, graph. That is to say, we are looking for a result of the following form:

Let $\mathcal{C}_{1}, \mathcal{C}_{2}, \ldots$, be a sequence of sets of Eulerian graphs. There exists some constant $\alpha>0$ such that the probability that $|\operatorname{ET}(G)| /|\operatorname{TS}(G)| \geq n^{\alpha}$ for $G$ chosen randomly from $\mathcal{C}_{n}$ tends to 1 as $n \rightarrow \infty$

In this chapter we investigate the above statement for the cases when $\mathcal{C}_{n}$ is taken to be equal to the class of $d$-in/d-out graphs or the class of $2 d$-regular graphs. In the case of $d$-in/d-out graphs, we are able to obtain a complete proof.

Definition 3.1. Let $d$ be some fixed positive integer and suppose we have a vector
$\mathbf{d}=\left(d_{1}, d_{2}, \cdots, d_{n}\right)$ such that

$$
d \geq d_{n} \geq d_{n-1} \geq \cdots \geq d_{1} \geq 1
$$

and let $m=\sum_{v} d_{v}$. We use $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$ to denote the set of Eulerian directed graphs with out-degree sequence $\mathbf{d}$, and $\overrightarrow{\mathbb{G}}^{\star}(n, \mathbf{d})$ to denote the set of Eulerian directed multigraphs with out-degree sequence $\mathbf{d}$. We use $\overrightarrow{\mathbb{G}}(n, d)$ and $\overrightarrow{\mathbb{G}}^{\star}(n, d)$ to denote the set of $d$-in/dout graphs and multigraphs, respectively.

Theorem 3.2. Let $d$ be some fixed positive integer greater than 1 , let $n \in \mathbb{N}$. If $\vec{G}$ is chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, d)$ we have

$$
\mathbb{P}\left[\frac{|\operatorname{ET}(G)|}{|\operatorname{TS}(G)|} \in \Omega\left(n^{-2}\right)\right] \rightarrow 1 \text { as } n \rightarrow \infty .
$$

In the case of $2 d$-regular graphs we are only able to make a conjecture but support this conjecture with a mix of empirical and theoretical evidence.

Definition 3.3. Let $d$ be some fixed positive integer and suppose we have a vector $\mathbf{d}=\left(d_{1}, d_{2}, \cdots, d_{n}\right)$ such that

$$
d \geq d_{n} \geq d_{n-1} \geq \cdots \geq d_{2} \geq d_{1} \geq 1
$$

and let $2 m=\sum_{v=1}^{n} d_{v}$. We use $\mathbb{G}(n, \mathbf{d})$ to denote the set of all undirected graphs with degree sequence $\mathbf{d}$ and $\mathbb{G}^{\star}(n, \mathbf{d})$ to denote the set of multigraphs with degree sequence d. We use $\mathbb{G}(n, d)$ and $\mathbb{G}^{\star}(n, d)$ to denote the sets of $d$-regular graphs and $d$-regular multigraphs with $n$ vertices, respectively.

Conjecture 3.4. Let $d$ be some fixed positive integer greater than 1 , let $n \in \mathbb{N}$. If $G$ is chosen uniformly at random from $\mathbb{G}(n, 2 d)$ we have

$$
\mathbb{P}\left[\frac{|\mathrm{ET}(G)|}{|\operatorname{TS}(G)|} \in \Omega\left(n^{-1}\right)\right] \rightarrow 1 \text { as } n \rightarrow \infty .
$$

The contents of this chapter are as follows. In §3.2, we describe the models we use to study random graphs and the techniques used in our analysis. Recall that, by the BEST Theorem (Theorem 1.57), counting arborescences is equivalent to counting Euler tours for any Eulerian directed graph. In $\S 3.3$ we analyse the distribution of the number of arborescences of a random $d$-in/d-out graph, and use this to infer results about the number of Euler tours of a random $d$-in/d-out graph. In $\S 3.4$, we analyse the distribution of the number of Eulerian orientations of a random $2 d$-regular graph. Each one of these orientations is a $d$-in/d-out graph, so, recalling the definition of an
orb of an Eulerian graph $G$ (Definition 1.69) as a pair $(\mathcal{E}, \mathcal{A})$, where $\mathcal{E}$ is an Eulerian orientation of $G$ and $\mathcal{A}$ is an arborescence of $\mathcal{E}$, we would hope that the results of §3.3 and S3.4 could be combined to say something about the number of orbs of a random $2 d$-regular graph. As observed by Brightwell and Winkler [15], the BEST theorem can be used to relate the number of orbs of an Eulerian graph to the number of Euler tours, so this would be equivalent to analysing the number of Euler tours of a random $2 d$-regular graph. Unfortunately, we were unable to bring this line of research to completion. In $\S 3.5$ we provide some initial results, and state a conjecture which, if true, will show that we can, with high probability, sample and approximately count the Euler tours of a random $2 d$-regular graph in expected polynomial time. Furthermore, we present some empirical evidence to support our conjecture. Finally, in §3.6, we give some examples of graphs on which Algorithm 4 and Algorithm 5 will not be effective.

### 3.2 Random graphs

In this chapter, we are interested in the random variables representing the number of Euler tours or Eulerian orientations of $G$ when $G$ is chosen uniformly at random from $\mathbb{G}(n, \mathbf{d})$ or $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$. In order to be able to study properties of random elements of $\mathbb{G}(n, \mathbf{d})$ or $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$, we need a model for generating random elements of both sets. In the next section we describe the configuration model $[9,11]$, which we will use to generate random graphs with a fixed degree sequence.

### 3.2.1 The configuration model for random graphs

The model we use to generate random graphs is the configuration model of Bollobás [11] (implicit in the earlier work of Bender and Canfield [9]). See the survey of Wormald [100] or the textbooks of Bollobás [12] and Janson, Luczak, and Rucinski [44] for more details of the contents of this section.

Definition 3.5. Suppose we have d as in Definition 3.3. For every $v=1,2, \ldots, n$ we define a set $W_{v}$ containing $d_{v}$ points. Each one of these represents an endpoint of an edge incident with $v$. We define the set $W$ to be formed as the union of these $n$ disjoint sets. We call a perfect matching on $W$ a configuration and we call a matching on $W$ a partial configuration. Let $\Omega_{n, \mathbf{d}}$ denote the sets of configurations for particular $\mathbf{d}$.

For each $\mathrm{F} \in \Omega_{n, \mathrm{~d}}$, we define the projection of $\mathrm{F}, \sigma(\mathrm{F})$, to be the multigraph in which each set $W_{v}$ is contracted to a single vertex: a pair $\{x, y\} \in \mathrm{F}$ with $x \in W_{u}$ and
$y \in W_{v}$ corresponds to an edge $\{u, v\}$ in $\sigma(\mathrm{F})$. For each partial configuration $\mathrm{F}^{\prime}$, we define the projection $\sigma\left(\mathrm{F}^{\prime}\right)$ to be the graph obtained by applying the same operation to the pairs in $\mathrm{F}^{\prime}$; that is, $\sigma\left(\mathrm{F}^{\prime}\right)$ is the subgraph of $\sigma(\mathrm{F})$ corresponding to the pairs in $\mathrm{F}^{\prime}$, for any $\mathrm{F} \supseteq \mathrm{F}^{\prime}$.

We say that F contains a loop if it contains a pair $\{x, y\} \in \mathrm{F}$ with $x, y \in W_{v}$ for some $v$, and that F contains a double edge if there are pairs $\{x, y\}$ and $\left\{x^{\prime}, y^{\prime}\right\}$ in F with $x, x^{\prime} \in W_{u}$ and $y, y^{\prime} \in W_{v}$ for some $u \neq v$. If F does not contain any loops or double edges then $\sigma(\mathrm{F})$ is a simple graph, i.e., $\sigma(\mathrm{F}) \in \mathbb{G}(n, \mathbf{d})$. Furthermore each graph $G \in \mathbb{G}(n, \mathbf{d})$ is the projection of exactly $\prod_{v=1}^{n} d_{v}$ ! different configurations; (non-simple) multigraphs are obtained as the projection of fewer configurations, with this number depending on the number of loops and double edges.

Hence, we can generate a uniformly random graph $G \in \mathbb{G}(n, \mathbf{d})$ by generating uniformly random configurations $\mathrm{F} \in \Omega_{n, \mathbf{d}}$ until we obtain an F with no loops or double edges, and then returning $G=\sigma(\mathrm{F})$. The probability of a configuration not containing any loops or double edges is, asymptotically,

$$
\begin{equation*}
\exp \left(-\lambda / 2-\lambda^{2} / 4\right) \tag{3.3}
\end{equation*}
$$

where $\lambda=\frac{1}{m} \sum_{1 \leq v \leq n}\binom{d_{v}}{2}$ [12, Theorem 2.16]. Since each $d_{v}$ is bounded by the constant $d$, it follows that this probability will always be a small constant. However, as we increase $d$ this probability gets smaller and smaller, so this approach will not be so useful for practical generation of random graphs. For example, if $d_{u}=d_{v}=10$ for all $u$ and $v$, then (3.3) is already $e^{-99 / 4}<10^{-10}$.

Our main interest in the model is as a tool for analysing properties of random graphs, and for this task the inefficiency of the random graph generation algorithm is irrelevant. In §3.5, we describe the result of some computational experiments. For this task we used one of the more practical algorithms for generating random regular graphs, due to Steger and Wormald [83]. This algorithm generates random graphs with a probability that is close to the uniform distribution, sacrificing accuracy for efficiency.

Note that the distribution on $\mathbb{G}^{\star}(n, \mathbf{d})$ that results from taking the projection of uniformly random $\mathrm{F} \in \Omega_{n, \mathbf{d}}$ is not uniform. This is because the probability of obtaining a particular element of $\mathbb{G}^{\star}(n, \mathbf{d})$ depends on the number of loops and double edges.

We use the directed configuration model for generating random Eulerian directed graphs with fixed out-degree sequence.

Definition 3.6. Suppose we have $\mathbf{d}$ as in Definition 3.1. For every $v=1,2, \ldots, n$ we define a pair of sets $S_{v}$ and $T_{v}$, each containing $d_{v}$ points. The points in $S_{v}$ represent the endpoints of arcs for which $v$ is the source, and the points in $T_{v}$ represent the endpoints of arcs for which $v$ is the target. We call a perfect matching from $S$ to $T$ a directed configuration, and we call a matching a partial directed configuration. We use $\Phi_{n, \mathbf{d}}$ to denote the set of all directed configurations, for particular $\mathbf{d}$.

For each $\vec{F} \in \Phi_{n, \mathbf{d}}$, we define the projection of $\overrightarrow{\mathrm{F}}, \sigma(\overrightarrow{\mathrm{F}})$, to be the directed multigraph in which the sets $S_{\nu} \cup T_{v}$ have been contracted to a single vertex. That is, we have an arc from $u$ to $v$ in $\sigma(\overrightarrow{\mathrm{F}})$ whenever we have a pair $(x, y)$ from $S_{u} \times T_{v}$ in $\overrightarrow{\mathrm{F}}$. For each partial configuration $\overrightarrow{\mathrm{F}}^{\prime}$, we define the projection $\sigma\left(\overrightarrow{\mathrm{F}}^{\prime}\right)$ to be graph obtained by applying the same operation to the pairs in $\overrightarrow{\mathrm{F}}$; that is, $\sigma\left(\overrightarrow{\mathrm{F}}^{\prime}\right)$ is the subgraph of $\sigma(\overrightarrow{\mathrm{F}})$ corresponding to the pairs in $\overrightarrow{\mathrm{F}}^{\prime}$, for any $\overrightarrow{\mathrm{F}} \supseteq \overrightarrow{\mathrm{F}}^{\prime}$.

We say a directed configuration $\overrightarrow{\mathrm{F}}$ has a loop if it contains a pair $(x, y) \in S_{v} \times T_{v}$ and a double arc if it contains pairs $(x, y)$ and $\left(x^{\prime}, y^{\prime}\right)$ from $S_{u} \times T_{v}$ for $u \neq v$. Note that we do not consider $(x, y) \in S_{u} \times T_{v}$ and $\left(x^{\prime}, y^{\prime}\right) \in S_{v} \times T_{u}$ to be a double arc. As in the undirected case, each simple directed graph is the projection of the same number of configurations, $\Pi_{v}\left(d_{v}!\right)^{2}$, and the probability that a uniformly random $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ projects to a simple directed graph is, asymptotically,

$$
\begin{equation*}
\exp \left(-\frac{m_{2}}{m}-\frac{\left(m_{2}-m\right)^{2}}{2 m^{2}}\right), \tag{3.4}
\end{equation*}
$$

where $m_{2}=\sum_{v} d_{v}^{2}$.
Remark 3.7. The probability (3.4) follows as a result of arguments given in Theorem 3.30. We believe this is a new result, though the calculations are similar to those used to prove (3.3). The directed configuration model has been studied in the literature [4], also in the context of Euler tours, but only for the case $d_{v}=2$ for all $v$.

### 3.2.2 Asymptotic Distributions

Any numerical characteristic of a graph $G$, e.g., the number of Euler tours, gives rise to a random variable when we consider $G$ as being drawn at random from some class of graphs. In this work, we would like to know when the probability of Algorithm 4 returning an Euler tour is bounded below by an inverse polynomial. For example, we want to show there exists some $\alpha>0$ such that

$$
\begin{equation*}
\mathbb{P}\left[\frac{|\mathrm{ET}(G)|}{|\operatorname{TS}(G)|} \geq n^{-\alpha}: G \in \mathbb{G}(n, 4)\right] \rightarrow 1, \tag{3.5}
\end{equation*}
$$

as $n \rightarrow \infty$. That is, we want to show that the probability Algorithm 4 terminates on any particular iteration is at least $n^{-\alpha}$ for almost every $G \in \mathbb{G}(n, 4)$. Let $\mathcal{T}_{n, 4}$ denote the random variable counting the number of Euler tours of a random 4-regular graph. In Section 3.5.1, we show that the expected number of Euler tours of a 4-regular graph satisfies

$$
\mathbb{E}\left[\mathcal{T}_{n, 4}\right] \rightarrow e^{3 / 4} \sqrt{\frac{\pi}{2 n}} 3^{n}
$$

There are 3 possible transition systems at each vertex of a 4-regular graph, so $|T S(G)|=3^{n}$ for each $G \in \mathbb{G}(n, 4)$. Thus, to prove (3.5) all we need to do is show $\mathcal{T}_{n, 4}$ is concentrated around its mean. In many cases, Chebyshev's inequality (see below) is sufficient to prove concentration results and it has seen a great deal of application to the analysis of randomised algorithms, see, e.g., [70].

Theorem 3.8 (Chebyshev's Inequality). Let $X$ be a random variable. For any $a>0$,

$$
\mathbb{P}[|X-\mathbb{E}[X]| \geq a] \leq \frac{\operatorname{Var}[X]}{a^{2}}
$$

However, in some cases Chebyshev's inequality is not strong enough and we need to look harder at the distribution of the random variable of interest.

Exact formulas for the distributions of interesting combinatorial quantities are rare and even when they exist are usually too complicated to be of use. Hence, the focus is shifted to the search for an asymptotic distribution that holds with increasing accuracy.

Definition 3.9. Let $X_{1}, X_{2}, \ldots$ be a sequence of integer-valued random variables. We say $\left(X_{n}\right)_{n \geq 0}$ converges in distribution to a random variable $Z$, or $Z$ is the asymptotic distribution of $X_{n}$, if

$$
\mathbb{P}\left[X_{n}=x\right] \rightarrow \mathbb{P}[Z=x], \quad \text { as } n \rightarrow \infty .
$$

We use $X_{n} \xrightarrow{d} Z$ to denote convergence in distribution.
If $X_{n} \xrightarrow{d} Z$ then, for large values of $n$, we can consider $X_{n}$ to have the same distribution as $Z$. A particularly important instance of this that arises in the study of random graphs with fixed degree sequence is convergence to a Poisson distribution.

Definition 3.10. A non-negative integer-valued random variable $X$ is said to have a Poisson distribution with mean $\lambda$ if

$$
\mathbb{P}[X=x]=e^{-\lambda} \frac{\lambda^{x}}{x!}, \quad x \in \mathbb{N} .
$$

Definition 3.11. Let $\lambda_{1}, \lambda_{2}, \ldots, \lambda_{k}$ be arbitrary non-negative real constants and let $X_{1, n}, X_{2, n}, \ldots, X_{k, n}$ be a set of non-negative integer random variables defined on the same probability space $\Omega_{n}$, for each $n$. We say these random variables are asymptotically independent Poisson random variables with means $\lambda_{i}$, if their joint distribution converges to that of independent Poisson variables with means $\lambda_{i}$. That is, for every fixed set of non-negative integers $\left\{x_{i}: 1 \leq i \leq k\right\}$ we have

$$
\mathbb{P}\left[X_{i, n}=x_{i}, \forall 1 \leq i \leq k\right] \rightarrow \prod_{i=1}^{k} e^{-\lambda_{i}} \frac{\lambda_{i}^{x_{i}}}{x_{i}!},
$$

as $n \rightarrow \infty$.

It is well-known that a Poisson random variable is completely determined by the values of its factorial moments.

Definition 3.12. For any non-negative integers $n$ and $k$, we define the factorial power (falling factorial) $(n)_{k}$ by

$$
(n)(n-1) \cdots(n-k+1) .
$$

If $k>n$ we have $(n)_{k}=0$.
Definition 3.13. Let $X$ be a random variable. The $k$-th factorial moment of $X$ is given by $\mathbb{E}\left[(X)_{k}\right]$.

The following lemma shows how the asymptotic values of the factorial moments can be used to show that a set of sequences of random variables converge to independent Poisson random variables.

Theorem 3.14 (Theorem 1.23 [12]). Let $\lambda_{1}, \lambda_{2}, \ldots, \lambda_{k}$ be arbitrary non-negative real constants and let $X_{1, n}, X_{2, n}, \ldots, X_{k, n}$ be a set of non-negative integer random variables defined on the same probability space $\Omega_{n}$, for each n. If

$$
\mathbb{E}\left[\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right] \rightarrow \prod_{i=1}^{k} \lambda_{i}^{j_{i}} .
$$

holds for each fixed set of non-negative integers $j_{1}, j_{2}, \ldots, j_{k}$ then the variables $X_{1, n}, \ldots, X_{k, n}$ are asymptotically independent Poisson random variables with means $\lambda_{i}$.

Often, when we apply Theorem 3.14 the random variables $X_{i, n}$ will be counting the number of occurrences of some sort of substructure. For example, suppose $X_{i, n}(G)$ counts the number of $i$-cycles in a random graph on $n$ vertices and let $\mathcal{C}_{i n}$ denote the set
of different $i$-cycles you can form on $n$ vertices. Then, we can write $X_{i, n}=\sum_{C \in \mathcal{C}_{i n}} I_{C}$, where $I_{C}$ is the indicator variable for the event $C \in G$. That is,

$$
I_{C}=\left\{\begin{array}{ll}
1 & C \in \operatorname{CYCLES}(G) \\
0 & C \notin \operatorname{CYCLES}(G)
\end{array} .\right.
$$

In this case, computing the factorial power $\left(X_{i, n}\right)_{k}$ amounts to summing over products of indicator functions for ordered tuples of distinct $i$-cycles, e.g.,

$$
\left(X_{i, n}\right)_{3}=\sum_{\substack{\left(C_{1}, C_{2}, C_{3}\right) \in C_{i n}^{3} \\ C_{1} \neq C_{2} \neq C_{3}}} I_{C_{1}} I_{C_{2}} I_{C_{3}} .
$$

Theorem 3.14 can be used to prove the following theorem, due to Bollobás [11] (or [12, Theorem 2.16]).

Theorem 3.15 (Bollobás [11]). Let $\lambda=\frac{1}{m} \sum_{v}\binom{d_{v}}{2}$ and for each integer i let $\lambda_{i}=\lambda^{i} / 2 i$ and let $X_{i, n}$ count the number of $i$-cycles in a multigraph $G$ obtained as the projection of a uniformly random $\mathrm{F} \in \Omega_{n, \mathbf{d}}$. For any fixed positive integer $k$, the set of variables $X_{i, n}$ for $i<k$ are asymptotically Poisson independent random variables with means $\lambda_{i}$.

The asymptotic independence of the random variables $X_{i, n}$ follows from a wellknown property of random regular graphs [44, 100].

Lemma 3.16. Let $d$ be some fixed, non-negative integer, and let $H$ be some graph with more edges than vertices. Then, for $G$ chosen uniformly at random from $\mathbb{G}(n, d)$ we have

$$
\mathbb{P}[H \text { is a subgraph of } G] \rightarrow 0,
$$

as $n \rightarrow \infty$.

Recall that a simple graph is a multigraph with no loops or double edges. Hence, when $F$ is chosen uniformly at random from $\Omega_{n, \mathbf{d}}$, we have

$$
\mathbb{P}[\boldsymbol{\sigma}(F) \text { is simple }]=\mathbb{P}\left[X_{1, n}=X_{2, n}=0\right],
$$

and thus we can infer (3.3) from Theorem 3.15 and Definition 3.11.

### 3.2.3 Conditioning on short cycle counts

There are many combinatorial quantities which do not have asymptotic Poisson distributions, e.g., the number of Hamiltonian cycles, the number of perfect matchings,
or the number of Euler tours of a random $2 d$-regular graph. However, it is still possible to compute asymptotic distributions for many of these quantities and Theorem 3.15 is the foundation of many of these results.

We will now discuss one of these results in more detail. As this result is on random regular graphs, where $d_{v}=d$ for all $v$, for the rest of this section we will speak of the set of $d$-regular graphs, $\mathbb{G}(n, d)$, and the regular configuration model, $\Omega_{n, d}$.

Definition 3.17. Let $G=(V, E)$ be a graph. We say a cycle in $G$ is a Hamiltonian cycle if it visits each vertex $v \in V$ exactly once.

Definition 3.18. Let $G=(V, E)$ be a graph. We say a subgraph $H$ of $G$ is a 2-factor if each vertex $v \in V$ has degree 2 in $H$; that is, $H$ is a set of vertex-disjoint cycles which contains every vertex of $G$.

In 1988, Frieze [35] showed that one could find a Hamiltonian cycle in almost every $d$-regular graph, for $d \geq 85$, though it was widely believed that this held for all $d \geq 3$. Then, Robinson and Wormald [75] showed that, for $d=3$, the number of Hamiltonian cycles of a random 3-regular graph, $H_{n}$, satisfied

$$
\begin{equation*}
\frac{\mathbb{E}\left[H_{n}^{2}\right]}{\mathbb{E}\left[H_{n}\right]^{2}} \rightarrow \frac{3}{e} \tag{3.6}
\end{equation*}
$$

Applying Chebyshev's inequality with $a=\mathbb{E}\left[H_{n}\right]$, we find

$$
\mathbb{P}\left[H_{n}=0\right] \leq \mathbb{P}\left[\left|H_{n}-\mathbb{E}\left[H_{n}\right]\right| \geq \mathbb{E}\left[H_{n}\right]\right] \leq\left(\frac{3}{e}-1\right)
$$

Hence, from (3.6) we can deduce that at least a $2-3 e^{-1}$ fraction of 3-regular graphs are Hamiltonian. However, proving the result people expected,

$$
\mathbb{P}[G \text { is Hamiltonian : } G \in \mathbb{G}(n, d)] \rightarrow 1,
$$

for any $d \geq 3$, was beyond the powers of Chebyshev's inequality.
Then, in a pair of papers from the early 90 's $[76,77]$ Robinson and Wormald managed to prove that for all $d \geq 3$, a random $d$-regular graph will contain a Hamiltonian cycle with high probability, using a technique we call conditioning on short cycle counts. Their proof involved partitioning $\Omega_{n, d}$ into a family of sets indexed by vectors of the numbers of occurrences of cycles of various lengths, namely

$$
\Omega_{\mathbf{x}}=\left\{\mathrm{F} \in \Omega_{n, d}: X_{i, n}(\mathrm{~F})=x_{i} \text { for all } 1 \leq i \leq k\right\} .
$$

Robinson and Wormald approached this result indirectly, by analysing the number of perfect matchings in a random $d$-regular graph. Subsequent work of Frieze et al. [36]
obtained the same result using the same technique, but this time working directly with the number of Hamiltonian cycles. Specifically, Frieze et al. [36] showed that the variance of the number of Hamiltonian cycles can be divided into two parts: the variance between the groups $\Omega_{\mathbf{x}}, \mathbf{x} \in \mathbb{N}^{k}$; and the variance within each of the different groups. The calculations in [76, 77, 36] showed that almost of all the variance comes from the variance between the groups, and so the variance within any particular group is negligible. Thus, with high probability the number of Hamilton cycles in a graph $G$ chosen randomly from $\Omega_{\mathbf{x}}$ will be close to the expected number of Hamilton cycles for graphs in that group,

$$
\mathbb{E}_{\mathbf{x}}=\mathbb{E}\left[H_{n}(F): F \in \Sigma_{\mathbf{x}}\right],
$$

Furthermore, with high probability the group $F$ lies in will have $\mathbb{E}_{\mathbf{x}}$ large enough to imply that the number of Hamilton cycles in $\sigma(F)$ is $>0$.

In fact, Frieze et al. [36] proved something stronger. The main contribution of their paper was to adapt the argument of Robinson and Wormald to show that, with high probability, the number of Hamiltonian cycles in a random $d$-regular graph was at most a linear factor less than the number of 2-factors. Since there exists an fpaus for the 2 -factors of any graph, using the Jerrum-Sinclair chain [49], it follows that we can sample (and so approximately count) the Hamiltonian cycles of almost every $d$-regular graph in polynomial time. Their proof is comprised of the following two steps:

1. They show that the expected number of Hamiltonian cycles in a random $d$ regular graph $G$ is within a linear factor of the number of 2-factors of $G$;
2. They show that the random variable counting the Hamiltonian cycles is concentrated around its mean.

A general version of this method was presented by Molloy et al. [71].
Janson [43] showed that the result of $[76,77,36]$ really amounts to obtaining an asymptotic distribution for the number of Hamiltonian cycles. He streamlined the general approach of Molloy et al. [71] to obtain the following general theorem.

Theorem 3.19 (Janson [43]). Let $\lambda_{i}>0$ and $\delta_{i} \geq-1, i=1,2, \ldots$, be constants and suppose that for each $n$ there are random variables $X_{i, n}, i=1,2, \ldots$, and $Y_{n}$ (defined on the same probability space $\Omega_{n}$ ) such that $X_{i, n}$ is non-negative integer valued and $\mathbb{E}\left[Y_{n}\right] \neq 0$ (at least for large $n$ ) and furthermore the following conditions are satisfied

1. $X_{i, n} \xrightarrow{d} Z_{i}$ as $n \rightarrow \infty$, jointly for all $i$, where $Z_{i}$ is a Poisson random variable with mean $\lambda_{i}$;
2. For any finite sequence $x_{1}, \ldots x_{k}$ of non-negative integers

$$
\frac{\mathbb{E}\left[Y_{n} \mid X_{1, n}=x_{1}, \ldots X_{k, n}=x_{k}\right]}{\mathbb{E}\left[Y_{n}\right]} \rightarrow \prod_{i=1}^{k}\left(1+\delta_{i}\right)^{x_{i}} e^{-\lambda_{i} \delta_{i}} \quad \text { as } n \rightarrow \infty ;
$$

3. $\sum_{i} \lambda_{i} \delta_{i}^{2}<\infty$;
4. $\lim _{n \rightarrow \infty} \frac{\mathbb{E}\left[Y_{n}^{2}\right]}{\mathbb{E}\left[Y_{n}\right]^{2}} \leq \exp \left(\sum_{i} \lambda_{i} \delta_{i}^{2}\right)+o(1)$.

Then

$$
\frac{Y_{n}}{\mathbb{E}\left[Y_{n}\right]} \xrightarrow{d} W=\prod_{i=1}^{\infty}\left(1+\delta_{i}\right)^{Z_{i}} e^{-\lambda_{i} \delta_{i}},
$$

as $n \rightarrow \infty$. Moreover, this and the convergence of the $X_{i, n}$ to the $Z_{i}$ in 1 hold jointly.
Remark 3.20. In the proof of Theorem 3.19 given in [43], it is shown that Conditions 1 and 2 imply an asymptotic lower bound of $\exp \left(\sum_{i} \lambda_{i} \delta_{i}^{2}\right)$ on $\mathbb{E}\left[Y_{n}^{2}\right] / \mathbb{E}\left[Y_{n}\right]^{2}$. Hence, whenever the conditions of Theorem 3.19 hold, we actually have

$$
\frac{\mathbb{E}\left[Y_{n}^{2}\right]}{\mathbb{E}\left[Y_{n}\right]^{2}} \rightarrow \exp \left(\sum_{i} \lambda_{i} \delta_{i}^{2}\right),
$$

as $n \rightarrow \infty$.
Remark 3.21. The convergence of $X_{i, n}$ to $Z_{i}$ and $Y_{n} / \mathbb{E}\left[Y_{n}\right]$ to $W$ implied by Theorem 3.19 holds jointly. Hence, if the conditions of Theorem 3.19 hold, and $Y_{n}^{\prime}$ has the distribution of $Y_{n}$ conditioned on $X_{1, n}=X_{2, n}=0$, then we can conclude that

$$
\begin{aligned}
\frac{\mathbb{E}\left[Y_{n}^{\prime}\right]}{\mathbb{E}\left[Y_{n}\right]} \rightarrow \exp \left(-\lambda_{1} \delta_{1}-\lambda_{2} \delta_{2}\right), \\
\frac{\mathbb{E}\left[\left(Y_{n}^{\prime}\right)^{2}\right]}{\mathbb{E}\left[Y_{n}^{\prime}\right]^{2}} \rightarrow \exp \left(-\lambda_{1} \delta_{1}^{2}-\lambda_{2} \delta_{2}^{2}\right) \frac{\mathbb{E}\left[Y_{n}^{2}\right]}{\mathbb{E}\left[Y_{n}\right]^{2}} .
\end{aligned}
$$

We will use the properties mentioned in Remark 3.20 and Remark 3.21 to simplify the proofs of §3.4. The following lemma gives a useful reformulation of Condition 2 of Theorem 3.19 that will be used several times in the proofs of this chapter.

Lemma 3.22 (Janson [43]). Suppose Condition 1 of Theorem 3.19 holds, $Y_{n} \geq 0$, and that

$$
\frac{\mathbb{E}\left[Y_{n}\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right]}{\mathbb{E}\left[Y_{n}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}},
$$

for some $\mu_{i} \geq 0$ and every finite sequence of non-negative integers $j_{1}, j_{2}, \ldots, j_{k}$. Then, Condition 2 of Theorem 3.19 holds with $\delta_{i}=\mu_{i} / \lambda_{i}-1$.

As in the work of Robinson and Wormald [76, 77] and Frieze et al. [36], the proof of Theorem 3.19 is basically an analysis of variance. The state space $\Omega_{n}$ is partitioned by conditioning on fixed values for the auxiliary variables $X_{i, n}$. The first three conditions of the theorem imply a lower bound on the variance between the partitions; the fourth condition gives a matching upper bound on the variance in the unconditioned space. Since this upper bound matches the lower bound, we can conclude that the variance within each component is negligible. Hence, the values of $Y_{n}$ are almost surely determined by the values of the auxiliary variables.

For example, in the problem studied by Frieze et al. [36], we have $Y_{n}$ equal to the number of Hamiltonian cycles in a random $G \in \mathbb{G}(n, d)$, for some fixed $d>2$, and $X_{i, n}$ equal to the number of $i$-cycles in a random $G \in \mathbb{G}(n, d)$. It is shown in [36] that, with high probability, the value of $W$ is sufficiently large so that for all $n$ sufficiently large we have $Y_{n}>0$.

In applications of Theorem 3.19 to random regular graphs, $X_{i, n}$ tends to be the number of $i$-cycles in a random configuration $\mathrm{F} \in \Omega_{n, d}, Z_{i}$ a Poisson random variable with mean $\lambda_{i}=(d-1)^{i} / 2 i$, and $Y_{n}$ the combinatorial quantity we are interested in. For example, this is the case in the original line of work investigating the number of Hamiltonian cycles [76, 77, 36], and also in subsequent work characterising the asymptotic distributions for several numerical characteristics of random $d$-regular graphs, e.g., the number of perfect matchings [43, 71], the number of 2-factors [74], and the number of 3 -star factors [5]. In most cases, the chief goal of the analysis is to show that the particular structure occurs with probability approaching 1 , in a random $G \in \mathbb{G}(n, d)$. However, as is the case in [36], this type of result may also have algorithmic implications.

In the remainder of this chapter, we derive asymptotic distributions for the number of Euler tours in a random $d$-in/d-out graph (§3.3), and the number of Eulerian orientations in a random $2 d$-regular Eulerian graph (§3.4), for small, fixed, $d$. By arguing along the lines of Frieze et al. [36], we are able to show that the number of Euler tours in random $d$-in/d-out graph is close to the number of transition systems, and consequently that Algorithm 4 and Algorithm 5 run in expected polynomial time on almost every $d$-in/d-out graph. We conjecture that a similar result holds for the Euler tours of a random $2 d$-regular graph and, in $\S 3.5$, we put forward some arguments to support this conjecture.

In $\S 3.4$, we also show that most $2 d$-regular graphs have few Eulerian orientations, in the sense that the probability that $\mathrm{EO}(G)$ is within a linear factor of the lower bound
of Schrijver (Theorem 1.56) tends to 1 as $n \rightarrow \infty$.

### 3.3 Euler tours of random $d$-in $/ d$-out graphs

In this section we analyse the number of Euler tours of random Eulerian directed graphs, using the directed configuration model as our model for generating random graphs. Recall that, given an integer vector $\mathbf{d}=\left(d_{1}, d_{2}, \ldots, d_{n}\right)$, this can be used to study properties of random elements of the set of directed graphs in which vertex $v \in\{1,2, \ldots, n\}$ has indeg $(v)=\operatorname{outdeg}(v)=d_{v}$.

Although the quantity we are interested in is the number of Euler tours of a random directed graph, it is easier to enumerate arborescences and, by the BEST theorem (Theorem 1.57), this is equivalent to enumerating Euler tours. Let $\mathcal{A}_{n, \mathbf{d}}$ count the number of arborescences of a directed graph $G$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$. In $\S 3.3 .1$ we obtain asymptotic expressions for the first two moments of $\mathcal{A}_{n, \mathbf{d}}$. Then, in §3.3.2, we use Janson's theorem (Theorem 3.19) to characterise the asymptotic distribution of $\mathcal{A}_{n, d}$, for the special case when $d_{v}=d$ for all $v$.

Given the asymptotic distribution of $\mathcal{A}_{n, d}$ it is straightforward to show that the value of $\mathcal{A}_{n, d}$ is almost always close to the mean, $\mathbb{E}\left[\mathcal{A}_{n, d}\right]$ (Theorem 3.35), from which we can immediately infer Theorem 3.2. Although we believe a similar result holds for Eulerian directed graphs with arbitrary fixed degree sequence d, proving this seems to be beyond the method used to prove Theorem 3.35.

### 3.3.1 Estimating the moments of $\mathcal{A}_{n, \mathrm{~d}}$

We obtain asymptotic estimates for the moments of $\mathcal{A}_{n, \mathbf{d}}$ by first computing the moments of a random variable counting arborescences of non-uniform directed multigraphs and then conditioning on the graph being simple. We will use the following two facts several times in the proofs of this section.

Fact 3.23. Falling factorials of sums obey the well known multinomial theorem

$$
\left(x_{1}+x_{2}+\cdots+x_{l}\right)_{k}=\sum_{\substack{\Sigma_{i}=k \\ \delta_{i} \geq 0}}\binom{k}{\delta_{1}, \ldots, \delta_{l}} \prod_{i=1}^{l}\left(x_{i}\right)_{\delta_{i}} .
$$

Definition 3.24. A rooted forest is a set of disjoint trees with distinguished root vertices. We call a rooted forest a rooted $k$-forest if it contains exactly $k$ components.

Fact 3.25. Let $V=\{1,2, \ldots, n\}$. The number of rooted $k$-forests on $V$ in which $v$ has $\delta_{v}$ children is

$$
\binom{n-1}{k-1}\binom{n-k}{\delta_{v}: v \in V} .
$$

See, e.g., [81, Theorem 5.3.4], for a proof of Fact 3.25. We also consider the natural generalisation of the configuration model to non-Eulerian directed graphs.

Definition 3.26. Suppose $\mathbf{s}=\left(s_{1}, s_{2}, \ldots, s_{n}\right)$ and $\mathbf{t}=\left(t_{1}, t_{2}, \ldots, t_{n}\right)$ are a pair of positive vectors. Now, suppose we have disjoint sets of points $S_{v}$ and $T_{v}$ satisfying $\left|S_{v}\right|=s_{v}$ and $\left|T_{v}\right|=t_{v}$, for each $v=1,2, \ldots, n$ and let $S=\bigcup_{v=1}^{n} S_{v}$ and $T=\bigcup_{v=1}^{n} T_{v}$. We call the set of perfect matchings from $S$ to $T$ directed configurations and denote the set of all directed configurations by $\Phi_{n, \mathbf{s}, \mathbf{t}}$. As before, we call the matchings from $S$ to $T$ the partial directed configurations of $\Phi_{n, \mathbf{s}, t}$.

We define projection in this generalised configuration model in the same way as in the original directed configuration model.

In Lemma 3.27 below, we use Fact 3.23 and Fact 3.25 to count the number of partial configurations that project to a directed forest in $\Phi_{n, \mathbf{s}, \mathbf{t}}$.

In the proof of the following lemma, and throughout this section we take the term directed forest to mean a forest in which every edge in a component has been directed towards the the root. Furthermore, we will often speak of a configuration for an arborescence or directed forest. We take this to mean a partial configuration that projects to an arborescence or directed forest.

Lemma 3.27. Let $\mathbf{s}$ and $\mathbf{t}$ be a pair of positive $n$-vectors and let $V=\{1,2, \ldots, n\}$. The number of ways to choose a configuration for a directed forest rooted at $R \subseteq V$ in $\Phi_{n, \mathbf{s}, \mathbf{t}}$ is

$$
\begin{equation*}
\prod_{v \notin R} s_{v}\left(\sum_{v \in R} t_{v}\right)\left(\sum_{v \in V} t_{v}-1\right)_{n-|R|-1} \tag{3.7}
\end{equation*}
$$

Proof. In counting the number of partial configurations that project to forests, we need to enumerate the forests that can occur in some directed graph $\vec{G} \in \overrightarrow{\mathbb{G}}(n, \mathbf{s}, \mathbf{t})$, and also the number of partial configurations in $\Phi_{n, \mathbf{s}, \mathbf{t}}$ that project to each forest. The first part of our analysis counts the number of such forests in which each vertex $v \in V$ has a particular number of children, $\delta_{v}$.

Let $\mathcal{F}$ be a forest on $V$ rooted at $R$ and, for each $v \in V$, let $\delta_{v}$ be the number of children of $v$ in $\mathcal{F}$. The number of ways to choose a point for the source and target of
each arc in $\mathcal{F}$ is

$$
\begin{equation*}
\prod_{v \notin R} s_{v} \prod_{v=1}^{n}\left(t_{v}\right)_{\delta_{v}}, \tag{3.8}
\end{equation*}
$$

since we must choose one of the points in $S_{v}$ for the source of the arc leaving each $v \notin R$ and choose one of the points in $T_{v}$ as the target of each arc entering $v$.

We now count the number of forests on $V$ which have roots at $R$, and which agree with the vector $\boldsymbol{\delta}=\left(\delta_{1}, \delta_{2}, \ldots, \delta_{n}\right)$ of child counts. One way to approach this task is to observe that we can construct a forest rooted at $R$ by first choosing a $k$-forest on $V-R$, where $k=\sum_{v \in R} \delta_{v}$, and then attaching each root of this forest as a child of some $v \in R$. By Fact 3.25, the number of $k$-forests on $V-R$ in which $v \in V-R$ has exactly $\delta_{v}$ children is

$$
\begin{equation*}
\binom{n-|R|-1}{k-1}\binom{n-|R|-k}{\delta_{v}: v \in V-R} . \tag{3.9}
\end{equation*}
$$

The number of ways we can divide the roots of this forest amongst the members of $R$ so that each $v \in R$ has $\delta_{v}$ children is

$$
\begin{equation*}
\binom{k}{\delta_{v}: v \in R} . \tag{3.10}
\end{equation*}
$$

Combining (3.9) and (3.10), we see that the number forests rooted at $R$ that agree with child count vector $\boldsymbol{\delta}$ is

$$
\binom{n-|R|-1}{k-1}\binom{n-|R|-k}{\delta_{v}: v \in V-R}\binom{k}{\delta_{v}: v \in R} .
$$

Hence, by (3.8), the number of partial configurations projecting to these forests is

$$
\begin{equation*}
\binom{n-|R|-1}{k-1}\binom{n-|R|-k}{\delta_{v}: v \in V-R}\binom{k}{\delta_{v}: v \in R} \prod_{v \notin R} s_{v} \prod_{v \in V}\left(t_{v}\right)_{\delta_{v}} . \tag{3.11}
\end{equation*}
$$

Now, summing (3.11) over all possible vectors $\boldsymbol{\delta}$ gives

$$
\begin{array}{r}
\prod_{v \notin R} s_{v} \sum_{k=1}^{n-|R|}\binom{n-|R|-1}{k-1}\left(\sum_{\sum_{v \in R} \delta_{v}=k}\binom{k}{\delta_{v}: v \in R} \prod_{v \in R}\left(t_{v}\right)_{\delta_{v}}\right) \\
\times\left(\sum_{\sum_{v \notin R} \delta_{v}=n-|R|-k}\binom{n-|R|-k}{\delta_{v}: v \notin R} \prod_{v \notin R}\left(t_{v}\right)_{\delta_{v}}\right) . \tag{3.12}
\end{array}
$$

We can use Fact 3.23 to simplify (3.12). The two sums over the different $\delta_{v}$ in (3.12) are the multinomial expansions of the falling factorial powers $\left(\sum_{v \in R} t_{v}\right)_{k}$ and $\left(\sum_{v \notin R} t_{v}\right)_{n-|R|-k}$, respectively. Hence, (3.12) is equal to

$$
\prod_{v \notin R} s_{v} \sum_{k=1}^{n-|R|}\binom{n-|R|-1}{k-1}\left(\sum_{v \in R} t_{v}\right)_{k}\left(\sum_{v \notin R} t_{v}\right)_{n-|R|-k}
$$

which is itself the multinomial expansion of (3.7).

We now use Lemma 3.27 to analyse the expectation and variance of the number of arborescences in $\sigma(F)$, when $F$ is chosen uniformly at random from $\Phi_{n, \mathbf{d}}$. In the following, we say $\mathcal{A} \subset \overrightarrow{\mathrm{F}}$ is an arborescence of $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ if $\sigma(\mathcal{A})$ is an arborescence of $\sigma(F)$, and denote the set of arborescences of $\vec{F}$ by $\operatorname{ARB}(\overrightarrow{\mathrm{F}})$. In the following proofs, we will abuse terminology slightly and switch between speaking of arborescences of configurations and directed graphs arbitrarily.

Theorem 3.28. Let $d$ be some fixed constant, $n \in \mathbb{N}$, and let $\mathbf{d}=\left(d_{1}, d_{2}, \ldots, d_{n}\right)$ be a sequence of integers satisfying

$$
d \geq d_{n} \geq d_{n-1} \geq \cdots \geq d_{1} \geq 1,
$$

and let $m=\sum_{v=1}^{n} d_{v}$. We define $\mathcal{A}_{n, \mathbf{d}}^{\star}$ to be the random variable counting the arborescences of uniformly random $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$. Then,

$$
\begin{gathered}
\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]=\frac{n}{m} \prod_{v=1}^{n} d_{v} ; \\
\frac{\mathbb{E}\left[\left(\mathscr{A}_{n, \mathbf{d}}^{\star}\right)^{2}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]^{2}}=\frac{m}{m-n+1} .
\end{gathered}
$$

Proof. We first consider the first moment of $\mathcal{A}_{n, \mathbf{d}}^{\star}$. To calculate the first moment of $\mathscr{A}_{n, \mathbf{d}}^{\star}$ we need to enumerate pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A})$, where $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ and $\mathcal{A}$ is an arborescence of $\overrightarrow{\mathrm{F}}$, and then divide this quantity by $\left|\Phi_{n, \mathbf{d}}\right|$. Given $\mathcal{A}$, it is easy to count the number of configurations $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ for which $\mathcal{A} \subset \overrightarrow{\mathrm{F}}$. In any directed graph $\vec{G}$ with $m$ arcs, there are exactly $m-n+1$ arcs not contained in any particular element of $\operatorname{ARB}(\vec{G})$. Hence, if we have a configuration for an arborescence, there are $(m-n+1)$ ! ways to extend this to a complete configuration. Applying Lemma 3.27 with $\mathbf{s}=\mathbf{t}=\mathbf{d}$, we see that the number of arborescences rooted at any particular vertex $v$ is

$$
\begin{equation*}
d_{v}\left(\prod_{u \neq v} d_{u}\right)(m-1)_{n-2} . \tag{3.13}
\end{equation*}
$$

By the BEST theorem (Theorem 1.57), there are an equal number of arborescences rooted at each vertex of any $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$. Hence, multiplying (3.13) by by $n(m-n+1)$ ! gives the number of pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ with $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ and $\mathcal{A} \in \operatorname{ARB}(\overrightarrow{\mathrm{F}})$ :

$$
n(m-1)!\left(\prod_{v \in V} d_{v}\right)
$$

Finally, dividing by the total number of configurations in $\Phi_{n, \mathbf{d}}$, which is $m!$, gives the claimed value for $\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]$.

To compute the second moment of $\mathscr{A}_{n, \mathbf{d}}^{\star}$ we need to evaluate the following expression

$$
\begin{equation*}
\frac{1}{m!} \sum_{\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathrm{~d}}}|\mathrm{ARB}(\overrightarrow{\mathrm{~F}})|^{2} \tag{3.14}
\end{equation*}
$$

We observe that the term $|\operatorname{ARB}(\overrightarrow{\mathrm{F}})|^{2}$ in (3.14) is equal to the number of elements in the set

$$
\left\{\left(\mathcal{A}, \mathcal{A}^{\prime}\right): \mathcal{A}, \mathcal{A}^{\prime} \in \operatorname{ARB}(\overrightarrow{\mathrm{F}})\right\} .
$$

That is,

$$
\mathbb{E}\left[\left(\mathscr{A}_{n, \mathbf{d}}^{\star}\right)^{2}\right]=\frac{\left|\widetilde{\Phi}_{n, \mathbf{d}}\right|}{\left|\Phi_{n, \mathbf{d}}\right|},
$$

where

$$
\widetilde{\Phi}_{n, \mathbf{d}}=\left\{\left(\overrightarrow{\mathrm{F}}, \mathscr{A}, \mathscr{A}^{\prime}\right): \overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}, \mathcal{A}, \mathscr{A}^{\prime} \in \mathrm{ARB}(\overrightarrow{\mathrm{~F}}) \mid\right\} .
$$

Hence, to evaluate $\mathbb{E}\left[\left(\mathscr{A}_{n, \mathbf{d}}^{\star}\right)^{2}\right]$ we need to count the number of elements of $\widetilde{\Phi}_{n, \mathbf{d}}$.
We compute $\left|\widetilde{\Phi}_{n, \mathbf{d}}\right|$ as follows. First, we count the number of ways to choose the intersection of a pair of arborescences $\mathcal{A}$ and $\mathcal{A}^{\prime}$. Then, we count the number of ways to extend this intersection to $\mathcal{A}$ and $\mathcal{A}^{\prime}$. Finally, we count the number of ways to choose the remainder of $\overrightarrow{\mathrm{F}}$ so that $\mathcal{A}$ and $\mathcal{A}^{\prime}$ are both in $\operatorname{ARB}(\overrightarrow{\mathrm{F}})$.

The last step is the easiest. Suppose we have a pair of arborescences $\left(\mathcal{A}, \mathcal{A}^{\prime}\right)$ of some configuration $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ and suppose $\mathcal{F}=\mathcal{A} \cap \mathcal{A}^{\prime}$ is a forest rooted at $R \subseteq V$. Since we need to add $|R|-1$ arcs to $\mathcal{F}$ for each arborescence, there will be $n+|R|-2$ edges in $\mathcal{A} \cup \mathcal{A}^{\prime}$ and, hence, there are $(m-n-|R|+2)$ ! ways to choose the remaining edges for $\vec{F}$.

We now proceed to enumerate the different pairs $\left(\mathcal{A}, \mathcal{A}^{\prime}\right)$ with $\mathcal{F}=\mathcal{A} \cap \mathcal{A}^{\prime}$ rooted at $R$. In fact, we overcount slightly, with the number of times $\left(\mathcal{A}, \mathscr{A}^{\prime}\right)$ is counted depending on the roots of $\mathcal{A}$ and $\mathcal{A}^{\prime}$. We use the BEST Theorem (Theorem 1.57) to get back to the correct number at the end of the proof.

We start by counting the number of ways we can choose $\mathcal{F}$, the edges in both arborescences, and then count the number of ways to choose the edges which are in one or the other arborescence. By Lemma 3.27, the number of ways to choose $\mathcal{F}$ rooted at $R$ is

$$
\begin{equation*}
\left(\prod_{v \notin R} d_{v}\right)\left(\sum_{v \in R} d_{v}\right)(m-1)_{n-|R|-1} . \tag{3.15}
\end{equation*}
$$

For each $v \in R$, let $\mathcal{F}_{v}$ denote the component of $\mathcal{F}$ with root $v$, and let $x_{v}$ be the number of points in $\bigcup_{u \in \mathcal{F}_{v}} T_{u}$ not used by arcs in $\mathcal{F}$. That is,

$$
x_{v}=\sum_{u \in \mathcal{F}_{v}} d_{u}-\left|\mathcal{F}_{v}\right|+1 .
$$

Note that this is the number of points available to add incoming arcs to vertices of $\mathcal{F}_{v}$ when we are completing $\mathcal{A}$ and $\mathcal{A}^{\prime}$. Moreover, we have

$$
\sum_{v \in R} x_{v}=m-n+|R| .
$$

We now turn our attention to the number of ways to choose $\mathcal{A} \backslash \mathcal{A}^{\prime}$ and $\mathcal{A}^{\prime} \backslash \mathcal{A}$. Choosing the remaining arcs for $\mathscr{A}$ and $\mathscr{A}^{\prime}$ is equivalent to choosing a pair of disjoint configurations for trees on $R$ in which there are $x_{v}$ points available for the targets of arcs entering $v$ and $d_{v}$ points available for the sources of the arcs leaving $v$, for each $v \in R$.

Suppose we have already chosen $\mathcal{A} \backslash \mathcal{A}^{\prime}$ such that the root of $\mathcal{A}$ is $r$ and suppose that there are $\delta_{v}$ additional arcs directed towards vertices in $\mathcal{F}_{v}$, for each $v \in R$. Now, suppose we want to choose $\mathcal{A}^{\prime} \backslash \mathcal{A}$ such that the root of $\mathcal{A}^{\prime}$ is $r^{\prime}$, and, for the moment, suppose $r \neq r^{\prime}$. Choosing $\mathcal{A}^{\prime} \backslash \mathcal{A}$ amounts to choosing a tree on $R$ rooted at $r^{\prime}$ in which there are $x_{v}-\delta_{v}$ points available for arcs directed towards each $v, d_{v}-1$ points available for the source of the arc directed away from each $v \neq r$, and $d_{r}$ points available for the source of the arc directed away from $r$. Hence, by Lemma 3.27, we see that the number of ways to choose $\mathfrak{A}^{\prime} \backslash \mathcal{A}$ is

$$
\frac{\left(x_{r^{\prime}}-\delta_{r^{\prime}}\right) d_{r}}{\left(d_{r}-1\right)\left(d_{r^{\prime}}-1\right)}\left(\prod_{v \in R}\left(d_{v}-1\right)\right)(m-n)_{|R|-2} .
$$

Using Fact 3.25 , we can deduce that the number of ways to choose $\mathcal{A}$ is

$$
\frac{\prod_{v \in R} d_{v}}{d_{r}} \sum_{\substack{|\delta|=|R|-1 \\ \delta_{r} \geq 1}}\binom{|R|-2}{\delta_{r}-1 ; \delta_{v}: v \in R \backslash\{r\}} \prod_{v \in R}\left(x_{v}\right)_{\delta_{v}} .
$$

Therefore, the number of ways to complete $\mathcal{F}$ to $\mathcal{A} \cup \mathcal{A}^{\prime}$ is equal to

$$
\frac{\prod_{v \in R} d_{v}\left(d_{v}-1\right)}{\left(d_{r}-1\right)\left(d_{r^{\prime}}-1\right)}(m-n)_{|R|-2}
$$

times

$$
\begin{equation*}
\sum_{\substack{|\delta|=|R|-1 \\ \delta_{r} \geq 1}}\left(x_{r^{\prime}}-\delta_{r^{\prime}}\right)\binom{|R|-2}{\delta_{r}-1 ; \delta_{v}: v \in R \backslash\{r\}} \prod_{v \in R}\left(x_{v}\right)_{\delta_{v}} . \tag{3.16}
\end{equation*}
$$

We can divide (3.16) into two sums:

$$
\begin{equation*}
x_{r^{\prime}} \sum_{\substack{|\delta|=|R|-1 \\ \delta_{r} \geq 1}}\binom{|R|-2}{\delta_{r}-1 ; \delta_{v}: v \in R \backslash\{r\}} \prod_{v \in R}\left(x_{v}\right)_{\delta_{v}} \tag{3.17}
\end{equation*}
$$

and

$$
\begin{equation*}
-\sum_{\substack{|\delta|=|R|-1 \\ \delta_{r} \geq 1}} \delta_{r^{\prime}}\binom{|R|-2}{\delta_{r}-1 ; \delta_{v}: v \in R \backslash\{r\}} \prod_{v \in R}\left(x_{v}\right)_{\delta_{v}} . \tag{3.18}
\end{equation*}
$$

Applying Fact 3.23, we see that (3.17) and (3.18) are equal to $x_{r} x_{r^{\prime}}(m-n+|R|-1)_{|R|-2}$ and $-x_{r} x_{r^{\prime}}(|R|-2)(m-n+|R|-2)_{|R|-3}$, respectively. Hence, the number of ways to complete $\mathcal{F}$ to $\mathcal{A} \cup \mathcal{A}^{\prime}$ is

$$
\begin{equation*}
\frac{x_{r} x_{r^{\prime}}}{\left(d_{r}-1\right)\left(d_{r^{\prime}}-1\right)}\left(\prod_{v \in R} d_{v}\left(d_{v}-1\right)\right)(m-n+|R|-2)_{2|R|-4} . \tag{3.19}
\end{equation*}
$$

If $r=r^{\prime}$, we can apply an almost identical argument to show that the number of ways to complete $\mathcal{F}$ to $\mathcal{A} \cup \mathcal{A}^{\prime}$ is

$$
\begin{equation*}
\frac{x_{r}\left(x_{r}-1\right)}{d_{r}\left(d_{r}-1\right)}\left(\prod_{v \in R} d_{v}\left(d_{v}-1\right)\right)(m-n+|R|-2)_{2|R|-4} \tag{3.20}
\end{equation*}
$$

Multiplying (3.19) and (3.20) by $\left(d_{r}-1\right)\left(d_{r^{\prime}}-1\right)$ and $d_{r}\left(d_{r}-1\right)$, respectively, and summing over $r$ and $r^{\prime}$ gives

$$
\begin{equation*}
\left(\sum_{r \neq r^{\prime}} x_{r} x_{r^{\prime}}+\sum_{r} x_{r}\left(x_{r}-1\right)\right)\left(\prod_{v \in R} d_{v}\left(d_{v}-1\right)\right)(m-n+|R|-2)_{2|R|-4} . \tag{3.21}
\end{equation*}
$$

Since $\sum_{r \in R} x_{r}=m-n+|R|$, we have

$$
\begin{aligned}
\sum_{r \neq r^{\prime}} x_{r} x_{r^{\prime}}+\sum_{r} x_{r}\left(x_{r}-1\right) & =\sum_{r \in R} x_{r}\left(x_{r}-1+\sum_{r^{\prime} \neq r}\left(x_{r^{\prime}}\right)\right) \\
& =(m-n+|R|)(m-n+|R|-1) .
\end{aligned}
$$

Hence, (3.21) is equal to

$$
\left(\prod_{v \in R} d_{v}\left(d_{v}-1\right)\right)(m-n+|R|)_{2|R|-2} .
$$

Multiplying by the number of ways to choose $\mathcal{F}$, given in (3.15), and the number of ways to choose the portion of $\overrightarrow{\mathrm{F}}$ not contained in $\mathcal{A} \cup \mathcal{A}^{\prime}$, which is $(m-n-|R|+2)$ !, yields the following expression

$$
\begin{equation*}
\left(\prod_{v \in V} d_{v}\right)(m-1)!\left(\prod_{v \in R}\left(d_{v}-1\right)\right)\left(\sum_{v \in R} d_{v}\right) . \tag{3.22}
\end{equation*}
$$

The expression (3.22) over-counts the number of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ in which the intersection $\mathcal{A} \cap \mathcal{A}^{\prime}$ is a forest rooted at $R$. Each triple $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ in which $\mathcal{A}$ and $\mathcal{A}^{\prime}$ are
rooted at different vertices $u$ and $v$ is counted $\left(d_{u}-1\right)\left(d_{v}-1\right)$ times, and each triple $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathfrak{A}^{\prime}\right)$ in which $\mathcal{A}$ and $\mathfrak{A}^{\prime}$ are rooted at the same vertex $v$ is counted $d_{v}\left(d_{v}-1\right)$ times.

Only the second two factors of (3.22) depend on $R$. Summing these over all $R \subseteq V$ gives

$$
\begin{equation*}
\sum_{R \subseteq V}\left(\sum_{v \in R} d_{v}\right)\left(\prod_{v \in R}\left(d_{v}-1\right)\right) \tag{3.23}
\end{equation*}
$$

We can evaluate (3.23) by separating it into $n$ separate sums, each corresponding to the sum over $R \ni v$, for $v \in V$,

$$
\begin{equation*}
d_{v} \sum_{R \ni v} \prod_{u \in R}\left(d_{u}-1\right)=\left(d_{v}-1\right)\left(\prod_{u \in V} d_{u}\right) . \tag{3.24}
\end{equation*}
$$

Then, summing the right-hand side of (3.24) over each $v \in V$, and combining with the rest of (3.22), gives

$$
\begin{equation*}
\left(\prod_{v \in V} d_{v}\right)^{2}(m-n)(m-1)! \tag{3.25}
\end{equation*}
$$

We cannot immediately obtain the quantity we are looking for from (3.25) as it over-counts different triples by different amounts. However, by the BEST theorem (Theorem 1.57), we know that the number of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ in which $\mathcal{A}$ is rooted at $u$ and $\mathcal{A}^{\prime}$ is rooted at $v$ does not depend on $u$ or $v$, since the projection $\sigma(\overrightarrow{\mathrm{F}})$ is always an Eulerian directed graph. Thus, it follows that the factor by which (3.25) over-counts the number of triples is

$$
\begin{equation*}
\frac{1}{n^{2}}\left(\sum_{u \neq v}\left(d_{u}-1\right)\left(d_{v}-1\right)+\sum_{v} d_{v}\left(d_{v}-1\right)\right)=\frac{(m-n+1)(m-n)}{n^{2}} . \tag{3.26}
\end{equation*}
$$

Then, dividing (3.25) by (3.26) and $m$ ! gives

$$
\mathbb{E}\left[\left(\mathscr{A}_{n, \mathbf{d}}^{\star}\right)^{2}\right]=\frac{n^{2}}{m(m-n+1)}\left(\prod_{v \in V} d_{v}\right)^{2} .
$$

An Eulerian directed graph is simple if and only if it does not contain any loops or double arcs; we do, however, allow the arcs $(u, v)$ and $(v, u)$ to be present. In Theorem 3.30 below, we estimate the first and second moments of $\mathcal{A}_{n, \mathbf{d}}$ using Theorem 3.28 and conditioning on there being no loops or double arcs in $\overrightarrow{\mathrm{F}}$.

Before proceeding with the proof of Theorem 3.30, we state a useful property of the directed configuration model. A similar property is used, implicitly or explicitly, in almost all results on the undirected configuration model [100].

Lemma 3.29. For any fixed directed graph $H$ with more arcs than vertices, and $\overrightarrow{\mathrm{F}}$ chosen uniformly at random from $\Phi_{n, \mathbf{d}}$, the probability $\sigma(\overrightarrow{\mathrm{F}})$ contains $H$ as a subgraph tends to 0 , as $m \rightarrow \infty$.

Proof. The reason for this is as follows. Suppose $H$ has $i$ vertices and $j$ edges. For each set $S \subset\{1,2, \ldots, n\}$ with $|S|=i$, the number of ways we can choose configuration pairs that will give $H$ as a subgraph of $\sigma(F)$ on the vertices $S$ is bounded by a constant. The probability of one of these sets of configuration pairs occurring in a random $F \in \Phi_{n, \mathbf{d}}$ is

$$
\frac{(m-j)!}{m!} \rightarrow m^{-j}
$$

since there are $(m-j)$ ! ways to choose the remainder of a configuration, given a set of $j$ edges. But the number of ways to choose $S$ is

$$
\binom{n}{i} \rightarrow n^{i} .
$$

Hence, we can, rather crudely, bound the probability of $H$ occurring as a subgraph of $\sigma(\overrightarrow{\mathrm{F}})$ by $O\left(m^{i-j}\right)$.

We now proceed with the proof of Theorem 3.30. We will use (the ideas of) Lemma 3.29 to show that the factorial moments of various random variables converge.

Theorem 3.30. Let $d$ be some fixed constant, $n \in \mathbb{N}$, and let $\mathbf{d}$ be a sequence of integers satisfying

$$
d \geq d_{n} \geq d_{n-1} \geq \cdots \geq d_{1} \geq 1
$$

and let $m=\sum_{v=1}^{n} d_{v}$. Furthermore, suppose $m-n \rightarrow \infty$. Let $\mathcal{A}_{n, \mathbf{d}}$ denote the number of arborescences of an Eulerian directed graph chosen randomly from $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$. Then,

$$
\begin{gathered}
\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right] \rightarrow e^{1} \frac{n}{m} \prod_{v} d_{v} \\
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{2}\right]}{\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]^{2}} \rightarrow e^{-n / m} \frac{m}{m-n},
\end{gathered}
$$

Proof. In the following we will use $m_{2}$ to denote $\sum_{v} d_{v}^{2}$.
The proof is as follows. Recall that $\overrightarrow{\mathrm{F}}$ contains a loop at $v$ if there is an edge from $S_{v} \times T_{v}$ in $\overrightarrow{\mathrm{F}}$ and that $\overrightarrow{\mathrm{F}}$ contains a double arc from $u$ to $v$ if there is a pair of edges from $S_{u} \times T_{v}$ in $\overrightarrow{\mathrm{F}}$, for some pair of vertices $u \neq v$. Let $L$ and $D$ denote the number of loops and double arcs in a random $\vec{F} \in \Phi_{n, \mathbf{d}}$. Then, the event " $\vec{F}$ is simple" is equivalent to the event $\{L=D=0\}$. We first analyse the distributions of $L$ and $D$, which we can
use to estimate the probability that $\overrightarrow{\mathrm{F}}$ is simple. Then, we consider two new random variables, $L^{(1)}$ and $D^{(1)}$, which count the number of loops and double arcs in $\overrightarrow{\mathrm{F}}$ when $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ is chosen randomly from the set

$$
\begin{equation*}
\bar{\Phi}_{n, \mathbf{d}}=\left\{(\overrightarrow{\mathrm{F}}, \mathcal{A}): \overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}, \mathcal{A} \in \operatorname{ARB}(\overrightarrow{\mathrm{~F}})\right\} . \tag{3.27}
\end{equation*}
$$

By analysing the distributions of $L^{(1)}$ and $D^{(1)}$ we can estimate

$$
\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]=\frac{\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right]}{\mathbb{P}[L=D=0]} \mathbb{E}\left[\mathscr{A}_{n . \mathbf{d}}^{\star}\right] .
$$

Finally, we consider random variables, $L^{(2)}$ and $D^{(2)}$, which count the number of loops and double arcs in $\overrightarrow{\mathrm{F}}$ when $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ is chosen randomly from the set

$$
\begin{equation*}
\widetilde{\Phi}_{n, \mathbf{d}}=\left\{\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right): \overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}, \mathcal{A}, \mathcal{A}^{\prime} \in \operatorname{ARB}(\overrightarrow{\mathrm{F}})\right\} . \tag{3.28}
\end{equation*}
$$

By analysing the distributions of $L^{(2)}$ and $D^{(2)}$ we can estimate

$$
\mathbb{E}\left[\left(\mathcal{A}_{n, \mathbf{d}}\right)^{2}\right]=\frac{\mathbb{P}\left[L^{(2)}=D^{(2)}=0\right]}{\mathbb{P}[L=D=0]} \mathbb{E}\left[\left(\mathscr{A}_{n . \mathbf{d}}^{\star}\right)^{2}\right]
$$

We first compute the expectation of $L$ and $D$. Suppose we have a loop edge $e \in S_{v} \times T_{v}$ in $\overrightarrow{\mathrm{F}}$ and let $I_{e}$ be the indicator variable for the event $e \in F$. Then, we can write $L=\sum_{v \in V} \sum_{e \in S_{v} \times T_{v}} I_{e}$ and, by linearity of expectation, we have

$$
\begin{equation*}
\mathbb{E}[L]=\sum_{v \in V} \sum_{e \in S_{v} \times T_{v}} \mathbb{E}\left[I_{e}\right]=\sum_{v \in V} \sum_{e \in S_{v} \times T_{v}} \mathbb{P}[e \in F] . \tag{3.29}
\end{equation*}
$$

Given $e$, the number of ways to choose $\overrightarrow{\mathrm{F}}$ with $e \in \overrightarrow{\mathrm{~F}}$ is $(m-1)$ !, so the probability of a random $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ containing $e$ is $1 / m$. For each $v \in V$, there are $d_{v}^{2}$ ways to choose an edge from $S_{v} \times T_{v}$. Hence,

$$
\begin{equation*}
\mathbb{E}[L]=\frac{1}{m} \sum_{v} d_{v}^{2}=\frac{m_{2}}{m} . \tag{3.30}
\end{equation*}
$$

Next, we compute the expectation of $D$. Here, for every pair of edges $e, f \in S_{u} \times T_{v}$, for some $u \neq v$, we define an indicator variable $I_{e, f}$ for the event $e, f \in \overrightarrow{\mathrm{~F}}$. Then $D=$ $\sum_{u \neq v} \sum_{e, f \in S_{u} \times T_{v}} I_{e, f}$. By linearity of expectation, we have

$$
\begin{equation*}
\mathbb{E}[D]=\sum_{u \in V} \sum_{v \in V \backslash\{u\}} \sum_{e, f \in S_{u} \times T_{v}} \mathbb{P}[e, f \in \overrightarrow{\mathrm{~F}}] . \tag{3.31}
\end{equation*}
$$

The probability of a particular pair of edges $e$ and $f$ occurring in a random configuration $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ is, asymptotically, $1 / \mathrm{m}^{2}$. Moreover, the number of ways to choose
$e, f \in S_{u} \times T_{v}$ is $2\binom{d_{u}}{2}\binom{d_{v}}{2}$. Hence, the sum in (3.31) becomes

$$
\begin{align*}
\mathbb{E}[D] & \rightarrow \frac{2}{m^{2}} \sum_{u \in V} \sum_{v \in V \backslash\{u\}}\binom{d_{u}}{2}\binom{d_{v}}{2} \\
& =\frac{1}{2 m^{2}}\left(\sum_{u \in V}\left(d_{u}\right)_{2}\right)^{2}-\frac{1}{2 m^{2}} \sum_{u \in V}\left(d_{u}\right)_{2}^{2} . \tag{3.32}
\end{align*}
$$

To finish the calculation we observe that the numerator of the negative term in (3.32) is $O(m)$ (each $d_{u}$ is bounded above by a constant $d$, so $\sum_{u}\left(d_{u}\right)_{2}^{2} \leq d^{3} m$ ). Hence, this part of the sum disappears as $m \rightarrow \infty$ and

$$
\begin{equation*}
\mathbb{E}[D] \rightarrow \frac{\left(m_{2}-m\right)^{2}}{2 m^{2}} \tag{3.33}
\end{equation*}
$$

We would like to use Theorem 3.14 to show that $L$ and $D$ are asymptotically Poisson, so we can use (3.30) and (3.33) to estimate $\mathbb{P}[L=D=0]$. To apply Theorem 3.14, we need to show that, for every pair of non-negative integers $i$ and $j$, we have

$$
\begin{equation*}
\mathbb{E}\left[(L)_{j}(D)_{k}\right] \rightarrow \mathbb{E}[L]^{j} \mathbb{E}[D]^{k} . \tag{3.34}
\end{equation*}
$$

Consider ordered tuples of the form $\mathbf{t}=\left(e_{1}, e_{2}, \ldots, e_{j},\left(f_{1}, g_{1}\right),\left(f_{2}, g_{2}\right), \ldots,\left(f_{k}, g_{k}\right)\right)$ where $e_{i}$ is a loop, for $i=1,2, \ldots, j,\left(f_{i}, g_{i}\right)$ is a double arc for $i=1,2, \ldots, k$, and all edges are distinct. For each such $\mathbf{t}$, we define an indicator variable $I_{\mathbf{t}}$ for the event $\mathbf{t} \in \overrightarrow{\mathrm{F}}(\overrightarrow{\mathrm{F}}$ contains all components of $\mathbf{t})$. Then, we can write

$$
\begin{equation*}
\mathbb{E}\left[(L)_{j}(D)_{k}\right]=\sum_{\mathbf{t}} \mathbb{E}\left[I_{\mathbf{t}}\right]=\sum_{\mathbf{t}} \mathbb{P}[\mathbf{t} \in \overrightarrow{\mathrm{F}}] . \tag{3.35}
\end{equation*}
$$

First, we consider the contribution to $\mathbb{E}\left[(L)_{j}(D)_{k}\right]$ from tuples where no vertex occurs more than once. We can write this as

$$
\begin{equation*}
\frac{1}{2^{k}} \sum_{v_{1} \in V_{1}} \frac{d_{v_{1}}^{2}}{m} \sum_{v_{2} \in V_{2}} \frac{d_{v_{2}}^{2}}{m-1} \cdots \sum_{v_{j+1} \in V_{j+1}} \frac{\left(d_{v_{j+1}}\right)_{2}}{m-j} \cdots \sum_{v_{2 k} \in V_{j+2 k}} \frac{\left(d_{v_{2 k}}\right)_{2}}{m-2 k}, \tag{3.36}
\end{equation*}
$$

where $V_{i}=V-\left\{v_{j}: j<i\right\}$. We can re-write the last sum as

$$
\begin{equation*}
\frac{1}{m-2 k}\left(m_{2}-m-\sum_{v \in V_{2 k}}\left(d_{v_{j+2 k}}\right)_{2}\right) . \tag{3.37}
\end{equation*}
$$

Since $k$ and $d$ are both fixed, it follows that (3.37) is, asymptotically, $\left(m_{2}-m\right) / m$. We can apply the same reasoning to show that each of the sums over $V_{i}$, for $i=j+1, \ldots, 2 k$, converge to $\left(m_{2}-m\right) / m$. Similarly, each of the sums over $V_{i}$, for $i=1,2, \ldots, j$, converges to $m_{2} / m$. Hence, the asymptotic value of (3.36) is

$$
\mathbb{E}[L]^{j} \mathbb{E}[D]^{k}
$$

By Lemma 3.29, we know that the probability of $\vec{F}$ containing any set of cycles on $j$ vertices, for fixed $j$, in which any vertex occurs more than once, is (asymptotically) 0 . Hence, the contribution to (3.35) from ordered tuples in which any vertex features more than once is negligible and, therefore, (3.34) holds.

We have shown that $L$ and $D$ converge to independent Poisson random variables and, therefore, the probability that $\vec{F}$ is simple when $\vec{F}$ is chosen uniformly at random from $\Phi_{n, \mathbf{d}}$ is (asymptotically)

$$
\begin{equation*}
\exp \left(-\frac{m_{2}}{m}-\frac{\left(m_{2}-m\right)^{2}}{2 m^{2}}\right) \tag{3.38}
\end{equation*}
$$

Next, we consider the distributions of $L^{(1)}$ and $D^{(1)}$. We first estimate $\mathbb{E}\left[L^{(1)}\right]$. Suppose we have a loop edge $e \in S_{v} \times T_{v}$, for some $v \in V$. A loop edge cannot be contained in any arborescence, and, thus, the number of pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}$ where $e \in \overrightarrow{\mathrm{~F}}$, is equal to the number of pairs $(\overrightarrow{\mathrm{F}}, \mathscr{A}) \in \bar{\Phi}_{n, \mathbf{d}^{\prime}}$, where $\mathbf{d}^{\prime}$ is equal to $\mathbf{d}$ with $d_{v}$ replaced by $d_{v}-1$. Hence, from Theorem 3.28, we can see that the number of elements of $\bar{\Phi}_{n, \mathbf{d}}$ with $e \in \overrightarrow{\mathrm{~F}}$ is equal to

$$
\begin{equation*}
n\left(d_{v}-1\right) \prod_{u \neq v} d_{u}(m-2)! \tag{3.39}
\end{equation*}
$$

Dividing (3.39) by the total number of elements in $\bar{\Phi}_{n . \mathrm{d}}$, which we can also obtain from Theorem 3.28, gives the probability

$$
\begin{equation*}
\mathbb{P}\left[e \in \overrightarrow{\mathrm{~F}}:(\overrightarrow{\mathrm{F}}, \mathscr{A}) \in \bar{\Phi}_{n, \mathbf{d}}\right]=\frac{d_{v}-1}{d_{v}(m-1)} \tag{3.40}
\end{equation*}
$$

Evaluating (3.29) with this probability in the place of $\mathbb{P}[e \in F]$ gives

$$
\mathbb{E}\left[L^{(1)}\right]=\frac{1}{m-1} \sum_{v} d_{v}\left(d_{v}-1\right) \rightarrow \frac{m_{2}-m}{m} .
$$

Next, we evaluate $\mathbb{E}\left[D^{(1)}\right]$. Suppose we have a pair of edges $e, f \in S_{u} \times T_{v}$ for some $u \neq v$. By Lemma 3.27, the number of arborescences rooted at $u$ in which each $w \notin\{u, v\}$ has $d_{w}$ points available for its incoming and outgoing arcs, $u$ has $d_{u}$ points available for incoming arcs, and $v$ has $d_{v}-2$ points available for incoming arcs and $d_{v}$ available for outgoing arcs is

$$
\begin{equation*}
\left(\prod_{w \in V} d_{w}\right)(m-3)_{n-2} . \tag{3.41}
\end{equation*}
$$

The expression in (3.41) counts the number of partial configurations which consist of the edges $e$ and $f$ along with $n-1$ configuration edges that project to an arborescence
rooted at $u$. There are $(m-n-1)$ ! ways to extend each of these partial configurations to some $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$. Hence, the following expression counts the number of pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}$ with $e, f \in \overrightarrow{\mathrm{~F}}$ and $\mathcal{A}$ rooted at $u$.

$$
\begin{equation*}
\left(\prod_{w=1}^{n} d_{w}\right)(m-3)!. \tag{3.42}
\end{equation*}
$$

By the BEST Theorem (Theorem 1.57), we know that each $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ has the same number of arborescences rooted at each vertex, so (3.42) counts exactly $1 / n$ of the pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}$ with $e, f \in \overrightarrow{\mathrm{~F}}$. Multiplying (3.42) by $n$ and dividing by $\left|\bar{\Phi}_{n, \mathbf{d}}\right|$ gives

$$
\begin{equation*}
\mathbb{P}\left[e, f \in \overrightarrow{\mathrm{~F}}:(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}\right] \rightarrow \frac{1}{m^{2}} . \tag{3.43}
\end{equation*}
$$

This is the same probability as when $\vec{F}$ is chosen uniformly at random from $\Phi_{n, \mathbf{d}}$, so evaluating (3.33) with (3.43) in place of $\mathbb{P}[e, f \in F]$ does not change the (asymptotic) value and we have

$$
\mathbb{E}\left[D^{(1)}\right] \rightarrow \mathbb{E}[D] .
$$

As described in Lemma 3.29, the probability that a random $\mathrm{F} \in \Phi_{n, \mathbf{d}}$ contains a particular $k$-vertex subgraph with more arcs than vertices tends to 0 . This is also true when we are sampling $(\tilde{\mathrm{F}}, \mathcal{A})$ uniformly at random from $\bar{\Phi}_{n, \mathbf{d}}$. Hence, the contribution to $\mathbb{E}\left[\left(L^{(1)}\right)_{j}\left(D^{(1)}\right)_{k}\right]$, from ordered tuples of loops and double arcs which overlap at any vertex is negligible. We can thus conclude that the factorial moments converge to

$$
\left.\mathbb{E}\left[\left(L^{(1)}\right)_{j}\left(D^{(1)}\right)_{k}\right] \rightarrow \mathbb{E}\left[L^{(1)}\right]^{j} \mathbb{E} D^{(1)}\right]^{k}
$$

Hence, the probability of $\overrightarrow{\mathrm{F}}$ being simple in a random $(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}$ is

$$
\begin{equation*}
\exp \left(-\frac{m_{2}-m}{m}-\frac{\left(m_{2}-m\right)^{2}}{2 m^{2}}\right) . \tag{3.44}
\end{equation*}
$$

Together (3.38) and (3.44) give the claimed estimate for $\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]$.
Finally, we consider the distributions of $L^{(2)}$ and $D^{(2)}$. Suppose we have a loop edge $e \in S_{v} \times T_{v}$. The number of elements of $\widetilde{\Phi}_{n, \mathbf{d}}$ with $e \in \overrightarrow{\mathrm{~F}}$ is equal to the number of elements of $\widetilde{\Phi}_{n, \mathbf{d}^{\prime}}$, where $\mathbf{d}^{\prime}$ is the out-degree vector we used to compute $\mathbb{E}\left[L^{(1)}\right]$. By Theorem 3.28, we have

$$
\left|\widetilde{\Phi}_{n, \mathbf{d}}\right|=\frac{\left(d_{v}-1\right)^{2}}{\left(d_{v}\right)^{2}} \frac{n^{2}}{m-n}\left(\prod_{w \in V} d_{w}\right)^{2}(m-2)!.
$$

Dividing by the number of elements in $\widetilde{\Phi}_{n, \mathbf{d}}$, which know from Theorem 3.28, we see that

$$
\mathbb{P}\left[e \in \overrightarrow{\mathrm{~F}}:\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right) \in \widetilde{\Phi}_{n, \mathbf{d}}\right] \rightarrow \frac{\left(d_{v}-1\right)^{2}}{\left(d_{v}\right)^{2} m}
$$

Evaluating (3.30) with this probability in the place of $\mathbb{P}[e \in F]$ gives

$$
\begin{equation*}
\mathbb{E}\left[L^{(2)}\right] \rightarrow \frac{m_{2}-2 m+n}{m} . \tag{3.45}
\end{equation*}
$$

We now evaluate $\mathbb{E}\left[D^{(2)}\right]$. Suppose we have a pair of edges $e, f \in S_{u} \times T_{v}$ for some $u \neq v$. There are three cases to consider: $e, f \in \mathcal{A} \cup \mathcal{A}^{\prime} ; e, f \notin \mathcal{A} \cup \mathcal{A}^{\prime} ;$ or exactly one of $e$ and $f$ is in $\mathcal{A} \cup \mathcal{A}^{\prime}$. We estimate $\mathbb{E}\left[D^{(2)}\right]$ as follows. Using slightly more general arguments than those used to compute the second moment in Theorem 3.28, we count the number of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ for each of these three cases, obtaining expressions which overcount in the same way as (3.25). Then, since the way in which triples are over-counted is the same in each of the three analyses, i.e., the number of times each triple $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathscr{A}^{\prime}\right)$ is counted is determined by the out-degrees of the roots of $\mathcal{A}$ and $\mathcal{A}^{\prime}$, we can add these three expressions together, apply the BEST theorem, and proceed as we did in the proof of Theorem 3.28.

Note that the fact that the in-degree and out-degree of each vertex $v$ are equal is only used at the last step of the analysis of the second moment of $\mathscr{A}_{n, \mathbf{d}}^{\star}$ (in Theorem 3.28). That is, if we are working in the general directed configuration model of Definition 3.3 (with the added condition that $\sum s_{v}=\sum t_{v}=m$ ) and follow the arguments of the second part of Theorem 3.28 we find that, for each $R \subseteq V$, the expression over-counting triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ where $\mathcal{A} \cap \mathcal{A}^{\prime}$ is a forest rooted at $R$ (given by (3.22) in the proof of Theorem 3.28) becomes

$$
\begin{equation*}
\left(\prod_{w \in V} s_{w}\right)(m-1)!\left(\sum_{w \in R} t_{w}\right)\left(\prod_{w \in R}\left(s_{w}-1\right)\right) . \tag{3.46}
\end{equation*}
$$

The factor by which (3.46) over-counts $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ is $\left(s_{r}-1\right)\left(s_{r^{\prime}}-1\right)$ if $\mathcal{A}$ and $\mathcal{A}^{\prime}$ are rooted at different vertices $r, r^{\prime} \in R$, and is $s_{r}\left(s_{r}-1\right)$ if both are rooted at the same vertex $r \in R$.

As we did for (3.22) in the proof of Theorem 3.28, we can separate (3.46) into two parts: the part that does not depend on $R$,

$$
\left(\prod_{w \in V} s_{w}\right)(m-1)!,
$$

and the part that does

$$
\begin{equation*}
\left(\sum_{w \in R} t_{w}\right)\left(\prod_{w \in R}\left(s_{w}-1\right)\right) . \tag{3.47}
\end{equation*}
$$

Summing (3.47) over all possibilities for $R$ gives

$$
\left(\sum_{w \in V} \frac{t_{w}\left(s_{w}-1\right)}{s_{w}}\right)\left(\prod_{w \in V} s_{w}\right)
$$

Hence, in the general directed configuration model, the expression which over-counts triples of the form $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathscr{A}^{\prime}\right)$ is

$$
\begin{equation*}
\left(\sum_{w \in V} \frac{t_{w}\left(s_{w}-1\right)}{s_{w}}\right)\left(\prod_{w \in V} s_{w}\right)^{2}(m-1)! \tag{3.48}
\end{equation*}
$$

Now, suppose $e, f \notin \mathcal{A} \cup \mathcal{A}^{\prime}$. In this case we want to count the number of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathscr{A}^{\prime}\right)$ where $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{s}, \mathbf{t}}$, for $\mathbf{s}=\mathbf{d}$, except $s_{u}=d_{u}-2$, and $\mathbf{t}=\mathbf{d}$, except $t_{v}=d_{v}-2$. Adding $e$ and $f$ to the $\overrightarrow{\mathrm{F}}$ part of each element of this set of triples gives the set of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right) \in \widetilde{\Phi}_{n, \mathbf{d}}$ with $e, f \in \overrightarrow{\mathrm{~F}}-\left(\mathcal{A} \cup \mathcal{A}^{\prime}\right)$. Evaluating (3.48) for the stated values of $\mathbf{s}$ and $\mathbf{t}$ gives

$$
\left(m-n-\frac{d_{u}}{d_{u}-2}-\frac{d_{v}-2}{d_{v}}\right) \frac{\left(d_{u}-2\right)^{2}}{d_{u}^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-3)!,
$$

or, asymptotically, as $m-n \rightarrow \infty$,

$$
\begin{equation*}
(m-n) \frac{\left(d_{u}-2\right)^{2}}{d_{u}^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-3)! \tag{3.49}
\end{equation*}
$$

Next, suppose $e, f \in \mathcal{A} \cup \mathcal{A}^{\prime}$. Since there can be at most one arc leaving $u$ in $\mathcal{A}$ or $\mathscr{A}^{\prime}$ it follows that we have an $\operatorname{arc}(u, v)$ in both $\mathcal{A}$ and $\mathcal{A}^{\prime}$. Hence, when we are choosing the pair of arborescences we must assume that $(u, v)$ is always present. This corresponds to replacing $u$ and $v$ by a single vertex $v^{\prime}$ which has $d_{v}$ points available for outgoing arcs and $d_{u}+d_{v}-2$ points available for incoming arcs. That is, in this instance we want to count triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ with $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{s}, \mathbf{t}}$, where $\mathbf{s}$ and $\mathbf{t}$ are vectors indexed by $V \backslash\{u, v\}$ and $v^{\prime}$ satisfying $s_{w}=t_{w}=d_{w}$ for $w \in V \backslash\{u, v\}, s_{v^{\prime}}=d_{v}$, and $t_{v^{\prime}}=d_{u}+d_{v}-2$. Evaluating (3.48) with these values for $\mathbf{s}$ and $\mathbf{t}$ yields, asymptotically,

$$
\begin{equation*}
(m-n) \frac{1}{\left(d_{u}\right)^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-3)!. \tag{3.50}
\end{equation*}
$$

We can extend each of the triples counted by (3.50) to an element of $\widetilde{\Phi}_{n, \mathbf{d}}$ in four ways, since we can put $e$ or $f$ into either arborescence. Hence, the expression over-counting triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right) \in \widetilde{\Phi}_{n, \mathbf{d}}$ with $e, f \in \mathscr{A} \cup \mathcal{A}^{\prime}$ is, asymptotically,

$$
\begin{equation*}
(m-n) \frac{4}{\left(d_{u}\right)^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-3)!. \tag{3.51}
\end{equation*}
$$

Finally, suppose exactly one of $e$ and $f$ is in $\mathcal{A} \cup \mathcal{A}^{\prime}$. This case is a little bit more complicated. Suppose $e \in \mathcal{A}$. We contract $u$ and $v$ to a single vertex $v^{\prime}$, as in the
previous case, and choose a forest on $V \backslash\{u, v\}$ and $v^{\prime}$, with root $R$. Then, we proceed as in the proof of Theorem 3.28, except we consider $u$ to be one of the roots of the components of $\mathcal{F}$ when choosing $\mathscr{A}^{\prime}$; that is, when choosing $\mathcal{A} \backslash \mathcal{A}^{\prime}$ we choose a tree on $R$, but when choosing $\mathcal{A}^{\prime} \backslash \mathcal{A}$ we choose a tree on $R \cup\{u\}$, where the vertices in the component of our initial forest have now been divided between a component rooted at $u$ and a component containing $v$. The final expression counting the number of ways to complete $\mathcal{F}$ to $\mathscr{A} \cup \mathscr{A}^{\prime}$ does not depend on how the vertices are distributed amongst the components of $\mathcal{F}$, so it is safe to do this. In this way, we obtain the expression

$$
\begin{equation*}
\left(d_{u}-2\right)\left(\sum_{w \in R} t_{w}\right)\left(\prod_{w \in R}\left(s_{w}-1\right)\right)\left(\prod_{w \in V \backslash\{u\}} s_{w}\right)(m-1)! \tag{3.52}
\end{equation*}
$$

which over-counts the number of triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathscr{A}^{\prime}\right)$ where $R \cup\{u\}$ are the roots of the components of $\mathcal{A} \cap \mathcal{A}^{\prime}, e \in \mathcal{A}$, and $f \in \overrightarrow{\mathrm{~F}} \backslash\left(\mathcal{A} \cup \mathcal{A}^{\prime}\right)$.

Proceeding as before, by summing over all possibilities for $R$, gives, asymptotically,

$$
(m-n) \frac{d_{u}-2}{\left(d_{u}\right)^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-1)!
$$

The cases where $e \in \mathcal{A}^{\prime}, f \in \mathcal{A}$, and $f \in \mathcal{A}^{\prime}$ are all equivalent, so the expression overcounting triples $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ with exactly one of $e$ and $f$ in $\mathcal{A} \cup \mathcal{A}^{\prime}$ is, asymptotically,

$$
\begin{equation*}
(m-n) \frac{4\left(d_{u}-2\right)}{\left(d_{u}\right)^{2}}\left(\prod_{w \in V} d_{w}\right)^{2}(m-1)! \tag{3.53}
\end{equation*}
$$

Adding (3.49), (3.51), and (3.53) gives

$$
\begin{equation*}
(m-n)\left(\prod_{w \in V} d_{w}\right)^{2}(m-3)!. \tag{3.54}
\end{equation*}
$$

Finally, we observe that the triples are over-counted consistently in the three separate constructions given above. Hence, we can conclude, by the same reasoning as was used in Theorem 3.28, that (3.54) over-counts the elements of $\widetilde{\Phi}_{n, \mathbf{d}}$ by a factor of

$$
\begin{equation*}
\frac{(m-n-1)(m-n-2)}{n^{2}} . \tag{3.55}
\end{equation*}
$$

Dividing (3.54) by (3.55) and the number of elements in $\widetilde{\Phi}_{n, \mathrm{~d}}$, which we can obtain using Theorem 3.28, gives

$$
\mathbb{P}\left[e, f \in \overrightarrow{\mathrm{~F}}:\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right) \in \widetilde{\Phi}_{n, \mathbf{d}}\right] \rightarrow \frac{1}{m^{2}} .
$$

This is the same probability for $e, f \in \overrightarrow{\mathrm{~F}}$ when $\overrightarrow{\mathrm{F}}$ is chosen uniformly at random from $\Phi_{n, \mathbf{d}}$, so we can conclude

$$
\mathbb{E}\left[D^{(2)}\right] \rightarrow \mathbb{E}[D]
$$

The factorial moments converge, again because we only need to consider the contribution from ordered tuples of non-overlapping loops and double arcs. Hence, the probability that $\overrightarrow{\mathrm{F}}$ is simple, when $\left(\overrightarrow{\mathrm{F}}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ is chosen uniformly at random from $\widetilde{\Phi}_{n, \mathbf{d}}$, is

$$
\begin{equation*}
\exp \left(-\frac{m_{2}-2 m+n}{m}-\frac{\left(m_{2}-m\right)^{2}}{2 m^{2}}\right) \tag{3.56}
\end{equation*}
$$

Combining (3.38) and (3.56) gives the claimed estimate for $\mathbb{E}\left(\mathcal{A}_{n, \mathbf{d}}\right)^{2}$.

### 3.3.2 The asymptotic distribution of $\mathcal{A}_{n, d}$

The primary goal in this section is to prove that $\overrightarrow{\mathcal{T}}_{n, \mathbf{d}}$ is concentrated around its mean. By the BEST Theorem, this is equivalent to proving a concentration result for $\mathcal{A}_{n, \mathbf{d}}$. The obvious first step is to try to apply Chebyshev's inequality (Theorem 3.8). Suppose we want to show

$$
\mathbb{P}\left\{\mathcal{A}_{n, \mathbf{d}} \leq n^{-\alpha} \mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]\right\} \rightarrow 0 .
$$

We can bound this probability by

$$
\begin{aligned}
\mathbb{P}\left\{\mathcal{A}_{n, \mathbf{d}} \leq n^{-\alpha} \mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]\right\} & \leq \mathbb{P}\left\{\left|\mathcal{A}_{n, \mathbf{d}}-\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]\right| \geq\left(1-n^{-\alpha}\right) \mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]\right\} \\
& \leq\left(\frac{n^{\alpha}-1}{n^{\alpha}}\right)^{2} \frac{\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}^{2}\right]-\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]^{2}}{\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]^{2}},
\end{aligned}
$$

where the last step is obtained by applying Chebyshev with $a=\left(1-n^{-\alpha}\right) \mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]$, and using the fact that $\operatorname{Var}[X]=\mathbb{E}\left[X^{2}\right]-\mathbb{E}[X]^{2}$ for any random variable $X$. Then, by Theorem 3.30, we can deduce that this upper bound is, asymptotically, equal to

$$
\frac{e^{-n / m} m}{m-n}-1 .
$$

This will be a constant, since we are assuming that $d_{v} \leq d$ for each $v \in V$. Thus, to prove the result we want, we will need something stronger.

In this section, we show how we can apply Janson's Theorem (Theorem 3.19) to determine the asymptotic distribution of $\mathcal{A}_{n, d}$ (the special case where $d_{v}=d$ for all $v \in V)$. In our application of Theorem 3.19 we will have $Y_{n}=\mathcal{A}_{n, d}$, and $X_{i, n}$ counting the number of directed $i$-cycles in $\sigma(\overrightarrow{\mathrm{F}})$ for $\overrightarrow{\mathrm{F}}$ chosen uniformly at random from $\Phi_{n, d}$.

In the following two lemmas (Lemma 3.31 and Lemma 3.32), we establish that Condition 1 and Condition 2 of Theorem 3.19 hold when we take $X_{i, n}$ to count directed $i$-cycles in a random $\vec{F} \in \Phi_{n, \mathbf{d}}$ and $Y_{n}$ to be equal to $\mathscr{A}_{n, \mathbf{d}}^{\star}$.

Lemma 3.31. For each fixed positive integer $i$, let $X_{i, n}$ count the number of directed $i$-cycles in a directed graph $\vec{G}$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, \mathbf{d})$, and let $\lambda_{i}=\lambda^{i} / i$, where $\lambda=m_{2} / m$. Then, for any fixed set of integers $j_{2}, \ldots, j_{k}$ we have

$$
\mathbb{E}\left[\prod_{i=2}^{k}\left(X_{i}\right)_{j_{i}}\right] \rightarrow \prod_{i=1}^{k} \lambda_{i}^{j_{i}} .
$$

Proof. We say a set of $i$ edges $e_{1}, e_{2}, \ldots, e_{i}$ in a directed configuration is an $i$-cycle if there is a sequence of distinct vertices $v_{0}, v_{1}, \ldots, v_{i-1}$ such that $e_{j} \in S_{v_{j}} \times T_{v_{j+1} \bmod i}$ for $j=0,1, \ldots, i-1$. Let $\mathcal{C}_{i, \mathbf{d}}$ denote the set of all possible $i$-cycles for a particular out-degree vector $\mathbf{d}$. For each $C \in \mathcal{C}_{i, \mathbf{d}}$, we define an indicator variable $I_{C}$ for the event $C \in \overrightarrow{\mathrm{~F}}$, so $X_{i, n}=\sum_{C \in \mathcal{C}_{i, n}} I_{C}$. Then, by linearity of expectation, we can write $\mathbb{E}\left[X_{i, n}\right]$ as

$$
\begin{align*}
\mathbb{E}\left[X_{i, n}\right] & =\sum_{C \in \mathcal{C}_{i, \mathrm{~d}}} \mathbb{E}\left[I_{C}\right] \\
& =\sum_{C \in \mathcal{C}_{i, \mathrm{~d}}} \mathbb{P}[C \in \overrightarrow{\mathrm{~F}}] . \tag{3.57}
\end{align*}
$$

There are $(m-i)$ ! configurations containing each $i$-cycle, and, therefore, for any $i$-cycle C,

$$
\mathbb{P}[C \in \overrightarrow{\mathrm{~F}}]=\frac{(m-i)!}{m!} \rightarrow \frac{1}{m^{i}}
$$

So, to estimate $\mathbb{E}\left[X_{i, n}\right]$, we only need to compute $\left|\mathcal{C}_{i, \mathbf{d}}\right|$ and then divide by $m^{i}$. Let $V=\{1,2, \ldots, n\}$ and suppose $S$ is an $i$-subset of $V$. There are $(i-1)!$ different ways to arrange $S$ into an $i$-cycle $\left(v_{0}, v_{1}, \ldots, v_{i-1}\right)$. Given a particular $i$-cycle $\left(v_{0}, v_{1}, \ldots, v_{i-1}\right)$ on $S$, there are $\prod_{v \in S} d_{v}^{2}$ ways to choose edges $e_{j} \in S_{v_{j}} \times T_{v_{(j+1) \text { mod } i}}$, for $j=0,1, \ldots, i-1$. Hence, (3.57) becomes

$$
\begin{equation*}
\mathbb{E}\left[X_{i, n}\right] \rightarrow \frac{(i-1)!}{m^{i}} \sum_{\substack{S \subset V \\|S|=i}} \prod_{v \in S} d_{v}^{2} \tag{3.58}
\end{equation*}
$$

Now, observe that

$$
\begin{align*}
\left(\frac{m_{2}}{m}\right)^{i} & =\frac{1}{m^{i}}\left(\sum_{v \in V}\left(d_{v}\right)^{2}\right)^{i} \\
& =\frac{1}{m^{i}} \sum_{\sum s_{v}=i}\binom{i}{s_{v}: v \in V} \prod_{v \in V} d_{v}^{2 s_{v}} . \tag{3.59}
\end{align*}
$$

Consider the contribution to (3.59) from terms in which $s_{v}=2$ for some $v \in V$. This can be crudely bounded as

$$
\frac{(i)_{2}}{2 m^{i}} \sum_{v \in V}\left(d_{v}\right)^{4}\left(\sum_{u \in V \backslash\{v\}} d_{u}^{2}\right)^{i-2} \leq \frac{(i)_{2}}{2} d^{4} m^{-1} \rightarrow 0
$$

Similarly, the contribution from all terms with $s_{v}=j>2$ for some $v \in V$ can be shown to be $O\left(m^{i-j+1}\right)$. Hence, we have

$$
\begin{equation*}
\frac{i!}{m^{i}} \sum_{\substack{S S V \\|S|=i}} \prod_{v \in S}\left(d_{v}\right)^{2} \rightarrow\left(\frac{m_{2}}{m}\right)^{i} \tag{3.60}
\end{equation*}
$$

Plugging the approximation given in (3.60) into (3.58) gives

$$
\mathbb{E}\left[X_{i, n}\right] \rightarrow \frac{1}{i}\left(\frac{m_{2}}{m}\right)^{i}=\lambda_{i} .
$$

Suppose we have a sequence of fixed non-negative integers $j_{1}, j_{2}, \ldots, j_{k}$. To evaluate the factorial moment

$$
\begin{equation*}
\mathbb{E}\left[\prod_{i=1}^{m}\left(X_{i, n}\right)_{j_{i}}\right] \tag{3.61}
\end{equation*}
$$

we need to enumerate ordered tuples of $j=j_{1}+j_{2}+\cdots+j_{k}$ directed cycles where the first $j_{1}$ cycles are loops, the next $j_{2}$ are 2 -cycles, and so on. By Lemma 3.29, we know that the probability of any particular partial configuration with more edges than vertices occurring in a random $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ tends to 0 . Hence, the only tuples of directed cycles which have non-negligible contribution to (3.61) are those in which the cycles are vertex-disjoint. Given a set $S \subset V$ with $|S|=j$, there are

$$
\frac{j!}{\prod_{i=1}^{k} i^{j_{i}}} \prod_{v \in S}\left(d_{v}\right)^{2}
$$

ways to choose configurations for an ordered tuple of vertex-disjoint $j$-cycles, with $j_{i}$ being the number of $i$-cycles. The probability of any one of these sets of cycles occurring in a random configuration is, asymptotically, $1 / \mathrm{m}^{j}$. Hence, the contribution to (3.61) from tuples of vertex-disjoint cycles, is, asymptotically,

$$
\begin{equation*}
\frac{j!}{m^{j} \prod_{i=1}^{k} i^{j_{i}}} \sum_{\substack{S \subset V \\|S|=j}} \prod_{v \in S}\left(d_{v}\right)^{2} \tag{3.62}
\end{equation*}
$$

By the same arguments used to prove (3.60), we can show that (3.62) is (asymptotically) equal to

$$
\frac{1}{\prod_{i=1}^{k} j_{i}}\left(\frac{m_{2}}{m}\right)^{j}=\prod_{i=1}^{k} \lambda_{i}^{j_{i}} .
$$

Lemma 3.32. For each $i$, let $X_{i, n}$ be as in Lemma 3.31, and let $\mu_{i}=\left(\lambda^{i}-1\right) / i$, where $\lambda=m_{2} / m$. Then, for any fixed set of integers $j_{1}, j_{2}, \ldots, j_{k}$ we have

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} \prod_{i=1}^{k}\left(X_{i, n}\right)_{j_{i}}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}} .
$$

Proof. We first establish

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]} \rightarrow \mu_{i} .
$$

The factorial moments then hold for the same reason as was given in Lemma 3.31: the contribution from ordered tuples of loops and double arcs with repeated vertices is asymptotically insignificant.

Let $\bar{\Phi}_{n, \mathbf{d}}$ be the set of pairs $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ where $\overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}$ and $\mathcal{A}$ is an arborescence of $\overrightarrow{\mathrm{F}}$. We let $\mathcal{C}_{i}(F)$ and $\mathcal{C}_{i, \mathbf{d}}$ have the same meaning as in the proof of Lemma 3.31. Then, as we did for $\mathbb{E}\left[X_{i, n}\right]$, we can expand

$$
\begin{aligned}
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]} & =\frac{1}{\left|\bar{\Phi}_{n, \mathbf{d}}\right|} \sum_{(\overrightarrow{\mathrm{F}}, \mathscr{A}) \in \bar{\Phi}_{n, \mathbf{d}}} \sum_{C \in \mathcal{C}_{i, \mathbf{d}}} I_{C}(F) \\
& =\frac{\mid\left\{(\overrightarrow{\mathrm{F}}, \mathcal{A}, C): \overrightarrow{\mathrm{F}} \in \Phi_{n, \mathbf{d}}, \mathcal{A} \in \operatorname{ARB}(\overrightarrow{\mathrm{~F}}), C \in \mathcal{C}_{i}(F) \mid\right.}{\left|\bar{\Phi}_{n, \mathbf{d}}\right|} \\
& =\sum_{C \in \mathcal{C}_{i, \mathbf{d}}} \mathbb{P}\left[C \subset \overrightarrow{\mathrm{~F}}:(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}\right] .
\end{aligned}
$$

Hence, we only need to estimate the probability that a particular $C \in \mathcal{C}_{i, \mathbf{d}}$ is contained in $\overrightarrow{\mathrm{F}}$ when $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ is chosen uniformly at random from $\bar{\Phi}_{n, \mathbf{d}}$. To do this, we enumerate the elements of $\bar{\Phi}_{n, \mathbf{d}}$ with $C \subset \overrightarrow{\mathrm{~F}}$. This is done by first choosing which of the edges of $C$ are going to be in $\mathcal{A}$, then choosing the remaining edges for $\mathcal{A}$, and finally choosing the remaining edges for $\overrightarrow{\mathrm{F}}$.

Let $S$ denote the set of vertices of $C$ and let $R \subseteq S$ be the vertices $v \in S$ for which the arc leaving $v$ on $C$ is not contained in $\mathcal{A}$. Then, $C \cap \mathcal{A}$ consists of a set of disjoint directed paths, each one ending at some $v \in R$. For each $v \in R$, let $P_{v}$ denote the directed path in $C \cap \mathcal{A}$ ending at $v$. Choosing the remainder of $\mathcal{A}$ is then equivalent to choosing an arborescence on $(V \backslash S) \cup R$, where we have collapsed each path to a single vertex. Each $v \in V \backslash S$ has $d_{v}$ points for arcs entering and leaving $v$. For each $v \in R$, the number of points available for arcs entering $v$ is equal to the number of points in $P_{v}$ not used by $C$. Therefore, for each $v \in R$, there are $\sum_{u \in P_{v}}\left(d_{u}-1\right)$ points for arcs entering $v$ and $d_{v}-1$ points for arcs leaving $v$. Using Lemma 3.27, we can deduce that the number of ways to choose the remainder of $\mathcal{A}$, and then $\overrightarrow{\mathrm{F}}$, is

$$
n \frac{\prod_{v \in R}\left(d_{v}-1\right)}{\prod_{v \in S} d_{v}}\left(\prod_{v} d_{v}\right)(m-i-1)!.
$$

Summing over $R$ and dividing by $\left|\bar{\Phi}_{n, \mathbf{d}}\right|$, which we know from Theorem 3.28, gives

$$
\begin{aligned}
\mathbb{P}\left[C \subset \overrightarrow{\mathrm{~F}}:(\overrightarrow{\mathrm{F}}, \mathcal{A}) \in \bar{\Phi}_{n, \mathbf{d}}\right] & \rightarrow \frac{1}{m^{i}} \sum_{\substack{R \subset S}} \frac{\prod_{v \in R}\left(d_{v}-1\right)}{\prod_{v \in S} d_{v}} \\
& =\frac{1}{m^{i} \mid \geq 1}\left(1-\frac{1}{\prod_{v \in S} d_{v}}\right) .
\end{aligned}
$$

For each $S \subset V$ with $|S|=i$, there are $(i-1)!\prod_{v \in S} d_{v}^{2}$ ways to form an $i$-cycle on $S$. Hence,

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]} \rightarrow \frac{(i-1)!}{m^{i}} \sum_{\substack{S \in V \\|S|=i}}\left(\prod_{v \in S} d_{v}^{2}-\prod_{v \in S} d_{v}\right)
$$

By a similar argument as was used in Lemma 3.31 we can show

$$
\sum_{\substack{S \subset V \\|S|=i}} \prod_{v \in S} d_{v} \rightarrow \frac{m^{i}}{i!}
$$

and so conclude

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} X_{i, n}\right]}{\mathbb{E} \mathscr{A}_{n, \mathbf{d}}^{\star}} \rightarrow \frac{1}{i}\left(\frac{m_{2}}{m}\right)^{i}-\frac{1}{i}=\mu_{i}
$$

We now have sufficient ammunition to apply Janson's theorem and obtain an asymptotic distribution for the number of arborescences of a random $d$-in/d-out graph.

Theorem 3.33. Let $d \geq 2$ be some fixed integer, let $n \in \mathbb{N}$, and let $\mathcal{A}_{n, d}$ denote the number of arborescences in a random directed graph $\vec{G}$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, d)$. Then,

$$
\frac{\mathcal{A}_{n, d}}{\mathbb{E}\left[\mathcal{A}_{n, d}\right]} \rightarrow \prod_{i=2}^{\infty}\left(1-\frac{1}{d^{i}}\right)^{Z_{i}} e^{1 / i}
$$

where the $Z_{i}$ are independent Poisson random variables with means $d^{i} / i$.
Proof. Let $X_{i, n}$ be the random variable counting $i$-cycles studied in Lemma 3.31 and Lemma 3.32. To apply Janson's Theorem (Theorem 3.19), we need to show that the four conditions in the hypothesis of the theorem are satisfied by $\mathcal{A}_{n, d}$ and $\left\{X_{i, n}: i \geq 2\right\}$. Restricted to the case where $d_{u}=d_{v}=d$ for all $u, v$, Lemma 3.31 provides Condition 1, by showing that the random variables $X_{i, n}$ converge jointly to independent Poisson random variables with means

$$
\lambda_{i}=\frac{d^{i}}{i}
$$

Lemma 3.32 tells us that the random variables $X_{i, n}$, taken with $\mathcal{A}_{n, d}^{\star}$, satisfy

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star} \prod_{i=1}^{k}\left(X_{i, n}\right)_{j_{i}}\right]}{\mathbb{E}\left[\mathscr{A}_{n, \mathbf{d}}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}},
$$

where

$$
\mu_{i}=\frac{d^{i}-1}{i} .
$$

Hence, by Lemma 3.22, Condition 2 is satisfied for $Y_{n}=\mathscr{A}_{n, d}^{\star}$ with

$$
\delta_{i}=-\frac{1}{d^{i}} .
$$

Since the $X_{i, n}$ are all considered to be independent, we can condition on $X_{1, n}=0$ to infer that Condition 2 of Theorem 3.19 is also satisfied, with the same $\delta_{i}$ values, when we take $Y_{n}=\mathcal{A}_{n, d}$ and $X_{i, n}, i \geq 2$. Evaluating the sum in Condition 3 gives

$$
\sum_{i=2}^{\infty} \frac{1}{i d^{i}}=-\frac{1}{d}+\log \left(\frac{d}{d-1}\right)
$$

since

$$
\log \left(\frac{d}{d-1}\right)=\log \left(\frac{1}{1-\frac{1}{d}}\right)=\sum_{i=1}^{\infty} \frac{1}{i d^{i}} .
$$

Finally, Theorem 3.30 provides Condition 4.
The short cycle conditioning method fails when we allow vertices of different degrees. In general, we have

$$
\frac{\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}^{2}\right]}{\mathbb{E}\left[\mathcal{A}_{n, \mathbf{d}}\right]^{2}}=e^{-n / m} \frac{m}{m-n}
$$

Consider the class of graphs containing $n / 2$ vertices of out-degree 2 and $n / 2$ vertices of out-degree 3 , which we denote $\overrightarrow{\mathbb{G}}(n, 2,3)$. If $\mathscr{A}_{n, 2,3}$ counts the number of Euler tours of $G$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, 2,3)$ then, by Theorem 3.30 , we have

$$
\frac{\mathbb{E}\left[\mathscr{A}_{n, 2,3}^{2}\right]}{\mathbb{E}\left[\mathcal{A}_{n, 2,3}\right]^{2}} \rightarrow e^{-2 / 5} \frac{5}{3} \approx 1.172
$$

since we have $m=5 n / 2$. From Lemma 3.31 and Lemma 3.32 we obtain

$$
\lambda_{i}=\frac{1}{i}\left(\frac{13}{5}\right)^{i} \quad \text { and } \quad \delta_{i}=-\frac{1}{i}\left(\frac{5}{13}\right)^{i} .
$$

Hence,

$$
\sum_{i=2}^{\infty} \lambda_{i} \delta_{i}^{2}=-\frac{5}{13}+\log (13 / 8) \approx 1.106
$$

and therefore Condition 4 of Theorem 3.19 is not satisfied. Proofs using the small subgraph conditioning method depend strongly on the fact that these values match, since it is this (asymptotic) equality that implies the variance within each class is negligible. We do not have a good explanation for why this breaks down when we allow different degrees, nor do we believe it to be strong evidence that a concentration result cannot be obtained for a more general case than $d_{u}=d_{v}=d$. Indeed, we can still conclude that a large proportion of directed graphs with fixed degree sequence have a large number of arborescences.

### 3.3.3 Bounding $\mathcal{T}_{n, d}$ with high probability

By Theorem 3.30 of §3.3.1, we know that the expected number of arborescences of a random $d$-in/d-out graph is

$$
\mathbb{E}\left[\mathcal{A}_{n, d}\right] \rightarrow e^{1} d^{n-1}
$$

Hence, if we let $\mathcal{T}_{n, d}$ denote the number of Euler tours of a random $d$-in/d-out graph we have

$$
\mathbb{E}\left[\mathcal{T}_{n, d}\right]=n^{-1} \mathbb{E}\left[\mathcal{A}_{n, d}\right](d-1)!^{n} \rightarrow \frac{e^{1}}{d n}(d!)^{n}
$$

Let $\mathrm{TS}(n, d)$ denote the number of transition systems of a $d$-in/d-out graph. There are $d$ ! ways to choose a pairing on in-arcs with out-arcs at each vertex, so

$$
\operatorname{TS}(n, d)=(d!)^{n} .
$$

Hence, to prove Theorem 3.2, we just need to show that the number of arborescences of a random $d$-in/d-out graph is concentrated around $\mathbb{E}\left[\mathcal{A}_{n, d}\right]$ (Theorem 3.35 below). We prove Theorem 3.35 by arguing along the lines of [36]. In the following proof we will use a Lemma from [77], (which was used in the proof of [36]).

Lemma 3.34 (Lemma 3 [77]). Let $\eta_{1}, \eta_{2}, \ldots$ be given, Suppose that $\eta_{1}>0$ and, for some $c>0, \eta_{i+1} / \eta_{i}>c$, for all $i \geq 1$. Then, uniformly over $x \geq 1$,

$$
\sum_{i=1}^{\infty} \sum_{t=\eta_{i}\left(1+y_{i}\right)}^{\infty} \frac{\eta_{i}^{t}!e_{i}^{\eta_{i}}}{}=O\left(e^{-c_{0} x}\right),
$$

where $y_{i}=x \eta_{i}^{-1 / 3}$ and $c_{0}=\min \left\{\eta_{1}^{1 / 3}, \eta_{1}^{2 / 3}\right\} / 4$
Theorem 3.35. Let $d \geq 2$ be some fixed integer, let $n \in \mathbb{N}$, and suppose $\vec{G}$ is a directed graph chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, d)$. Then,

$$
\mathbb{P}\left[\frac{|\mathrm{ET}(\vec{G})|}{|\operatorname{TS}(\vec{G})|} \geq \frac{e^{1}}{d} n^{-2}\right] \rightarrow 1
$$

as $n \rightarrow \infty$.
Proof. Let $\mathcal{A}_{n, d}$ denote the number of arborescences of a directed graph $\vec{G}$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, d)$. We have shown, in Theorem 3.30 , that the expectation of $\mathcal{A}_{n, d}$ satisfies

$$
\mathbb{E}\left[\mathcal{A}_{n, d}\right] \rightarrow e^{1} d^{n-1}
$$

The number of Euler tours of any directed graph $\vec{G} \in \mathbb{G}(n, d)$ is equal to

$$
\frac{1}{n}|\operatorname{ARB}(\vec{G})|(d-1)!^{n},
$$

and the number of transition systems of $\vec{G}$ is equal to $d!^{n}$. Hence, the statement of the theorem is equivalent to saying

$$
\mathbb{P}\left[\frac{\mathcal{A}_{n, d}}{\mathbb{E}\left[\mathcal{A}_{n, d}\right]} \geq \frac{1}{n}\right] \rightarrow 1
$$

as $n \rightarrow \infty$.
For $\mathbf{x}=\left(x_{1}, x_{2}, \ldots, x_{k}\right)$ we define $\mathcal{G}_{\mathbf{x}}$ to be the set of all $d$-regular Eulerian directed graphs containing exactly $x_{i}$ directed cycles of length $i$ for each $i \leq k$, and

$$
W^{(k)}(\mathbf{x})=\prod_{i=2}^{k}\left(1-\frac{1}{d^{i}}\right)^{x_{i}} e^{1 / i} .
$$

For each fixed $\gamma>0$ we define

$$
S(\gamma)=\left\{\mathbf{x}: x_{i} \leq \lambda_{i}+\gamma \lambda_{i}^{2 / 3} \text { for } 2 \leq i \leq k\right\} .
$$

From Lemma 3.31, the probability that $\vec{G}$ chosen uniformly at random from $\overrightarrow{\mathbb{G}}(n, d)$ has exactly $x_{i} i$-cycles is

$$
\frac{\lambda_{i}^{x_{i}}}{e^{\lambda_{i} x_{i}!}} .
$$

Hence, by Lemma 3.34, the probability that a random $\vec{G}$ is not contained in $\mathcal{G}_{\mathbf{x}}$ for some $\mathbf{x} \in S(\gamma)$ is

$$
\begin{aligned}
\sum_{\mathbf{x} \notin S(\gamma)} \mathbb{P}\left[\vec{G} \in \mathcal{G}_{\mathbf{x}}\right] & \leq \sum_{i=2}^{\infty} \mathbb{P}\left[X_{i, n} \geq \lambda_{i}+\gamma \lambda_{i}^{2 / 3}\right] \\
& =\sum_{i=2}^{\infty} \sum_{t=\lambda_{i}\left(1+\gamma \lambda_{i}^{-1 / 3}\right)}^{\infty} \frac{\lambda_{i}^{t}}{e^{\lambda_{i} t!}} \\
& \in e^{-a \gamma},
\end{aligned}
$$

where $a=d^{1 / 3} / 4$; i.e., $a$ is an absolute constant greater than 0 . Hence, to verify the theorem all we need do is show that

$$
\begin{equation*}
W^{(k)}(\mathbf{x}) \geq e^{-(b+c \gamma)} \quad \forall \mathbf{x} \in S(\gamma), \tag{3.63}
\end{equation*}
$$

where $b$ and $c$ are absolute constants independent of $\gamma$. Then, by Theorem 3.33, we have

$$
\mathbb{P}\left[\frac{\mathcal{A}_{n, d}}{\mathbb{E}\left[\mathcal{A}_{n, d}\right]} \geq e^{-(b+c \gamma)}\right] \geq 1-e^{-a \gamma} .
$$

For any $\varepsilon>0$, we can choose $\gamma$ so that $e^{-a \gamma}<\varepsilon$. Then, we can choose $n$ large enough so that $e^{-(b+c) \gamma}>n^{-1}$. That is, if we can show (3.63) holds, we will have shown that for every $\varepsilon>0$, there exists $n \in \mathbb{N}$ such that

$$
\mathbb{P}\left[\frac{\mathcal{A}_{n, d}}{\mathbb{E}\left[\mathcal{A}_{n, d}\right]} \geq \frac{1}{n}\right]>1-\varepsilon,
$$

So, it remains to prove (3.63). For $x \in S(\gamma)$ we have $W^{(k)}(\mathbf{x}) \geq A B^{\gamma}$, where

$$
\begin{aligned}
A & =\prod_{i=2}^{k}\left(1-\frac{1}{d^{i}}\right)^{\lambda_{i}} e^{1 / i} \\
B & =\prod_{i=2}^{k}\left(1-\frac{1}{d^{i}}\right)^{\lambda_{i}^{2 / 3}} .
\end{aligned}
$$

Using the elementary inequality $1-x \geq e^{-x /(1-x)}$, we can bound $A$ and $B$ as

$$
\begin{aligned}
A & \geq \prod_{i=2}^{\infty} \exp \left(-\frac{d^{i}}{i\left(d^{i}-1\right)}+\frac{1}{i}\right) \\
& =\exp \left(\sum_{i=2}^{k}-\frac{1}{i\left(d^{i}-1\right)}\right),
\end{aligned}
$$

and

$$
\begin{aligned}
B & \geq \prod_{i=2}^{\infty} \exp \left(-\frac{d^{2 i / 3}}{i^{2 / 3}\left(d^{i}-1\right)}\right) \\
& \geq \exp \left(-\sum_{i=2}^{\infty} \frac{1}{i^{2 / 3} d^{i / 3}}\right) .
\end{aligned}
$$

The sums inside the exponentials are clearly convergent, so we can find absolute constants $b$ and $c$ so that $A \geq e^{-b}$ and $B \geq e^{-c}$.

### 3.4 The number of Eulerian orientations of random regular graphs

In this section we turn our attention to the distribution of the number of Eulerian orientations of a random regular graph. Recalling the relationship between the number of orbs and Euler tours, we would hope that this will take us closer to our main goal: obtaining an asymptotic distribution for the number of Euler tours of a random regular graph. Although we did not manage to achieve this, we believe the analysis in this section, particularly the analysis of the second moment of the number of Eulerian orientations of a random regular graph, will be a key part of the analysis required to obtain the corresponding result for orbs. Moreover, as the number of Eulerian orientations is a quantity of interest in its own right, the results of this section are of independent interest.

In this section we restrict our attention to regular graphs. The reason for this is that we cannot find a closed form for the number of Eulerian orientations of a random configuration $\mathrm{F} \in \Omega_{n, \mathbf{d}}$, as we could for arborescences of $\Phi_{n, \mathbf{d}}$, and the asymptotic analysis carried out in the proof of Theorem 3.47 does not immediately carry over to the general configuration model.

For fixed positive integer $d$, and $n \in \mathbb{N}$, we define $\mathcal{E}_{n, 2 d}$ to be the random variable counting Eulerian orientations of a random $2 d$-regular graph. In §3.4.2 we analyse the number of Eulerian orientations of a random $2 d$-regular graph, obtaining asymptotic results for constant $d$. As we did for arborescences in §3.3, we first analyse the moments of an auxiliary variable, $\mathcal{E}_{n, 2 d}^{\star}$, which counts Eulerian orientations of random configurations $\mathrm{F} \in \Omega_{n, 2 d}$, and then condition on F projecting to a simple graph. The asymptotic analysis of §3.4.2 uses Laplace's method for estimating integrals. We briefly describe the details of this method in §3.4.1.

We apply these estimates in two ways. Firstly, in §3.4.3, we use these estimates and Theorem 3.19 to characterise the asymptotic distribution of $\mathcal{E}_{n, 2 d}$. Then, in §3.4.4, we use Chebyshev's inequality (Theorem 3.8) to show that almost every $2 d$-regular graph has few Eulerian orientations; that is, we show that the probability the number of Eulerian orientations of a random $G \in \mathbb{G}(n, 2 d)$ exceeds Schrijver's lower bound [78] (see Theorem 1.56) by more than a linear factor tends to 0 as $n$ goes to infinity.

In the previous section, the asymptotic distribution of the number of arborescences was used to show a simple algorithm for sampling Euler tours runs in expected polynomial time for almost every directed graph $\vec{G} \in \vec{G}(n, d)$. We remark that we
cannot obtain a similar result here. Firstly, it is not clear if the set of Eulerian orientations of a graph $G$ sit nicely inside some other set of orientations, apart from the set of all orientations of $G$. Moreover, our results imply that most regular graphs have few Eulerian orientations, i.e., close to the minimum possible, so even if we had a "nice" set of orientations that the Eulerian orientations sat inside, it is still unlikely that the relative size of the set of Eulerian orientations (of a particular graph) within this set would be large enough for a dart-throwing approach to work.

### 3.4.1 Laplace's Method

In our analysis of the second moment of the number of Eulerian orientations, we will need to compute an asymptotic estimate of an integral. The approach we use is Laplace's method [25, Chapter 4], also known as saddle-point analysis. This is a general approach for estimating an integral of the form $\int e^{n f(\mathbf{x})} d \mathbf{x}$ whose value is concentrated around a single maximum term. The first step in an application of this method is to locate the maximum.

Definition 3.36. Let $f(x)$ be a real-valued function. We say $f$ has a critical point at $a$, if the value of the first derivative of $f$ at $a$ is 0 : $f^{\prime}(a)=0$.

A critical point $a$ is a local maximum if and only if the second derivative of $f$ is negative at $a, f^{\prime \prime}(a)<0$, and a critical point $a$ is a local minimum if and only if the second derivative is positive at $a, f^{\prime \prime}(a)>0$.

The local maxima and minima of a multi-dimensional function can be defined in a similar way. First, we need to define positive definite and negative definite matrices.

Definition 3.37. We say an $n \times n$ symmetric, real-valued matrix $A$ is negative definite if

$$
\mathbf{x}^{T} A \mathbf{x}<0, \quad \forall \mathbf{x} \in \mathbb{R}^{n}
$$

We say $A$ is positive definite if $-A$ is negative definite.
Definition 3.38. Let $f\left(x_{1}, x_{2}, \ldots, x_{d}\right)$ be an $d$-ary real-valued function. A point $\mathbf{a} \in \mathbb{R}^{d}$ is a critical point of $f$ if all first order partial derivatives of $f$ are 0 at $\mathbf{a}$ :

$$
\frac{\partial f}{\partial x_{i}}(\mathbf{a})=0, \quad \text { for } i=1,2, \ldots, d
$$

We define the Hessian matrix of second order partial derivatives of $f$ :

$$
D^{2} f(\mathbf{x})_{i, j}=\frac{\partial f}{\partial x_{i} x_{j}}, \quad \text { for } i, j=1,2, \ldots, d
$$

A critical point a is a local maximum of $f$ if and only if $D^{2} f(\mathbf{a})$ is a negative definite matrix, and a local minimum if and only if $D^{2} f(\mathbf{a})$ is a positive definite matrix.

To apply Laplace's method to estimate an integral, $\int e^{n f(\mathbf{x})} d \mathbf{x}$, we search for the local maxima on the range of integration and consider the value of $f(\mathbf{x})$ in the neighbourhoods of these points. We want to show that the asymptotic value of the integral is determined by the value within a small neighbourhood of the maximum value $\mathbf{a}$, and that in this neighbourhood $f(\mathbf{x})$ can be approximated by a Gaussian integral; i.e., an integral of the form

$$
\begin{equation*}
\int_{\mathbf{R}^{n}} e^{-\frac{1}{2} \mathbf{x}^{T} A \mathbf{x}} d \mathbf{x} \tag{3.64}
\end{equation*}
$$

for positive definite $n \times n$ matrix $A$.
Definition 3.39. Let $A$ be an $n \times n$ matrix. The determinant of $A$, denoted $|A|$, is the quantity

$$
\sum_{\sigma \in S_{n}} \operatorname{sgn}(\sigma) \prod_{i=1}^{n} A_{i, \sigma(i)}
$$

where $S_{n}$ is taken to be the set of all permutations of $\{1,2, \ldots, n\}$. The signature of a permutation is +1 (resp. -1 ) if $\sigma$ can be obtained from the identity permutation,

$$
\sigma(i)=i, \text { for all } i \in\{1,2, \ldots, n\},
$$

by performing an even (resp. odd) number of transpositions of adjacent numbers.
Example 3.40. The determinant of the $2 \times 2$ matrix

$$
A=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

is $a d-b c$. This is the only determinant we will deal with explicitly in the proofs of this section.

The following result, which can be found in, e.g., [25, Chapter 4], shows how we can compute the value of (3.64) by computing the determinant of $A$.

Theorem 3.41. Let $A$ be a positive definite $n \times n$ matrix. Then,

$$
\int_{\mathbf{R}^{n}} e^{-\frac{1}{2} \mathbf{x}^{T} A \mathbf{x}} d \mathbf{x}=\frac{(2 \pi)^{d / 2}}{\sqrt{|A|}},
$$

where $|A|$ denotes the determinant of $A$.

### 3.4.2 Estimating the moments of $\mathcal{E}_{n, d}$

Our goal in this section is to find asymptotic estimates of the first and second moments of the number of Eulerian orientations of a random graph $G \in \mathbb{G}(n, 2 d)$. We start by estimating the moments of $\mathscr{E}_{n, 2 d}^{\star}$, the number of Eulerian orientations of $\sigma(\mathrm{F})$ where F is chosen uniformly at random from $\Omega_{n, 2 d}$. These estimates can then be turned into estimates for the moments of $\mathcal{E}_{n, 2 d}$ by conditioning on F not containing any loops or double edges.

Before proving the theorem, we provide some technical lemmas. We make no claim to the novelty of the identities presented in Lemma 3.44 or Lemma 3.45 but include proofs here due to not knowing a suitable reference.

For a square matrix $A$ we denote the determinant of $A$ by $|A|$ and write $A[\alpha, \beta]$ for the submatrix with rows indexed by $\alpha$ and columns indexed by $\beta$. In Lemma 3.44, we show how to compute the determinant of a sum of two matrices, using Laplace's formula. Recall the definition of a matrix minor.

Definition 3.42. Let $M$ be a $n \times n$ matrix. For any row and column indices $i$ and $j$, we denote by $M_{i j}$ the matrix obtained by removing the row indexed by $i$ and the column indexed by $j$. This is known as the $(i, j)$-minor of $M$. We define the $(i, j)$-cofactor of $M$ to be $(-1)^{i+j}\left|M_{i j}\right|$.

One approach to computing determinants of matrices, is to use Laplace Expansion, which expresses the value of the determinant of a matrix in terms of its cofactors.

Fact 3.43 (Laplace, see Chapter 5 of [84]). Let $A$ be an $n \times n$ matrix and let $a_{i j}$ denote the entry in row $i$ and column $j$. Then,

$$
|A|=\sum_{i=1}^{n}(-1)^{i+j} a_{i j}\left|A_{i j}\right|
$$

for any $j \in\{1,2, \ldots, n\}$ (expanding along column $j$ ) and

$$
|A|=\sum_{j=1}^{n}(-1)^{i+j} a_{i j}\left|A_{i j}\right|
$$

for any $i \in\{1,2, \ldots, n\}$ (expanding along row $i$ ).
Lemma 3.44. Let $A, B$ be $n \times n$ square matrices. Then

$$
\begin{equation*}
|A+B|=\sum_{k=0}^{n} \sum_{\alpha, \beta \in S_{n, k}}(-1)^{s(\alpha)+s(\beta)}|A[\alpha, \beta]||B[\bar{\alpha}, \bar{\beta}]|, \tag{3.65}
\end{equation*}
$$

where $S_{n, k}$ is the set of $k$-subsequences of $(1,2, \ldots, n), \bar{\alpha}$ is the subsequence obtained by removing $\alpha$ from $(1,2, \ldots, n)$, and $s(\alpha)=\sum \alpha_{i}$.

Proof. We prove by induction. For $n=2$ the lemma can be checked by expanding all the determinants involved. Let $n>2$ and suppose the lemma is true for all $k<n$. Using Fact 3.43 to expand $|A+B|$ gives

$$
|A+B|=\sum_{i=1}^{n}(-1)^{i+1}\left(a_{1, i}+b_{1, i}\right)\left|A_{1, i}+B_{1, i}\right| .
$$

Each term in the sum is itself a sum of two $(n-1) \times(n-1)$ matrices, and so, by induction, we can write

$$
\begin{equation*}
|A+B|=\sum_{i=1}^{n} \sum_{k=1}^{n} \sum_{\substack{\alpha, \beta \in S_{n, k} \\ \alpha \ni 1, \beta \ni i}}(-1)^{s(\alpha)+s(\beta)+|\{j \in \beta: j<i\}|}\left(a_{1, i}+b_{1, i}\right)\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right||B[\bar{\alpha}, \bar{\beta}]|, \tag{3.66}
\end{equation*}
$$

where $\gamma_{-k}$ is the sequence $\gamma$ with $k$ removed. We will now explain where the terms in (3.66) come from. Consider the expansion of $\left|A_{1, i}+B_{1, i}\right|$ by applying the Lemma:

$$
\left|A_{1, i}+B_{1, i}\right|=\sum_{k=0}^{n-1} \sum_{\substack{\alpha^{\prime} \cdot \beta^{\prime} \in S_{n, k} \\ 1 \notin \alpha^{\prime}, i \notin \beta^{\prime}}}(-1)^{s_{1}\left(\alpha^{\prime}\right)+s_{i}\left(\beta^{\prime}\right)}\left|A\left[\alpha^{\prime}, \beta^{\prime}\right]\right| \mid B\left[\bar{\alpha}^{\prime}{ }_{-1}, \bar{\beta}^{\prime}{ }_{-i} \mid,\right.
$$

where $s_{i}(\gamma)=s(\gamma)-\mid\{j \in \gamma: j>i\}$. The functions $s_{i}$ come from the fact that any $j>1$ in $\alpha^{\prime}$ is actually indexing the $(j-1)$-th row of $A_{1, i}+B_{1, i}$, and any $j>i$ in $\beta^{\prime}$ is actually indexing the $(j-1)$-th column of $A_{1, i}+B_{1, i}$. For the sake of presentation, we have mapped the subsequences $\alpha^{\prime}$ and $\beta^{\prime}$ used in the inductive application of (3.65) to $\left|A_{1, i}+B_{1, i}\right|$ to subsequences $\alpha$ and $\beta$ of $(1,2, \ldots, n)$ satisfying $1 \in \alpha$ and $i \in \beta$. By considering this mapping, we can see that the power of -1 in the term corresponding to $\alpha$ and $\beta$ in the inductive application of (3.65) to $\left|A_{1, i}+B_{1, i}\right|$ is

$$
\begin{align*}
(-1)^{s(\alpha)-k+s(\beta)-i-|\{j \in \beta: j>i\}|} & =(-1)^{s(\alpha)-k+s(\beta)-i-(k-1-|\{j \in \beta: j<i\}|)} \\
& =(-1)^{s(\alpha)+s(\beta)-i+1+|\{j \in \beta: j<i\}|} . \tag{3.67}
\end{align*}
$$

Multiplying (3.67) by $(-1)^{i+1}$ gives the power of -1 used in (3.66).
We will now show the equivalence of (3.66) and the RHS of (3.65) by showing that we can obtain (3.66) by expanding each term in the RHS of (3.65) using Laplace expansion. Each term in (3.65) is expanded as follows: if $1 \in \alpha$ then we expand $|A[\alpha, \beta]|$ along row 1 ; if $1 \notin \alpha$ then we expand $|B[\bar{\alpha}, \bar{\beta}]|$ along row 1 .

Let $i \in[n]$ and suppose we have a pair of sequences $\alpha$ and $\beta$ with $1 \in \alpha$ and $i \in \beta$. We obtain the term $a_{1, i} A\left[\alpha_{-1}, \beta_{-i}\right] B[\bar{\alpha}, \bar{\beta}]$ by expanding $A[\alpha, \beta] B[\bar{\alpha}, \bar{\beta}]$ and obtain the term $b_{1, i} A\left[\alpha_{-1}, \beta_{-i}\right] B[\bar{\alpha}, \bar{\beta}]$ by expanding $A\left[\alpha_{-1}, \beta_{-i}\right] B\left[\overline{\alpha_{-1}}, \overline{\beta_{-i}}\right]$, as described above.

Thus, the terms in the sum obtained by expanding each term of (3.65) are identical to the terms in the RHS of (3.66). It remains to show that each term has the correct sign.

We first consider terms of the form $a_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right||B[\bar{\alpha}, \bar{\beta}]|$. Using Fact 3.43, expanding $|A[\alpha, \beta]|$ along row 1 gives

$$
\begin{align*}
(-1)^{s(\alpha)+s(\beta)}|A[\alpha, \beta]| & =\sum_{i \in \beta}(-1)^{s(\alpha)+s(\beta)+|\{j \in \beta: j \leq i\}|+1} a_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right| \\
& =\sum_{i \in \beta}(-1)^{s(\alpha)+s(\beta)+|\{j \in \beta: j<i\}|} a_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right| \tag{3.68}
\end{align*}
$$

so $a_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right||B[\bar{\alpha}, \bar{\beta}]|$ has the same sign in (3.66) and the sum obtained by applying Laplace expansion to each term in (3.65)..

Next, we consider terms of the form $b_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right||B[\bar{\alpha}, \bar{\beta}]|$. Again using Fact 3.43, we expand $\left|B\left[\overline{\alpha_{-1}}, \overline{\beta_{-i}}\right]\right|$ along row 1 . Consider the term in this expansion involving $b_{1, i}$ :

$$
\begin{equation*}
(-1)^{s\left(\alpha_{-1}\right)+s\left(\beta_{-i}\right)+1+|\{j \notin \beta: j \leq i\}|} b_{1, i}|B[\bar{\alpha}, \bar{\beta}]| . \tag{3.69}
\end{equation*}
$$

The sign of $b_{1, i}|B[\bar{\alpha}, \bar{\beta}]|$ in (3.69) is

$$
(-1)^{s(\alpha)+s(\beta)-i+|\{j \notin \beta: j \leq i\}|} .
$$

Note that

$$
|\{j \notin \beta: j \leq i\}|+|\{j \in \beta: j<i\}|=i
$$

this implies that $|\{j \in \beta: j<i\}|$ is even if and only if $-i+\mid\{j \notin \beta: j \leq i\}$ is even. Hence, $b_{1, i}\left|A\left[\alpha_{-1}, \beta_{-i}\right]\right||B[\bar{\alpha}, \bar{\beta}]|$ has the same sign in (3.66) and the sum obtained by applying Laplace expansion to each term in (3.65).

Lemma 3.45. For each positive integer $d$,

$$
\begin{aligned}
& \sum_{i=0}^{d}\binom{d}{i}^{2}=\binom{2 d}{d} \\
& \sum_{i=1}^{d} i\binom{d}{i}^{2}=\frac{d}{2}\binom{2 d}{d} \\
& \sum_{i=1}^{d} i^{2}\binom{d}{i}^{2}=\frac{d^{3}}{2(2 d-1)}\binom{2 d}{d} .
\end{aligned}
$$

Proof. We prove the first identity by observing that the number of ways to choose $d$ values from $\{1,2, \ldots, 2 d\}$, is the same as the total number of ways to choose $i$ values from $\{1,2, \ldots, d\}$ and $d-i$ values from $\{d+1, d+2, \ldots, 2 d\}$, for $i=0,1, \ldots, d$. Hence,

$$
\binom{2 d}{d}=\sum_{i=0}^{d}\binom{d}{i}\binom{d}{d-i}=\sum_{i=0}^{d}\binom{d}{i}^{2} .
$$

The other two follow by straightforward calculations:

$$
\begin{aligned}
& \sum_{i=1}^{d} i\binom{d}{i}^{2}=d \sum_{i=1}^{d}\binom{d-1}{i-1}\binom{d}{d-i}=d\binom{2 d-1}{d-1}=\frac{d}{2}\binom{2 d}{d} \\
& \sum_{i=1}^{d} i^{2}\binom{d}{i}^{2}=d^{2} \sum_{i=0}^{d-1}\binom{d-1}{i}=\frac{d^{3}}{2(2 d-1)}\binom{2 d}{d} .
\end{aligned}
$$

We will now use Lemma 3.44 and Lemma 3.45 to prove the following Lemma, which gives some properties of a particular matrix. This matrix will arise in the proof of Theorem 3.47 as the Hessian matrix of a function.

Lemma 3.46. Suppose $d$ is an integer greater than 1 and let $M$ be the $d \times d$ symmetric matrix given by

$$
M_{i j}= \begin{cases}\frac{4 i j}{d}-\binom{2 d}{d} & i \neq j \\ \frac{4 i^{2}}{d}-\binom{2 d}{d}-\frac{\binom{2 d}{d}}{\binom{d}{i}^{2}} & i=j\end{cases}
$$

This matrix is negative definite and has determinant

$$
|M|=(-1)^{d} \frac{\binom{2 d}{d}^{d+1}}{\prod_{i=1}^{d}\binom{d}{i}^{2}} \frac{d-1}{2 d-1} .
$$

Proof. We first show that $M$ is negative definite. Recall that a symmetric real-valued matrix is negative definite if and only if $\mathbf{x}^{T} M \mathbf{x}<0$ for every $\mathbf{x} \in \mathbb{R}^{d}$. In other words, we can prove $M$ is negative definite by showing the following strict inequality holds for all $\mathbf{x} \in \mathbb{R}^{d}$ :

$$
\left(\sum_{i=1}^{d} 2 i x_{i}\right)^{2}<d\binom{2 d}{d}\left(\sum_{i=1}^{d} x_{i}\right)^{2}+d\binom{2 d}{d} \sum_{i=1}^{d} \frac{x_{i}^{2}}{\binom{d}{i}^{2}}
$$

We will use the following elementary inequality,

$$
(a+b)^{2} \leq \frac{2 d-1}{d} a^{2}+\frac{2 d-1}{d-1} b^{2}
$$

which holds for any $a$ and $b$. Then, by writing

$$
a=\sum_{i=0}^{d}(2 i-d) x_{i}, \text { and } b=\sum_{i=0}^{d} d x_{i},
$$

we have

$$
\left(\sum_{i=1}^{d} 2 i x_{i}\right)^{2} \leq \frac{2 d-1}{d}\left(\sum_{i=0}^{d}(2 i-d) x_{i}\right)^{2}+\frac{d^{2}(2 d-1)}{d-1}\left(\sum_{i=0}^{d} x_{i}\right)^{2} .
$$

Next, we apply the Cauchy-Schwarz inequality [82],

$$
\left(\sum_{i=1}^{d} a_{i} b_{i}\right)^{2} \leq\left(\sum_{i=1}^{d} a_{i}^{2}\right)\left(\sum_{i=1}^{d} b_{i}^{2}\right)
$$

giving

$$
\left(\sum_{i=1}^{d}(2 i-d) x_{i}\right)^{2} \leq\left(\sum_{i=1}^{d}(2 i-d)^{2}\binom{d}{i}^{2}\right)\left(\sum_{i=1}^{d} \frac{x_{i}^{2}}{\binom{d}{i}^{2}}\right) .
$$

We can use Lemma 3.45 to show

$$
\sum_{i=0}^{d}(2 i-d)^{2}\binom{d}{i}^{2}=\frac{d^{2}}{2 d-1}\binom{2 d}{d}
$$

Hence, we have shown

$$
\left(\sum_{i=1}^{d} 2 i x_{i}\right)^{2} \leq \frac{d^{2}(2 d-1)}{d-1}\left(\sum_{i=1}^{d} x_{i}\right)^{2}+d\binom{2 d}{d} \sum_{i=1}^{d} \frac{x_{i}^{2}}{\binom{d}{i}^{2}} .
$$

Finally, we observe that for any $d \geq 2$, we have $\frac{d^{2}(2 d-1)}{d-1}<d\binom{2 d}{d}$.
We now turn our attention to evaluating the determinant. To evaluate $|M|$ using Lemma 3.44, we write $M$ as $A+B$ where

$$
A_{i, j}=\left\{\begin{array}{cl}
-\frac{\binom{2 d}{d}}{\binom{d}{i}^{2}} & i=j \\
0 & \text { otherwise }
\end{array}\right.
$$

and

$$
B_{i, j}=\frac{4 i j}{d}-\binom{2 d}{d} .
$$

We first show that the determinant of every square sub-matrix of $B$ with more than 2 rows is 0 . Recall that a square matrix has determinant 0 if we can write any row as a linear combination of other rows [84]. Let $B_{k}$ denote the $k$-th row of $B$,

$$
B_{k}=\left(\frac{4 k}{d}-\binom{2 d}{d}, \frac{8 k}{d}-\binom{2 d}{d}, \ldots, \frac{4 d k}{d}-\binom{2 d}{d}\right)
$$

It is straightforward to see that $B_{k}$ can be written as the sum

$$
B_{k}=\frac{k-j}{i-j} B_{i}+\frac{i-k}{i-j} B_{j}
$$

for any $i$ and $j$. Hence, any square submatrix of $B$ with more than two rows will have determinant 0 . Moreover, the fact that $A$ is a diagonal matrix means that $\left|A_{\alpha, \beta}\right| \neq 0$ if
and only if $\alpha=\beta$. Thus, using Lemma 3.44 and Lemma 3.45, we can compute the determinant of $M$ as follows

$$
\begin{aligned}
|M| & =\left(\prod_{j=1}^{d} A_{j j}\right)\left(1+\sum_{j=1}^{k} \frac{B_{j j}}{A_{j j}}+\sum_{1 \leq i<j \leq d} \frac{\left(B_{i i} B_{j j}-B_{i j}^{2}\right)}{A_{i i} A_{j j}}\right) \\
& =(-1)^{d}\left(\frac{\binom{2 d}{d}^{d}}{\prod_{i=1}^{d}\binom{d}{i}^{2}}\right)\left(\sum_{j=0}^{d}\binom{d}{j}^{2}-\frac{2}{d\binom{2 d}{d}} \sum_{0 \leq i, j \leq d}(j-i)^{2}\binom{d}{i}^{2}\binom{d}{j}^{2}\right) \\
& =(-1)^{d}\left(\frac{\binom{2 d}{d}^{d}}{\prod_{i=1}^{d}\binom{d}{i}^{2}}\right)\left(\binom{2 d}{d}-\frac{d}{2 d-1}\binom{2 d}{d}\right) \\
& =(-1)^{d} \frac{\binom{2 d}{d}^{d+1}}{\prod_{i=1}^{d}\binom{d}{i}^{2}} \frac{d-1}{2 d-1} .
\end{aligned}
$$

In the following theorem, Theorem 3.47, we estimate the first and second moment of the number of Eulerian orientations of $G$, when $G$ is obtained as the projection of a random $\mathrm{F} \in \Omega_{n, 2 d}$. In the following, we will refer to Eulerian orientations of graphs and configurations. We define an Eulerian orientation of a configuration $\mathrm{F} \in \Omega_{n, 2 d}$ to be a partitioning of $W$ into a pair of equally sized disjoint subsets $S$ and $T$, in which each $W_{v}$ is partitioned into a pair of $d$-element subsets, $S_{v}$ and $T_{v}$, such that each edge in F has one endpoint in $S$ and the other in $T$. For each vertex $v, S_{v}$ corresponds to the endpoints of arcs for which $v$ is the source and $T_{v}$ corresponds to the endpoints of arcs for which $v$ is the target. We will use $\mathrm{EO}(\mathrm{F})$ to denote the set of Eulerian orientations of F . The following proof, and the proofs of several subsequent theorems, use Stirling's formula [25] to approximate factorials: as $n \rightarrow \infty$,

$$
\begin{equation*}
n!\rightarrow\left(\frac{n}{e}\right)^{2} \sqrt{2 \pi n} \tag{3.70}
\end{equation*}
$$

Theorem 3.47. Let $n \in \mathbb{N}$ and let $d$ be some fixed positive integer satisfying $d \geq 2$. Let $\mathcal{E}_{n, 2 d}^{\star}$ be the number of Eulerian orientations in a random $2 d$-regular multigraph obtained as the projection of a uniformly random $\mathrm{F} \in \Omega_{n, 2 d}$. Then, as $n \rightarrow \infty$,

$$
\begin{aligned}
& \mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right] \rightarrow\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} \sqrt{\pi d n} \\
& \\
& \frac{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}} \rightarrow \sqrt{\frac{2 d-1}{2 d-2}} .
\end{aligned}
$$

Proof. We first evaluate $\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]$. The value of $\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]$ is given by

$$
\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]=\frac{\left|\bar{\Omega}_{n, 2 d}\right|}{\left|\Omega_{n, 2 d}\right|},
$$

where

$$
\bar{\Omega}_{n, 2 d}=\left\{(\mathrm{F}, \mathcal{E}): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E} \in \mathrm{EO}(\mathrm{~F})\right\} .
$$

Let $V=\{1,2, \ldots, n\}$. We can enumerate the elements of $\bar{\Omega}_{n, 2 d}$ as follows: for each vertex $v \in V$ partition the set $W_{v}$ into two equally sized sets $S_{v}$ and $T_{v}$; choose a perfect matching from $S=\bigcup_{v \in V} S_{v}$ to $T=\bigcup_{v \in V} T_{v}$. There are

$$
\binom{2 d}{d}^{n}
$$

different ways to choose the sets $S$ and $T$ and (dn)! ways to choose a perfect matching from $S$ to $T$, for each choice of $S$ and $T$. Hence, the number of elements in $\bar{\Omega}_{n, 2 d}$ is

$$
\left|\bar{\Omega}_{n, 2 d}\right|=\binom{2 d}{d}^{n}(d n)!.
$$

Dividing this by the number of configurations in $\Omega_{n, 2 d}$,

$$
\left|\Omega_{n, 2 d}\right|=\binom{2 d n}{d n}(d n)!2^{-d n},
$$

and applying Stirling's formula (3.70) we obtain

$$
\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]=\frac{\binom{2 d}{d}^{n} 2^{d n}}{\binom{2 d n}{d n}} \rightarrow\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} \sqrt{\pi d n} .
$$

We now turn our attention to computing $\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right] / \mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}$; in fact, we only compute an asymptotic upper bound. The corresponding lower bound follows from the results in §3.4.3 and the comments in Remark 3.20. The second moment of $\mathcal{E}_{n, 2 d}^{\star}$ can be written as

$$
\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]=\frac{\left|\widetilde{\Omega}_{n, 2 d}\right|}{\left|\Omega_{n, 2 d}\right|},
$$

where

$$
\widetilde{\Omega}_{n, 2 d}=\left\{\left(\mathrm{F}, \mathcal{E}, \mathcal{E}^{\prime}\right): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E}, \mathcal{E}^{\prime} \in \mathrm{EO}(\mathrm{~F})\right\}
$$

We now describe how we can enumerate $\widetilde{\Omega}_{n, 2 d}$. Any pair of Eulerian orientations $\left(\mathcal{E}, \mathcal{E}^{\prime}\right)$ of a graph $G$ defines a decomposition of the edges of $G$ into two sets: $\mathcal{E} \cap \mathcal{E}^{\prime}$, the edges that have the same orientation in $\mathcal{E}$ and $\mathcal{E}^{\prime}$; and $\mathcal{E} \oplus \mathcal{E}^{\prime}$, the edges that are oriented differently in $\mathcal{E}$ and $\mathcal{E}^{\prime}$. Given this decomposition and the orientation $\mathcal{E}$, we can immediately determine the orientation $\mathcal{E}^{\prime}$. Hence, to enumerate triples $\left(F, \mathcal{E}, \mathscr{E}^{\prime}\right)$, we start by partitioning $W$ into $S$ and $T$, as we did for the first moment. Then, for each $v \in V$ we choose an equal number of vertices from $S_{v}$ and $T_{v}$ to be swapped between the two sets. This gives two new sets $S^{\prime}$ and $T^{\prime}$, where $S_{v}^{\prime}$ corresponds to endpoints
of arcs for which $v$ is the source in $\mathcal{E}^{\prime}$ and $T_{v}^{\prime}$ corresponds to the endpoints of arcs for which $v$ is the target in $\mathcal{E}^{\prime}$. Choosing $\left(\mathrm{F}, \mathcal{E}, \mathcal{E}^{\prime}\right)$ is then equivalent to choosing two perfect matchings, one from $S \cap S^{\prime}$ to $T \cap T^{\prime}$ and one from $S \backslash S^{\prime}$ to $T \backslash T^{\prime}$.

We start by choosing the out-degree in $\mathcal{E} \cap \mathcal{E}^{\prime}$ for each $v \in V$; that is, we choose the number of points that will be contained in $S_{v} \cap S_{v}^{\prime}$. Suppose $d_{v}=\left|S_{v} \cap S_{v}^{\prime}\right|$. The number of ways to choose $S$ and $T$ is, as before,

$$
\binom{2 d}{d}^{n}
$$

Given $S$ and $T$ we choose, for each $v \in V$, a subset of size $d_{v}$ from each of $S_{v}$ and $T_{v}$; these will be the points in $S_{v} \cap S_{v}^{\prime}$ and $T_{v} \cap T_{v}^{\prime}$, respectively. The number of ways in which we can do this is

$$
\prod_{v \in V}\binom{d}{d_{v}}^{2}
$$

Once we have chosen $S, T, S^{\prime}$, and $T^{\prime}$ the number of ways to choose a pair of perfect matchings from $S \cap S^{\prime}$ to $T \cap T^{\prime}$ and from $S \backslash S^{\prime}$ to $T \backslash T^{\prime}$ is

$$
\left(\sum_{v \in V} d_{v}\right)!\left(d n-\sum_{v \in V} d_{v}\right)!.
$$

For each $i=0,1, \ldots, d$, we define $V_{i}=\left\{v \in V:\left|S_{v} \cap S_{v}^{\prime}\right|=i\right\}$ and let $c_{i}=\left|V_{i}\right|$. Then, the number of ways to choose $\mathcal{E}$ and $\mathcal{E}^{\prime}$, satisfying $\left|S_{v} \cap S_{v}^{\prime}\right|=d_{v}$ for all $v \in V$, is

$$
\begin{equation*}
\binom{2 d}{d}^{n}\left(\prod_{i=0}^{d}\binom{d}{i}^{2 c_{i}}\right)\left(\sum_{i=1}^{d} i c_{i}\right)!\left(d n-\sum_{i=1}^{d} i c_{i}\right)!. \tag{3.71}
\end{equation*}
$$

Summing over all choices for $d_{v}$ gives $\left|\bar{\Omega}_{n, 2 d}\right|$, and then dividing by the total number of configurations in $\Omega_{n, d}$ gives an expression for $\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]$ :

$$
\begin{align*}
\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]= & \frac{\binom{2 d}{d}^{n}}{\binom{2 d n}{d n}(d n)!2^{-d n}} \\
& \times \sum_{c_{0}+c_{1}+\cdots+c_{d}=n}\binom{n}{c_{0}, c_{1}, \ldots, c_{d}}\left(\prod_{i=0}^{d}\binom{d}{i}^{2 c_{i}}\right)\left(\sum_{i=1}^{d} i c_{i}\right)!\left(d n-\sum_{i=1}^{d} i c_{i}\right)! \\
= & \frac{\binom{2 d}{d}^{n} 2^{d n}}{\binom{2 d n}{d n}} \sum_{c_{0}+c_{1}+\cdots+c_{d}=n}\left(\prod_{i=0}^{d}\binom{d}{i}^{2 c_{i}}\right) \frac{n!\left(\sum_{i=1}^{d} i c_{i}\right)!\left(d n-\sum_{i=1}^{d} i c_{i}\right)!}{(d n)!\prod_{i=0}^{d} c_{i}!} . \tag{3.72}
\end{align*}
$$

By setting $c_{0}=n-\sum_{i=1}^{d} c_{i}$ and applying Stirling's formula (3.70) to the factorials, $(d n)!,\left(\sum i c_{i}\right)!$, and $\left(d n-\sum i c_{i}\right)!$, we can approximate the summand corresponding to
$\mathbf{c}=\left(c_{1}, c_{2}, \ldots, c_{d}\right)$ in (3.72) by

$$
\begin{align*}
&\left(\prod_{i=0}^{d}\binom{d}{i}^{2 c_{i}}\right) \frac{n^{n}\left(\sum_{i=1}^{d} i c_{i}\right)^{\sum_{i=1}^{d} i c_{i}}\left(d n-\sum_{i=1}^{d} i c_{i}\right)^{d n-\sum_{i=1}^{d} i c_{i}}}{(d n)^{d n}\left(n-\sum_{i=1}^{d} c_{i}\right)^{n-\sum_{i=1}^{d} c_{i}} \prod_{i=1}^{d} c_{i}^{c_{i}}} \\
& \times\left(\frac{\left(\sum_{i=1}^{d} i c_{i}\right)\left(d n-\sum_{i=1}^{d} i c_{i}\right)}{(2 \pi)^{d-1} d\left(n-\sum_{i=1}^{d} c_{i}\right) \prod_{i=1}^{d} c_{i}}\right)^{1 / 2} \tag{3.73}
\end{align*}
$$

We can also use Stirling's formula (3.70) to approximate the first part of (3.72) as

$$
\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} \sqrt{\pi d n}
$$

Next, we divide the top and bottom of the upper term in (3.73) by $n^{d n+1}$ to get

$$
\left[\left(\prod_{i=1}^{d}\binom{d}{i}^{2 c_{i} / n}\right) \frac{\left(\sum_{i=1}^{d} i c_{i} / n\right)^{\sum_{i=1}^{d} i_{i} / n}\left(d-\sum_{i=1}^{d} i c_{i} / n\right)^{d-\sum_{i=1}^{d} i_{i} / n}}{d^{d}\left(1-\sum_{i=1}^{d} c_{i} / n\right)^{1-\sum_{i=1}^{d} c_{i} / n} \prod_{i=1}^{d}\left(c_{i} / n\right)^{c_{i} / n}}\right]^{n} .
$$

Similarly, we replace $c_{i}$ by $c_{i} / n$ in the lower term to get

$$
\left(\frac{\left(\sum_{i=1}^{d} i c_{i} / n\right)\left(d-\sum_{i=1}^{d} i c_{i} / n\right)}{(2 \pi)^{d-1} d n^{d-1}\left(1-\sum_{i=1}^{d} c_{i} / n\right) \prod_{i=1}^{d}\left(c_{i} / n\right)}\right)^{1 / 2}
$$

Hence, we can write

$$
\begin{equation*}
\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right] \rightarrow\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{2 n} \frac{1}{(2 \pi n)^{d / 2-1}} \sum_{c_{1}+c_{2}+\cdots c_{d} \leq n} e^{n g(\mathbf{c} / n)} h(\mathbf{c} / n), \tag{3.74}
\end{equation*}
$$

where

$$
\begin{aligned}
g(\mathbf{x})= & d \log (2)-d \log (d)-\log \binom{2 d}{d} \\
& +\left(\sum_{i=1}^{d} i x_{i}\right) \log \left(\sum_{i=1}^{d} i x_{i}\right)+\left(d-\sum_{i=1}^{d} i x_{i}\right) \log \left(d-\sum_{i=1}^{d} i x_{i}\right) \\
& +\sum_{i=1}^{d} x_{i} \log \binom{d}{i}^{2}-\sum_{i=1}^{d} x_{i} \log \left(x_{i}\right)-\left(1-\sum_{i=1}^{d} x_{i}\right) \log \left(1-\sum_{i=1}^{d} x_{i}\right) .
\end{aligned}
$$

and

$$
h(\mathbf{x})=\left(\frac{\left(\sum_{i=1}^{d} x_{i}\right)\left(d-\sum_{i=1}^{d} x_{i}\right)}{2\left(\prod_{1 \leq i \leq d} x_{i}\right)\left(1-\sum_{i=1}^{d} x_{i}\right)}\right)^{1 / 2}
$$

As we are only looking for an asymptotic upper bound on $\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right] / \mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}$, we reduce our problem to estimating the following upper bound

$$
\begin{equation*}
\frac{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}} \leq \frac{2}{d(2 \pi n)^{d / 2}} \int_{\mathbf{x} \in S(d)} e^{n g(\mathbf{x})} h(\mathbf{x}) d \mathbf{x} \tag{3.75}
\end{equation*}
$$

where $S(d)=\left\{\mathbf{x} \in[0,1]^{d}:|\mathbf{x}| \leq 1\right\}$; we ensure $g(\mathbf{x})$ and $h(\mathbf{x})$ are both defined for all of $S(d)$ by setting $g(\mathbf{x})=h(\mathbf{x})=0$ if any of the logarithms would evaluate to $\log (0)$ (or any of the terms in the denominator of $h(\mathbf{x})$ are 0 ). We observe that the contribution to individual terms of the integral in (3.75) from $h(\mathbf{x})$ is very small in comparison with $e^{n g(\mathbf{x})}$. Hence, the value of the integral is dominated by the terms where $g(\mathbf{x})$ is close to the maxima of $g$ on $S(d)$.

We will now estimate the integral in (3.75) using Laplace's method. First, we show that $g(\mathbf{x})$ has a unique local maximum at a point $\mathbf{a}$; hence, the value of $e^{n g(\mathbf{x})}$ is negligible, except in a small neighbourhood of $\mathbf{a}$. By expanding $g(\mathbf{x})$ around a using a Taylor series, we are able to approximate (3.75) by a Gaussian integral. We can then evaluate this using Theorem 3.41 and Lemma 3.46 to obtain an asymptotic estimate for (3.75).

First, we find the unique local maximum of $g$ in $S(d)$. We write $g(\mathbf{x})$ as the sum

$$
d \log (2)-\log \binom{2 d}{d}-d \log (d)+g_{1}(\mathbf{x})+g_{2}(\mathbf{x}),
$$

where

$$
g_{1}(\mathbf{x})=\left(\sum_{i=1}^{d} i x_{i}\right) \log \left(\sum_{i=1}^{d} i x_{i}\right)+\left(d-\sum_{i=1}^{d} i x_{i}\right) \log \left(d-\sum_{i=1}^{d} i x_{i}\right),
$$

and

$$
g_{2}(\mathbf{x})=\sum_{i=1}^{d} x_{i} \log \binom{d}{i}^{2}-\sum_{i=1}^{d} x_{i} \log \left(x_{i}\right)-\left(1-\sum_{i=1}^{d} x_{i}\right) \log \left(1-\sum_{i=1}^{d} x_{i}\right) .
$$

The value of $g_{1}(\mathbf{x})$ is

$$
m \log (m)+(d-m) \log (d-m)
$$

for $m=\sum_{i} i x_{i}$, and, hence, the maxima of $g_{1}$ are given by the set of points satisfying

$$
\begin{equation*}
\sum_{i=1}^{d} i x_{i}=d-\sum_{i=1}^{d} i x_{i} . \tag{3.76}
\end{equation*}
$$

Taking the partial derivatives of $g_{2}$ gives

$$
\frac{\partial g_{2}}{\partial x_{i}}=\log \binom{d}{i}^{2}-\log x_{i}+\log \left(1-\sum_{j=1}^{d} x_{j}\right),
$$

for each $i=1,2, \ldots, d$. Hence, any critical point of $g_{2}$ must satisfy

$$
\begin{equation*}
x_{i}=\binom{d}{i}^{2}\left(1-\sum_{j=1}^{d} x_{j}\right), \quad i=1,2, \ldots, d \tag{3.77}
\end{equation*}
$$

This is a system of $d$ equations in $d$ unknowns, so $g_{2}$ has a unique critical point. Using Lemma 3.45, we can verify that a particular point a, defined by

$$
\mathbf{a}=\left(\frac{\binom{d}{i}^{2}}{\binom{2 d}{d}}: i=1 \ldots d\right),
$$

satisfies (3.77). Hence, a is the unique critical point of $g_{2}$.
Recall the condition for a to be a local maximum of $g_{2}$ : we must have that the Hessian matrix, $D^{2} g_{2}(\mathbf{a})$, of second order partial derivatives evaluated at a is negative definite. The second order derivatives of $g_{2}(\mathbf{x})$ are

$$
\begin{array}{ll}
\frac{\partial g}{\partial x_{i} x_{j}}=-\frac{1}{1-\sum_{i=1}^{d} x_{i}}, & i \neq j, \\
\frac{\partial g}{\partial x_{i} x_{j}}=-\frac{1}{1-\sum_{i=1}^{d} x_{i}}-\frac{1}{x_{i}} & i=j . \tag{3.79}
\end{array}
$$

Evaluating (3.78) and (3.79) at a gives the values for the Hessian matrix

$$
D^{2} g_{2}(\mathbf{a})_{i, j}=\left\{\begin{array}{ll}
-\binom{2 d}{d} & i \neq j \\
-\binom{2 d}{d}-\frac{\binom{2 d}{d}}{\binom{d}{i}^{2}} & i=j
\end{array} .\right.
$$

$D^{2} g_{2}(\mathbf{a})$ is certainly negative definite, so $\mathbf{a}$ is the unique local maximum of $g_{2}$ on $S(d)$.
Moreover, since a satisfies (3.76), a is also a local maximum of $g_{1}$. Recall that $g(\mathbf{x})$ is equal to the sum of $g_{1}(\mathbf{x})$ and $g_{2}(\mathbf{x})$ (plus some terms not dependent on $\mathbf{x}$ ). Since a is the unique point which is a local maximum of both functions, it follows that a must be the unique local maximum of $g(\mathbf{x})$. Furthermore, using Lemma 3.45, we can see that $g(\mathbf{a})=0$.

Next we consider the region

$$
\bar{S}(d)=\left\{\mathbf{x} \in[0,1]^{d}:\left|x_{i}-a_{i}\right| \leq n^{-1 / 2} \log (n), \text { for each } i=1,2, \ldots, d\right\} .
$$

We note that within $\bar{S}(d)$, we have that $n g(\mathbf{x})$ is approximated by its second order Taylor expansion

$$
n g(\mathbf{x}) \rightarrow \frac{n}{2}(\mathbf{x}-\mathbf{a})^{T} D^{2} g(\mathbf{a})(\mathbf{x}-\mathbf{a}) .
$$

To see this consider the contribution from any third order term to $n g(\mathbf{x})$ :

$$
\begin{equation*}
n\left(x_{i}-a_{i}\right)\left(x_{j}-a_{j}\right)\left(x_{k}-a_{k}\right) \frac{\partial g}{\partial x_{i} x_{j} x_{k}}(\mathbf{a}) . \tag{3.80}
\end{equation*}
$$

Within the region $\bar{S}(d)$, the absolute value of (3.80) is $O\left(n^{-1 / 2} \log (n)^{3}\right)$, which is equal to 0 in the limit.

Moreover, the contribution to (3.75) from points outside $\bar{S}(d)$ will be negligible. To see this, we note that the point $\mathbf{x}$ with $\left|x_{i}-a_{i}\right| \geq n^{-1 / 2} \log (n)$ that maximizes $g(\mathbf{x})$ will, necessarily, be the point closest to a, i.e., one of the two points with $\left|x_{i}-a_{i}\right|=n^{-1 / 2} \log (n)$ and $x_{j}=a_{j}$ for $j \neq i$. Then, by considering the Taylor expansion of $g(\mathbf{x})$ around $\mathbf{a}$, we can see that, for all points with $x_{i}-a_{i} \geq n^{-1 / 2} \log (n)$, for some $i=1,2, \ldots, d$, we have

$$
e^{n g(\mathbf{x})} \leq e^{-\gamma \log (n)^{2}},
$$

where $\gamma$ is some absolute constant independent of $n$. That is, in the limit, we will have $e^{n g(\mathbf{x})} h(\mathbf{x}) \rightarrow 0$ for $\mathbf{x} \notin \bar{S}(d)$. Hence, we can approximate (3.75) as

$$
\begin{equation*}
\frac{2}{d(2 \pi n)^{d / 2}} \int_{a_{1}-n^{-1 / 2} \log (n)}^{a_{1}+n^{-1 / 2} \log (n)} \cdots \int_{a_{d}-n^{-1 / 2} \log (n)}^{a_{d}+n^{-1 / 2} \log (n)} e^{n g(\mathbf{x})} h(\mathbf{x}) d \mathbf{x} . \tag{3.81}
\end{equation*}
$$

We want to replace the integral in (3.81) by a Gaussian. To achieve this, we apply the change of variable

$$
\mathbf{x}=\mathbf{a}+\mathbf{z} n^{-1 / 2},
$$

and integrate over $\mathbf{z} \in[-\log (n), \log (n)]^{d}$. As $n \rightarrow \infty$, we have

$$
h\left(\mathbf{a}+\mathbf{z} n^{-1 / 2}\right) \rightarrow h(\mathbf{a})=\frac{d}{2}\left(\frac{\binom{2 d}{d}^{d+1}}{2 \prod_{i=0}^{d}\binom{d}{i}^{2}}\right)^{1 / 2},
$$

and

$$
g\left(\mathbf{a}+\mathbf{z} n^{-1 / 2}\right) \rightarrow \frac{1}{2} \mathbf{z}^{T} M \mathbf{z},
$$

where $M=-D^{2} g(\mathbf{a})$. Furthermore, we have

$$
d \mathbf{x}=n^{d / 2} d \mathbf{z}
$$

Hence, by dominated convergence ${ }^{1}$, (3.81) is asymptotically equal to

$$
\begin{equation*}
\left(\frac{\binom{2 d}{d}^{d+1}}{2 \prod_{i=0}^{d}\binom{d}{i}^{2}}\right)^{1 / 2} \frac{1}{(2 \pi)^{d / 2}} \int_{[-\log (n), \log (n)]^{d}} e^{-\frac{1}{2} \mathbf{x}^{T} M \mathbf{z}} d \mathbf{z} \tag{3.82}
\end{equation*}
$$

[^6]We can extend the range of integration to $\mathbb{R}^{d}$ without changing the asymptotic value of (3.82), since the contribution from all values outside of $[-\log (n), \log (n)]^{d}$ will be negligible, giving

$$
\begin{equation*}
\left(\frac{\binom{2 d}{d}^{d+1}}{2 \prod_{i=0}^{d}\binom{d}{i}^{2}}\right)^{1 / 2} \frac{1}{(2 \pi)^{d / 2}} \int_{\mathbb{R}^{d}} e^{-\frac{1}{2} \mathbf{z}^{T} M \mathbf{z}} d \mathbf{z} \tag{3.83}
\end{equation*}
$$

Recall that $D^{2} g(\mathbf{a})$ is a negative definite matrix. Hence, the integral in (3.83) is a $d$-dimensional Gaussian, which we can evaluate using Theorem 3.41 as

$$
\frac{(2 \pi)^{d / 2}}{\sqrt{|M|}}
$$

From Lemma 3.46, we see that

$$
|M|=\frac{\binom{2 d}{d}^{d+1}}{\prod_{i=1}^{d}\binom{d}{i}^{2}} \frac{d-1}{2 d-1}
$$

and so we can conclude that

$$
\lim _{n \rightarrow \infty} \frac{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}} \leq \sqrt{\frac{2 d-1}{2 d-2}}
$$

Remark 3.48. It would also be possible to obtain an exact asymptotic estimate for $\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]$ by using the Euler-Maclaurin summation formula [25, Chapter 3]. This is a technique that can be used to approximate certain, smooth, summations by integrals. The terms in the summation (3.74) vary smoothly enough for this to be applicable. However, the asymptotic upper bound computed in Theorem 3.47 is sufficient for our purposes, as we can obtain the matching asymptotic lower bound as a side-effect of our application of Theorem 3.19 in $\S 3.4 .3$, thus avoiding the need to introduce more analytical machinery.

The next theorem estimates the expected number of Eulerian orientations in a random $2 d$-regular graph. We use the same idea as the proof of Theorem 3.30, except we are now conditioning on there being no loops or double edges. However, where we had to do a tedious analysis of the number of loops and double arcs that can occur on configurations sampled in weighted distributions in the proof of Theorem 3.30, we can automatically infer Theorem 3.49 from the joint convergence of the random variables in Janson's Theorem (Theorem 3.19; see also Remark 3.21) and the results presented in the next section.

When we were analysing the moments of the number of arborescences of a random $d$-in/d-out graph, we needed to condition on a random variable not involved in the application of Janson's Theorem, the number of double arcs. Hence, we could not use the joint convergence of the random variables implied by Janson's Theorem to obtain Theorem 3.30.

Theorem 3.49. Let $n \in \mathbb{N}$ and let $d$ be some fixed positive integer. Let $\mathcal{E}_{n, 2 d}$ denote the number of Eulerian orientations of a graph $G$ chosen uniformly at random from $\mathbb{G}(n, 2 d)$. Then

$$
\begin{aligned}
& \mathbb{E}\left[\mathcal{E}_{n, 2 d}\right] \rightarrow e^{-3 / 4}\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} \sqrt{\pi d n}, \text { and } \\
& \frac{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}\right)^{2}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]^{2}} \rightarrow \exp \left(-\frac{1}{2(2 d-1)}-\frac{1}{4(2 d-1)^{2}}\right) \sqrt{\frac{2 d-1}{2 d-2}} .
\end{aligned}
$$

Proof. Follows from Theorem 3.47, Remark 3.21, Theorem 3.15, and Lemma 3.51.
In $\S 3.4 .3$ we will prove that $\mathcal{E}_{n, 2 d}^{\star}$ and $X_{i, n}$ (the random variables counting the number of $i$-cycles in $\sigma(\mathrm{F})$ when F is chosen uniformly at random from $\Omega_{n, 2 d}$ ) satisfy the conditions of Janson's Theorem (Theorem 3.19) with $\lambda_{i}=(2 d-1)^{i} / 2 i$ and $\delta_{i}=(2 d-1)^{-i}$. Hence, in light of Remark 3.21, we have

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \exp (-1 / 2-1 / 4),
$$

and

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{2}\right]}{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]} \rightarrow \exp \left(-\frac{3}{2}-\frac{1}{2(2 d-1)}-\frac{1}{4(2 d-1)^{2}}\right)
$$

### 3.4.3 The asymptotic distribution of $\mathcal{E}_{n, 2 d}$

In this section we characterise the asymptotic distribution of the number of Eulerian orientations of a random $2 d$-regular graph.

Theorem 3.50. Let d be some fixed integer greater than 1 , and let $\mathcal{E}_{n, 2 d}$ count the number of Eulerian orientations of a random $G \in \mathbb{G}(n, 2 d)$. Then

$$
\frac{\mathcal{E}_{n, 2 d}}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]} \rightarrow \prod_{i=3}^{\infty}\left(1+\frac{1}{(2 d-1)^{i}}\right)^{Z_{i}} e^{-1 / 2 i}
$$

as $n \rightarrow \infty$, where $Z_{i}$ are independent Poisson random variables with $\mathbb{E}\left[Z_{i}\right]=\frac{(2 d-1)^{i}}{2 i}$.

Although we do not have an immediate algorithmic application of this result, we believe it to be a milestone on the road to characterising the asymptotic distribution of the number of Euler tours of a random $2 d$-regular graph, a result that would have algorithmic value (see §3.5).

To obtain an asymptotic distribution of the number of Eulerian orientations, $\mathcal{E}_{n, 2 d}$, we will use Janson's Theorem (Theorem 3.19). In our application of Janson's Theorem, we will have $Y_{n}=\mathcal{E}_{n, 2 d}$, and $X_{i, n}$ counting simple cycles of length $i$ in random $G \in \mathbb{G}(n, 2 d)$. For Condition 1 of Theorem 3.19, we need to show that the random variables $X_{i, n}$ are asymptotic to Poisson random variables. It is a classic result in the theory of random graphs (due to Bollobás [11]; given as Theorem 3.15 in this thesis) that, for any fixed $k$, the random variables counting $i$-cycles, for $i \leq k$, are asymptotic to independent random variables with means

$$
\lambda_{i}=\frac{(2 d-1)^{i}}{2 i} .
$$

The following lemma (see also Lemma 3.22) establishes Condition 2 of Theorem 3.19.
Lemma 3.51. Let d be some fixed positive integer, let $\mathcal{E}_{n, 2 d}^{\star}$ denote the number of Eulerian orientations in $\sigma(\mathrm{F})$ for $\mathrm{F} \in \Omega_{n, 2 d}$, and let $X_{i, n}$ denote the number of cycles of length $i$ in $\sigma(\mathrm{F})$ when F is chosen uniformly at random from $\Omega_{n, 2 d}$. Then, for any fixed set of non-negative integers $j_{1}, j_{2}, \ldots, j_{k}$ we have

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}},
$$

where $\mu_{i}=\frac{(2 d-1)^{i}+1}{2 i}$.
Proof. We start by establishing

$$
\begin{equation*}
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \mu_{i} \tag{3.84}
\end{equation*}
$$

for $i \geq 1$. Convergence of the factorial moments then follows for similar reasons as before (see Lemma 3.29 or Lemma 3.16).

In this proof, we consider cycles in configurations $\mathrm{F} \in \Omega_{n, 2 d}$. We say a partial configuration $\mathrm{F}^{\prime} \subseteq \mathrm{F}$ is a cycle, if the projection of $\mathrm{F}^{\prime}, \sigma\left(\mathrm{F}^{\prime}\right)$, is a cycle in $\sigma(\mathrm{F})$. That is, for each $i \geq 1$, we call a set of edges $e_{0}, e_{1}, \ldots, e_{i-1}$ in $\mathrm{F} \in \Omega_{n, d}$ an $i$-cycle if there is a sequence of vertices $v_{0}, v_{1}, \ldots, v_{i-1}$ such that $e_{j}$ has one endpoint in $W_{v_{j}}$ and one in $W_{v_{j+1 \bmod i}}$ for $i=0,1, \ldots, i-1$. We will use $\mathcal{C}_{i}(\mathrm{~F})$ to denote the set of $i$-cycles in F , and use $\mathcal{C}_{i, n}$ to denote the set of all $i$-cycles that can occur in some $\mathrm{F} \in \Omega_{n, 2 d}$.

Recalling the definition of $\bar{\Omega}_{n, 2 d}$ from Theorem 3.47, we can expand the left-hand side of (3.84) as

$$
\begin{aligned}
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} & =\frac{1}{\left|\bar{\Omega}_{n, 2 d}\right|} \sum_{\mathrm{F} \in \Omega_{n, 2 d}}|\mathrm{EO}(\mathrm{~F})|\left|\mathcal{C}_{i}(\mathrm{~F})\right| \\
& =\sum_{\mathrm{F} \in \Omega_{n, 2 d}} \sum_{\mathcal{E} \in \mathrm{EO}(\mathrm{~F})} \sum_{C \in \mathcal{C}_{i}(\mathrm{~F})} \frac{1}{\left|\bar{\Omega}_{n, 2 d}\right|} \\
& =\sum_{C \in \mathcal{C}_{i, n}} \frac{\left|\left\{(\mathrm{~F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}: C \subseteq \mathrm{~F}\right\}\right|}{\left|\bar{\Omega}_{n, 2 d}\right|} \\
& =\left|\mathcal{C}_{i, n}\right| \mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}\right],
\end{aligned}
$$

since the probability $\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}\right]$ is the same for each $C \in \mathcal{C}_{i, n}$. Hence, to prove (3.84), it suffices to estimate $\left|\mathcal{C}_{i, n}\right|$ and $\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}\right]$.

For $i \geq 3$, the number of ways to choose an $i$-cycle on $V$ is $\frac{(n)_{i}}{2 i}$, since there are $(n)_{i}$ ways to choose a sequence of $i$ vertices, and sequences which are the reverse of one another or equivalent under cyclic permutation are considered to be the same cycle. For each $i$-cycle $C$, there are $\left((2 d)_{2}\right)^{i}$ ways to choose a partial configuration that projects to $C$, so

$$
\begin{equation*}
\left|\mathcal{C}_{i, n}\right|=\frac{(n)_{i}(2 d)^{i}(2 d-1)^{i}}{2 i} \rightarrow \frac{(n)^{i}(2 d)^{i}(2 d-1)^{i}}{2 i} \tag{3.85}
\end{equation*}
$$

for $i \geq 3$. In fact, (3.85) also describes the number of elements in $\mathcal{C}_{i, n}$ for $i=1$ and $i=2$. There are $n$ ways to choose a loop vertex $v$, and $\binom{2 d}{2}$ ways to choose the endpoints for a loop at $v$. There are $\binom{n}{2}$ ways to choose a pair of vertices $u$ and $v$ and $2\binom{2 d}{2}^{2}$ ways to choose a partial configuration projecting to a double edge between $u$ and $v$.

To estimate $\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}\right]$ we need to count the number of elements of $\bar{\Omega}_{n, 2 d}$ with $C \subseteq \mathrm{~F}$. Recall how we enumerate the elements of $\bar{\Omega}_{n, 2 d}$ when computing $\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]$ in the proof of Theorem 3.47. First, we choose a partition of $W$ into $S$ and $T$ by dividing each $W_{v}$ into equally sized disjoint subsets $S_{v}$ and $T_{v}$. Then, we choose a perfect matching from $S$ to $T$. We follow the same approach here, by first assigning the endpoints of the edges in $C$ to $S$ and $T$ and then choosing the remainder of $S$ and $T$. A vertex $v$ on the cycle $C$ is said to be balanced if the edges incident with $W_{v}$ on $C$ contain one point from $S_{v}$ and one point from $T_{v}$, negatively imbalanced if the edges contain two points from $S_{v}$ and positively imbalanced if the edges contain two points from $T_{v}$. Positively and negatively imbalanced vertices must alternate around $C$; therefore there must be an equal number of each. Hence, assigning the endpoints of the edges in $C$ to $S$ and $T$ is equivalent to choosing an even subset of the vertices of $C$ to be imbalanced, and then fixing one vertex in that set to be positively imbalanced.

Given an assignment of the endpoints of the edges of $C$ to $S$ and $T$, we now want to partition the rest of the points in each $W_{v}$ into $S_{v} \cup T_{v}$. If $v$ is an imbalanced vertex of $C$, then we have already chosen two points from $W_{v}$ to lie in $S_{v}$ or in $T_{v}$, and, hence, there are $\binom{2 d-2}{d-2}$ ways to complete the partitioning of $W_{v}$ into $S_{v} \cup T_{v}$. If $v$ is a balanced vertex on $C$ then there are $\binom{2 d-2}{d-1}$ ways to complete the partition of $W_{v}$ into $S_{v} \cup T_{v}$, since we have already chosen one element of $S_{v}$ and one element of $T_{v}$. Finally, if $v$ is not on $C$ there are $\binom{2 d}{d}$ ways to partition $W_{v}$ into $S_{v} \cup T_{v}$. Once we have chosen $S$ and $T$, there are $(d n-i)$ ! ways to choose the remainder of the perfect matching from $S$ to $T$. Hence, the number of pairs $(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}$ where F contains a particular cycle $C \in \mathcal{C}_{i, n}$ with $2 j$ imbalanced vertices is

$$
2\binom{i}{2 j}\binom{2 d}{d}^{n-i}\binom{2 d-2}{d-2}^{2 j}\binom{2 d-2}{d-1}^{i-2 j}(d n-i)!
$$

We note that

$$
\binom{2 d-2}{d-2}=\frac{(d-1)}{2(2 d-1)}\binom{2 d}{d}, \text { and }\binom{2 d-2}{d-1}=\frac{d}{2(2 d-1)}\binom{2 d}{d} .
$$

Hence, the number of pairs $(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}$ where F contains a particular cycle $C \in \mathcal{C}_{i, n}$ with $2 j$ imbalanced vertices is

$$
2\binom{2 d}{d}^{n}(d n-i)!\left(\frac{d}{2(2 d-1)}\right)^{i}\binom{i}{2 j}\left(\frac{d-1}{d}\right)^{2 j} .
$$

Using the following identity

$$
\sum_{j=0}^{\lfloor i / 2\rfloor}\binom{i}{2 j} x^{2 j}=\frac{(1+x)^{i}}{2}+\frac{(1-x)^{i}}{2}
$$

we see that

$$
\sum_{j=0}^{\lfloor i / 2\rfloor}\binom{i}{2 j}\left(\frac{d-1}{d}\right)^{2 j}=\frac{1}{2}\left(\frac{2 d-1}{d}\right)^{i}+\frac{1}{2 d^{i}} .
$$

Hence, we can see that the number of pairs $(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}$ with $C \subseteq \mathrm{~F}$ is equal to

$$
\binom{2 d}{d}^{n} \frac{(d n-i)!}{2^{i}}\left(1+\frac{1}{(2 d-1)^{i}}\right) .
$$

Dividing by the number of elements in $\bar{\Omega}_{n, 2 d}$, which we know to be $\binom{2 d}{d}^{n}(d n)$ !, gives

$$
\begin{aligned}
\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}) \in \bar{\Omega}_{n, 2 d}\right] & =\frac{1}{2^{i}(d n)_{i}}\left(1+\frac{1}{(2 d-1)^{i}}\right) \\
& \rightarrow \frac{1}{2^{i} d^{i} n^{i}}\left(1+\frac{1}{(2 d-1)^{i}}\right) .
\end{aligned}
$$

Finally, multiplying by the number of elements of $\mathcal{C}_{i, n}$, which is given in (3.85), we obtain

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star} X_{i, n}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \frac{(2 d-1)^{i}+1}{2 i} .
$$

The factorial moments converge as stated for similar reasons as were given in Lemma 3.29, see also, e.g, [100, Lemma 2.7] (given as Lemma 3.16 of this thesis). Suppose we have a graph $H$ on $r$ vertices with $r+s$ edges, for $r$ and $s$ fixed and $s>0$. The number of ways to choose a partial configuration $\mathrm{F}^{\prime}$ projecting to $H$ is $O\left(n^{r}\right)$, and the probability of $\mathrm{F}^{\prime} \subseteq \mathrm{F}$ when $(\mathrm{F}, \mathcal{E})$ is chosen uniformly at random from $\bar{\Omega}_{n, 2 d}$ is $O\left(n^{-r-s}\right)$. Thus, we can conclude that the contribution to

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]}
$$

from sets of cycles in which any vertex occurs on more than one cycle is $O\left(n^{-1}\right)$, and so can be treated as 0 as $n \rightarrow 0$. Hence, cycles occur (asymptotically) independently and we can conclude

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}} .
$$

We will now prove Theorem 3.50 by applying Janson's Theorem (Theorem 3.19) with Lemma 3.51, Theorem 3.49, and Theorem 3.15 providing the required values.

Proof of Theorem 3.50. Our goal is to show

$$
\frac{\mathcal{E}_{n, 2 d}}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]} \rightarrow \prod_{i=3}^{\infty}\left(1+\frac{1}{(2 d-1)^{i}}\right)^{Z_{i}} e^{-1 / 2 i}
$$

as $n \rightarrow \infty$, where $Z_{i}$ are independent Poisson random variables with $\mathbb{E}\left[Z_{i}\right]=\frac{(2 d-1)^{i}}{2 i}$. We first apply Theorem 3.19 with $Y_{n}=\mathcal{E}_{n, 2 d}^{\star}$ (from Theorem 3.47) and $X_{i, n}$ the random variables counting $i$-cycles from Lemma 3.51 and Theorem 3.15 to obtain

$$
\frac{\mathcal{E}_{n, 2 d}^{\star}}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow W=\prod_{i=1}^{\infty}\left(1+\frac{1}{(2 d-1)^{i}}\right)^{Z_{i}} e^{-1 / 2 i}
$$

as $n \rightarrow \infty$. The convergence of $\mathcal{E}_{n, 2 d}^{\star} / \mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]$ to $W$ implied by Theorem 3.19 is joint with the convergence of the random variables $X_{i, n}$ to $Z_{i}$. Hence, the distribution of $\mathcal{E}_{n, 2 d}^{\star}$ conditioned on $X_{1, n}=X_{2, n}=0$ is asymptotic to that of $W$ conditioned on $Z_{1}=Z_{2}=0$ (see Remark 3.21). Thus, it suffices to show that Theorem 3.19 holds with $Y_{n}=\mathcal{E}_{n, 2 d}^{\star}$.

To apply Theorem 3.19, we need to verify the conditions 1 to 4 (of Theorem 3.19). Theorem 3.15, due to Bollobás [11, 12], tells us that the $X_{i, n}$ converge jointly to Poisson random variables with means

$$
\lambda_{i}=\frac{(2 d-1)^{i}}{2 i} .
$$

Hence, Condition 1 holds.
Lemma 3.51 tells us that the random variables $X_{i, n}$, taken together with $\mathcal{E}_{n, 2 d}^{\star}$, satisfy, for any sequence of non-negative integers $j_{1}, j_{2}, \ldots, j_{k}$,

$$
\frac{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\left(X_{1, n}\right)_{j_{1}}\left(X_{2, n}\right)_{j_{2}} \cdots\left(X_{k, n}\right)_{j_{k}}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}},
$$

as $n \rightarrow \infty$. Hence, by Lemma 3.22, we see that Condition 2 is satisfied with

$$
\delta_{i}=\frac{\mu_{i}}{\lambda_{i}}-1=\frac{1}{(2 d-1)^{i}} .
$$

For Condition 3, we evaluate

$$
\sum_{i=1}^{\infty} \lambda_{i} \delta_{i}^{2}=\frac{1}{2} \log \left(\frac{2 d-1}{2 d-2}\right),
$$

since

$$
\log \left(\frac{2 d-1}{2 d-2}\right)=\log \left(\frac{1}{1-\frac{1}{2 d-1}}\right)=\sum_{i=1}^{\infty} \frac{1}{i(2 d-1)^{i}} .
$$

Finally, Theorem 3.49 gives

$$
\lim _{n \rightarrow \infty} \frac{\mathbb{E}\left[\left(\mathcal{E}_{n, 2 d}^{\star}\right)^{2}\right]}{\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{\star}\right]^{2}} \leq \sqrt{\frac{2 d-1}{2 d-2}}
$$

so Condition 4 is also satisfied.

### 3.4.4 Bounding $\mathcal{E}_{n, 2 d}$ with high probability

Recall that Schrijver ([78], given as Theorem 1.56 of this thesis) has shown that the number of Eulerian orientations in any $2 d$-regular graph is at least

$$
\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n} .
$$

We can use Theorem 3.49 to infer that almost every $2 d$-regular graph has few Eulerian orientations, in the sense that the number of Eulerian orientations is at most a linear factor larger than Schrijver's lower bound.

Theorem 3.52. Let $d$ be a fixed positive integer satisfying $d \geq 2$. Suppose $G$ is a random $2 d$-regular graph. Then, with probability tending to 1 as $n \rightarrow \infty$, we have

$$
\mathbb{P}\left[\# \mathrm{EO}(G) \leq O(n)\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n}\right] \rightarrow 1
$$

as $n \rightarrow \infty$.

Proof of Theorem 3.52. Let $\mathcal{E}_{n, 2 d}$ denote the number of Eulerian orientations of a graph chosen uniformly at random from $\mathbb{G}(n, 2 d)$. We have shown, in Theorem 3.49, that the expectation of $\mathcal{E}_{n, 2 d}$ satisfies

$$
\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right] \rightarrow e^{-3 / 4} \sqrt{\pi d n}\left(\frac{\binom{2 d}{d}}{2^{d}}\right)^{n}
$$

Hence, the statement of the theorem is equivalent to saying

$$
\mathbb{P}\left\{\mathcal{E}_{n, 2 d} \leq \sqrt{n} \mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]\right\} \rightarrow 1,
$$

as $n \rightarrow \infty$.
Recall Chebyshev's inequality (Theorem 3.8): for any random variable $X$,

$$
\mathbb{P}\{|X-\mathbb{E}[X]| \geq a \mathbb{E}[X]\} \leq \frac{\operatorname{Var}[X]}{a^{2} \mathbb{E}[X]^{2}},
$$

By Theorem 3.49, we can assume the existence of a function $c_{d}(n)$ such that

$$
\mathbb{E}\left[\mathcal{E}_{n, 2 d}^{2}\right]=c_{d}(n) \mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]^{2},
$$

and

$$
c_{d}(n) \rightarrow \exp \left(-\frac{4 d-1}{4(2 d-1)^{2}}\right) \sqrt{\frac{2 d-1}{2 d-2}}
$$

Then, we have

$$
\begin{aligned}
\mathbb{P}\left\{\mathcal{E}_{n, 2 d} \geq \sqrt{n} \mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]\right\} & \leq \mathbb{P}\left\{\left|\mathcal{E}_{n, 2 d}-\mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]\right| \geq(\sqrt{n}-1) \mathbb{E}\left[\mathcal{E}_{n, 2 d}\right]\right\} \\
& \leq \frac{c_{d}(n)-1}{(\sqrt{n}-1)^{2}} \\
& \rightarrow 0,
\end{aligned}
$$

where we are using Chebyshev's inequality with $a=\sqrt{n}-1$ and the fact that, for any random variable $X, \operatorname{Var}[X]=\mathbb{E}\left[X^{2}\right]-\mathbb{E}[X]^{2}$.

Theorem 3.52 tells us that almost every $2 d$-regular graph has few Eulerian orientations, in the sense that the number of Eulerian orientations is within a linear factor of Schrijver's lower bound. An example of a regular graph $G$ for which the number of Eulerian orientations greatly exceeds Schrijver's lower bound is the toroidal grid (see §1.3.1). Lieb [61] showed that the number of Eulerian orientations of the $\sqrt{n} \times \sqrt{n}$ toroidal grid is

$$
\left(\frac{4}{3}\right)^{3 n / 2} \gg\left(\frac{\binom{4}{2}}{2^{2}}\right)^{n}
$$

In $\S 3.6$ we show a similar (but weaker) result for the number of Euler tours of toroidal grids.

Let $G$ be a particular $2 d$-regular graph, let $\vec{G}$ be a random orientation of $G$, and let $I_{v}$ denote the indicator variable for $v$ satisfying the Eulerian condition in $\vec{G}$. There are $2^{2 d}$ different orientations for the edges incident with $v$, of which $\binom{2 d}{d}$ satisfy the Eulerian condition. Hence,

$$
\mathbb{P}\left(I_{v}=1\right)=\frac{\binom{2 d}{d}}{2^{2 d}}
$$

Now, suppose all the $I_{v}$ were independent. Then the total number of Eulerian orientations of $G$ would be

$$
2^{d n}\left(\frac{\binom{2 d}{d}}{2^{2 d}}\right)^{n}=\left(\frac{\binom{\binom{d}{d}}{2^{d}}^{n} . . . . . . .}{}\right.
$$

Theorem 3.52 tells us that this rather crude estimate actually manages to capture the exponential factor in the number of Eulerian orientations of almost every $2 d$ regular graph. Of course, in reality the orientations of the arcs at different vertices are not independent since choosing an orientation for the arcs incident with $v$ implies a particular orientation on the arc $v$ shares with each of its neighbours. However, Theorem 3.52 confirms the intuition that this dependence cannot be very strong, at least for large graphs.

### 3.5 The number of Euler tours of random Eulerian graphs

As was the case in the previous section, we restrict our attention to the class of graphs with fixed even degrees, and generate our random graphs using the configuration model. In this section, as was the case in $\S 3.4$, we restrict ourselves to the regular case. That is, we are sampling graphs from $\mathbb{G}(n, 2 d)$. Recall the BEST Theorem
(Theorem 1.57). This gives the following relationship between the sizes of the sets of Euler tours and arborescences of an Eulerian directed graph $\vec{G}$ :

$$
|\operatorname{ET}(\vec{G})|=\frac{1}{n}(d-1)!^{n}|\operatorname{ARB}(\vec{G})|
$$

Brightwell and Winkler used this relationship to count Euler tours of an undirected graph by counting the orbs of $G$. We recall (Definition 1.69) that the set of orbs of a graph is the set of pairs

$$
\operatorname{ORB}(G)=\{(\mathcal{E}, \mathcal{A}): \mathcal{E} \in \operatorname{EO}(G), \mathscr{A} \in \operatorname{ARB}(\mathcal{E})\} .
$$

In previous sections, we were able to obtain asymptotic estimates for the second moments of random variables counting arborescences (and so also Euler tours) of random $d$-in/d-out graphs and Eulerian orientations of random $2 d$-regular graphs. Unfortunately, finding an estimate for the second moment of the number of orbs (and so also the number of tours) of undirected graphs has proved elusive. However, we are able to make a conjecture, and in $\S 3.5 .1$ we provide some evidence, both empirical and theoretical, as to why we think it to be true. We then explain, in §3.5.2, how, if this conjecture holds, we will be able to obtain an asymptotic distribution for the number of orbs of a random $2 d$-regular graph. Then, by almost identical arguments to those used to prove Theorem 3.35, we would be able to show the probabilistic inequality of Conjecture 3.4 holds.

### 3.5.1 Estimating the moments of $O_{n, 2 d}$

Recall the definition of an orb: the set of orbs of a graph $G$ is the set of pairs $(\mathcal{E}, \mathcal{A})$ where $\mathcal{E}$ is an Eulerian orientation of $G$ and $\mathcal{A}$ is an arborescence of $\mathcal{E}$. It is more straightforward to analyse the moments of the number of orbs of a random regular graph than the number of Euler tours. Given an estimate of any moment of the number of orbs of a random $2 d$-regular graph we can immediately infer an estimate of the corresponding moment for the number of Euler tours.

Theorem 3.53. Let $d$ be some fixed positive integer and let $n \in \mathbb{N}$. Let $O_{n, 2 d}^{\star}$ be the random variable counting the orbs of a $2 d$-regular multigraph obtained as the projection of a random $\mathrm{F} \in \Omega_{n, 2 d}$. Then,

$$
\mathbb{E}\left[O_{n, 2 d}^{\star}\right] \rightarrow \sqrt{\frac{\pi n}{d}}\left(\frac{\binom{2 d}{d} d}{2^{d}}\right)^{n}
$$

Proof. As in the proof of Theorem 3.47, we define an Eulerian orientation of a configuration $\mathrm{F} \in \Omega_{n, 2 d}$ to be a partition $W=S \cup T$, where each $W_{v}$ is split into equally sized sets $S_{v}$ and $T_{v}$, such that each edge in F has one endpoint in $S$ and one endpoint in $T$. Hence, each Eulerian orientation of a configuration F corresponds to a directed configuration $\overrightarrow{\mathrm{F}}$ from an instance of the directed configuration model $\Phi_{n, d}$. Recall, from §3.3.1, that we defined an arborescence of a directed configuration $\vec{F}$ to be a partial directed configuration $\overrightarrow{\mathrm{F}}^{\prime} \subset \overrightarrow{\mathrm{F}}$ such that $\sigma\left(\overrightarrow{\mathrm{F}}^{\prime}\right)$ is an arborescence of $\sigma(\overrightarrow{\mathrm{F}})$. We now define an orb of a configuration to be a pair $(\mathcal{E}, \mathcal{A})$, where $\mathcal{E} \in \mathrm{EO}(\mathrm{F})$ and $\mathcal{A} \in \operatorname{ARB}(\mathcal{E})$. The expected number of orbs of a random configuration can be computed as

$$
\mathbb{E}\left[O_{n, 2 d}^{\star}\right]=\frac{\left|\bar{\Omega}_{n, 2 d}\right|}{\left|\Omega_{n, 2 d}\right|},
$$

where

$$
\bar{\Omega}_{n, 2 d}=\left\{(\mathrm{F}, \mathcal{E}, \mathcal{A}): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E} \in \mathrm{EO}(\mathrm{~F}), \mathcal{A} \in \operatorname{ARB}(\mathcal{E})\right\} .
$$

To enumerate the elements of $\bar{\Omega}_{n, 2 d}$, we first partition $W$ into $S$ and $T$. Then, we enumerate the triples $(\mathrm{F}, \mathcal{E}, \mathcal{A})$ corresponding to that partitioning. Each partitioning of $W$ into $S$ and $T$ gives rise to an instance of the directed configuration model $\Phi_{n, d}$, and, therefore, we can compute the number of triples $(\mathrm{F}, \mathcal{E}, \mathcal{A})$ with $\mathcal{E}$ consistent with $S$ and $T$ using Lemma 3.27. Note that we never need to consider the generation of F , as this is implicit in $\mathcal{E}$. From Lemma 3.27, we see that the number of partial directed configurations that project to arborescences in an instance of $\Phi_{n, d}$ is

$$
n d^{n}(d n-1)_{n-2} .
$$

Given an arborescence $\mathcal{A}$, there are $d n-n+1$ arcs that need to be added to complete $\mathcal{E}$, and $(d n-n+1)$ ! ways to choose configuration edges for these arcs. Finally, we note that there are $\binom{2 d}{d}^{n}$ ways to partition $W$ into $S$ and $T$. Hence, the number of elements in $\bar{\Omega}_{n, 2 d}$ is

$$
\left|\bar{\Omega}_{n, 2 d}\right|=\binom{2 d}{d}^{n} n d^{n}(d n-1)!.
$$

Dividing by

$$
\left|\Omega_{n, 2 d}\right|=\frac{(2 d n)!}{(d n)!2^{d n}}
$$

and applying Stirling's formula (3.70) we find

$$
\mathbb{E}\left[O_{n, 2 d}^{\star}\right] \rightarrow\left(\binom{2 d}{d} \frac{d}{2^{d}}\right)^{n} \sqrt{\frac{\pi n}{d}}
$$

As in the case of Eulerian orientations, we can infer an estimation for $\mathbb{E}\left[O_{n, 2 d}\right]$ from the estimation of $\mathbb{E}\left[O_{n, 2 d}^{\star}\right]$ by conditioning on there being no loops or double edges.

Theorem 3.54. Let $d$ be some fixed positive integer and let $n \in \mathbb{N}$. Let $O_{n, 2 d}$ denote the random variable counting the orbs of a random $G \in \mathbb{G}(n, 2 d)$. Then,

$$
\mathbb{E}\left[O_{n, 2 d}\right] \rightarrow e^{3 / 4} \sqrt{\frac{\pi n}{d}}\left(\frac{\binom{2 d}{d} d}{2^{d}}\right)^{n} .
$$

Proof. The theorem follows from the result of Bollobás [11] (see Theorem 3.15) and from calculations already carried out in the proofs of Theorem 3.30 and Lemma 3.32.

Let $L$ and $D$ count the number of loops and double edges in a random configuration $\mathrm{F} \in \Omega_{n, 2 d}$, and let $L^{(1)}$ and $D^{(1)}$ count the number of loops and double edges in F when ( $\mathrm{F}, \mathcal{E}, \mathcal{A}$ ) is chosen randomly from

$$
\bar{\Omega}_{n, 2 d}=\left\{(\mathrm{F}, \mathcal{E}, \mathcal{A}): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E} \in \mathrm{EO}(\mathrm{~F}), \mathcal{A} \in \operatorname{ARB}(\mathcal{E})\right\} .
$$

We can estimate $\mathbb{E}\left[O_{n, 2 d}\right]$ as

$$
\mathbb{E}\left[O_{n, 2 d}\right]=\frac{\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right]}{\mathbb{P}[L=D=0]} \mathbb{E}\left[O_{n, 2 d}^{\star}\right]
$$

From the classic result of Bollobás [11] (Theorem 3.15), we have

$$
\begin{equation*}
\mathbb{P}[L=D=0] \rightarrow \exp \left(-\frac{2 d-1}{2}-\frac{(2 d-1)^{2}}{4}\right)=\exp \left(-d^{2}+1 / 4\right) . \tag{3.86}
\end{equation*}
$$

Thus, it remains to estimate $\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right]$.
Recall how we enumerated the elements of $\bar{\Omega}_{n, 2 d}$ in Theorem 3.47 by first partitioning $W$ into $S$ and $T$. Each partition of $W$ into $S$ and $T$ gives rise to an instance of the directed configuration model $\Phi_{n, d}$, and an equal number of elements of $\bar{\Omega}_{n, 2 d}$ arise from each choice of $S$ and $T$. Hence, the expected number of loops in F, when $(\mathrm{F}, \mathcal{E}, \mathcal{A})$ is chosen uniformly at random from $\bar{\Omega}_{n, 2 d}$, is equal to the expected number of loops in $\overrightarrow{\mathrm{F}}$, when $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ is chosen randomly from

$$
\bar{\Phi}_{n, d}=\left\{(\overrightarrow{\mathrm{F}}, \mathcal{A}): \overrightarrow{\mathrm{F}} \in \Phi_{n, d}, \mathcal{A} \in \operatorname{ARB}(\mathcal{E})\right\},
$$

as defined in the proof of Theorem 3.30 and Lemma 3.32. Hence, by arguments already presented in the proof of Theorem 3.30

$$
\mathbb{E}\left[L^{(1)}\right] \rightarrow \frac{m_{2}-m}{m}=d-1,
$$

since now $m_{2}=d^{2} n$ and $m=d n$. Similarly, the expected number of double edges can be obtained as the expected number of double arcs and directed 2-cycles in $\vec{F}$ when $(\overrightarrow{\mathrm{F}}, \mathcal{A})$ is chosen uniformly at random from $\bar{\Phi}_{n, d}$. Hence, by arguments presented in Theorem 3.30 and Lemma 3.32, we have

$$
\begin{aligned}
\mathbb{E}\left[D^{(1)}\right] & \rightarrow \frac{\left(m_{2}-m\right)^{2}}{2 m^{2}}+\frac{\left(m_{2} / m\right)^{2}-1}{2} \\
& =d(d-1) .
\end{aligned}
$$

It is straightforward to check that the factorial moments of $L^{(1)}$ and $D^{(1)}$ converge jointly. Hence, $L^{(1)}$ and $D^{(1)}$ converge to a pair of independent Poisson random variables with means $d-1$ and $d(d-1)$, respectively. We can now estimate

$$
\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right] \rightarrow \exp \left(-d^{2}+1\right) .
$$

Combining our estimate of $\mathbb{P}[L=D=0]$ and $\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right]$ gives

$$
\frac{\mathbb{E}\left[O_{n, 2 d}\right]}{\mathbb{E}\left[O_{n, 2 d}^{\star}\right]}=\frac{\mathbb{P}\left[L^{(1)}=D^{(1)}=0\right]}{\mathbb{P}[L=D=0]} \rightarrow \exp (3 / 4) .
$$

Let $\mathcal{T}_{n, 2 d}$ be the random variable counting Euler tours of a random $2 d$-regular graph. Applying the BEST theorem to Theorem 3.53 and Theorem 3.54, i.e., multiplying by $n^{-1}(d-1)!^{n}$, we get

$$
\begin{aligned}
& \mathbb{E}\left[\mathcal{T}_{n, 2 d}^{\star}\right] \rightarrow \sqrt{\frac{\pi}{d n}}\left(\frac{\binom{2 d}{d} d!}{2^{d}}\right)^{n} ; \\
& \mathbb{E}\left[\mathcal{T}_{n, 2 d}\right] \rightarrow e^{3 / 4} \sqrt{\frac{\pi}{d n}}\left(\frac{\binom{2 d}{d} d!}{2^{d}}\right)^{n}
\end{aligned}
$$

We have not been able to obtain an asymptotic estimate for the second moment, but make the following conjecture.

Conjecture 3.55. Let $d$ be a fixed integer satisfying $d \geq 2$. For each $n \in \mathbb{N}$, we define $O_{n, 2 d}^{\star}$ to be the random variable counting the number of orbs of a graph $G$ obtained as the projection of a random $\mathrm{F} \in \Omega_{n, 2 d}$ and define $O_{n, 2 d}$ to be the random variable counting the number of orbs of a random $2 d$-regular graph. Then,

$$
\begin{aligned}
& \frac{\mathbb{E}\left[\left(O_{n, 2 d}^{\star}\right)^{2}\right]}{\mathbb{E}\left[O_{n, 2 d}^{\star}\right]^{2}} \rightarrow \sqrt{\frac{2 d-1}{2 d-2}} ; \\
& \frac{\mathbb{E}\left[O_{n, 2 d}^{2}\right]}{\mathbb{E}\left[O_{n, 2 d}\right]^{2}} \rightarrow \exp \left(-\frac{1}{2(2 d-1)}-\frac{1}{4(2 d-1)^{2}}\right) .
\end{aligned}
$$

We now discuss the motivation behind Conjecture 3.55, and present some empirical evidence that seems to support it. The principal reason we believe (or, rather, hope) Conjecture 3.55 to be true comes from the next section: if Conjecture 3.55 holds then we can apply Janson's Theorem (Theorem 3.19) with $Y_{n}=\mathcal{T}_{n, 2 d}$ and $X_{i, n}$ counting $i$-cycles to obtain an asymptotic distribution for $\mathcal{T}_{n, 2 d}$ (and, hence, confirm Conjecture 3.4).

The key to establishing Conjecture 3.55 is enumerating the elements of the set
$\widetilde{\Omega}_{n, 2 d}=\left\{\left(\mathrm{F}, \mathcal{E}, \mathcal{E}^{\prime}, \mathcal{A}, \mathcal{A}^{\prime}\right): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E}, \mathcal{E}^{\prime} \in \mathrm{EO}(\mathrm{F}), \mathcal{A} \in \operatorname{ARB}(\mathcal{E}), \mathcal{A}^{\prime} \in \operatorname{ARB}\left(\mathcal{E}^{\prime}\right)\right\}$.
This should be achievable by combining the proofs of the two previous sections. First, we partition $W$ into $S \cup T$ and $S^{\prime} \cup T^{\prime}$, as per the proof of Theorem 3.47. Then, we want to enumerate the set of tuples $\left(\mathrm{F}, \mathcal{E}, \mathcal{E}^{\prime}, \mathcal{A}, \mathcal{A}^{\prime}\right)$ with $S, T, S^{\prime}$, and $T^{\prime}$ defining $\mathcal{E}$ and $\mathcal{E}^{\prime}$. For example, by Theorem 3.28 , we can compute that the number of elements of $\widetilde{\Omega}_{n, 2 d}$ in which $\mathcal{E} \cap \mathcal{E}^{\prime}$ or $\mathfrak{E} \oplus \mathcal{E}^{\prime}$ is empty is

$$
\frac{d^{2 n-1}}{d-1}(2 n)!.
$$

As another example, consider the number of elements of $\widetilde{\Omega}_{n, 2 d}$ satisfying $\mathcal{A} \cup \mathcal{A}^{\prime} \subseteq$ $\mathcal{E} \cap \mathcal{E}^{\prime}$. That is, the case when both the arborescences are contained in the portion of edges the Eulerian orientations agree on. Suppose we have $\left|S_{v} \cap S_{v}^{\prime}\right|=d_{v}$, for some vector of non-negative integers $\mathbf{d}=\left(d_{1}, d_{2}, \ldots, d_{n}\right)$ satisfying

$$
d \geq d_{n} \geq d_{n-1} \geq \cdots \geq 1
$$

Since $\mathscr{A} \cup \mathcal{A}^{\prime} \subseteq \mathcal{E} \cap \mathcal{E}^{\prime}$, the number of ways we can choose $\mathcal{A}, \mathscr{A}^{\prime}$ and $\mathcal{E} \cap \mathcal{E}^{\prime}$ is, by Theorem 3.28,

$$
\frac{n^{2}}{m(m-n)}\left(\prod_{v} d_{v}\right)^{2} m!
$$

where $m=\sum_{v} d_{v}$. Then, since there are no edges from $\mathcal{A}$ or $\mathcal{A}^{\prime}$ in $\mathcal{E} \oplus \mathcal{E}^{\prime}$, the number of ways we can choose the remaining edges of $\mathcal{E}$ (and so also F and $\mathcal{E}^{\prime}$ ) is $(d n-m)$ !. Letting $c_{i}=\left|\left\{v: d_{v}=i\right\}\right|$, we can see that the number of elements of $\widetilde{\Omega}_{n, 2 d}$ satisfying $\mathcal{A} \cup \mathcal{A}^{\prime} \subseteq \mathcal{E} \cap \mathcal{E}^{\prime}$ is

$$
\begin{gathered}
\binom{2 d}{d}^{2} \sum_{c_{0}+c_{1}+\cdots+c_{d}=n}\binom{n}{c_{0}, c_{1}, \ldots, c_{d}} \prod_{i=1}^{d}\left(i^{2 c_{i}}\binom{d}{i}^{2 c_{i}}\right) \\
\times \frac{n^{2}\left(\sum_{i} i c_{i}\right)!\left(d n-\sum_{i} i c_{i}\right)!}{\left(\sum_{i} i_{i}\right)!\left(n-\sum_{i} c_{i}\right)!}
\end{gathered}
$$



Table 3.1: Critical ratio for $d=2$

It is clear that computing this value alone, which could be seen as one of the "easy cases" in the computation of $\left|\widetilde{\Omega}_{n, 2 d}\right|$ takes as much work as proving Theorem 3.28 and Theorem 3.47.

We make the conjecture that, in general, the number of triples, for a particular $\mathbf{c}$ is something like

$$
\begin{equation*}
d^{2 n}\left(\sum_{i=1}^{d} i c_{i}\right)!\left(d n-\sum_{i=1}^{d} i c_{i}\right)!f(\mathbf{c}), \tag{3.87}
\end{equation*}
$$

for some rational function $f(\mathbf{c})$. If this is the case, then Conjecture 3.55 could be proven (or disproven) by the argument used to prove Theorem 3.47 with $f(\mathbf{c})$ incorporated into $h(\mathbf{c})$. However, we have not been able to derive a nice expression in the form of (3.87). Once we allow edges from $\mathcal{A}$ and $\mathscr{A}^{\prime}$ to occur in $\mathcal{E} \cap \mathcal{E}^{\prime}$ or $\mathcal{E} \oplus \mathcal{E}^{\prime}$, the situation grows more complicated as we now have two types of edges (those in $\mathcal{E} \cap \mathcal{E}^{\prime}$ and those in $\mathcal{E} \oplus \mathcal{E}^{\prime}$ ) and the argument used to prove Theorem 3.28 does not work. Despite this, it seems an expression of the form given in (3.87) should exist.

Although we have not been able to prove Conjecture 3.55 we have obtained some empirical evidence to support it, see Table 3.1. This data was generated as follows: for each $n$ we generated $n^{2}$ different 4-regular graphs using the random generation algorithm of Steger and Wormald [83]; then, for each $G$ we computed an estimate for
$|\operatorname{ET}(G)|, Z_{G}$, using Algorithm 5 with $N=n^{2}$. We then computed two values

$$
\begin{aligned}
Z^{(1)} & =\frac{1}{n^{2}} \sum_{G} Z_{G} ; \\
Z^{(2)} & =\frac{1}{n^{2}} \sum_{G} Z_{G}^{2} .
\end{aligned}
$$

The ratio $Z_{2} /\left(Z_{1}\right)^{2}$ was taken as an approximation of the critical ratio of $O_{n, 4}$,

$$
\frac{\mathbb{E}\left[O_{n, 4}^{2}\right]}{\mathbb{E}\left[O_{n, 4}\right]^{2}}
$$

In Table 3.1 we have graphed it against the number of vertices $n$, with the conjectured critical ratio marked by a green line at

$$
\exp \left(-\frac{1}{6}-\frac{1}{36}\right) \sqrt{\frac{3}{2}} \approx 1.0083225
$$

### 3.5.2 The asymptotic distribution of $O_{n, 2 d}$

The main result of this section is contingent on a positive answer to Conjecture 3.55. Assuming Conjecture 3.55, we proceed to show how we can obtain an asymptotic distribution for the number of orbs (or Euler tours) of a random $2 d$-regular graph. Let $X_{i, n}$ denote the number of $i$-cycles in a random $\mathrm{F} \in \Omega_{n, d}$ and let $\lambda_{i}=\frac{(2 d-1)^{i}}{2 i}$. As stated previously, it is well known that the variables $X_{i, n}$ are asymptotically independent Poisson random variables with means $\lambda_{i}$. Hence, Condition 1 of Theorem 3.19 is satisfied by $X_{i, n}$. The next lemma establishes Condition 2 of Theorem 3.19 for $Y_{n}=$ $O_{n, 2 d}$.

Lemma 3.56. Let $X_{i, n}$ count the number of directed $i$-cycles in a graph $G \in \mathbb{G}^{\star}(n, 2 d)$. Then, for any fixed set of integers $j_{1}, j_{2}, \ldots, j_{k}$ we have

$$
\frac{\mathbb{E}\left[O_{n, 2 d}^{\star} \prod_{i=1}^{k}\left(X_{i, n}\right)_{j_{i}}\right]}{\mathbb{E}\left[O_{n, 2 d}^{\star}\right]} \rightarrow \prod_{i=1}^{k} \mu_{i}^{j_{i}},
$$

where $\mu_{i}=\frac{(2 d-1)^{i}-1}{2 i}$.
Proof. We start establishing

$$
\begin{equation*}
\frac{\mathbb{E}\left[O_{n, 2 d}^{\star} X_{i, n}\right]}{\mathbb{E}\left[O_{n, 2 d}^{\star}\right]} \rightarrow \mu_{i} \tag{3.88}
\end{equation*}
$$

for $i \geq 1$. Convergence of the factorial moments then follows for similar reasons as before (see Lemma 3.51).

Let $\mathcal{C}_{i, n}$ denote the number of $i$-cycles which can occur in a configuration $\Omega_{n, 2 d}$, as defined in Lemma 3.51. Then, for the same reasons as were given in the proof of Lemma 3.51 we have

$$
\frac{\mathbb{E}\left[O_{n, 2 d}^{\star} X_{i, n}\right]}{\mathbb{E}\left[O_{n, 2 d}^{\star}\right]}=\left|\mathcal{C}_{i, n}\right| \mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}, \mathcal{A}) \in \bar{\Omega}_{n, 2 d}\right],
$$

where we now define $\bar{\Omega}_{n, 2 d}$ to be the set

$$
\bar{\Omega}_{n, 2 d}=\left\{(\mathrm{F}, \mathcal{E}, \mathcal{A}): \mathrm{F} \in \Omega_{n, 2 d}, \mathcal{E} \in \mathrm{EO}(\mathrm{~F}), \mathcal{A} \in \operatorname{ARB}(\mathcal{E})\right\} .
$$

From the proof of Lemma 3.51, we already know

$$
\left|\mathcal{C}_{i, n}\right|=\frac{(n)_{i}(2 d)^{i}(2 d-1)^{i}}{2 i} \rightarrow \frac{(n)^{i}(2 d)^{i}(2 d-1)^{i}}{2 i}
$$

Thus, it remains to estimate the probability $\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}, \mathcal{A}) \in \bar{\Omega}_{n, 2 d}\right]$.
Let $C \in \mathcal{C}_{i, n}$. We want to count the number of elements of $\bar{\Omega}_{n, 2 d}$ with $C \subset \mathrm{~F}$. We can achieve this by following a similar approach to that used in the proof of Lemma 3.51. We start by choosing a partition of $W$ into $S$ and $T$ as per the proof of Lemma 3.51. First, we assign the endpoints of the edges of $C$ to $S$ and $T$. A vertex $v$ on $C$ is said to be balanced if one of the edges incident with $v$ on $C$ has a point from $S_{v}$ and one has a point from $T_{v}$; we say a vertex $v$ is negatively (resp. positively) imbalanced if both edges incident with $v$ on $C$ contain a point from $S_{v}$ (resp. $T_{v}$ ). Given a particular choice of $j$ positively and negatively imbalanced vertices, the number of ways to choose $S$ and $T$ is (see proof of Lemma 3.51)

$$
\begin{equation*}
\binom{2 d}{d}^{n}\left(\frac{d}{2(2 d-1)}\right)^{i}\left(\frac{d-1}{d}\right)^{2 j} \tag{3.89}
\end{equation*}
$$

Next, we choose which of the edges of $C$ are to lie in $\mathcal{A}$. For each balanced vertex we have two choices: the edge with a point from $S_{v}$ is either contained in $\mathcal{A}$ or it is not. For each negatively balanced vertex there are three choices: we could have exactly one of the two edges contained in $\mathcal{A}$, or we could have that neither of them are. Finally, the edge corresponding to the arc leaving any positively imbalanced vertex in $\mathcal{A}$ cannot be contained in $C$. The intersection of any arborescence $\mathcal{A}$ with $C$ is a set of directed paths. Suppose this intersection contains $k$ paths ending in negatively imbalanced vertices (these must necessarily be paths of length 1 ), and $l$ paths ending in balanced vertices. Then $\mathcal{A} \cap C$ contains $j+k+l$ components, since each positively imbalanced vertex must necessarily be the endpoint of a path in the intersection. Suppose $R$ is the set of endpoints of these paths and let $P_{v}$ denote the path ending in $v$, for each $v \in R$.

Next, we want to choose the remaining edges in $\mathcal{A}$. The argument is similar to that in Lemma 3.32: choosing the remainder of $\mathcal{A}$ is equivalent to choosing a arborescence on $(V \backslash C) \cup R$, where we have contracted the path ending in $v$ to a single vertex, for each $v \in R$. For $v \in R$, let $x_{v}$ denote the points in $\bigcup_{u \in P_{v}} T_{u}$ that are not used by edges in $C$ and let $y_{v}$ be equal to the number of points from $\bigcup_{u \in P_{v}} S_{u}$ not used in edges of $C$. Then, for each $v \in R$, we have

$$
x_{v}=(d-1)\left|P_{v}\right|,
$$

if the first and last vertices of $P_{v}$ are both balanced or the first vertex of $P_{v}$ is negatively imbalanced and the last vertex is positively imbalanced,

$$
x_{v}=(d-1)\left|P_{v}\right|+1,
$$

if the first vertex of $P_{v}$ is negatively imbalanced and either $\left|P_{v}\right|=1$ or the last vertex is balanced, and

$$
x_{v}=(d-1)\left|P_{v}\right|-1,
$$

if the last vertex is positively imbalanced and either $\left|P_{v}\right|=1$ or the first vertex of $P_{v}$ is balanced. We have $y_{v}$ equal to either $d-2, d-1$, or $d$, depending on whether $v$ is negatively imbalanced, balanced, or positively imbalanced.

Choosing the remainder of $\mathcal{A}$ is equivalent to choosing a tree on $(V \backslash C) \cup R$ in which there are $d$ points available for incoming and outgoing arcs of each $v \notin C$. For each $v \in R$, there are $x_{v}$ points available for incoming arcs and $y_{v}$ points available for outgoing arcs.

By Lemma 3.27, the number of ways to choose the remaining edges for $\mathcal{A}$ and $\mathcal{E}$ such that $R$ gives the roots of the components of $\mathcal{A} \cap C$, is

$$
n(d-2)^{k}(d-1)^{l} d^{n-i+j}(d n-i-1)!,
$$

since we have $\sum_{v \in R} x_{v}+d(n-i)=d n-i$. Hence, the number of arborescences $\mathcal{A}$ with $R$ as the roots of the components in $\mathcal{A} \cap C$ is

$$
n 2^{j-k}(d-2)^{k}(d-1)^{l} d^{n-i+j-1}(d n-i)!,
$$

since we must choose which of the two arcs leaving each imbalanced vertex in $R$ is contained in $\mathcal{A}$.

Suppose $j=0$. That is, $C$ is a directed cycle in $\mathcal{E}$. We note that $C \cap \mathcal{A}$ cannot be equal to $C$. Hence, in this case, we must have $l>0$. Summing over all choices for $R$
gives

$$
\sum_{l=1}^{i}\binom{i}{l}(d-1)^{l} d^{n-i}(d n-i-1)!=\left(d^{n}-d^{n-i}\right)(d n-i-1)!
$$

Then, by (3.89), the number of elements of $\bar{\Omega}_{n, 2 d}$ with $C \subset \mathrm{~F}$ in which $C$ has no imbalanced vertices in $\mathcal{E}$ is

$$
\begin{equation*}
2 n\binom{2 d}{d}^{n}\left(\frac{d}{2(2 d-1)}\right)^{i}\left(d^{n}-d^{n-i}\right)(d n-i-1)! \tag{3.90}
\end{equation*}
$$

On the other hand, suppose that $j>0$. Each of the positively imbalanced vertices of $C$ will be the endpoint of a path in $C \cap \mathcal{A}$. Hence, in this case we can have $l=0$. First, we sum over $l$, to get

$$
\sum_{l=0}^{i-2 j}\binom{i-2 j}{l}(d-1)^{l}=d^{i-2 j}
$$

Then, we sum over $k$ to get

$$
2^{j} \sum_{k=0}^{j}\binom{j}{d}\left(\frac{d-2}{2}\right)^{k}=d^{j}
$$

Hence, the number of elements of $\bar{\Omega}_{n, 2 d}$ with $C \subset \mathrm{~F}$ in which $C$ has a particular set of $j$ imbalanced vertices in $\mathcal{E}$ is, for each $j>0$,

$$
\begin{equation*}
2 n\binom{2 d}{d}^{n}\left(\frac{d}{2(2 d-1)}\right)^{i}\left(\frac{d-1}{d}\right)^{2 j} d^{n}(d n-i-1)! \tag{3.91}
\end{equation*}
$$

Combining (3.90) and (3.91), and summing over all choices for the sets of imbalanced vertices we get that the number of $(\mathrm{F}, \mathcal{E}, \mathcal{A}) \in \bar{\Omega}_{n, 2 d}$ with $C \subset \mathrm{~F}$ is

$$
2 n\binom{2 d}{d}^{n}\left(\frac{d}{2(2 d-1)}\right)^{i} d^{n}(d n-i-1)!\left(\sum_{j=0}^{\lfloor i / 2\rfloor}\binom{i}{2 j}\left(\frac{d-1}{d}\right)^{2 j}-\frac{1}{d^{i}}\right)
$$

We have shown, in Lemma 3.51, that

$$
\sum_{j=0}^{\lfloor i / 2\rfloor}\binom{i}{2 j}\left(\frac{d-1}{d}\right)^{2 j}=\frac{1}{2}\left(\frac{2 d-1}{d}\right)^{i}+\frac{1}{2 d^{i}} .
$$

Hence, the number of $(\mathrm{F}, \mathcal{E}, \mathcal{A}) \in \bar{\Omega}_{n, 2 d}$ with $C \subset \mathrm{~F}$ is

$$
\binom{2 d}{d}^{n} \frac{d^{n} n(d n-i-1)!}{2^{i}}\left(1-\frac{1}{(2 d-1)^{i}}\right) .
$$

Dividing by the number of elements in $\bar{\Omega}_{n, 2 d}$,

$$
\left|\bar{\Omega}_{n, 2 d}\right|=\binom{2 d}{d}^{n} n d^{n}(d n-1)!
$$

gives

$$
\mathbb{P}\left[C \subseteq \mathrm{~F}:(\mathrm{F}, \mathcal{E}, \mathcal{A}) \in \bar{\Omega}_{n, 2 d}\right] \rightarrow \frac{1}{2^{i} d^{i} n^{i}}\left(1-\frac{1}{(2 d-1)^{i}}\right) .
$$

Finally, multiplying by $\left|\mathcal{C}_{i, n}\right|$ gives (3.88).
Now, if Conjecture 3.55 was true, we could apply Janson's theorem to obtain an asymptotic distribution for $\mathcal{T}_{n}$. For $i \geq 3$ we have

$$
\delta_{i}=-\frac{1}{(2 d-1)^{i}} .
$$

Hence,

$$
\sum_{i \geq 3} \lambda_{i} \delta_{i}^{2}=\frac{1}{2} \sum_{i \geq 3} \frac{1}{i(2 d-1)^{i}}=\frac{1}{2} \log \left(\frac{2 d-1}{2 d-2}\right)-\frac{1}{2(2 d-1)}-\frac{1}{4(2 d-1)^{2}} .
$$

Thus, $O_{n, 2 d}$ and $X_{i, n}$ would satisfy the conditions of Janson's Theorem if Conjecture 3.55 was true, in which case we would have

$$
\frac{O_{n, 2 d}}{\mathbb{E}\left[O_{n, 2 d}\right]} \rightarrow \prod_{i \geq 3}\left(1+\frac{1}{(2 d-1)^{i}}\right)^{Z_{i}}
$$

where the $Z_{i}$ are Poisson random variables with mean $\lambda_{i}=\frac{(2 d-1)^{i}}{2 i}$. This could then be used to infer a concentration result of the form

$$
\mathbb{P}\left[\frac{O_{n, 2 d}}{\mathbb{E}\left[O_{n, 2 d}\right]} \geq n^{-1 / 2}\right] \rightarrow 1
$$

as $n \rightarrow \infty$, of which Conjecture 3.4 would be an immediate corollary.

### 3.6 Euler tours of the grid

We close this chapter by presenting some examples of simple 4-regular graphs for which Algorithm 4 and Algorithm 5 have exponential running time. The graphs we consider are the $m \times n$ square grids with toroidal boundaries.

Definition 3.57. The $m \times n$ toroidal grid, denoted $G(m, n)$, is the 4-regular graph with vertex set $\{(i, j): 0 \leq i<m, 0 \leq j<n\}$ and an edge joining each $(i, j)$ to $(i \pm 1 \bmod m, j \pm 1 \bmod n)$.

Remark 3.58. In the following, for ease of presentation, we assume all addition of row and column indices is done modulo $m$ and $n$, respectively. Hence, the set of edges is $G(n, m)$ is given by $\{\{(i, j),(i \pm 1, j \pm 1)\}: 0 \leq i \leq m-1,0 \leq j \leq n-1\}$.

Golin et al. [39] claim that we can exactly count Euler tours of $G(m, n)$ using the transfer matrix method. This is the same method used in [61] and [6] to obtain the exponential growth rate of the number of Eulerian orientations of the square and triangular lattices, respectively. This is certainly true - we can define a transfer matrix for counting many combinatorial structures defined on $G(m, n)$ - but Golin et al. do not provide any calculations or even a description of the transfer matrix used. As we shall see, just defining an appropriate transfer matrix is not exactly straightforward. Moreover, the (naive) algorithm based on the transfer matrix we define is sure to be very inefficient, even for relatively small values of $m$.

We now discuss the problem of counting the Euler tours of $G(m, n)$ and devise a transfer matrix for the problem. We illustrate how this approach can be used to derive exact expressions for $|\operatorname{ET}(G(m, n))|$ by computing the number of Euler tours for $G(2, n)$ and $G(3, n)$. In general, this leads to a polynomial-time algorithm for counting the Euler tours of $G(m, n)$, for fixed $m$, albeit one with an exponential dependence on $m$. Finally, we discuss why Algorithms 4 and Algorithm 5 will not behave like an fpaus or an fpras for $\operatorname{ET}(G(m, n))$, even when one of $m$ and $n$ is bounded.

Recall the definition of a transition system from §1.3.2. Each transition system $T$ of $G(m, n)$ defines a decomposition of $G(m, n)$ into a set of disjoint cycles, which we denote by $\mathcal{C}(T)$. If $\mathcal{C}(T)$ has a single element, then $T$ defines an Euler tour of $G(m, n)$. Let $V_{k}=\{(i, j): 0 \leq i \leq m-1,0 \leq j \leq k-1\}$. We say a partial transition system, defined only on $V_{k}$, is legal if it can be extended to a transition system defining an Euler tour of $G(m, n)$. Each partial transition system $T$ defined on $V_{k}$ decomposes the subgraph induced by $V_{k}$ into a set of disjoint paths and cycles. The paths in this decomposition will include the trailing horizontal edges joining column 0 and column $n-1$, and column $k-1$ and column $k$. A necessary condition for a partial transition system $T$ defined on $V_{k}$ to be legal is for there to be no cycles in the decomposition of $V_{k}$ induced by $T$. Hence, each legal partial transition system $T$ decomposes $V_{k}$ into a set of $m$ paths, where each one of the end-edges of these paths is contained in the set

$$
\{\{(i, 0),(i, n-1)\}: 0 \leq i \leq m-1\} \cup\{\{(i, k-1),(i, k)\}: 0 \leq i \leq m-1\} .
$$

We classify the legal transition systems on $V_{k}$ by the set of perfect matchings on $\left\{l_{0}, l_{1}, \ldots, l_{m-1}, r_{0}, r_{1}, \ldots, r_{m-1}\right\}$, denoted $\mathcal{P}(m)$, by identifying $l_{i}$ with the edge $\{(i, 0),(i, n-1)\}$ and $r_{i}$ with the edge $\{(i, k-1),(i, k)\}$. Then, to each legal $T$ we assign the class $\mathcal{C} \in \mathcal{P}(m)$, where the edges of $\mathcal{C}$ correspond to the endpoints of the paths in the decomposition of $V_{k}$ induced by $T$. Clearly, the set of classes is the same
for any $k$.
For $\mathcal{C}, \mathcal{C}^{\prime} \in \mathcal{P}(m)$, we define $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right)$ to be the the number of transition systems on $V_{k+1} \backslash V_{k}=\{(i, k): 0 \leq i \leq m-1\}$ which extend a transition system on $V_{k}$ with class $\mathcal{C}$ to a transition system on $V_{k+1}$ with class $\mathcal{C}^{\prime}$. Then $A$ is a $P(2 m) \times P(2 m)$ matrix, where $P(2 m)$ is the number of perfect matchings on $K_{2 m}$ :

$$
P(2 m)=\binom{2 m}{m} \frac{m!}{2^{m}} .
$$

Consider the effect of choosing a transition system on

$$
V_{k+1} \backslash V_{k}=\{(i, j): 0 \leq i \leq m-1\} .
$$

The only illegal transition system on $V_{k+1} \backslash V_{k}$ is the one in which $\{(i, k),(i+1 \bmod m)\}$ is paired with $\{(i, k),(i-1 \bmod m, k)\}$, for each $i=0,1, \ldots, m-1$. Each other transition system defines a set of edge-disjoint paths, where each path has its end-edges in the set

$$
\{\{(i, k-1),(i, k)\}: 0 \leq i \leq m-1\} \cup\{\{(i, k),(i, k+1)\}: 0 \leq i \leq m-1\}
$$

Hence, we can identify the legal transition systems on $V_{k+1} \backslash V_{k}$ with the set of perfect matchings on $\left\{l_{0}^{\prime}, l_{1}^{\prime}, \ldots, l_{m-1}^{\prime}, r_{0}^{\prime}, r_{1}^{\prime}, \ldots, r_{m-1}^{\prime}\right\}$, where $l_{i}^{\prime}$ corresponds to the edge $\{(i, k-1),(i, k)\}$ and $r_{i}^{\prime}$ corresponds to the edge $\{(i, k),(i, k+1)\}$. Suppose the perfect matching contains $l_{i}^{\prime} r_{j}^{\prime}$ for $i>j$. Then exactly one of the following must hold (see Figure 3.1 for an example):

1. The horizontal edge $\{(i, k-1),(i, k)\}$ is paired with the upwards vertical edge $\{(i, k),(i+1, k)\}$ at $(i, k)$, the downwards vertical edge $\left.\left\{\left(i^{\prime}-1, k\right),\left(i^{\prime}, k\right)\right)\right\}$ is paired with the horizontal edge $\left\{\left(i^{\prime}, k\right),\left(i^{\prime}, k+1\right\}\right.$ at $\left(i^{\prime}, k\right)$, and the horizontal edge $\{(t, k-1),(t, k)\}$ is paired with the horizontal edge $\{(t, k),(t, k+1)$ at $(t, k)$ for each $t \in[0 . . j-1] \cup[i+1 . . m-1]$;
2. The horizontal edge $\{(i, k-1),(i, k)\}$ is paired with the downwards vertical edge $\{(i, k),(i-1, k)\}$ at $(i, k)$, the downwards vertical edge $\left.\left\{\left(i^{\prime}+1, k\right),\left(i^{\prime}, k\right)\right)\right\}$ is paired with the horizontal edge $\left\{\left(i^{\prime}, k\right),\left(i^{\prime}, k+1\right\}\right.$ at $\left(i^{\prime}, k\right)$, and the horizontal edge $\{(t, k-1),(t, k)\}$ is paired with the horizontal edge $\{(t, k),(t, k+1)$ at $(t, k)$ for each $t \in[j+1 . . i-1]$;

A similar situation holds for $i<j$ or edges $l_{i}^{\prime} l_{j}^{\prime}$ or $r_{i}^{\prime} r_{j}^{\prime}$. Hence, we can enumerate the set of matchings which can correspond to transition systems on a particular column


Figure 3.1: Two possibilities for transition systems extending $V_{k}$ to $V_{k+1}$ in $G(n, 6)$ with edge $l_{4}^{\prime} r_{1}^{\prime}$. Solid (resp. broken) edges are paired together.
of $G(n, m)$ as follows. Let $M$ represent an arbitrary matching on $\left\{l_{0}^{\prime}, l_{1}^{\prime}, \ldots, l_{m-1}^{\prime}\right\}$ and $\left\{r_{0}^{\prime}, r_{1}^{\prime}, \ldots, r_{m-1}^{\prime}\right\}$. Choose a decomposition of the cycle $(0,1, \ldots, m-1)$ into a number of edge disjoint paths. Then, for each $i$ which is not the endpoint of one of these paths we set $l_{i}^{\prime} r_{i}^{\prime} \in M$. Each other $i$ occurs as the endpoint of exactly two paths: label one of these $l_{i}^{\prime}$ and the other $r_{i}^{\prime}$. Collapsing each of these paths to a single edge-pairing gives the remaining edges in $M$.

Given $\mathcal{C} \in \mathcal{P}(m)$ for $V_{k}$ and $M$, a matching for $V_{k+1} \backslash V_{k}$, we obtain $\mathcal{C}^{\prime}$ by identifying $r_{i}$ with $l_{i}^{\prime}$ in $M$, and renaming $r_{i}^{\prime}$ as $r_{i}$. The result is either a set of 2-paths that contract to give the perfect matching $\mathcal{C}^{\prime}$, or some set of edges including at least one cycle. In this latter case, the transition system defining $M$ does not extend a legal transition system on $V_{k}$ of class $\mathcal{C}$ to a legal transition system on $V_{k+1}$. Finally, we note that each $M$ yields a different $\mathcal{C}^{\prime}$ when applied to $\mathcal{C}$, so $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right)$ is just the number of transition systems on $V_{k+1} \backslash V_{k}$ that yield the $M$ which extends $\mathcal{C}$ to $\mathcal{C}^{\prime}$, if it exists, and 0 otherwise.

Suppose $l_{i}^{\prime} r_{i}^{\prime} \in M$ for all $i$. There are $2 m$ distinct transition systems which yield $M$ : $m$ ways to decompose $(0,1, \ldots, m-1)$ into a single path and then 2 ways to label the endpoints. Alternatively, for each $i$, there is a pair of transition systems where $l_{i}^{\prime}$ is connected to $r_{i}^{\prime}$ by a path that uses all the vertical edges, and such that for $j \neq i, l_{i}^{\prime}$ is connected to $r_{i}^{\prime}$ by a path consisting of two vertical edges. These are the only transition systems that gives rise to the matching $M$ with $l_{i}^{\prime} r_{i}^{\prime} \in M$ for all $i$. Hence, $A(\mathcal{C}, \mathcal{C})=2 m$.

Now, suppose there is exactly one pair $i \neq j$ with $l_{i} r_{j}, l_{j} r_{i} \in M$. This can arise as a
result of two transition systems: choose a decomposition of $(0,1, \ldots, m-1)$ into two paths by breaking the cycle at $i$ and $j$ and then label the endpoints so that $l_{i}$ and $r_{j}$ are opposite endpoints of one path. Similarly, there are two transition systems yielding each matching $M$ with $l_{i} l_{j}, r_{j} r_{i} \in M$ for exactly one pair $i \neq j$. Hence, in this case we have $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right)=2$.

In all other cases, there is a single transition system on $V_{k+1} \backslash V_{k}$ that gives rise to $M$. Hence, $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right) \in\{0,1,2,2 m\}$.

Let $x(\mathcal{C})=1$ if $l_{i} r_{i} \in \mathcal{C}$ for all $i$ and $x(\mathcal{C})=0$ otherwise, and let $y(\mathcal{C})=1$ if identifying $l_{i}$ with $r_{i}$ in $\mathcal{C}$ gives rise to a single $m$-cycle and $y(\mathcal{C})=0$ otherwise. Then,

$$
|\mathrm{ET}(G(m, n))|=x A^{n} y^{T}
$$

For any fixed constant $m, A$ is a constant-sized matrix so, in theory, we can compute $|\mathrm{ET}(G(m, n))|$ in polynomial time. However, the number of rows in $A$,

$$
\frac{(2 m)!}{m!2^{m}}
$$

grows very quickly, e.g., for $m=10, A$ has 654729075 rows and the same number of columns, making computation impractical. It may be possible to exploit structure in $A$ to simplify the computation, as Lieb [61] and Baxter [6] did for the transfer matrix counting Eulerian orientations.

For example, if $l_{i} l_{j} \in \mathcal{C}$, then $A\left(\mathcal{C}, \mathcal{C}^{\prime}\right) \neq 0$ only if $l_{i} l_{j} \in \mathcal{C}^{\prime}$. Hence, we could exploit the use of an ordering on $\mathcal{P}(m)$ such that $\mathcal{C} \prec \mathcal{C}^{\prime}$ if there are fewer edges of the form $l_{i} l_{j}$ in $\mathcal{C}$ than are in $\mathcal{C}^{\prime}$. If the rows and columns of $A$ satisfy such an ordering then $A$ has a structure similar to a block-diagonal matrix; see Proposition 3.59 and Proposition 3.59 below for examples. With such a matrix, we know there will be large contiguous blocks consisting entirely of 0s. Indeed, it was a structure similar to this that enabled Lieb and Baxter to compute asymptotic estimates of the number of Eulerian orientations of the grid and triangular lattice.

In the following we use $A[S]$ to denote the square sub-matrix of $S$ with rows and columns indexed by $S$. We use $I_{n}$ to denote the $n \times n$ matrix which has value 1 in each entry on the diagonal, and is 0 everywhere else, and use $J_{n}$ to denote the $n \times n$ matrix with all entries equal to 1 .

Proposition 3.59. The number of Euler tours of the $2 \times n$ toroidal grid, $G(2, n)$, is, asymptotically,

$$
(2 n+3) 6^{n-1}-2^{n-1}
$$

Proof. There are 3 perfect matchings on $K_{4}$, so the transfer matrix for counting Euler tours of the $2 \times n$ toroidal grid has 3 entries. These are

$$
A=\left[\begin{array}{lll}
4 & 2 & 2 \\
2 & 4 & 2 \\
0 & 0 & 6
\end{array}\right]
$$

It is straightforward to calculate $A^{n}$. Firstly, we observe that $A^{n}[0,1]=(A[0,1])^{n}$ and

$$
\left[\begin{array}{ll}
4 & 2 \\
2 & 4
\end{array}\right]^{n}=\left[\begin{array}{ll}
2^{n-1}\left(3^{n}+1\right) & 2^{n-1}\left(3^{n}-1\right) \\
2^{n-1}\left(3^{n}-1\right) & 2^{n-1}\left(3^{n}+1\right)
\end{array}\right] .
$$

Moreover, it is clear that we also have $A_{3,3}^{n}=6^{n}$ since $A_{3,3}$ is the only non-zero value in the third row. Finally, we can calculate $A_{3,1}^{n}=A_{3,2}^{n}=2 n 6^{n-1}$. Hence, we have

$$
A^{n}=\left[\begin{array}{ccc}
2^{n-1}\left(3^{n}+1\right) & 2^{n-1}\left(3^{n}-1\right) & 2 n 6^{n-1} \\
2^{n-1}\left(3^{n}-1\right) & 2^{n-1}\left(3^{n}+1\right) & 2 n 6^{n-1} \\
0 & 0 & 6^{n}
\end{array}\right]
$$

Finally, $x=(1,0,0)$ and $y=(0,1,1)$, from which we can deduce the claimed value for $|\mathrm{ET}(G(2, n))|$.

The number of transition systems of $G(2, n)$ is $9^{n}$. Hence, the probability that Algorithm 4 will generate an Euler tour on a particular orientation is $\sim\left(\frac{2}{3}\right)^{n}$.

Proposition 3.60. The number of Euler tours of the $3 \times n$ toroidal grid $G(3, n)$ is

$$
\frac{4}{3} 20^{n}-14^{n}-\frac{4}{3} 5^{n}+2^{n} .
$$

Proof. There are 15 pairings on $\left\{l_{1}, l_{2}, l_{3}, r_{1}, r_{2}, r_{3}\right\}$, each of which can arise from the transition system on a single column of $G(n, 3)$. Table 3.2 enumerates these pairings, along with the number of transition systems on $V_{k+1} \backslash V_{k}$ that each arise from, and whether each of them extends to an Euler tour. With the ordering given in Table 3.2 we can write the transfer matrix $A$ as

$$
A=\left[\begin{array}{ccccc}
A_{1} & B_{1} & C_{1} & C_{1} & C_{1} \\
B_{1} & A_{1} & C_{1} & C_{1} & C_{1} \\
0 & 0 & D_{1} & 0 & 0 \\
0 & 0 & 0 & D_{1} & 0 \\
0 & 0 & 0 & 0 & D_{1}
\end{array}\right]
$$

| No. | $\sigma$ | \#TS | ET | No. | $\sigma$ | \#TS | ET |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | $\because$ | 6 | No | 2 | $X$ | 1 | Yes |
| 3 | $X$ | 1 | Yes | 4 | $X$ | 2 | No |
| 5 | $\%$ | 2 | No | 6 | $\underset{\sim}{X}$ | 2 | No |
| 7 | $\because$ | 2 | No | 8 | ! $!$ | 1 | Yes |
| 9 | $!!$ | 1 | Yes | 10 | $\because$ | 2 | No |
| 11 | $\therefore$ | 1 | Yes | 12 | $\because$ | 1 | Yes |
| 13 | $\square!$ | 2 | No | 14 | $\because$ | 1 | Yes |
| 15 | $!$ | 1 | Yes |  |  |  |  |

Table 3.2: Pairings for $m=3$
where $A_{1}=5 I_{3}+J_{3}, B_{1}=2 J_{3}, C_{1}=I_{3}+J_{3}$ and $D_{1}=2 I_{3}+6 J_{3}$. Moreover, this block structure is preserved over taking powers of $A$; that is, there exist $3 \times 3$ matrices $A_{n}, B_{n}, C_{n}$ and $D_{n}$ such that

$$
A^{n}=\left[\begin{array}{ccccc}
A_{n} & B_{n} & C_{n} & C_{n} & C_{n} \\
B_{n} & A_{n} & C_{n} & C_{n} & C_{n} \\
0 & 0 & D_{n} & 0 & 0 \\
0 & 0 & 0 & D_{n} & 0 \\
0 & 0 & 0 & 0 & D_{n}
\end{array}\right] .
$$

Observe that

$$
\left[\begin{array}{ll}
A_{n} & B_{n} \\
B_{n} & A_{n}
\end{array}\right]=\left[\begin{array}{ll}
A_{1} & B_{1} \\
B_{1} & A_{1}
\end{array}\right]^{n} \quad \text { and } \quad D_{n}=\left(D_{1}\right)^{n}
$$

so we can calculate the values of these entries independently.

## Claim 3.61.

$$
A_{n}=5^{n} I_{3}+\frac{14^{n}+2^{n}-2 \cdot 5^{n}}{6} J_{3} \quad \text { and } \quad B_{n}=\frac{14^{n}-2^{n}}{6} J_{3}
$$

Proof. We write

$$
\left[\begin{array}{ll}
A_{1} & B_{1} \\
B_{1} & A_{1}
\end{array}\right]=5 I_{6}+\left[\begin{array}{cc}
J_{3} & 2 J_{3} \\
2 J_{3} & J_{3}
\end{array}\right]
$$

and expand the powers of this sum by the binomial theorem ${ }^{2}$. Observing that

$$
\left[\begin{array}{cc}
J_{3} & 0 \\
0 & J_{3}
\end{array}\right]^{k-j}\left[\begin{array}{cc}
0 & 2 J_{3} \\
2 J_{3} & 0
\end{array}\right]^{j}=\left\{\begin{array}{cc}
2^{j}\left[\begin{array}{cc}
J_{3}^{k} & 0 \\
0 & J_{3}^{k}
\end{array}\right] \quad & \text { if } j \bmod 2=0 \\
2^{j}\left[\begin{array}{cc}
0 & J_{3}^{k} \\
J_{3}^{k} & 0
\end{array}\right] \quad \text { if } j \bmod 2=1
\end{array}\right.
$$

we find

$$
\left[\begin{array}{cc}
J_{3} & 2 J_{3} \\
2 J_{3} & J_{3}
\end{array}\right]^{k}=\left[\begin{array}{cc}
\frac{3^{k}+(-1)^{k}}{2} J_{3}^{k} & \frac{3^{k}-(-1)^{k}}{2} J_{3}^{k} \\
\frac{3^{k}-(-1)^{k}}{2} J_{3}^{k} & \frac{3^{k}+(-1)^{k}}{2} J_{3}^{k}
\end{array}\right],
$$

from which the claimed values for $A_{n}$ and $B_{n}$ follow.
By a similar calculation we can obtain

$$
D_{n}=2^{n} I_{3}+\frac{20^{n}-2^{n}}{3} J_{3}
$$

Now, observe that $C_{n}=A_{1} C_{n-1}+B_{1} C_{n-1}+C_{1} D_{n-1}$. Hence, if we let $C_{n}=x_{n} I_{3}+y_{n} J_{3}$ we can derive linear recurrences for $x_{n}$ and $y_{n}$, which can be solved for

$$
\begin{aligned}
& x_{n}=\frac{1}{3}\left(5^{n}-2^{n}\right) \\
& y_{n}=\frac{1}{9}\left(2 \cdot 20^{n}-2 \cdot 14^{n}-5^{n}+2^{n}\right)
\end{aligned}
$$

From Table 3.2 we obtain

$$
y=(0,1,1,0,0,0,0,1,1,0,1,1,0,1,1),
$$

from which we can calculate $|\operatorname{ET}(G(3, n))|$.
The number of transition systems of $G(3, n)$ is $27^{n}$. Hence, the probability of Algorithm 4 generating an Euler tour of any iteration is $\sim\left(\frac{20}{27}\right)^{n}$. For $m=4$ the transfer matrix has 105 rows and 105 columns, making direct computations of the $n$-th power impractical. In both [61] and [6] the authors obtain the exponential growth rate of $|E O(G(n, n))|$ by calculating the maximum eigenvalue of the transfer matrix $A$. This is made possible by the fact that the number of Eulerian orientations is equal to the trace of $A^{n}$. Moreover, the transfer matrix for Eulerian orientations is block diagonal, so it is sufficient to restrict analysis to a single block. In the case of the transfer matrix for Euler tours things become more difficult because (a) the number of

[^7]Euler tours is not obtainable in terms of something simple like the trace and (b) the matrix is not block diagonal so we need to compute all/more entries of $A^{n}$. Although this complexity prevented us from deriving exact values or asymptotic estimates for $|\operatorname{ET}(G(m, n))|$, we were able to obtain a crude upper bound on the number of Euler tours of $G(m, n)$, which is sufficient to show that Algorithm 4 and Algorithm 5 require exponential running time on $G(m, n)$, for any $m$ and $n$.

## Proposition 3.62.

$$
\mid \operatorname{ET}\left(G(n, m) \mid \leq 80^{n m / 4}\right.
$$

Proof. The set of cycles of the form $((2 i, 2 j),(2 i+1,2 j),(2 i+1,2 j+1),(2 i, 2 j+1))$ for $0 \leq i \leq m / 2$ and $0 \leq j \leq n / 2$ is a set of $n m / 4$ vertex-disjoint 4 -cycles. The number of transition systems containing any particular $k$-subset of these cycles is $3^{n m-4 k}$. Hence, by inclusion-exclusion, the number of transition systems containing none of these cycles is

$$
\sum_{k=0}^{n m / 4}\binom{n m / 4}{k}(-1)^{k} 3^{n m-4 k}=80^{n m / 4}
$$

and this is certainly greater than the number of Euler tours of $G(n, m)$.
There are 3 choices for the transition system at each vertex of $G(n, m)$, so $|\operatorname{TS}(G(n, m))|$ is equal to $3^{n m}$. Hence, we can bound the probability Algorithm 4 generates an Euler tour of $G(n, m)$ on any particular iteration as

$$
\begin{equation*}
\frac{|\mathrm{ET}(G(m, n))|}{|\operatorname{TS}(G(m, n))|} \leq\left(\frac{80}{81}\right)^{n m / 4}, \tag{3.92}
\end{equation*}
$$

which tends to 0 as either $m$ or $n$ goes to infinity. Hence, we can conclude that Algorithm 4 does not have the behaviour of an fpaus, and Algorithm 5 does not have the behaviour of an fpras, on $G(m, n)$.

In light of the conjectured asymptotic distribution of $\mathcal{T}_{n, 4}$ it is not wholly surprising that $G(m, n)$ has exponentially fewer Euler tours than transition systems. There are at least $n m$ 4-cycles in $G(m, n)$ : for $0 \leq i \leq m-1$ and $0 \leq j \leq n-1$ we have the cycle

$$
((i, j),(i+1, j),(i+1, j+1),(i, j+1)),
$$

where addition of the first and second coordinate is taken modulo $n$ and modulo $m$, respectively. Then, the factor of $W$ corresponding to $i=4$ is

$$
\left(\frac{80}{81}\right)^{n m}
$$

for the class $\mathcal{G}_{\mathbf{X}}$ containing $G(m, n)$, and thus we can expect that the value of $\mathcal{T}_{n, 4}$ for graphs in this class will almost surely be exponentially smaller than $\mathbb{E}\left[\mathcal{I}_{n, 4}\right]$.

## Chapter 4

## Conclusions

In this thesis, we addressed the complexity of generating and counting Eulerian orientations and Euler tours. Firstly, we analysed the complexity of exactly counting random Eulerian orientations of planar graphs. We then investigated the running time of a particular Markov chain Monte Carlo algorithm for generating random Eulerian orientations of planar graphs with a distribution that is close to uniform. Secondly, we studied the problem of generating and counting Euler tours of Eulerian graphs and digraphs. Although there are many positive results for the directed case of this problem, almost nothing is known for the undirected case. Indeed, the structure of this problem seems to make it particularly difficult to analyse Markov chain Monte Carlo algorithms. Hence, we shifted our focus to investigating the running time of a simple algorithm on random Eulerian graphs and digraphs.

In Chapter 2 we showed that the complexity of counting Eulerian orientations of planar graphs is \#P-complete. Although there already exists an algorithm which can approximately count the Eulerian orientations of any Eulerian graph, we considered an alternative approach to this problem for planar graphs. We showed that a different approach, using a well-known Markov chain (the face-reversal Markov chain $\mathcal{M}_{\mathcal{F}}$, a.k.a. Glauber Dynamics) mixes rapidly on the Eulerian orientations of triangular lattice, which are of practical importance as configurations of an ice-type model studied in statistical physics. These results complement existing results for the square lattice and are of a practical value as it is this algorithm that physicists tend to use in practice. However, we also showed that this chain is torpidly mixing on the set of Eulerian orientations of certain planar graphs containing a bounded face with a large number of edges (linear in the size of the graph). The only rapid mixing results for this chain are on the Eulerian orientations of the square and triangular lattices; however, one
would expect that it mixes rapidly for a larger class of graphs. We were not able to determine the class of planar graphs for which it should be rapidly mixing. However, we were able to develop some intuition. For example, suppose that, in every planar graph $G$, the cut with the worst (smallest) conductance is of the form used in the proof of Theorem 2.27. That is, there exists some face $\gamma \in \mathcal{F}(G)$, such that we can partition $\Omega=\mathrm{EO}(G)$ into two sets $S$ and $\bar{S}$ satisfying

1. $\forall \mathcal{E} \in S, \wp_{\mathcal{E}}(\gamma) \leq \wp_{\max }(\gamma) / 2$;
2. Every transition (of $\mathcal{M}_{\mathcal{F}}$ ) from $S$ to $\bar{S}$ reverses the edges of $\gamma$;
3. $\Phi(S)=\Phi\left(\mathcal{M}_{\mathcal{F}}\right)$.

Let $k=|\eta|$. There are $2^{k}$ orientations of the edges of $\eta$, and, therefore, the fraction of $\Omega$ in which $\gamma$ is clockwise oriented is $>2^{-k}$. Thus, if $k$ is a constant, we would expect the conductance of $S$ to be at least bounded below by some inverse polynomial. Thus, we make the following conjecture:

Conjecture 4.1. Let $k$ be some fixed constant and suppose $G$ is a planar graph that can be embedded in the plane such that none of the bounded faces have more than $k$ edges. Then the face-reversal chain is rapidly mixing.

In Chapter 3 we analysed the number of Eulerian orientations and Euler tours of random graphs. In particular, we were able to obtain asymptotic distributions for the number of Euler tours of a random $d$-in/ $d$-out directed graph and the number of Eulerian orientations of a random $2 d$-regular graph. Intuitively, one would expect that these results would combine to enable us to find an asymptotic characterisation of the distribution of the number of Euler tours of a random $2 d$-regular graph. However, this result proved elusive. We did, however, make a conjecture regarding the ratio of the second moment and the square of the first (for the random variable counting Euler tours of a random $2 d$-regular graph). In $\S 3.5$, we showed how, if this conjecture is true, we would be able to obtain an asymptotic distribution for the number of Euler tours of a random $2 d$-regular graph. Although we did not prove this conjecture, we did perform an empirical investigation which provided some evidence to support it; see Table 3.1. Not only do our empirical studies support the conjecture upon which the proof of the asymptotic distribution depends, but the corresponding ratio of the random variable counting Eulerian orientations of $2 d$-regular graphs (obtained in §3.4) matches the conjectured ratio for the random variable counting Euler orientation of random $2 d$ regular graphs.

In the case of Euler tours of $d$-in/d-out directed graphs we were able to show that almost every digraph $G \in \overrightarrow{\mathbb{G}}(n, d)$ has close to the maximum number of Euler tours possible. A consequence of this is that a pair of simple algorithms, Algorithm 4 and Algorithm 5, satisfy the conditions of an fpras and an fpaus respectively, for the Euler tours of almost every $d$-regular Eulerian digraph. It is already known that we can sample Euler tours of every Eulerian orientation in polynomial time so this result is not so interesting in itself. However, unlike other polynomial time algorithms for this problem, Algorithm 4 and Algorithm 5 generalise naturally to the undirected case. That is, if the conjectured asymptotic distribution for the number of Euler tours of undirected $2 d$-regular graphs is true, and we have provided (reasonably) convincing evidence of this, then Algorithm 4 and Algorithm 5 will be an fpaus and an fpras for the Euler tours of almost every $2 d$-regular graph.

The asymptotic distributions of Chapter 3 suggest a strong connection between the number of short cycles of different lengths and the number of Euler tours of graphs. This matches the results of $\S 3.6$, where we showed that the toroidal grid, which has many short cycles, has much less Euler tours than the expected value. This also relates to the work of Lieb [61] and Baxter [6] on the Eulerian orientations of the square and triangular lattices. Our analysis of the number of Eulerian orientations of random $2 d$ regular graphs, presented in $\S 3.4$ shows that the number of Eulerian orientations of each of these lattices is very far from typical. It would be interesting to try to prove a non-probabilistic result along the lines of the asymptotic distribution results, e.g., showing that graphs which do not have many short cycles, or possibly graphs with large girth, have a number of Euler tours (resp. Eulerian orientations) that is close the expected value we can obtain from Theorem 3.53 (resp. Theorem 3.49).

To conclude, we have, in the work presented this thesis, extended the body of knowledge regarding the problems of counting Euler tours and Eulerian orientations in several ways. We have analysed the complexity of exactly counting Eulerian orientations of planar graphs. We have made some further progress towards classifying the planar graphs on which the natural face-reversal Markov chain is rapidly mixing. Finally, in the most significant part of this thesis, we have made progress towards obtaining an algorithmic result on the complexity of sampling and approximately counting Euler tours of random graphs. The problem of sampling (and approximately counting) Euler tours in polynomial time is, as explained in §1.3.2, a hard open problem in the area of sampling algorithms. We hope that the work presented in Chapter 3 of this thesis will prove useful to future students of this problem.

## Appendix A

## The interlace Matrix approach to counting Euler tours

An alternative approach to the determinant method (Theorem 1.57 and Theorem 1.59) for counting Euler tours of directed graphs was provided by Macris and Pulé in [64]. The justification for the claimed result in [64] was by complicated analytic calculation, but Lauri attempted to provide a combinatorial proof in [59]. Unfortunately, both these attempted proofs are false. In this section we explain the approach and provide a counter-example.

Definition A.1. Let $G$ be a 2-regular Eulerian digraph on $V=\{1,2, \ldots, n\}$ and let $\mathcal{T}$ be some arbitrary Euler tour of $G$. We say $u$ and $v$ interlace on $\mathcal{T}$ if they alternate on $\mathcal{T}$; that is, $\mathcal{T}=(u \ldots v \ldots u \ldots v \ldots)$. We use $u \sim v$ to denote " $u$ and $v$ interlace on $\mathcal{T}$ " The interlace matrix of $\mathcal{T}$ is then defined as

$$
I(\mathcal{T})_{u, v}=\left\{\begin{aligned}
1 & \text { if } u \sim v \text { and } u<v \\
-1 & \text { if } u \sim v \text { and } u>v \\
0 & \text { otherwise }
\end{aligned}\right.
$$

Macris and Pulé make following claim.
Claim A.2. Let $G$ be a 2 -regular Eulerian digraph on $V=\{1,2, \ldots, n\}$, and let $I_{n}$ be the $n \times n$ identity matrix. Then, for any Euler tour $\mathcal{T}$ of $G$,

$$
|E T(G)|=\left|I_{n}+I(\mathcal{T})\right| .
$$

However, there exists a simple example for which the above claim does not hold! Suppose we have the Eulerian orientation $\mathcal{E}$ of $K_{5}$ given in Figure A.1. The Laplacian

Figure A.1: Counter-example to Claim A. 2
matrix of this Eulerian orientation is

$$
L=\left[\begin{array}{ccccc}
2 & -1 & 0 & 0 & -1 \\
0 & 2 & -1 & 0 & -1 \\
-1 & 0 & 2 & -1 & 0 \\
-1 & -1 & 0 & 2 & 0 \\
0 & 0 & -1 & -1 & 2
\end{array}\right]
$$

By the BEST theorem, we can easily compute the number of Euler tours of $\mathcal{E}$ to be 11. Now, consider the Euler tour $\mathcal{T} \in E T(\mathcal{E}):(1,5,4,2,3,4,1,2,5,3)$. The interlace matrix of $\mathcal{T}$ is

$$
I(\mathcal{T})=\left[\begin{array}{ccccc}
0 & 1 & 1 & 0 & 1 \\
-1 & 0 & 1 & 1 & 0 \\
-1 & -1 & 0 & 1 & 1 \\
0 & -1 & -1 & 0 & 0 \\
-1 & 0 & -1 & 0 & 0
\end{array}\right]
$$

Calculating the determinant $\left|I_{5}+I(\mathcal{T})\right|$ gives 15 ! Thus, even for this small example Claim A. 2 does not hold.

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[^0]:    ${ }^{1}$ In [32], $\alpha$-orientations were defined as orientations satisfying outdeg $(v)=\alpha(v)$. However, for our purposes, it is more convenient to work with this definition

[^1]:    ${ }^{2}$ Recall from the discussion in Remark 1.6 that not every Eulerian directed graph is an Eulerian orientation.

[^2]:    ${ }^{3}$ It is worth mentioning that the constant term hidden by the $O(\cdot)$ here is so large that the algorithm is only worthwhile when used for matrices that are so large modern computers would struggle to handle them. Hence, the best practical algorithm is still Strassen's algorithm [85] (see [98] for a general reference), which has running time $O\left(n^{2.807}\right)$

[^3]:    ${ }^{4}$ As one might have expected, given the difficultly in finding a polynomial-time algorithm.

[^4]:    ${ }^{1}$ This situation cannot occur in the subgraphs of the triangular lattice we consider in §2.6

[^5]:    ${ }^{2}$ The expression (2.9) can be obtained by standard generating function calculations, see e.g., [40, §2.3.22]

[^6]:    ${ }^{1}$ If $f_{n}$ converges point-wise on $S$ and is dominated by some integrable function $F$, then

    $$
    \lim _{n \rightarrow \infty} \int_{S} f_{n} d \mu=\int_{S} \lim _{n \rightarrow \infty} f_{n} d \mu
    $$

[^7]:    ${ }^{2}$ This is possible since all the matrices involved commute multiplicatively

