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# Probability and Real Trees 

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## Foreword

The Saint-Flour Probability Summer School was founded in 1971. It is supported by CNRS, the "Ministère de la Recherche", and the "Université Blaise Pascal".

Three series of lectures were given at the 35 th School (July 6-23, 2005) by the Professors Doney, Evans and Villani. These courses will be published separately, and this volume contains the course of Professor Evans. We cordially thank the author for the stimulating lectures he gave at the school, and for the redaction of these notes.

53 participants have attended this school. 36 of them have given a short lecture. The lists of participants and of short lectures are enclosed at the end of the volume.

Here are the references of Springer volumes which have been published prior to this one. All numbers refer to the Lecture Notes in Mathematics series, except S-50 which refers to volume 50 of the Lecture Notes in Statistics series.

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Further details can be found on the summer school web site
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Jean Picard
Clermont-Ferrand, December 2006

For Ailan Hywel, Ciaran Leuel and Huw Rhys

## Preface

These are notes from a series of ten lectures given at the Saint-Flour Probability Summer School, July 6 - July 23, 2005.

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I particularly acknowledge my wonderful collaborators over the years whose work with me appears here in some form: David Aldous, Martin Barlow, Peter Donnelly, Klaus Fleischmann, Tom Kurtz, Jim Pitman, Richard Sowers, Anita Winter, and Xiaowen Zhou. Lastly, I thank my friend and collaborator Persi Diaconis for advice on what to include in these notes.

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## Introduction

The Oxford English Dictionary provides the following two related definitions of the word phylogeny :

1. The pattern of historical relationships between species or other groups resulting from divergence during evolution.
2. A diagram or theoretical model of the sequence of evolutionary divergence of species or other groups of organisms from their common ancestors.

In short, a phylogeny is the "family tree" of a collection of units designated generically as taxa. Figure 1.1 is a simple example of a phylogeny for four primate species. Strictly speaking, phylogenies need not be trees. For instance, biological phenomena such as hybridization and horizontal gene transfer can lead to non-tree-like reticulate phylogenies for organisms. However, we will only be concerned with trees in these notes.

Phylogenetics (that is, the construction of phylogenies) is now a huge enterprise in biology, with several sophisticated computer packages employed extensively by researchers using massive amounts of DNA sequence data to study all manner of organisms. An introduction to the subject that is accessible to mathematicians is [67], while many of the more mathematical aspects are surveyed in [125].

It is often remarked that a tree is the only illustration Charles Darwin included in The Origin of Species. What is less commonly noted is that Darwin acknowledged the prior use of trees as representations of evolutionary relationships in historical linguistics - see Figure 1.2. A recent collection of papers on the application of computational phylogenetic methods to historical linguistics is [69].

The diversity of life is enormous. As J.B.S Haldane often remarked ${ }^{1}$ in various forms:

[^0]

Fig. 1.1. The phylogeny of four primate species. Illustrations are from the Tree of Life Web Project at the University of Arizona

I don't know if there is a God, but if He exists He must be inordinately fond of beetles.

Thus, phylogenetics leads naturally to the consideration of very large trees - see Figure 1.3 for a representation of what the phylogeny of all organisms might look like and browse the Tree of Life Web Project web-site at http://www.tolweb.org/tree/ to get a feeling for just how large the phylogenies of even quite specific groups (for example, beetles) can be.

Not only can phylogenetic trees be very large, but the number of possible phylogenetic trees for even a moderate number of taxa is enormous. Phylogenetic trees are typically thought of as rooted bifurcating trees with only the leaves labeled, and the number of such trees for $n$ leaves is $(2 n-3) \times(2 n-5) \times$ $\cdots \times 7 \times 5 \times 3 \times 1$ - see, for example, Chapter 3 of [67]. Consequently, if we try to use statistical methods to find the "best" tree that fits a given set of data, then it is impossible to exhaustively search all possible trees and we must use techniques such as Bayesian Markov Chain Monte Carlo and simulated annealing that randomly explore tree space in some way. Hence phylogenetics leads naturally to the study of large random trees and stochastic processes that move around spaces of large trees.


Fig. 1.2. One possible phylogenetic tree for the Indo-European family of languages from [118]

Although the investigation of random trees has a long history stretching back to the eponymous work of Galton and Watson on branching processes, a watershed in the area was the sequence of papers by Aldous [12, 13, 10]. Previous authors had considered the asymptotic behavior of numerical features of an ensemble of random trees such as their height, total number of vertices, average branching degree, etc. Aldous made sense of the idea of a sequence of trees converging to a limiting "tree-like object", so that many such limit results could be read off immediately in a manner similar to the way that limit theorems for sums of independent random variables are straightforward consequences of Donsker's invariance principle and known properties of Brownian motion. Moreover, Aldous showed that, akin to Donsker's invariance principle, many different sequences of random trees have the same limit, the Brownian continuum random tree, and that this limit is essentially the standard Brownian excursion "in disguise".

We briefly survey Aldous's work in Chapter 2, where we also present some of the historical development that appears to have led up to it, namely the probabilistic proof of the Markov chain tree theorem from [21] and the algorithm of $[17,35]$ for generating uniform random trees that was inspired by that proof. Moreover, the asymptotic behavior of the tree-generating algo-


Fig. 1.3. A somewhat impressionistic depiction of the phylogenetic tree of all life produced by David M. Hillis, Derrick Zwickl, and Robin Gutell, University of Texas
rithm when the number of vertices is large is the subject of Chapter 5 , which is based on [63].

Perhaps the key conceptual difficulty that Aldous had to overcome was how to embed the collection of finite trees into a larger universe of "tree-like objects" that can arise as re-scaling limits when the number of vertices goes to infinity. Aldous proposed two devices for doing this. Firstly, he began with a classical bijection, due to Dyck, between rooted planar trees and suitable lattice paths (more precisely, the sort of paths that can appear as the "positive excursions" of a simple random walk). He showed how such an encoding of trees as continuous functions enables us to make sense of weak convergence of random trees as just weak convergence of random functions (in the sense of weak convergence with respect to the usual supremum norm). Secondly, he
noted that a finite tree with edge lengths is naturally isomorphic to a compact subset of $\ell^{1}$, the space of absolutely summable sequences. This enabled him to treat weak convergence of random trees as just weak convergence of random compact sets (where compact subsets of $\ell^{1}$ are equipped with the Hausdorff distance arising from the usual norm on $\ell^{1}$ ).

Although Aldous's approaches are extremely powerful, the identification of trees as continuous functions or compact subsets of $\ell^{1}$ requires, respectively, that they are embedded in the plane or leaf-labeled. This embedding or labeling can be something of an artifact when the trees we are dealing with don't naturally come with such a structure. It can be particularly cumbersome when we are considering tree-valued stochastic processes, where we have to keep updating an artificial embedding or labeling as the process evolves. Aldous's perspective is analogous to the use of coordinates in differential geometry: explicit coordinates are extremely useful for many calculations but they may not always offer the smoothest approach. Moreover, it is not clear a priori that every object we might legitimately think of as tree-like necessarily has a representation as an excursion path or a subset of $\ell^{1}$. Also, the topologies inherited from the supremum norm or the Hausdorff metric may be too strong for some purposes.

We must, therefore, seek more intrinsic ways of characterizing what is meant by a "tree-like object". Finite combinatorial trees are just graphs that are connected and acyclic. If we regard the edges of such a tree as intervals, so that a tree is a cell complex (and, hence, a particular type of topological space), then these two defining properties correspond respectively to connectedness in the usual topological sense and the absence of subspaces that are homeomorphic to the circle. Alternatively, a finite combinatorial tree thought of as a cell complex has a natural metric on it: the distance between two points is just the length of the unique "path" through the tree connecting them (where each edge is given unit length). There is a well-known characterization of the metrics that are associated with trees that is often called (Buneman's) four point condition - see Chapter 3. Its significance seems to have been recognized independently in $[149,130,36]$ - see [125] for a discussion of the history.

These observations suggest that the appropriate definition of a "tree-like object" should be a general topological or metric space with analogous properties. Such spaces are called $\mathbb{R}$-trees and they have been studied extensively see $[46,45,137,39]$. We review some of the relevant theory and the connection with 0 -hyperbolicity (which is closely related to the four point condition) in Chapter 3.

We note in passing that $\mathbb{R}$-trees, albeit ones with high degrees of symmetry, play an important role in geometric group theory - see, for example, $[126,110$, $127,30,39]$. Also, 0-hyperbolic metric spaces are the simplest example of the $\delta$-hyperbolic metric spaces that were introduced in [79] as a class of spaces with global features similar to those of complete, simply connected manifolds of negative curvature. For more on the motivation and subsequent history of this notion, we refer the reader to $[33,39,80]$. Groups with a natural $\delta$ -
hyperbolic metric have turned out to be particularly important in a number of areas of mathematics, see $[79,20,40,76]$.

In order to have a nice theory of random $\mathbb{R}$-trees and $\mathbb{R}$-tree-valued stochastic processes, it is necessary to metrize a collection of $\mathbb{R}$-trees, and, since $\mathbb{R}$-trees are just metric spaces with certain special properties, this means that we need a way of assigning a distance between two metric spaces (or, more correctly, between two isometry classes of metric spaces). The Gromov-Hausdorff distance - see [80, 37, 34] - does exactly this and turns out to be very pleasant to work with. The particular properties of the Gromov-Hausdorff distance for collections of $\mathbb{R}$-trees have been investigated in $[63,65,78]$ and we describe some of the resulting theory in Chapter 4.

Since we introduced the idea of using the formalism of $\mathbb{R}$-trees equipped with the Gromov-Hausdorff metric to study the asymptotics of large random trees and tree-valued processes in [63, 65], there have been several papers that have adopted a similar point of view - see, for example, [49, 101, 102, 103, 50, 81, 78].

As we noted above, stochastic processes that move through a space of finite trees are an important ingredient for several algorithms in phylogenetic analysis. Usually, such chains are based on a set of simple rearrangements that transform a tree into a "neighboring" tree. One standard set of moves that is implemented in several phylogenetic software packages is the set of subtree prune and re-graft (SPR) moves that were first described in [134] and are further discussed in [67, 19, 125]. Moreover, as remarked in [19],

The SPR operation is of particular interest as it can be used to model biological processes such as horizontal gene transfer and recombination.

Section 2.7 of [125] provides more background on this point as well as a comment on the role of SPR moves in the two phenomena of lineage sorting and gene duplication and loss. Following [65], we investigate in Chapter 9 the behavior when the number of vertices goes to infinity of the simplest Markov chain based on SPR moves.

Tree-valued Markov processes appear in contexts other than phylogenetics. For example, a number of such processes appear in combinatorics associated with the random graph process, stochastic coalescence, and spanning trees see [115]. One such process is the wild chain, a Markov process that appears as a limiting case of tree-valued Markov chains arising from pruning operations on Galton-Watson and conditioned Galton-Watson trees in [16, 14].

The state space of the wild chain is the set $\mathbf{T}^{*}$ consisting of rooted $\mathbb{R}$ trees such that each edge has length 1 , each vertex has finite degree, and if the tree is infinite there is a single path of infinite length from the root. The wild chain is reversible (that is, symmetric). Its equilibrium measure is the distribution of the critical Poisson Galton-Watson branching process (we denote this probability measure on rooted trees by PGW(1)). When started in a state that is a finite tree, the wild chain holds for an exponentially distributed
amount of time and then jumps to a state that is an infinite tree. Then, as must be the case given that the PGW(1) distribution assigns all of its mass to finite trees, the process instantaneously re-enters the set of finite trees. In other words, the sample-paths of the wild chain bounce backwards and forwards between the finite and infinite trees.

As we show in Chapter 6 following [15], the wild chain is a particular instance of a general class of symmetric Markov processes that spend Lebesgue almost all of their time in a countable, discrete part of their state-space but continually bounce back and forth between this region and a continuous "boundary". Other processes in this general class are closely related to the Markov processes on totally disconnected Abelian groups considered in [59]. A special case of these latter group-valued processes, where the group is the additive group of a local field such as the $p$-adic numbers, is investigated in $[4,5,7,6,2,8,9,87,131,68]$.

Besides branching models such as Galton-Watson processes, another familiar source of random trees is the general class of coalescing models - see [18] for a recent survey and bibliography.

Kingman's coalescent was introduced in $[90,89]$ as a model for genealogies in the context of population genetics and has since been the subject of a large amount of applied and theoretical work - see [136, 144, 83] for an indication of some of the applications of Kingman's coalescent in genetics.

Families of coalescing Markov processes appear as duals to interacting particle systems such as the voter model and stepping stone models. Motivated by this connection, [22] investigated systems of coalescing Brownian motions and the closely related coalescing Brownian flow . Coalescing Brownian motion has recently become a topic of renewed interest, primarily in the study of filtrations and "noises" - see, for example, [140, 132, 138, 55].

In Chapter 8 we show, following [60, 44], how Kingman's coalescent and systems of coalescing Brownian motions on the circle are each naturally associated with random compact metric spaces and we investigate the fractal properties of those spaces. A similar study was performed in [28] for trees arising from the beta-coalescents of [116]. There has been quite a bit of work on fractal properties of random trees constructed in various ways from GaltonWatson branching processes; for example, [82] computed the Hausdorff dimension of the boundary of a Galton-Watson tree equipped with a natural metric - see also [104, 96].

We observe that Markov processes with continuous sample paths that take values in a space of continuous excursion paths and are reversible with respect to the distribution of standard Brownian excursion have been investigated in $[148,147,146]$. These processes can be thought of as $\mathbb{R}$-tree valued diffusion processes that are reversible with respect to the distribution of the Brownian continuum random tree.

Moving in a slightly different but related direction, there is a large literature on random walks with state-space a given infinite tree: [145, 105] are excellent bibliographical references. In particular, there is a substantial
amount of research on the Martin boundary of such walks beginning with [52, 38, 122].

The literature on diffusions on tree-like or graph-like structures is more modest. A general construction of diffusions on graphs using Dirichlet form methods is given in [141]. Diffusions on tree-like objects are studied in [42, 93] using excursion theory ideas, local times of diffusions on graphs are investigated in $[53,54]$, and an averaging principle for such processes is considered in [71]. One particular process that has received a substantial amount of attention is the so-called Walsh's spider. The spider is a diffusion on the tree consisting of a finite number of semi-infinite rays emanating from a single vertex - see [142, 26, 139, 25].

A higher dimensional diffusion with a structure somewhat akin to that of the spider, in which regions of higher dimensional spaces are "glued" together along lower dimensional boundaries, appears in the work of Sowers [133] on Hamiltonian systems perturbed by noise - see also [111]. A general construction encompassing such processes is given in [64]. This construction was used in [24] to build diffusions on the interesting fractals introduced in [95] to answer a question posed in [84].

In Chapter 7 we describe a particular Markov process with state-space an $\mathbb{R}$-tree that does not have any leaves (in the sense that any path in the tree can be continued indefinitely in both directions). The initial study of this process in [61] was motivated by Le Gall's Brownian snake process - see, for example, [97, 98, 99, 100]. One agreeable feature of this process is that it serves as a new and convenient "test bed" on which we can study many of the objects of general Markov process theory such as Doob $h$-transforms, the classification of entrance laws, the identification of the Martin boundary and representation of excessive functions, and the existence of non-constant harmonic functions and the triviality of tail $\sigma$-fields.

We use Dirichlet form methods in several chapters, so we have provided a brief summary of some of the more salient parts of the theory in Appendix A. Similarly, we summarize some results on Hausdorff dimension, packing dimension and capacity that we use in various places in Appendix B.

## Around the continuum random tree

### 2.1 Random trees from random walks

### 2.1.1 Markov chain tree theorem

Suppose that we have a discrete time Markov chain $X=\left\{X_{n}\right\}_{n \in \mathbb{N}_{0}}$ with state space $V$ and irreducible transition matrix $P$. Let $\pi$ be the corresponding stationary distribution. The Markov chain tree theorem gives an explicit formula for $\pi$, as opposed to the usual implicit description of $\pi$ as the unique probability vector that solves the equation $\pi P=\pi$. In order to describe this result, we need to introduce some more notation.

Let $G=(V, E)$ be the directed graph with vertex set $V$ and directed edges consisting of pairs of vertices $(i, j)$ such that $p_{i j}>0$. We call $p_{i j}$ the weight of the edge $(i, j)$.

A rooted spanning tree of $G$ is a directed subgraph of $G$ that is a spanning tree as an undirected graph (that is, it is a connected subgraph without any cycles that has $V$ as its vertex set) and is such that each vertex has out-degree 1 , except for a distinguished vertex, the root, that has out-degree 0 . Write $\mathcal{A}$ for the set of all rooted spanning trees of $G$ and $\mathcal{A}_{i}$ for the set of rooted spanning trees that have $i$ as their root.

The weight of a rooted spanning tree $T$ is the product of its edge weights, which we write as weight $(T)$.

Theorem 2.1. The stationary distribution $\pi$ is given by

$$
\pi_{i}=\frac{\sum_{T \in \mathcal{A}_{i}} \operatorname{weight}(T)}{\sum_{T \in \mathcal{A}} \operatorname{weight}(T)}
$$

Proof. Let $\bar{X}=\left\{\bar{X}_{n}\right\}_{n \in \mathbb{Z}}$ be a two-sided stationary Markov chain with the transition matrix $P$ (so that $\bar{X}_{n}$ has distribution $\pi$ for all $n \in \mathbb{Z}$ ).

Define a map $f: V^{\mathbb{Z}} \rightarrow \mathcal{A}$ as follows - see Figure 2.2.

- The root of $f(x)$ is $x_{0}$.


Fig. 2.1. A rooted spanning tree. The solid directed edges are in the tree, whereas dashed directed edges are edges in the underlying graph that are not in the tree. The tree is rooted at $d$. The weight of this tree is $p_{a d} p_{e d} p_{b e} p_{f e} p_{c b}$.

- For $i \neq x_{0}$, the unique edge in $f(x)$ with tail $i$ is $\left(i, x_{\tau(i)+1}\right)=$ $\left(x_{\tau(i)}, x_{\tau(i)+1}\right)$ where $\tau(i):=\sup \left\{m<0: x_{m}=i\right\}$.

It is clear that $f$ is well-defined almost surely under the distribution of $\bar{X}$ and so we can define a stationary, $\mathcal{A}$-valued, $\mathbb{Z}$-indexed stochastic process $\bar{Y}=\left\{\bar{Y}_{n}\right\}_{n \in \mathbb{Z}}$ by

$$
\bar{Y}_{n}:=f\left(\theta^{n}(\bar{X})\right), \quad n \in \mathbb{Z}
$$

where $\theta: V^{\mathbb{Z}} \rightarrow V^{\mathbb{Z}}$ denotes the usual shift operator defined by $\theta(x)_{n}:=x_{n+1}$.
It is not hard to see that $\bar{Y}$ is Markov. More specifically, consider the following forward procedure that produces a spanning tree rooted at $j$ from a spanning tree $S$ rooted at $i$ - see Figure 2.3.

- Attach the directed edge $(i, j)$ to $S$.
- This creates a directed graph with unique directed loop that contains $i$ and $j$ (possibly a self loop at $i$ ).
- Delete the unique directed edge out of $j$.
- This deletion breaks the loop and produces a spanning tree rooted at $j$.


Fig. 2.2. The construction of the rooted tree $f(x)$ for $V=\{a, b, c, d, e, f\}$ and $\left(\ldots, x_{-2}, x_{-1}, x_{0}\right)=(\ldots, e, f, c, a, c, d, d, a, f, b, f, a, c, c, f, c)$

Then, given $\left\{\bar{Y}_{m}: m \leqslant n\right\}$, the tree $\bar{Y}_{n+1}$ is obtained from the tree $\bar{Y}_{n}$ with root $i$ by choosing the new root $j$ in the forward procedure with conditional probability $p_{i j}$.

It is easy to see that a rooted spanning tree $T \in \mathcal{A}$ can be constructed from $S \in \mathcal{A}$ by the forward procedure if and only if $S$ can be constructed from $T$ by the following reverse procedure for a suitable vertex $k$.

- Let $T$ have root $j$.
- Attach the directed edge $(j, k)$ to $T$.
- This creates a directed graph with unique directed loop containing $j$ and $k$ (possibly a self loop at $j$ ).
- Delete the unique edge, say $(i, j)$, directed into $j$ that lies in this loop.
- This deletion breaks the loop and produces a rooted spanning tree rooted at $i$.

Moving up rather than down the page in Figure 2.3 illustrates the reverse procedure.

Let $S$ and $T$ be rooted spanning trees such that $T$ can be obtained from $S$ by the forward procedure, or, equivalently, such that $S$ can be obtained from $T$ by the reverse procedure. Write $i$ and $j$ for the roots of $S$ and $T$,


Fig. 2.3. The forward procedure. The dashed line represents a directed path through the tree that may consist of several directed edges.
respectively, and write $k$ for the (unique) vertex appearing in the description of the reverse procedure. Denote by $Q$ the transition matrix of the $\mathcal{A}$-valued process $\bar{Y}$. We have observed that

- If $S$ has root $i$ and $T$ has root $j$, then $Q_{S T}=p_{i j}$.
- To get $T$ from $S$ we first attached the edge $(i, j)$ and then deleted the unique outgoing edge $(j, k)$ from $j$.
- To get $S$ from $T$ we would attach the edge $(j, k)$ to $T$ and then delete the edge $(i, j)$.

Thus, if we let $\rho$ be the probability measure on $\mathcal{A}$ such that $\rho_{U}$ is proportional to the weight of $U$ for $U \in \mathcal{A}$, then we have

$$
\rho_{S} Q_{S T}=\rho_{T} R_{T S}
$$

where $R_{T S}:=p_{j k}$. In particular,

$$
\sum_{S} \rho_{S} Q_{S T}=\sum_{S} \rho_{T} R_{T S}=\rho_{T}
$$

since $R$ is a stochastic matrix. Hence $\rho$ is the stationary distribution corresponding to the irreducible transition matrix $Q$. That is, $\rho$ is the one-
dimensional marginal of the stationary chain $\bar{Y}$. We also note in passing that $R$ is the transition matrix of the time-reversal of $\bar{Y}$.

Thus,

$$
\begin{aligned}
\pi_{i} & =\sum\left\{\rho_{T}: \operatorname{root} \text { of } T \text { is } i\right\} \\
& =\frac{\sum_{T \in \mathcal{A}_{i}} \operatorname{weight}(T)}{\sum_{T \in \mathcal{A}} \operatorname{weight}(T)},
\end{aligned}
$$

as claimed.
The proof we have given of Theorem 2.1 is from [21], where there is a discussion of the history of the result.

### 2.1.2 Generating uniform random trees

Proposition 2.2. Let $\left(X_{j}\right)_{j \in \mathbb{N}_{0}}$ be the natural random walk on the complete graph $K_{n}$ with transition matrix $P$ given by $P_{i j}:=\frac{1}{n-1}$ for $i \neq j$ and $X_{0}$ uniformly distributed. Write

$$
\tau_{\nu}=\min \left\{j \geqslant 0: X_{j}=\nu\right\}, \quad \nu=1,2, \ldots, n
$$

Let $T$ be the directed subgraph of $K_{n}$ with edges

$$
\left(X_{\tau_{\nu}}, X_{\tau_{\nu}-1}\right), \quad \nu \neq X_{0}
$$

Then $T$ is uniformly distributed over the rooted spanning trees of $K_{n}$.
Proof. The argument in the proof of Theorem 2.1 plus the time-reversibility of $X$.

Remark 2.3. The set of rooted spanning trees of the complete graph $K_{n}$ is just the set of of $n^{n-1}$ rooted trees with vertices labeled by $\{1,2, \ldots, n\}$, and so the random tree $T$ produced in Proposition 2.2 is nothing other than a uniform rooted random tree with $n$ labeled vertices.

Proposition 2.2 suggests a procedure for generating uniformly distributed rooted random trees with $n$ labeled vertices. The most obvious thing to do would be to run the chain $X$ until all $n$ states had appeared and then construct the tree $T$ from the resulting sample path. The following algorithm, presented independently in [17, 35], improves on this naive approach by, in effect, generating $X_{0}$ and the pairs $\left(X_{\tau_{\nu}-1}, X_{\tau_{\nu}}\right), \nu \neq X_{0}$ without generating the rest of the sample path of $X$.

Algorithm 2.4. Fix $n \geqslant 2$. Let $U_{2}, \ldots, U_{n}$ be independent and uniformly distributed on $1, \ldots, n$, and let $\Pi$ be an independent uniform random permutation of $1, \ldots, n$.


Fig. 2.4. Step (i) of Algorithm 2.4 for $n=6$ and $\left(V_{2}, V_{3}, V_{4}, V_{5}, V_{6}\right)=(1,1,3,2,3)$
(i) For $2 \leqslant i \leqslant n$ connect vertex $i$ to vertex $V_{i}=(i-1) \wedge U_{i}$ (that is, build a tree rooted at 1 with edges $\left(i, V_{i}\right)$ ).
(ii)Relabel the vertices $1, \ldots, n$ as $\Pi_{1}, \ldots, \Pi_{n}$ to produce a tree rooted at $\Pi_{1}$.

See Figure 2.4 for an example of Step (i) of the algorithm.

Proposition 2.5. The rooted random tree with $n$ labeled vertices produced by Algorithm 2.4 is uniformly distributed.

Proof. Let $Z_{0}, Z_{1}, \ldots$ be independent and uniform on $1,2, \ldots, n$. Define $\pi_{1}, \pi_{2}, \ldots, \pi_{n}, \xi_{1}, \xi_{2}, \ldots, \xi_{n} \in \mathbb{N}_{0}$ and $\lambda_{2}, \ldots, \lambda_{n} \in\{1,2, \ldots, n\}$ by

$$
\begin{gathered}
\xi_{1}:=0 \\
\pi_{1}:=Z_{0} \\
\xi_{i+1}:=\min \left\{m>\xi_{i}: Z_{m} \notin\left\{\pi_{1}, \ldots, \pi_{i}\right\}\right\}, \quad 1 \leqslant i \leqslant n-1, \\
\pi_{i+1}:=Z_{\xi_{i+1}}, \quad 1 \leqslant i \leqslant n-1 \\
\lambda_{i+1}:=Z_{\xi_{i+1}-1}, \quad 1 \leqslant i \leqslant n-1 .
\end{gathered}
$$

Consider the random tree $T$ labeled by $\{1,2, \ldots, n\}$ with edges $\left(\pi_{i}, \lambda_{i}\right), 2 \leqslant$ $i \leqslant n$.

Note that this construction would give the same tree if the sequence $Z$ was replaced by the subsequence $Z^{\prime}$ in which terms $Z_{i}$ identical with their predecessor $Z_{i-1}$ were deleted. The process $Z^{\prime}$ is just the natural random walk on the complete graph. Thus, the construction coincides with the construction of Proposition 2.2. Hence, the tree $T$ is a uniformly distributed tree on $n$ labeled vertices. To complete the proof, we need only argue that this construction is equivalent to Algorithm 2.4.

It is clear that $\pi$ is a uniform random permutation. The construction of the tree of $T$ can be broken into two stages.
(i) Connect $i$ to $\pi_{\lambda_{i}}^{-1}, i=2, \ldots, n$.
(ii) Relabel $1, \ldots, n$ as $\pi_{1}, \ldots, \pi_{n}$.

Thus, it will suffice to show that the conditional joint distribution of the random variables $\pi_{\lambda_{i}}^{-1}, i=2, \ldots, n$, given $\pi$ is always the same as the (unconditional) joint distribution of the random variables $V_{i}, i=2, \ldots, n$, in Algorithm 2.4 no matter what the value of $\pi$ is.

To see that this is so, first fix $i$ and condition on $Z_{1}, \ldots, Z_{\xi_{i}}$ as well as $\pi$. Note the following two facts.

- With probability $1-i / n$ we have $\xi_{i+1}=\xi_{i}+1$, which implies that $\lambda_{i+1}=$ $Z_{\xi_{i}}$ and, hence, $\pi_{\lambda_{i+1}}^{-1}=i$.
- Otherwise, $\xi_{i+1}=\xi_{i}+M+1$ for some random integer $M \geqslant 1$. Conditioning on the event $\{M=m\}$, we have that the random variables $Z_{\xi_{i}+1}, \ldots, Z_{\xi_{i}+m}$ are independent and uniformly distributed on the previously visited states $\left\{\pi_{1}, \ldots, \pi_{i}\right\}$. In particular, $\lambda_{i+1}=Z_{\xi_{i}+m}$ is uniformly distributed on $\left\{\pi_{1}, \ldots, \pi_{i}\right\}$, and so $\pi_{\lambda_{i+1}}^{-1}$ is uniformly distributed on $\{1, \ldots, i\}$.
Combining these two facts, we see that

$$
\begin{aligned}
\mathbb{P}\left\{\pi_{\lambda_{i+1}}^{-1}=u \mid Z_{1}, \ldots, Z_{\xi_{i}}, \pi\right\} & =1 / n=\mathbb{P}\left\{V_{i+1}=u\right\}, \quad 1 \leqslant u \leqslant i-1, \\
\mathbb{P}\left\{\pi_{\lambda_{i+1}}^{-1}=i \mid Z_{1}, \ldots, Z_{\xi_{i}} \pi\right\} & =1-(i-1) / n=\mathbb{P}\left\{V_{i+1}=i\right\}
\end{aligned}
$$

as required.

### 2.2 Random trees from conditioned branching processes

If we were to ask most probabilists to propose a natural model for generating random trees, they would first think of the family tree of a Galton-Watson branching process. Such a tree has a random number of vertices and if we further required that the random tree had a fixed number $n$ of vertices, then they would suggest simply conditioning the total number of vertices in the Galton-Watson tree to be $n$. Interestingly, special cases of this mechanism for generating random trees produce trees that are also natural from a combinatorial perspective, as we shall soon see.

Let $\left(p_{i}\right)_{i \in \mathbb{N}_{0}}$ be a probability distribution on the non-negative integers that has mean one. Write $T$ for the family tree of the Galton-Watson branching process with offspring distribution $\left(p_{i}\right)_{i \in \mathbb{N}_{0}}$ started with 1 individual in generation 0 . For $n \geqslant 1$ denote by $T_{n}$ a random tree that arises by conditioning on the total population size $|T|$ being $n$ (we suppose that the event $\{|T|=n\}$ has positive probability). More precisely, we think of the trees $T$ and $T_{n}$ as rooted ordered trees: a rooted tree is ordered if we distinguish the offspring of a vertex according with a "birth order". Equivalently, a rooted ordered tree is a rooted planar tree: the birth order is given by the left-to-right ordering of offspring in the given embedding of the tree in the plane. The distribution of the random tree $T$ is then

$$
\begin{aligned}
\mathbb{P}\{T=t\} & =\prod_{v \in t} p_{d(v, t)} \\
& =\prod_{i \geqslant 0} p_{i}^{D_{i}(t)} \\
& =: \omega(t)
\end{aligned}
$$

where $d(v, t)$ is the number of offspring of vertex $v$ in $t$, and $D_{i}(t)$ is the number of vertices in $t$ with $i$ offspring. Thus, $\mathbb{P}\left\{T_{n}=t\right\}$ is proportional to $\omega(t)$.

Example 2.6. If the offspring distribution $\left(p_{i}\right)_{i \in \mathbb{N}_{0}}$ is the geometric distribution $p_{i}=2^{-(i+1)}, i \in \mathbb{N}_{0}$, then $T_{n}$ is uniformly distributed (on the set of rooted ordered trees with $n$ vertices).

Example 2.7. Suppose that the offspring distribution $\left(p_{i}\right)_{i \in \mathbb{N}_{0}}$ is the Poisson distribution $p_{i}=\frac{e^{-1}}{i!}, i \in \mathbb{N}_{0}$. If we randomly assign the labels $\{1,2, \ldots, n\}$ to the vertices of $T_{n}$ and ignore the ordering, then $T_{n}$ is a uniformly distributed rooted labeled tree with $n$ vertices.

### 2.3 Finite trees and lattice paths

Although rooted planar trees are not particularly difficult to visualize, we would like to have a quite concrete way of "representing" or "coordinatizing" the planar trees with $n$ vertices that is amenable to investigating the behavior of a random such trees as the number of vertices becomes large. The following simple observation is the key to the work of Aldous, Le Gall and many others on the connections between the asymptotics of large random trees and models for random paths.

Given a rooted planar tree with $n$ vertices, start from the root and traverse the tree as follows. At each step move away from the root along the leftmost edge that has not been walked on yet. If this is not possible then step back along the edge leading toward the root. We obtain a with steps of
$\pm 1$ by plotting the height (that is, the distance from the root) at each step. Appending $a+1$ step at the beginning and $a-1$ step at the end gives a lattice excursion path with $2 n$ steps that we call the Harris path of the tree, although combinatorialists usually call this object a Dyck path. We can reverse this procedure and obtain a rooted planar tree with $n$ vertices from any lattice excursion path with $2 n$ steps.


Fig. 2.5. Harris path of a rooted combinatorial tree (figure courtesy of Jim Pitman).

### 2.4 The Brownian continuum random tree

Put $S_{n}=\sum_{i=1}^{n} X_{i}$, where the random variables $X_{i}$ are independent with $\mathbb{P}\left\{X_{i}= \pm 1\right\}=1 / 2$. Conditional on $S_{1}=+1$, the path $S_{0}, S_{1}, \ldots, S_{N}$, where $N=\min \left\{k>0: S_{k}=0\right\}$, is the Harris path of the Galton-Watson branching process tree with offspring distribution $p_{i}=2^{-(i+1)}, i \in \mathbb{N}_{0}$. Therefore, if we condition on $S_{1}=+1$ and $N=2 n$, we get the Harris path of the GaltonWatson branching process tree conditioned on total population size $n$, and we have observed is the uniform rooted planar tree on $n$ vertices.

We know that suitably re-scaled simple random walk converges to Brownian motion. Similarly, suitably re-scaled simple random walk conditioned to be positive on the first step and return to zero for the first time at time $2 n$ converges as $n \rightarrow \infty$ to the standard Brownian excursion. Of course, simple random walk is far from being the only process that has Brownian motion as a scaling limit, and so we might hope that there are other random trees with Harris paths that converge to standard Brownian excursion after re-scaling. The following result of Aldous [10] shows that this is certainly the case.

Theorem 2.8. Let $T_{n}$ be a conditioned Galton-Watson tree, with offspring mean 1 and variance $0<\sigma^{2}<\infty$. Write $H_{n}(k), 0 \leqslant k \leqslant 2 n$ for the Harris path associated with $T_{n}$, and interpolate $H_{n}$ linearly to get a continuous process
real-valued indexed by the interval $[0,2 n]$ (which we continue to denote by $H_{n}$ ). Then, as $n \rightarrow \infty$ through possible sizes of the unconditioned Galton-Watson tree,

$$
\left(\sigma \frac{H_{n}(2 n u)}{\sqrt{n}}\right)_{0 \leqslant u \leqslant 1} \Rightarrow\left(2 B_{u}^{e x}\right)_{0 \leqslant u \leqslant 1}
$$

where $B^{\text {ex }}$ is the standard Brownian excursion and $\Rightarrow$ is the usual weak convergence of probability measures on $C[0,1]$.

The Harris path construction gives a bijection between excursion-like lattice paths with steps of $\pm 1$ and rooted planar trees. We will observe in Example 3.14 that any continuous excursion path gives rise to a tree-like object via an analogy with one direction of this bijection. Hence Theorem 2.8 shows that, in some sense, any conditioned finite-variance Galton-Watson tree converges after re-scaling to the tree-like object associated with $2 B^{\text {ex }}$. Aldous called this latter object the Brownian continuum random tree .

### 2.5 Trees as subsets of $\ell^{1}$

We have seen in Sections 2.3 and 2.4 that representing trees as continuous paths allows us to use the metric structure on path space to make sense of the idea of a family of random trees converging to some limit random object. In this section we introduce an alternative "coordinatization" of tree-space as the collection of compact subsets of the Banach space $\ell^{1}:=\left\{\left(x_{1}, x_{2}, \ldots\right)\right.$ : $\left.\sum_{i}\left|x_{i}\right|<\infty\right\}$. This allows us to use the machinery that has been developed for describing random subsets of a metric space to give another way of expressing such convergence results.

Equip $\ell^{1}$ with the usual norm. Any finite tree with edge lengths can be embedded isometrically as a subset of $\ell^{1}$ (we think of such a tree as a onedimensional cell complex, that is, as a metric space made up of the vertices of the tree and the connecting edges - not just as the finite metric space consisting of the vertices themselves). For example, the tree of Figure 2.6 is isometric to the set

$$
\begin{aligned}
\left\{t e_{1}\right. & : 0 \leqslant t \leqslant d(\rho, a)\} \\
& \cup\left\{d(\rho, d) e_{1}+t e_{2}: 0 \leqslant t \leqslant d(d, b)\right\} \\
& \cup\left\{d(\rho, d) e_{1}+d(d, e) e_{2}+t e_{3}: 0 \leqslant t \leqslant d(e, c)\right\}
\end{aligned}
$$

where $e_{1}=(1,0,0, \ldots), e_{2}=(0,1,0, \ldots)$, etc.
Recall Algorithm 2.4 for producing a uniform tree on $n$ labeled vertices. Let $\mathcal{S}^{n}$ be the subset of $\ell^{1}$ that corresponds to the tree produced by the algorithm. We think of this tree as having edge lengths all equal to 1. More precisely, define a random length random sequence $\left(C_{j}^{n}, B_{j}^{n}\right), 0 \leqslant j \leqslant J^{n}$, as follows:

- $C_{0}^{n}=B_{0}^{n}:=0$,


Fig. 2.6. A rooted tree with edge lengths

- $C_{j}^{n}$ is the $j$ th element of $\left\{i: U_{i}<i-1\right\}$,
- $B_{j}^{n}:=U_{C_{j}^{n}}$.

Define $\rho^{n}:\left[0, C_{J^{n}}^{n}\right] \rightarrow \ell^{1}$ by $\rho^{n}(0):=0$ and

$$
\rho^{n}(x):=\rho^{n}\left(B_{j}^{n}\right)+\left(x-C_{j}^{n}\right) e_{j+1} \text { for } C_{j}^{n}<x \leqslant C_{j+1}^{n}, \quad 0 \leqslant j \leqslant J^{n}-1 .
$$

Put $\mathcal{S}^{n}:=\rho^{n}\left(\left[0, C_{J^{n}}^{n}\right]\right)$.
It is not hard to show that

$$
\begin{aligned}
& \left(\left(n^{-1 / 2} C_{1}^{n}, n^{-1 / 2} B_{1}^{n}\right),\left(n^{-1 / 2} C_{2}^{n}, n^{-1 / 2} B_{2}^{n}\right), \ldots\right) \\
& \quad \Rightarrow\left(\left(C_{1}, B_{1}\right),\left(C_{2}, B_{2}\right), \ldots\right)
\end{aligned}
$$

where $\Rightarrow$ denotes weak convergence and $\left(\left(C_{1}, B_{1}\right),\left(C_{2}, B_{2}\right), \ldots\right)$ are defined as follows. Put $C_{0}=B_{0}:=0$. Let $\left(C_{1}, C_{2}, \ldots\right)$ be the arrival times in an inhomogeneous Poisson process on $\mathbb{R}_{+}$with intensity $t d t$. Let $B_{j}:=\xi_{j} C_{j}$, where the $\left\{\xi_{j}\right\}_{j \in \mathbb{N}}$ are independent, identically distributed uniform random variables on $[0,1]$, independent of $\left\{C_{j}\right\}_{j \in \mathbb{N}}$.

Define $\rho: \mathbb{R}_{+} \rightarrow \ell^{1}$ by $\rho(0):=0$ and

$$
\rho(x):=\rho\left(B_{j}\right)+\left(x-C_{j}\right) e_{j+1} \text { for } C_{j}<x \leqslant C_{j+1} .
$$

Set

$$
\mathcal{S}:=\overline{\bigcup_{t \geqslant 0} \rho([0, t])}
$$

It seems reasonable that

$$
n^{-1 / 2} \mathcal{S}^{n} \Rightarrow \mathcal{S}
$$

in some sense. Aldous [12] showed that $\mathcal{S}$ is almost surely a compact subset of $\ell^{1}$ and that there is convergence in the sense of weak convergence of random compact subsets of $\ell^{1}$ equipped with the Hausdorff metric that we will discuss in Section 4.1. Aldous studied $\mathcal{S}$ further in [13, 10]. In particular, he showed that $\mathcal{S}$ is tree-like in various senses: for example, for any two points $x, y \in \mathcal{S}$ there is a unique path connecting $x$ and $y$ (that is, a unique homeomorphic image of the unit interval), and this path has length $\|x-y\|_{1}$.

Because the uniform rooted tree with $n$ labeled vertices is a conditioned Galton-Watson branching process (for the Poisson offspring distribution), we see from Theorem 2.8 that the Poisson line-breaking "tree" $\mathcal{S}$ is essentially the same as the Brownian continuum random tree, that is, the random tree-like object associated with the random excursion path $2 B^{e x}$. In fact, the random tree $\rho\left(C_{n}\right)$ has the same distribution as the subtree of the Brownian CRT spanned by $n$ i.i.d. uniform points on the unit interval.

## $\mathbb{R}$-trees and 0-hyperbolic spaces

### 3.1 Geodesic and geodesically linear metric spaces

We follow closely the development in [39] in this section and leave some of the more straightforward proofs to the reader.

Definition 3.1. $A$ segment in a metric space $(X, d)$ is the image of an isometry $\alpha:[a, b] \rightarrow X$. The end points of the segment are $\alpha(a)$ and $\alpha(b)$.

Definition 3.2. A metric space $(X, d)$ is geodesic if for all $x, y \in X$, there is a segment in $X$ with endpoints $\{x, y\}$, and $(X, d)$ is geodesically linear if, for all $x, y \in X$, there is a unique segment in $X$ with endpoints $\{x, y\}$.

Example 3.3. Euclidean space $\mathbb{R}^{d}$ is geodesically linear. The closed annulus $\left\{z \in \mathbb{R}^{2}: 1 \leqslant|z| \leqslant 2\right\}$ is not geodesic in the metric inherited from $\mathbb{R}^{2}$, but it is geodesic in the metric defined by taking the infimum of the Euclidean lengths of piecewise-linear paths between two points. The closed annulus is not geodesically linear in this latter metric: for example, a pair of points of the form $z$ and $-z$ are the endpoints of two segments - see Figure 3.1. The open annulus $\left\{z \in \mathbb{R}^{2}: 1<|z|<2\right\}$ is not geodesic in the metric defined by taking the infimum over all piecewise-linear paths between two points: for example, there is no segment that has a pair of points of the form $z$ and $-z$ as its endpoints.

Lemma 3.4. Consider a metric space $(X, d)$. Let $\sigma$ be a segment in $X$ with endpoints $x$ and $z$, and let $\tau$ be a segment in $X$ with endpoints $y$ and $z$.
(a) Suppose that $d(u, v)=d(u, z)+d(z, v)$ for all $u \in \sigma$ and $v \in \tau$. Then $\sigma \cup \tau$ is a segment with endpoints $x$ and $y$.
(b) Suppose that $\sigma \cap \tau=\{z\}$ and $\sigma \cup \tau$ is a segment. Then $\sigma \cup \tau$ has endpoints $x$ and $y$.


Fig. 3.1. Two geodesics with the same endpoints in the intrinsic path length metric on the annulus

Lemma 3.5. Let $(X, d)$ be a geodesic metric space such that if two segments of $(X, d)$ intersect in a single point, which is an endpoint of both, then their union is a segment. Then $(X, d)$ is a geodesically linear metric space.

Proof. Let $\sigma, \tau$ be segments, both with endpoints $u, v$. Fix $w \in \sigma$, and define $w^{\prime}$ to be the point of $\tau$ such that $d(u, w)=d\left(u, w^{\prime}\right)$ (so that $\left.d(v, w)=d\left(v, w^{\prime}\right)\right)$. We have to show $w=w^{\prime}$.

Let $\rho$ be a segment with endpoints $w, w^{\prime}$. Now $\sigma=\sigma_{1} \cup \sigma_{2}$, where $\sigma_{1}$ is a segment with endpoints $u, w$, and $\sigma_{2}$ is a segment with endpoints $w, v-$ see Figure 3.2.

We claim that either $\sigma_{1} \cap \rho=\{w\}$ or $\sigma_{2} \cap \rho=\{w\}$. This is so because if $x \in \sigma_{1} \cap \rho$ and $y \in \sigma_{2} \cap \rho$, then $d(x, y)=d(x, w)+d(w, y)$, and either $d(w, y)=d(w, x)+d(x, y)$ or $d(w, x)=d(w, y)+d(x, y)$, depending on how $x, y$ are situated in the segment $\rho$. It follows that either $x=w$ or $y=w$, establishing the claim.

Now, if $\sigma_{1} \cap \rho=\{w\}$, then, by assumption, $\sigma_{1} \cup \rho$ is a segment, and by Lemma 3.4(b) its endpoints are $u, w^{\prime}$. Since $w \in \sigma_{1} \cup \rho, d\left(u, w^{\prime}\right)=d(u, w)+$ $d\left(w, w^{\prime}\right)$, so $w=w^{\prime}$. Similarly, if $\sigma_{2} \cap \rho=\{w\}$ then $w=w^{\prime}$.

Lemma 3.6. Consider a geodesically linear metric space $(X, d)$.


Fig. 3.2. Construction in the proof of Lemma 3.5
(i) Given points $x, y, z \in X$, write $\sigma$ for the segment with endpoints $x, y$. Then $z \in \sigma$ if and only if $d(x, y)=d(x, z)+d(z, y)$.
(ii) The intersection of two segments in $X$ is also a segment if it is nonempty.
(iii) Given $x, y \in X$, there is a unique isometry $\alpha:[0, d(x, y)] \rightarrow X$ such that $\alpha(0)=x$ and $\alpha(d(x, y))=y$. Write $[x, y]$ for the resulting segment. If $u, v \in[x, y]$, then $[u, v] \subseteq[x, y]$.

### 3.2 0-hyperbolic spaces

Definition 3.7. For $x, y, v$ in a metric space $(X, d)$, set

$$
(x \cdot y)_{v}:=\frac{1}{2}(d(x, v)+d(y, v)-d(x, y))
$$

- see Figure 3.3.

Remark 3.8. For $x, y, v, t \in X$,

$$
0 \leqslant(x \cdot y)_{v} \leqslant d(x, v) \wedge d(y, v)
$$



Fig. 3.3. $(x \cdot y)_{v}=d(w, v)$ in this tree
and

$$
(x \cdot y)_{t}=d(t, v)+(x \cdot y)_{v}-(x \cdot t)_{v}-(y \cdot t)_{t} .
$$

Definition 3.9. A metric space $(X, d)$ is 0-hyperbolic with respect to $v$ if for all $x, y, z \in X$

$$
(x \cdot y)_{v} \geqslant(x \cdot z)_{v} \wedge(y \cdot z)_{v}
$$

- see Figure 3.4.

Lemma 3.10. If the metric space $(X, d)$ is 0 -hyperbolic with respect to some point of $X$, then $(X, d)$ is 0 -hyperbolic with respect to all points of $X$.

Remark 3.11. In light of Lemma 3.10, we will refer to a metric space that is 0 -hyperbolic with respect to one, and hence all, of its points as simply being 0 -hyperbolic. Note that any subspace of a 0 -hyperbolic metric space is also 0 -hyperbolic.

Lemma 3.12. The metric space $(X, d)$ is 0 -hyperbolic if and only if

$$
d(x, y)+d(z, t) \leqslant \max \{d(x, z)+d(y, t), d(y, z)+d(x, t)\}
$$

for all $x, y, z, t \in X$,


Fig. 3.4. The 0-hyperbolicity condition holds for this tree. Here $(x \cdot y)_{v}$ and $(y \cdot z)_{v}$ are both given by the length of the dotted segment, and $(x \cdot z)_{v}$ is the length of the dashed segment. Note that $(x \cdot y)_{v} \geqslant(x \cdot z)_{v} \wedge(y \cdot z)_{v}$, with similar inequalities when $x, y, z$ are permuted.

Remark 3.13. The set of inequalities in Lemma 3.12 is usually called the fourpoint condition - see Figure 3.5.

Example 3.14. Write $C\left(\mathbb{R}_{+}\right)$for the space of continuous functions from $\mathbb{R}_{+}$ into $\mathbb{R}$. For $e \in C\left(\mathbb{R}_{+}\right)$, put $\zeta(e):=\inf \{t>0: e(t)=0\}$ and write

$$
U:=\left\{\begin{array}{c}
e(0)=0, \zeta(e)<\infty \\
e \in C\left(\mathbb{R}_{+}\right): \quad e(t)>0 \text { for } 0<t<\zeta(e) \\
\text { and } e(t)=0 \text { for } t \geqslant \zeta(e)
\end{array}\right\}
$$

for the space of positive excursion paths. Set $U^{\ell}:=\{e \in U: \zeta(e)=\ell\}$.
We associate each $e \in U^{\ell}$ with a compact metric space as follows. Define an equivalence relation $\sim_{e}$ on $[0, \ell]$ by letting

$$
u_{1} \sim_{e} u_{2}, \quad \text { iff } \quad e\left(u_{1}\right)=\inf _{u \in\left[u_{1} \wedge u_{2}, u_{1} \vee u_{2}\right]} e(u)=e\left(u_{2}\right)
$$

Consider the following semi-metric on $[0, \ell]$


Fig. 3.5. The four-point condition holds on a tree: $d(x, z)+d(y, t) \leqslant d(x, y)+$ $d(z, t)=d(x, t)+d(y, z)$

$$
d_{T_{e}}\left(u_{1}, u_{2}\right):=e\left(u_{1}\right)-2 \inf _{u \in\left[u_{1} \wedge u_{2}, u_{1} \vee u_{2}\right]} e(u)+e\left(u_{2}\right),
$$

that becomes a true metric on the quotient space $T_{e}:=\left.[0, \ell]\right|_{\sim_{e}}-$ see Figure 3.6.

It is straightforward to check that the quotient map from $[0, \ell]$ onto $T_{e}$ is continuous with respect to $d_{T_{e}}$. Thus, $\left(T_{e}, d_{T_{e}}\right)$ is path-connected and compact as the continuous image of a metric space with these properties. In particular, $\left(T_{e}, d_{T_{e}}\right)$ is complete. It is not difficult to check that $\left(T_{e}, d_{T_{e}}\right)$ satisfies the four-point condition, and, hence, is 0-hyperbolic.

## $3.3 \mathbb{R}$-trees

### 3.3.1 Definition, examples, and elementary properties

Definition 3.15. An $\mathbb{R}$-tree is a metric space $(X, d)$ with the following properties.

Axiom (a) The space $(X, d)$ is geodesic.


Fig. 3.6. An excursion path on $[0,1]$ determines a distance between the points $a$ and $b$

Axiom (b) If two segments of $(X, d)$ intersect in a single point, which is an endpoint of both, then their union is a segment.

Example 3.16. Finite trees with edge lengths (sometimes called weighted trees) are examples of $\mathbb{R}$-trees. To be a little more precise, we don't think of such a tree as just being its finite set of vertices with a collection of distances between them, but regard the edges connecting the vertices as also being part of the metric space.

Example 3.17. Take $X$ to be the plane $\mathbb{R}^{2}$ equipped with the metric

$$
d\left(\left(x_{1}, x_{2}\right),\left(y_{1}, y_{2}\right)\right):= \begin{cases}\left|x_{2}-y_{2}\right|, & \text { if } x_{1}=y_{1} \\ \left|x_{1}-y_{1}\right|+\left|x_{2}\right|+\left|y_{2}\right|, & \text { if } x_{1} \neq y_{1}\end{cases}
$$

That is, we think of the plane as being something like the skeleton of a fish, in which the horizontal axis is the spine and vertical lines are the ribs. In order to compute the distance between two points on different ribs, we use the length of the path that goes from the first point to the spine, then along the spine to the rib containing the second point, and then along that second rib - see Figure 3.7.


Fig. 3.7. The distance between two points of $\mathbb{R}^{2}$ in the metric of Example 3.17 is the (Euclidean) length of the dashed path

Example 3.18. Consider the collection $\mathcal{T}$ of bounded subsets of $\mathbb{R}$ that contain their supremum. We can think of the elements of $\mathcal{T}$ as being arrayed in a tree-like structure in the following way. Using genealogical terminology, write $h(B):=\sup B$ for the real-valued generation to which $B \in \mathcal{T}$ belongs and $B \mid t:=(B \cap]-\infty, t]) \cup\{t\} \in \mathcal{T}$ for $t \leqslant h(B)$ for the ancestor of $B$ in generation $t$. For $A, B \in \mathcal{T}$ the generation of the most recent common ancestor of $A$ and $B$ is $\tau(A, B):=\sup \{t \leqslant h(A) \wedge h(B): A|t=B| t\}$. That is, $\tau(A, B)$ is the generation at which the lineages of $A$ and $B$ diverge. There is a natural genealogical distance on $\mathcal{T}$ given by

$$
D(A, B):=[h(A)-\tau(A, B)]+[h(B)-\tau(A, B)] .
$$

See Figure 3.8.
It is not difficult to show that the metric space $(\mathcal{T}, D)$ is a $\mathbb{R}$-tree. For example, the segment with end-points $A$ and $B$ is the set $\{A \mid t: \tau(A, B) \leqslant$ $t \leqslant h(A)\} \cup\{B \mid t: \tau(A, B) \leqslant t \leqslant h(B)\}$.

The metric space $(\mathcal{T}, D)$ is essentially "the" real tree of $[47,137]$ (the latter space has as its points the bounded subsets of $\mathbb{R}$ that contain their infimum and the corresponding metric is such that the map from $(\mathcal{T}, D)$ into this latter space given by $A \mapsto-A$ is an isometry). With a slight abuse of


Fig. 3.8. The set $C$ is the most recent common ancestor of the sets $A, B \subset \mathbb{R}$ thought of as points of "the" real tree of Example 3.18. The distance $D(A, B)$ is $[s-u]+[t-u]$.
nomenclature, we will refer here to $(\mathcal{T}, D)$ as the real tree. Note that $(\mathcal{T}, D)$ is huge: for example, the removal of any point shatters $\mathcal{T}$ into uncountably many connected components.

Example 3.19. We will see in Example 3.37 that the compact 0-hyperbolic metric space $\left(T_{e}, d_{T_{e}}\right)$ of Example 3.14 that arises from an excursion path $e \in U$ is a $\mathbb{R}$-tree.

The following result is a consequence of Axioms (a) and (b) and Lemma 3.5.
Lemma 3.20. An $\mathbb{R}$-tree is geodesically linear. Moreover, if $(X, d)$ is a $\mathbb{R}$-tree and $x, y, z \in X$ then $[x, y] \cap[x, z]=[x, w]$ for some unique $w \in X$.

Remark 3.21. It follows from Lemma 3.4, Lemma 3.6 and Lemma 3.20 that Axioms (a) and (b) together imply following condition that is stronger than Axiom (b):
Axiom (b') If $(X, d)$ is a $\mathbb{R}$-tree, $x, y, z \in X$ and $[x, y] \cap[x, z]=\{x\}$, then $[x, y] \cup[x, z]=[y, z]$

Lemma 3.22. Let $x, y, z$ be points of a $\mathbb{R}$-tree $(X, d)$, and write $w$ for the unique point such that $[x, y] \cap[x, z]=[x, w]$.
(i) The points $x, y, z, w$ and the segments connecting them form a $Y$ shape, with $x, y, z$ at the tips of the $Y$ and $w$ at the center. More precisely, $[y, w] \cap[w, z]=\{w\},[y, z]=[y, w] \cup[w, z]$ and $[x, y] \cap[w, z]=\{w\}$.
(ii) If $y^{\prime} \in[x, y]$ and $z^{\prime} \in[x, z]$, then

$$
d\left(y^{\prime}, z^{\prime}\right)=\left\{\begin{array}{l}
\left|d\left(x, y^{\prime}\right)-d\left(x, z^{\prime}\right)\right|, \text { if } d\left(x, y^{\prime}\right) \wedge d\left(x, z^{\prime}\right) \leqslant d(x, w) \\
d\left(x, y^{\prime}\right)+d\left(x, z^{\prime}\right)-2 d(x, w), \text { otherwise }
\end{array}\right.
$$

(iii) The "centroid" $w$ depends only on the set $\{x, y, z\}$, not on the order in which the elements are written.

Proof. (i) Since $y, w \in[x, y]$, we have $[y, w] \subseteq[x, y]$. Similarly, $[w, z] \subseteq[x, z]$. So, if $u \in[y, w] \cap[w, z]$, then $u \in[x, y] \cap[x, z]=[x, w]$. Hence $u \in[x, w] \cap$ $[y, w]=\{w\}$ (because $w \in[x, y]$ ). Thus, $[y, w] \cap[w, z]=\{w\}$, and $[y, z]=$ $[y, w] \cup[w, z]$ by Axiom (b').

Now, since $w \in[x, y]$, we have $[x, y]=[x, w] \cup[w, y]$, so $[x, y] \cap[w, z]=$ $([x, w] \cap[w, z]) \cup([y, w] \cap[w, z])$, and both intersections are equal to $\{w\}$ ( $w \in[x, z]$ ).
(ii) If $d\left(x, y^{\prime}\right) \leqslant d(x, w)$ then $y^{\prime}, z^{\prime} \in[x, z]$, and so $d\left(y^{\prime}, z^{\prime}\right)=\left|d\left(x, y^{\prime}\right)-d\left(x, z^{\prime}\right)\right|$. Similarly, if $d\left(x, z^{\prime}\right) \leqslant d(x, w)$, then $y^{\prime}, z^{\prime} \in[x, y]$, and once again $d\left(y^{\prime}, z^{\prime}\right)=$ $\left|d\left(x, y^{\prime}\right)-d\left(x, z^{\prime}\right)\right|$.

If $d\left(x, y^{\prime}\right)>d(x, w)$ and $d\left(x, z^{\prime}\right)>d(x, w)$, then $y^{\prime} \in[y, w]$ and $z^{\prime} \in[z, w]$. Hence, by part (i),

$$
\begin{aligned}
d\left(y^{\prime}, z^{\prime}\right) & =d\left(y^{\prime}, w\right)+d\left(w, z^{\prime}\right) \\
& =\left(d\left(x, y^{\prime}\right)-d(x, w)\right)+\left(d\left(x, z^{\prime}\right)-d(x, w)\right. \\
& =d\left(x, y^{\prime}\right)+d\left(x, z^{\prime}\right)-2 d(x, w)
\end{aligned}
$$

(iii) We have by part (i) that

$$
\begin{aligned}
{[y, x] \cap[y, z] } & =[y, x] \cap([y, w] \cup[w, z]) \\
& =[y, w] \cup([y, x] \cap[w, z]) \\
& =[y, w] \cup([y, w] \cap[w, z]) \cup([w, x] \cap[w, z])
\end{aligned}
$$

Now $[y, w] \cap[w, z]=\{w\}$ by part (1) and $[w, x] \cap[w, z]=\{w\}$ since $w \in[x, z]$. Hence, $[y, x] \cap[y, z]=[y, w]$. Similarly, $[z, x] \cap[z, y]=[z, w]$, and part (iii) follows.
Definition 3.23. In the notation of Lemma 3.22, write $Y(x, y, z):=w$ for the centroid of $\{x, y, z\}$.
Remark 3.24. Note that we have

$$
[x, y] \cap[w, z]=[x, z] \cap[w, y]=[y, z] \cap[w, x]=\{w\}
$$

Also, $d(x, w)=(y \cdot z)_{x}, d(y, w)=(x \cdot z)_{y}$, and $d(z, w)=(x \cdot y)_{z}$. In Figure 3.3, $Y(x, y, v)=w$.

Corollary 3.25. Consider $a \mathbb{R}$-tree $(X, d)$ and points $x_{0}, x_{1}, \ldots, x_{n} \in X$. The segment $\left[x_{0}, x_{n}\right]$ is a subset of $\bigcup_{i=1}^{n}\left[x_{i-1}, x_{i}\right]$.
Proof. If $n=2$, then, by Lemma 3.22,

$$
\left[x_{0}, x_{2}\right]=\left[x_{0}, Y\left(x_{0}, x_{1}, x_{2}\right)\right] \cup\left[Y\left(x_{0}, x_{1}, x_{2}\right), x_{2}\right] \subseteq\left[x_{0}, x_{1}\right] \cup\left[x_{1}, x_{2}\right]
$$

If $n>2$, then $\left[x_{0}, x_{n}\right] \subseteq\left[x_{0}, x_{n-1}\right] \cup\left[x_{n-1}, x_{n}\right]$ by the case $n=2$, and the result follows by induction on $n$.

Lemma 3.26. Consider a $\mathbb{R}$-tree $(X, d)$. Let $\alpha:[a, b] \rightarrow X$ be a continuous map. If $x=\alpha(a)$ and $y=\alpha(b)$, then $[x, y]$ is a subset of the image of $\alpha$.
Proof. Let $A$ denote the image of $\alpha$. Since $A$ is a closed subset of $X$ (being compact as the image of a compact interval by a continuous map), it is enough to show that every point of $[x, y]$ is within distance $\epsilon$ of $A$, for all $\epsilon>0$.

Given $\epsilon>0$, the collection $\left\{\alpha^{-1}(B(x, \epsilon / 2)): x \in A\right\}$ is an open covering of the compact metric space $[a, b]$, so there is a number $\delta>0$ such that any two points of $[a, b]$ that are distance less than $\delta$ apart belong to some common set in the cover.

Choose a partition of $[a, b]$, say $a=t_{0}<\cdots<t_{n}=b$, so that for $1 \leqslant i \leqslant n$ we have $t_{i}-t_{i-1}<\delta$, and, therefore, $d\left(\alpha\left(t_{i-1}\right), \alpha\left(t_{i}\right)\right)<\epsilon$. Then all points of $\left[\alpha\left(t_{i-1}\right), \alpha\left(t_{i}\right)\right]$ are at distance less than $\epsilon$ from $\left\{\alpha\left(t_{i-1}\right), \alpha\left(t_{i}\right)\right\} \subseteq A$ for $1 \leqslant i \leqslant n$. Finally, $[x, y] \subseteq \bigcup_{i=1}^{n}\left[\alpha\left(t_{i-1}\right), \alpha\left(t_{i}\right)\right]$, by Corollary 3.25 .
Definition 3.27. For points $x_{0}, x_{1}, \ldots, x_{n}$ in a $\mathbb{R}$-tree $(X, d)$, write $\left[x_{0}, x_{n}\right]=$ $\left[x_{0}, x_{1}, \ldots, x_{n}\right]$ to mean that, if $\alpha:\left[0, d\left(x_{0}, x_{n}\right)\right] \rightarrow X$ is the unique isometry with $\alpha(0)=x_{0}$ and $\alpha\left(d\left(x_{0}, x_{n}\right)\right)=x_{n}$, then $x_{i}=\alpha\left(a_{i}\right)$, for some $a_{0}, a_{1}, a_{2}, \ldots, a_{n}$ with $0=a_{0} \leqslant a_{1} \leqslant a_{2} \leqslant \cdots \leqslant a_{n}=d\left(x_{0}, x_{n}\right)$.
Lemma 3.28. Consider a $\mathbb{R}$-tree $(X, d)$. If $x_{0}, \ldots, x_{n} \in X, x_{i} \neq x_{i+1}$ for $1 \leqslant i \leqslant n-2$ and $\left[x_{i-1}, x_{i}\right] \cap\left[x_{i}, x_{i+1}\right]=\left\{x_{i}\right\}$ for $1 \leqslant i \leqslant n-1$, then $\left[x_{0}, x_{n}\right]=\left[x_{0}, x_{1}, \ldots, x_{n}\right]$.
Proof. There is nothing to prove if $n \leqslant 2$. Suppose $n=3$. We can assume $x_{0} \neq x_{1}$ and $x_{2} \neq x_{3}$, otherwise there is again nothing to prove. Let $w=$ $Y\left(x_{0}, x_{2}, x_{3}\right)$.

Now $w \in\left[x_{0}, x_{2}\right]$ and $x_{1} \in\left[x_{0}, x_{2}\right]$, so $\left[x_{2}, w\right] \cap\left[x_{2}, x_{1}\right]=\left[x_{2}, v\right]$, where $v$ is either $w$ or $x_{1}$, depending on which is closer to $x_{2}$. But $\left[x_{2}, w\right] \cap\left[x_{2}, x_{1}\right] \subseteq$ $\left[x_{2}, x_{3}\right] \cap\left[x_{2}, x_{1}\right]=\left\{x_{2}\right\}$, so $v=x_{2}$.

Since $x_{1} \neq x_{2}$, we conclude that $w=x_{2}$. Hence $\left[x_{0}, x_{2}\right] \cap\left[x_{2}, x_{3}\right]=\left\{x_{2}\right\}$, which implies $\left[x_{0}, x_{3}\right]=\left[x_{0}, x_{2}, x_{3}\right]=\left[x_{0}, x_{1}, x_{2}, x_{3}\right]$.

Now suppose $n>3$. By induction,

$$
\left[x_{0}, x_{n-1}\right]=\left[x_{0}, x_{1}, \ldots, x_{n-2}, x_{n-1}\right]=\left[x_{0}, x_{n-2}, x_{n-1}\right] .
$$

By the $n=3$ case,

$$
\left[x_{0}, x_{n}\right]=\left[x_{0}, x_{n-2}, x_{n-1}, x_{n}\right]=\left[x_{0}, x_{1}, \ldots, x_{n-2}, x_{n-1}, x_{n}\right]
$$

as required.

### 3.3.2 $\mathbb{R}$-trees are 0-hyperbolic

Lemma 3.29. $A \mathbb{R}$-tree $(X, d)$ is 0 -hyperbolic.
Proof. Fix $v \in X$. We have to show

$$
\begin{aligned}
& (x \cdot y)_{v} \geqslant(x \cdot z)_{v} \wedge(y \cdot z)_{v} \\
& (x \cdot z)_{v} \geqslant(x \cdot y)_{v} \wedge(y \cdot z)_{v} \\
& (y \cdot z)_{v} \geqslant(x \cdot y)_{v} \wedge(x \cdot z)_{v}
\end{aligned}
$$

for all $x, y, z$. Note that if this is so, then one of $(x \cdot y)_{v},(x \cdot z)_{v},(y \cdot z)_{v}$ is at least as great as the other two, which are equal.

Let $q=Y(x, v, y), r=Y(y, v, z)$, and $s=Y(z, v, x)$. We have $(x \cdot y)_{v}=$ $d(v, q),(y \cdot z)_{v}=d(v, s)$, and $(z \cdot x)_{v}=d(v, r)$. We may assume without loss of generality that

$$
d(v, q) \leqslant d(v, r) \leqslant d(v, s)
$$

in which case have to show that $q=r$ - see Figure 3.9.


Fig. 3.9. The configuration demonstrated in the proof of Lemma 3.29

Now $r, s \in[v, z]$ by definition, and $d(v, r) \leqslant d(v, s)$, so that $[v, s]=[v, r, s]$. Also, by definition of $s,[v, x]=[v, s, x]=[v, r, s, x]$. Hence $r \in[v, x] \cap[v, y]=$ $[v, q]$. Since $d(v, q) \leqslant d(v, r)$, we have $q=r$, as required.

Remark 3.30. Because any subspace of a 0-hyperbolic space is still 0hyperbolic, we can't expect that the converse to Lemma 3.29 holds. However, we will see in Theorem 3.38 that any 0-hyperbolic space is isometric to a subspace of a $\mathbb{R}$-tree.

### 3.3.3 Centroids in a 0-hyperbolic space

Definition 3.31. A set $\{a, b, c\} \subset \mathbb{R}$ is called an isosceles triple if

$$
a \geqslant b \wedge c, b \geqslant c \wedge a, \text { and } c \geqslant a \wedge b .
$$

(This means that at least two of $a, b, c$ are equal, and not greater than the third.)

Remark 3.32. The metric space $(X, d)$ is 0 -hyperbolic if and only if $(x \cdot y)_{v},(x$. $z)_{v},(y \cdot z)_{v}$ is an isosceles triple for all $x, y, z, v \in X$.

Lemma 3.33. (i) If $\{a, b, c\}$ is any triple then

$$
\{a \wedge b, b \wedge c, c \wedge a\}
$$

is an isosceles triple.
(ii) If $\{a, b, c\}$ and $\{d, e, f\}$ are isosceles triples then so is

$$
\{a \wedge d, b \wedge e, c \wedge f\}
$$

Lemma 3.34. Consider a 0-hyperbolic metric space $(X, d)$. Let $\sigma, \tau$ be segments in $X$ with endpoints $v, x$ and $v, y$ respectively. Write $x \cdot y:=(x \cdot y)_{v}$.
(i) If $x^{\prime} \in \sigma$, then $x^{\prime} \in \tau$ if and only if $d\left(v, x^{\prime}\right) \leqslant x \cdot y$.
(ii) If $w$ is the point of $\sigma$ at distance $x \cdot y$ from $v$, then $\sigma \cap \tau$ is a segment with endpoints $v$ and $w$.

Proof. If $d\left(x^{\prime}, v\right)>d(y, v)$ then $x^{\prime} \notin \tau$, and $d\left(x^{\prime}, v\right)>x \cdot y$, so we can assume that $d\left(x^{\prime}, v\right)<d(y, v)$. Let $y^{\prime}$ be the point in $\tau$ such that $d\left(v, x^{\prime}\right)=d\left(v, y^{\prime}\right)$. Define

$$
\alpha=x \cdot y, \beta=x^{\prime} \cdot y, \gamma=x \cdot x^{\prime}, \alpha^{\prime}=x^{\prime} \cdot y^{\prime}
$$

Since $x^{\prime} \in \sigma$ and $y^{\prime} \in \tau$, we have $\gamma=d\left(v, x^{\prime}\right)=d\left(v, y^{\prime}\right)=y \cdot y^{\prime}$. Hence, $(\alpha, \beta, \gamma)$ and $\left(\alpha^{\prime}, \beta, \gamma\right)$ are isosceles triples. We have to show that $x^{\prime} \in \tau$ if and only if $\alpha \geqslant \gamma$. The two cases $\alpha<\gamma$ and $\alpha \geqslant \gamma$ are illustrated in Figure 3.10 and Figure 3.11 respectively.

Now,

$$
\beta=x^{\prime} \cdot y \leqslant d\left(v, x^{\prime}\right)=x \cdot x^{\prime}=\gamma
$$

Also,

$$
\alpha^{\prime}=d\left(v, x^{\prime}\right)-\frac{1}{2} d\left(x^{\prime}, y^{\prime}\right)=\gamma-\frac{1}{2} d\left(x^{\prime}, y^{\prime}\right) \leqslant \gamma
$$

and


Fig. 3.10. First case of the construction in the proof of Lemma 3.34. Here $\gamma$ is either of the two equal dashed lengths and $\alpha=\beta=\alpha^{\prime}$ is the dotted length. As claimed, $\alpha<\gamma$ and $x^{\prime} \notin \tau$.

$$
x^{\prime} \in \tau \Leftrightarrow x^{\prime}=y^{\prime} \Leftrightarrow d\left(x^{\prime}, y^{\prime}\right)=0 \Leftrightarrow \alpha^{\prime}=\gamma .
$$

Moreover, $\alpha^{\prime}=\gamma$ if and only if $\beta=\gamma$, because $\left(\alpha^{\prime}, \beta, \gamma\right)$ is an isosceles triple and $\alpha^{\prime}, \beta \leqslant \gamma$. Since $(\alpha, \beta, \gamma)$ is also an isosceles triple, the equality $\beta=\gamma$ is equivalent to the inequality $\alpha \geqslant \gamma$. This proves part (i). Part (ii) of the lemma follows immediately.

Lemma 3.35. Consider a 0-hyperbolic metric space $(X, d)$. Let $\sigma, \tau$ be segments in $X$ with endpoints $v, x$ and $v, y$ respectively. Set $x \cdot y:=(x \cdot y)_{v}$. Write $w$ for the point of $\sigma$ at distance $x \cdot y$ from $v$ (so that $w$ is an endpoint of $\sigma \cap \tau$ by Lemma 3.34). Consider two points $x^{\prime} \in \sigma, y^{\prime} \in \tau$, and suppose $d\left(x^{\prime}, v\right) \geqslant x \cdot y$ and $d\left(y^{\prime}, v\right) \geqslant x \cdot y$. Then

$$
d\left(x^{\prime}, y^{\prime}\right)=d\left(x^{\prime}, w\right)+d\left(y^{\prime}, w\right)
$$

Proof. The conclusion is clear if $d\left(x^{\prime}, v\right)=x \cdot y$ (when $\left.x^{\prime}=w\right)$ or $d\left(y^{\prime}, v\right)=x \cdot y$ (when $y^{\prime}=w$ ), so we assume that $d\left(x^{\prime}, v\right)>x \cdot y$ and $d\left(y^{\prime}, v\right)>x \cdot y$. As in the proof of Lemma 3.34, we put

$$
\alpha=x \cdot y, \beta=x^{\prime} \cdot y, \gamma=x \cdot x^{\prime}, \alpha^{\prime}=x^{\prime} \cdot y^{\prime}
$$



Fig. 3.11. Second case of the construction in the proof of Lemma 3.34. Here $\gamma=$ $\beta=\alpha^{\prime}$ is the dashed length and $\alpha$ is the dotted length. As claimed, $\alpha \geqslant \gamma$ and $x^{\prime} \in \tau$.
and we also put $\gamma^{\prime}=y \cdot y^{\prime}$, so that $\gamma=d\left(v, x^{\prime}\right)$ and $\gamma^{\prime}=d\left(v, y^{\prime}\right)$. Thus, $\alpha<\gamma$. Hence, $\alpha=\beta$ since $(\alpha, \beta, \gamma)$ is an isosceles triple. Also, $\alpha<\gamma^{\prime}$, so that $\beta<\gamma^{\prime}$. Hence, $\alpha=\alpha^{\prime}=\beta$ because $\left(\alpha^{\prime}, \beta, \gamma\right)$ is an isosceles triple.

By definition of $\alpha^{\prime}$,

$$
\begin{aligned}
d\left(x^{\prime}, y^{\prime}\right) & =d\left(v, x^{\prime}\right)+d\left(v, y^{\prime}\right)-2 \alpha^{\prime} \\
& =d\left(v, x^{\prime}\right)+d\left(v, y^{\prime}\right)-2 \alpha .
\end{aligned}
$$

Since $w \in \sigma \cap \tau, \alpha=d(v, w)<d\left(v, x^{\prime}\right), d\left(v, y^{\prime}\right)$ and $\sigma, \tau$ are segments, it follows that

$$
d\left(x^{\prime}, w\right)=d\left(v, x^{\prime}\right)-\alpha
$$

and

$$
d\left(y^{\prime}, w\right)=d\left(v, y^{\prime}\right)-\alpha
$$

and the lemma follows on adding these equations.

### 3.3.4 An alternative characterization of $\mathbb{R}$-trees

Lemma 3.36. Consider a 0 -hyperbolic metric space ( $X, d$ ). Suppose that there is a point $v \in X$ such that for every $x \in X$ there is a segment with endpoints $v, x$. Then $(X, d)$ is a $\mathbb{R}$-tree.

Proof. Take $x, y \in X$ and let $\sigma, \tau$ be segments with endpoints $v, x$ and $v, y$ respectively.

By Lemma 3.34, if $w$ is the point of $\sigma \cap \tau$ at distance $(x \cdot y)_{v}$ from $v$, then $\sigma$ is the union $(\sigma \cap \tau) \cup \sigma_{1}$, where

$$
\sigma_{1}:=\left\{u \in \sigma: d(v, u) \geqslant(x \cdot y)_{v}\right\}
$$

is a segment with endpoints $w, x$. Similarly, $\tau$ is the union $(\sigma \cap \tau) \cup \tau_{1}$, where

$$
\tau_{1}:=\left\{u \in \tau: d(v, u) \geqslant(x \cdot y)_{v}\right\}
$$

is a segment with endpoints $w, y$.
By Lemma 3.35 and Lemma 3.4, $\sigma_{1} \cup \tau_{1}$ is a segment with endpoints $x, y$. Thus, $(X, d)$ is geodesic.

Note that by Lemma 3.34, $\sigma \cap \tau$ is a segment with endpoints $v, w$. Also, by Lemma 3.34, if $\sigma \cap \tau=\{w\}$ then $(x \cdot y)_{v}=0$ and $\sigma_{1}=\sigma, \tau_{1}=\tau$. Hence, $\sigma \cup \tau$ is a segment. Now, by Lemma 3.10, we may replace $v$ in this argument by any other point of $X$. Hence, $(X, d)$ satisfies the axioms for a $\mathbb{R}$-tree.

Example 3.37. We noted in Example 3.14 that the compact metric space $\left(T_{e}, d_{T_{e}}\right)$ that arises from an excursion path $e \in U$ is 0 -hyperbolic. We can use Lemma 3.36 to show that $\left(T_{e}, d_{T_{e}}\right)$ is a $\mathbb{R}$-tree. Suppose that $e \in U^{\ell}$. Take $x \in T_{e}$ and write $t$ for a point in $[0, \ell]$ such that $x$ is the image of $t$ under the quotient map from $[0, \ell]$ onto $T_{e}$. Write $v \in T_{e}$ for the image of $0 \in[0, \ell]$ under the quotient map from $[0, \ell]$ onto $T_{e}$. Note that $v$ is also the image of $\ell \in[0, \ell]$. For $h \in[0, e(t)]$, set $\lambda_{h}:=\sup \{s \in[0, t]: e(s)=h\}$. Then the image of the set $\left\{\lambda_{h}: h \in[0, e(t)]\right\} \subseteq[0, \ell]$ under the quotient map is a segment in $T_{e}$ that has endpoints $v$ and $x$.

### 3.3.5 Embedding 0-hyperbolic spaces in $\mathbb{R}$-trees

Theorem 3.38. Let $(X, d)$ be a 0-hyperbolic metric space. There exists a $\mathbb{R}$ tree $\left(X^{\prime}, d^{\prime}\right)$ and an isometry $\phi: X \rightarrow X^{\prime}$.

Proof. Fix $v \in X$. Write $x \cdot y:=(x \cdot y)_{v}$ for $x, y \in X$. Let

$$
Y=\{(x, m): x \in X, m \in \mathbb{R} \text { and } 0 \leqslant m \leqslant d(v, x)\}
$$

Define, for $(x, m),(y, n) \in Y$,

$$
(x, m) \sim(y, n) \text { if and only if } x \cdot y \geqslant m=n
$$

This is an equivalence relation on $Y$. Let $X^{\prime}=Y / \sim$, and let $\langle x, m\rangle$ denote the equivalence class of $(x, m)$. We define the metric by

$$
d^{\prime}(\langle x, m\rangle,\langle y, n\rangle)=m+n-2[m \wedge n \wedge(x \cdot y)]
$$

The construction is illustrated in Figure 3.12.


Fig. 3.12. The embedding of Theorem 3.38. Solid lines represent points that are in $X$, while dashed lines represent points that are added to form $X^{\prime}$.

It follows by assumption that $d^{\prime}$ is well defined. Note that

$$
d^{\prime}(\langle x, m\rangle,\langle x, n\rangle)=|m-n|
$$

and $\langle x, 0\rangle=\langle v, 0\rangle$ for all $x \in X$, so $d^{\prime}(\langle x, m\rangle,\langle v, 0\rangle)=m$. Clearly $d^{\prime}$ is symmetric, and it is easy to see that $d^{\prime}(\langle x, m\rangle,\langle y, n\rangle)=0$ if and only if $\langle x, m\rangle=\langle y, n\rangle$. Also, in $X^{\prime}$,

$$
(\langle x, m\rangle \cdot\langle y, n\rangle)_{\langle v, 0\rangle}=m \wedge n \wedge(x \cdot y) .
$$

If $\langle x, m\rangle,\langle y, n\rangle$ and $\langle z, p\rangle$ are three points of $X^{\prime}$, then

$$
\{m \wedge n, n \wedge p, p \wedge m\}
$$

is an isosceles triple by Lemma 3.33(1). Hence, by Lemma 3.33(2), so is $\{m \wedge$ $n \wedge(x \cdot y), n \wedge p \wedge(y \cdot z), p \wedge m \wedge(z \cdot x)\}$. It follows that $\left(X^{\prime}, d^{\prime}\right)$ is a 0 -hyperbolic metric space.

If $\langle x, m\rangle \in X^{\prime}$, then the mapping $\alpha:[0, m] \rightarrow X^{\prime}$ given by $\alpha(n)=\langle x, n\rangle$ is an isometry, so the image of $\alpha$ is a segment with endpoints $\langle v, 0\rangle$ and $\langle x, m\rangle$. It now follows from Lemma 3.36 that $\left(X^{\prime}, d^{\prime}\right)$ is a $\mathbb{R}$-tree. Further, the mapping $\phi: X \rightarrow X^{\prime}$ defined by $\phi(x)=\langle x, d(v, x)\rangle$ is easily seen to be an isometry. $\quad \square$

### 3.3.6 Yet another characterization of $\mathbb{R}$-trees

Lemma 3.39. Let $(X, d)$ be a $\mathbb{R}$-tree. Fix $v \in X$.
(i) For $x, y \in X \backslash\{v\},[v, x] \cap[v, y] \neq\{v\}$ if and only if $x, y$ are in the same path component of $X \backslash\{v\}$.
(ii) The space $X \backslash\{v\}$ is locally path connected, the components of $X \backslash\{v\}$ coincide with its path components, and they are open sets in $X$.

Proof. (i) Suppose that $[v, x] \cap[v, y] \neq\{v\}$. It can't be that $v \in[x, y]$, because that would imply $[x, v] \cap[v, y]=\{v\}$. Thus, $[x, y] \subseteq X \backslash\{v\}$ and $x, y$ are in the same path component of $X \backslash\{v\}$. Conversely, if $\alpha:[a, b] \rightarrow X \backslash\{v\}$ is a continuous map, with $x=\alpha(a), y=\alpha(b)$, then $[a, b]$ is a subset of the image of $\alpha$ by Lemma 3.26, so $v \notin[x, y]$, and $[v, x] \cap[v, y] \neq\{v\}$ by Axiom (b') for a $\mathbb{R}$-tree.
(ii) For $x \in X \backslash\{v\}$, the set $U:=\{y \in X: d(x, y)<d(x, v)\}$ is an open set in $X, U \subseteq X \backslash\{v\}, x \in U$, and $U$ is path connected. For if $y, z \in U$, then $[x, y] \cup[x, z] \subseteq U$, and so $[y, z] \subseteq U$ by Corollary 3.25 . Thus, $X \backslash\{v\}$ is locally path connected. It follows that the path components of $X \backslash\{v\}$ are both open and closed, and (ii) follows easily.

Theorem 3.40. A metric space $(X, d)$ is a $\mathbb{R}$-tree if and only if it is connected and 0-hyperbolic.

Proof. An $\mathbb{R}$-tree is geodesic, so it is path connected. Hence, it is connected. Therefore, it is 0-hyperbolic by Lemma 3.29.

Conversely, assume that a metric space $(X, d)$ is connected and 0 hyperbolic. By Theorem 3.38 there is an embedding of $(X, d)$ in a $\mathbb{R}$-tree $\left(X^{\prime}, d^{\prime}\right)$. Let $x, y \in X$, suppose $v \in X^{\prime} \backslash X$ and $v \in[x, y]$. Then $[v, x] \cap[v, y]=$ $\{v\}$ and so by Lemma 3.39, $x, y$ are in different components of $X \backslash\{v\}$.

Let $C$ be the component of $X \backslash\{v\}$ containing $x$. By Lemma 3.39, $C$ is open and closed, so $X \cap C$ is open and closed in $X$. Since $x \in X \cap C, y \notin X \cap C$, this contradicts the connectedness of $X$. Thus, $[x, y] \subseteq X$ and $(X, d)$ is geodesic. It follows that $(X, d)$ is a $\mathbb{R}$-tree by Lemma 3.36.

Example 3.41. Let $\mathcal{P}$ denote the collection of partitions of the positive integers $\mathbb{N}$. There is a natural partial order $\leqslant$ on $\mathcal{P}$ defined by $P \leqslant Q$ if every block of $Q$ is a subset of some block of $P$ (that is, the blocks of $P$ are unions of
blocks of $Q)$. Thus, the partition $\{\{1\},\{2\}, \ldots\}$ consisting of singletons is the unique largest element of $\mathcal{P}$, while the partition $\{\{1,2, \ldots\}\}$ consisting of a single block is the unique smallest element. Consider a function $\Pi: \mathbb{R}_{+} \mapsto \mathcal{P}$ that is non-increasing in this partial order. Suppose that $\Pi(0)=\{\{1\},\{2\}, \ldots\}$ and $\Pi(t)=\{\{1,2, \ldots\}\}$ for all $t$ sufficiently large. Suppose also that if $\Pi$ is right-continuous in the sense that if $i$ and $j$ don't belong to the same block of $\Pi(t)$ for some $t \in \mathbb{R}_{+}$, then they don't belong to the same block of $\Pi(u)$ for $u>t$ sufficiently close to $t$.

Let $T$ denote the set consisting of points of the form $(t, B)$, where $t \in \mathbb{R}_{+}$ and $B \in \Pi(t)$. Given two point $(s, A),(t, B) \in T$, set

$$
\begin{aligned}
& m((s, A),(t, B)) \\
& \quad:=\inf \{u>s \wedge t: A \text { and } B \text { subsets of a common block of } \Pi(u)\}
\end{aligned}
$$

and put

$$
d((s, A),(t, B)):=[m((s, A),(t, B))-s]+[m((s, A),(t, B))-t]
$$

It is not difficult to check that $d$ is a metric that satisfies the four point condition and that the space $T$ is connected. Hence, $(T, d)$ is a $\mathbb{R}$-tree by Theorem 3.40. The analogue of this construction with $\mathbb{N}$ replaced by $\{1,2,3,4\}$ is shown in Figure 3.13.

Moreover, if we let $\bar{T}$ denote the completion of $T$ with respect to the metric $d$, then $\bar{T}$ is also a $\mathbb{R}$-tree. It is straightforward to check that $\bar{T}$ is compact if and only if $\Pi(t)$ has finitely many blocks for all $t>0$.

Write $\delta$ for the restriction of $d$ to the positive integers $\mathbb{N}$, so that

$$
\delta(i, j)=2 \inf \{t>0: i \text { and } j \text { belong to the same block of } \Pi(t)\}
$$

The completion $\mathbb{S}$ of $\mathbb{N}$ with respect to $\delta$ is isometric to the closure of $\mathbb{N}$ in $\bar{T}$, and $\mathbb{S}$ is compact if and only if $\Pi(t)$ has finitely many blocks for all $t>0$. Note that $\delta$ is an ultrametric, that is, $\delta(x, y) \leqslant \delta(x, z) \vee \delta(z, y)$ for $x, y, z \in \mathbb{S}$. This implies that at least two of the distances are equal and are no smaller than the third. Hence, all triangles are isosceles. When $\mathbb{S}$ is compact, the open balls for the metric $\delta$ coincide with the closed balls and are obtained by taking the closure of the blocks of $\Pi(t)$ for $t>0$. In particular, $\mathbb{S}$ is totally disconnected

The correspondence between coalescing partitions, tree structures and ultrametrics is a familiar idea in the physics literature - see, for example, [109].

## $3.4 \mathbb{R}$-trees without leaves

### 3.4.1 Ends

Definition 3.42. An $\mathbb{R}$-tree without leaves is a $\mathbb{R}$-trees $(T, d)$ that satisfies the following extra axioms.


Fig. 3.13. The construction of a $\mathbb{R}$-tree from a non-increasing function taking values in the partitions of $\{1,2,3,4\}$.

Axiom (c) The metric space $(T, d)$ is complete.
Axiom (d) For each $x \in T$ there is at least one isometric embedding $\theta: \mathbb{R} \rightarrow T$ with $x \in \theta(\mathbb{R})$.

Example 3.43. "The" real tree $(\mathcal{T}, D)$ of Example 3.18 satisfies Axioms (c) and (d).

We will suppose in this section that we are always working with a $\mathbb{R}$-tree $(T, d)$ that is without leaves.

Definition 3.44. An end of $T$ is an equivalence class of isometric embeddings from $\mathbb{R}_{+}$into $T$, where we regard two such embeddings $\phi$ and $\psi$ as being equivalent if there exist $\alpha \in \mathbb{R}$ and $\beta \in \mathbb{R}_{+}$such that $\alpha+\beta \geqslant 0$ and $\phi(t)=$ $\psi(t+\alpha)$ for all $t \geqslant \beta$. Write $E$ for the set of ends of $T$.

By Axiom (d), $E$ has at least 2 points. Fix a distinguished element $\dagger$ of $E$. For each $x \in T$ there is a unique isometric embedding $\kappa_{x}: \mathbb{R}_{+} \rightarrow T$ such that $\kappa_{x}(0)=x$ and $\kappa_{x}$ is a representative of the equivalence class of $\dagger$. Similarly, for each $\xi \in E_{+}:=E \backslash\{\dagger\}$ there is at least one isometric embedding $\theta: \mathbb{R} \rightarrow T$ such that $t \mapsto \theta(t), t \geqslant 0$, is a representative of the equivalence class of $\xi$
and $t \mapsto \theta(-t), t \geqslant 0$, is a representative of the equivalence class of $\dagger$. Denote the collection of all such embeddings by $\Theta_{\xi}$. If $\theta, \theta^{\prime} \in \Theta_{\xi}$, then there exists $\gamma \in \mathbb{R}$ such that $\theta(t)=\theta^{\prime}(t+\gamma)$ for all $t \in \mathbb{R}$. Thus, it is possible to select an embedding $\theta_{\xi} \in \Theta_{\xi}$ for each $\xi \in E_{+}$in such a way that for any pair $\xi, \zeta \in E_{+}$ there exists $t_{0}$ (depending on $\xi, \zeta$ ) such that $\theta_{\xi}(t)=\theta_{\zeta}(t)$ for all $t \leqslant t_{0}$ (and $\left.\theta_{\xi}(] t_{0}, \infty[) \cap \theta_{\zeta}(] t_{0}, \infty[)=\varnothing\right)$. Extend $\theta_{\xi}$ to $\mathbb{R}^{*}:=\mathbb{R} \cup\{ \pm \infty\}$ by setting $\theta_{\xi}(-\infty):=\dagger$ and $\theta_{\xi}(+\infty):=\xi$.

Example 3.45. The ends of the real tree $(\mathcal{T}, D)$ of Example 3.18 can be identified with the collection consisting of the empty set and the elements of $\mathcal{E}_{+}$, where $\mathcal{E}_{+}$consists of subsets $B \subset \mathbb{R}$ such that $-\infty<\inf B$ and $\sup B=+\infty$. If we choose $\dagger$ to be the empty set so that $\mathcal{E}_{+}$plays the role of $E_{+}$, then we can define the isometric embedding $\theta_{A}$ for $A \in \mathcal{E}_{+}$by $\left.\left.\theta_{A}(t):=(A \cap]-\infty, t\right]\right) \cup\{t\}=A \mid t$, in the notation of Example 3.18.

The $\operatorname{map}(t, \xi) \mapsto \theta_{\xi}(t)$ from $\mathbb{R} \times E_{+}\left(\right.$resp. $\left.\mathbb{R}^{*} \times E_{+}\right)$into $T($ resp. $T \cup E)$ is surjective. Moreover, if $\eta \in T \cup E$ is in $\theta_{\xi}\left(\mathbb{R}^{*}\right) \cap \theta_{\zeta}\left(\mathbb{R}^{*}\right)$ for $\xi, \zeta \in E_{+}$, then $\theta_{\xi}^{-1}(\eta)=\theta_{\zeta}^{-1}(\eta)$. Denote this common value by $h(\eta)$, the height of $\eta$. In genealogical terminology, we think of $h(\eta)$ as the generation to which $\eta$ belongs. In particular, $h(\dagger):=-\infty$ and $h(\xi)=+\infty$ for $\xi \in E_{+}$. For the real tree $(\mathcal{T}, D)$ of Example 3.18 with corresponding isometric embeddings defined as above, $h(B)$ is just $\sup B$, with the usual convention that $\sup \varnothing:=-\infty$ (in accord with the notation of Example 3.18).

Define a partial order $\leqslant$ on $T \cup E$ by declaring that $\eta \leqslant \rho$ if there exists $-\infty \leqslant s \leqslant t \leqslant+\infty$ and $\xi \in E_{+}$such that $\eta=\theta_{\xi}(s)$ and $\rho=\theta_{\xi}(t)$. In genealogical terminology, $\eta \leqslant \rho$ corresponds to $\eta$ being an ancestor of $\rho$ (note that individuals are their own ancestors). In particular, $\dagger$ is the unique point that is an ancestor of everybody, while points of $E_{+}$are characterized by being only ancestors of themselves. For the real tree $(\mathcal{T}, D)$ of Example 3.18, $A \leqslant B$ if and only if $A=(B \cap]-\infty, \sup A]) \cup\{\sup A\}$. In particular, this partial order is not the usual inclusion partial order (for example, the singleton $\{0\}$ is an ancestor of the singleton $\{1\})$.

Each pair $\eta, \rho \in T \cup E$ has a well-defined greatest common lower bound $\eta \wedge \rho$ in this partial order, with $\eta \wedge \rho \in T$ unless $\eta=\rho \in E_{+}, \eta=\dagger$ or $\rho=\dagger$. In genealogical terminology, $\eta \wedge \rho$ is the most recent common ancestor of $\eta$ and $\rho$. For $x, y \in T$ we have

$$
\begin{align*}
d(x, y) & =h(x)+h(y)-2 h(x \wedge y) \\
& =[h(x)-h(x \wedge y)]+[h(y)-h(x \wedge y)] \tag{3.1}
\end{align*}
$$

Therefore, $h(x)=d(x, y)-h(y)+2 h(x \wedge y) \leqslant d(x, y)+h(y)$ and, similarly, $h(y) \leqslant d(x, y)+h(x)$, so that

$$
\begin{equation*}
|h(x)-h(y)| \leqslant d(x, y) \tag{3.2}
\end{equation*}
$$

with equality if $x, y \in T$ are comparable in the partial order (that is, if $x \leqslant y$ or $y \leqslant x$ ).

If $x, x^{\prime} \in T$ are such that $h(x \wedge y)=h\left(x^{\prime} \wedge y\right)$ for all $y \in T$, then, by (3.1), $d\left(x, x^{\prime}\right)=\left[h(x)-h\left(x \wedge x^{\prime}\right)\right]+\left[h\left(x^{\prime}\right)-h\left(x \wedge x^{\prime}\right)\right]=[h(x)-h(x \wedge x)]+\left[h\left(x^{\prime}\right)-\right.$ $\left.h\left(x^{\prime} \wedge x^{\prime}\right)\right]=0$, so that $x=x^{\prime}$. Slight elaborations of this argument show that if $\eta, \eta^{\prime} \in T \cup E$ are such that $h(\eta \wedge y)=h\left(\eta^{\prime} \wedge y\right)$ for all $y$ in some dense subset of $T$, then $\eta=\eta^{\prime}$.

For $x, x^{\prime}, z \in T$ we have that if $h(x \wedge z)<h\left(x^{\prime} \wedge z\right)$, then $x \wedge x^{\prime}=x \wedge z$ and a similar conclusion holds with the roles of $x$ and $x^{\prime}$ reversed; whereas if $h(x \wedge z)=h\left(x^{\prime} \wedge z\right)$, then $x \wedge z=x^{\prime} \wedge z \leqslant x \wedge x^{\prime}$. Using (3.1) and (3.2) and checking the various cases we find that

$$
\begin{equation*}
\left|h(x \wedge z)-h\left(x^{\prime} \wedge z\right)\right| \leqslant d\left(x \wedge z, x^{\prime} \wedge z\right) \leqslant d\left(x, x^{\prime}\right) \tag{3.3}
\end{equation*}
$$

For $\eta \in T \cup E$ and $t \in \mathbb{R}^{*}$ with $t \leqslant h(\eta)$, let $\eta \mid t$ denote the unique $\rho \in T \cup E$ with $\rho \leqslant \eta$ and $h(\rho)=t$. Equivalently, if $\eta=\theta_{\xi}(u)$ for some $u \in \mathbb{R}^{*}$ and $\xi \in E_{+}$, then $\eta \mid t=\theta_{\xi}(t)$ for $t \leqslant u$. For the real tree of Example 3.18, this definition coincides with the one given in Example 3.18.

The metric space $\left(E_{+}, \delta\right)$, where

$$
\delta(\xi, \zeta):=2^{-h(\xi \wedge \zeta)}
$$

is complete. Moreover, the metric $\delta$ is actually an ultrametric ; that is, $\delta(\xi, \zeta) \leqslant$ $\delta(\xi, \eta) \vee \delta(\eta, \zeta)$ for all $\xi, \zeta, \eta \in E_{+}$.

### 3.4.2 The ends compactification

Suppose in this subsection that the metric space $\left(E_{+}, \delta\right)$ is separable. For $t \in \mathbb{R}$ consider the set

$$
\begin{equation*}
T_{t}:=\{x \in T: h(x)=t\}=\left\{\xi \mid t: \xi \in E_{+}\right\} \tag{3.4}
\end{equation*}
$$

of points in $T$ that have height $t$. For each $x \in T_{t}$ the set $\left\{\zeta \in E_{+}: \zeta \mid t=x\right\}$ is a ball in $E_{+}$of diameter at most $2^{-t}$ and two such balls are disjoint. Thus, the separability of $E_{+}$is equivalent to each of the sets $T_{t}$ being countable. In particular, separability of $E_{+}$implies that $T$ is also separable, with countable dense set $\left\{\xi \mid t: \xi \in E_{+}, t \in \mathbb{Q}\right\}$, say.

We can, via a standard Stone-Cech-like procedure, embed $T \cup E$ in a compact metric space in such a way that for each $y \in T \cup E$ the map $x \mapsto$ $h(x \wedge y)$ has a continuous extension to the compactification (as an extended real-valued function).

More specifically, let $S$ be a countable dense subset of $T$. Let $\pi$ be a strictly increasing, continuous function that maps $\mathbb{R}$ onto $] 0,1[$. Define an injective map $\Pi$ from $T$ into the compact, metrizable space $[0,1]^{S}$ by $\Pi(x):=(\pi(h(x \wedge$ $y))_{y \in S}$. Identify $T$ with $\Pi(T)$ and write $\bar{T}$ for the closure of $T(=\Pi(T))$ in $[0,1]^{T}$. In other words, a sequence $\left\{x_{n}\right\}_{n \in \mathbb{N}} \subset T$ converges to a point in $\bar{T}$ if $h\left(x_{n} \wedge y\right)$ converges (possibly to $-\infty$ ) for all $y \in S$, and two such sequences $\left\{x_{n}\right\}_{n \in \mathbb{N}}$ and $\left\{x_{n}^{\prime}\right\}_{n \in \mathbb{N}}$ converge to the same point if and only if $\lim _{n} h\left(x_{n} \wedge y\right)=\lim _{n} h\left(x_{n}^{\prime} \wedge y\right)$ for all $y \in S$.

We can identify distinct points in $T \cup E$ with distinct points in $\bar{T}$. If $\left\{x_{n}\right\}_{n \in \mathbb{N}} \subset T$ and $\xi \in E_{+}$are such that for all $t \in \mathbb{R}$ we have $\xi \mid t \leqslant x_{n}$ for all sufficiently large $n$, then $\lim _{n} h\left(x_{n} \wedge y\right)=h(\xi \wedge y)$ for all $y \in S$. We leave the identification of $\dagger$ to the reader.

In fact, we have $\bar{T}=T \cup E$. To see this, suppose that $\left\{x_{n}\right\}_{n \in \mathbb{N}} \subset T$ converges to $x_{\infty} \in \bar{T}$. Put $h_{\infty}:=\sup _{y \in S} \lim _{n} h\left(x_{n} \wedge y\right)$. Assume for the moment that $h_{\infty} \in \mathbb{R}$. We will show that $x_{\infty} \in T$ with $h\left(x_{\infty}\right)=h_{\infty}$. For all $k \in \mathbb{N}$ we can find $y_{k} \in S$ such that

$$
h_{\infty}-\frac{1}{k} \leqslant \lim _{n} h\left(x_{n} \wedge y_{k}\right) \leqslant h\left(y_{k}\right) \leqslant h_{\infty}+\frac{1}{k} .
$$

Observe that

$$
\begin{aligned}
d\left(y_{k}, y_{\ell}\right) \leqslant & \limsup _{n}\left(d\left(y_{k}, x_{n} \wedge y_{k}\right)+d\left(x_{n} \wedge y_{k}, x_{n} \wedge y_{\ell}\right)\right. \\
& \left.+d\left(x_{n} \wedge y_{\ell}, y_{\ell}\right)\right) \\
= & \limsup _{n}\left(\left[h\left(y_{k}\right)-h\left(x_{n} \wedge y_{k}\right)\right]+\left|h\left(x_{n} \wedge y_{k}\right)-h\left(x_{n} \wedge y_{\ell}\right)\right|\right. \\
& \left.+\left[h\left(y_{\ell}\right)-h\left(x_{n} \wedge y_{\ell}\right)\right]\right) \\
\leqslant & \frac{2}{k}+\left(\frac{1}{k}+\frac{1}{\ell}\right)+\frac{2}{\ell}
\end{aligned}
$$

Therefore, $\left(y_{k}\right)_{k \in \mathbb{N}}$ is a $d$-Cauchy sequence and, by Axiom (c), this sequence converges to $y_{\infty} \in T$. Moreover, by (3.2) and (3.3), $\lim _{n} h\left(x_{n} \wedge y_{\infty}\right)=h\left(y_{\infty}\right)=$ $h_{\infty}$.

We claim that $y_{\infty}=x_{\infty}$; that is, $\lim _{n} h\left(x_{n} \wedge z\right)=h\left(y_{\infty} \wedge z\right)$ for all $z \in S$. To see this, fix $z \in T$ and $\epsilon>0$. If $n$ is sufficiently large, then

$$
\begin{equation*}
h\left(x_{n} \wedge z\right) \leqslant h\left(y_{\infty}\right)+\epsilon \tag{3.5}
\end{equation*}
$$

and

$$
\begin{equation*}
h\left(y_{\infty}\right)-\epsilon \leqslant h\left(x_{n} \wedge y_{\infty}\right) \leqslant h\left(y_{\infty}\right) . \tag{3.6}
\end{equation*}
$$

If $h\left(y_{\infty} \wedge z\right) \leqslant h\left(y_{\infty}\right)-\epsilon$, then (3.6) implies that $y_{\infty} \wedge z=x_{n} \wedge z$. On the other hand, if $h\left(y_{\infty} \wedge z\right) \geqslant h\left(y_{\infty}\right)-\epsilon$, then (3.6) implies that

$$
\begin{equation*}
h\left(x_{n} \wedge z\right) \geqslant h\left(y_{\infty}\right)-\epsilon, \tag{3.7}
\end{equation*}
$$

and so, by (3.5) and (3.6),

$$
\begin{align*}
& \left|h\left(y_{\infty} \wedge z\right)-h\left(x_{n}, z\right)\right| \\
& \quad \leqslant\left[h\left(y_{\infty}\right)-\left(h\left(y_{\infty}\right)-\epsilon\right)\right] \vee\left[\left(h\left(y_{\infty}\right)+\epsilon\right)-\left(h\left(y_{\infty}\right)-\epsilon\right)\right]  \tag{3.8}\\
& \quad=2 \epsilon
\end{align*}
$$

We leave the analogous arguments for $h_{\infty}=+\infty$ (in which case $x_{\infty} \in E_{+}$) and $h_{\infty}=-\infty$ (in which case $x_{\infty}=\dagger$ ) to the reader.

We have just seen that the construction of $\bar{T}$ does not depend on $T$ (more precisely, any two such compactifications are homeomorphic). Moreover, a sequence $\left\{x_{n}\right\}_{n \in \mathbb{N}} \subset T \cup E$ converges to a limit in $T \cup E$ if and only if $\lim _{n} h\left(x_{n} \wedge y\right)$ exists for all $y \in T$, and two convergent sequences $\left\{x_{n}\right\}_{n \in \mathbb{N}}$ and $\left\{x_{n}^{\prime}\right\}_{n \in \mathbb{N}}$ converge to the same limit if and only if $\lim _{n} h\left(x_{n} \wedge y\right)=\lim _{n} h\left(x_{n}^{\prime} \wedge\right.$ $y)$ for all $y \in T$.

### 3.4.3 Examples of $\mathbb{R}$-trees without leaves

Fix a prime number $p$ and constants $r_{-}, r_{+} \geqslant 1$. Let $\mathbb{Q}$ denote the rational numbers. Define an equivalence relation $\sim$ on $\mathbb{Q} \times \mathbb{R}$ as follows. Given $a, b \in \mathbb{Q}$ with $a \neq b$ write $a-b=p^{v(a, b)}(m / n)$ for some $v(a, b), m, n \in \mathbb{Z}$ with $m$ and $n$ not divisible by $p$. For $v(a, b) \geqslant 0$ put $w(a, b)=\sum_{i=0}^{v(a, b)} r_{+}^{i}$, and for $v(a, b)<0$ put $w(a, b):=1-\sum_{i=0}^{-v(a, b)} r_{-}^{i}$. Set $w(a, a):=+\infty$. Given $(a, s),(b, t) \in \mathbb{Q} \times \mathbb{R}$ declare that $(a, s) \sim(b, t)$ if and only if $s=t \leqslant w(a, b)$. Note that

$$
\begin{equation*}
v(a, c) \geqslant v(a, b) \wedge v(b, c) \tag{3.9}
\end{equation*}
$$

so that

$$
\begin{equation*}
w(a, c) \geqslant w(a, b) \wedge w(b, c) \tag{3.10}
\end{equation*}
$$

and $\sim$ is certainly transitive (reflexivity and symmetry are obvious).
Let $T$ denote the collection of equivalence classes for this equivalence relation. Define a partial order $\leqslant$ on $T$ as follows. Suppose that $x, y \in T$ are equivalence classes with representatives $(a, s)$ and $(b, t)$. Say that $x \leqslant y$ if and only if $s \leqslant w(a, b) \wedge t$. It follows from (3.10) that $\leqslant$ is indeed a partial order. A pair $x, y \in T$ with representatives $(a, s)$ and $(b, t)$ has a unique greatest common lower bound $x \wedge y$ in this order given by the equivalence class of $(a, s \wedge t \wedge w(a, b))$, which is also the equivalence class of $(b, s \wedge t \wedge w(a, b))$.

For $x \in T$ with representative $(a, s)$, put $h(x):=s$. Define a metric $d$ on $T$ by setting $d(x, y):=h(x)+h(y)-2 h(x \wedge y)$. We leave it to the reader to check that $(T, d)$ is a $\mathbb{R}$-tree satisfying Axioms (a)-(d), and that the definitions of $x \leqslant y, x \wedge y$ and $h(x)$ fit into the general framework of Section 3.4, with the set $E_{+}$corresponding to $\mathbb{Q} \times \mathbb{R}$-valued paths $s \mapsto(a(s), s)$ such that $s \leqslant w(a(s), a(t)) \wedge t$.

Note that there is a natural Abelian group structure on $E_{+}$: if $\xi$ and $\zeta$ correspond to paths $s \mapsto(a(s), s)$ and $s \mapsto(b(s), s)$, then define $\xi+\zeta$ to correspond to the path $s \mapsto(a(s)+b(s), s)$. We mention in passing that there is a bi-continuous group isomorphism between $E_{+}$and the additive group of the $p$-adic integers $\mathbb{Q}_{p}$. (This map is, however, not an isometry if $E_{+}$is equipped with the $\delta$ metric and $\mathbb{Q}_{p}$ is equipped with the usual $p$-adic metric.)

## Hausdorff and Gromov-Hausdorff distance

### 4.1 Hausdorff distance

We follow the presentation in [37] in this section and omit some of the more elementary proofs.

Definition 4.1. Denote by $U_{r}(S)$ the $r$-neighborhood of a set $S$ in a metric space $(X, d)$. That is, $U_{r}(S):=\{x \in X: d(x, S)<r\}$, where $d(x, S):=$ $\inf \{d(x, y): y \in S\}$. Equivalently, $U_{r}(S):=\bigcup_{x \in S} B_{r}(x)$, where $B_{r}(x)$ is the open ball of radius $r$ centered at $x$.

Definition 4.2. Let $A$ and $B$ be subsets of a metric space $(X, d)$. The Hausdorff distance between $A$ and $B$, denoted by $d_{H}(A, B)$, is defined by

$$
d_{H}(A, B):=\inf \left\{r>0: A \subset U_{r}(B) \text { and } B \subset U_{r}(A)\right\}
$$

See Figure 4.1

Proposition 4.3. Let $(X, d)$ be a metric space. Then
(i) $d_{H}$ is a semi-metric on the set of all subsets of $X$.
(ii) $d_{H}(A, \bar{A})=0$ for any $A \subseteq X$, where $\bar{A}$ denotes the closure of $A$.
(iii) If $A$ and $B$ are closed subsets of $X$ and $d_{H}(A, B)=0$, then $A=B$.

Let $\mathfrak{M}(X)$ denote the set of non-empty closed subsets of $X$ equipped with Hausdorff distance. Proposition 4.3 says that $\mathfrak{M}(X)$ is a metric space (provided we allow the metric to take the value $+\infty$ ).

Proposition 4.4. If the metric space $(X, d)$ is complete, then the metric space $\left(\mathfrak{M}(X), d_{H}\right)$ is also complete.

Proof. Consider a Cauchy sequence $\left\{S_{n}\right\}_{n \in \mathbb{N}}$ in $\mathfrak{M}(X)$. Let $S$ denote the set of points $x \in X$ such that any neighborhood of $x$ intersects with infinitely many of the $S_{n}$. That is, $S:=\overline{\bigcap_{m=1}^{\infty} \bigcup_{n=m}^{\infty} S_{n}}$. By definition of the Hausdorff metric,


Fig. 4.1. The Hausdorff distance between the sets $A$ and $B$ is $d$
we can find a sequence $\left\{y_{n}\right\}_{n \in \mathbb{N}}$ such that $y_{n} \in S_{n}$ and $d\left(y_{m}, y_{n}\right) \leqslant d_{H}\left(S_{m}, S_{n}\right)$ for all $m, n \in \mathbb{N}$. Since $X$ is complete, $\lim _{n \rightarrow y_{n}=y \text { exists. Note that } y \in S}$ and so $S$ is non-empty. By definition, $S$ is closed, and so $S \in \mathfrak{M}(X)$.

We will show that

$$
S_{n} \rightarrow S
$$

Fix $\epsilon>0$ and let $n_{0}$ be such that $d_{H}\left(S_{n}, S_{m}\right)<\epsilon$ for all $m, n \geqslant n_{0}$. It suffices to show that $d_{H}\left(S, S_{n}\right)<2 \epsilon$ for any $n \geqslant n_{0}$, and this is equivalent to showing that:

$$
\begin{equation*}
\text { For } x \in S \text { and } n \geqslant n_{0}, d\left(x, S_{n}\right)<2 \epsilon \tag{4.1}
\end{equation*}
$$

For $x \in S_{n}$ and $n \geqslant n_{0}, d(x, S)<2 \epsilon$.
To establish (4.1), note first that there exists an $m \geqslant n_{0}$ such that $B_{\epsilon}(x) \cap$ $S_{m} \neq \varnothing$. In other words, there is a point $y \in S_{m}$ such that $d(x, y)<\epsilon$. Since $d_{H}\left(S_{n}, S_{m}\right)<\epsilon$, we also have $d\left(y, S_{n}\right)<\epsilon$, and, therefore, $d\left(x, S_{n}\right)<2 \epsilon$.

Turning to (4.2), let $n_{1}=n$ and for every integer $k>1$ choose an index $n_{k}$ such that $n_{k}>n_{k+1}$ and $d_{H}\left(S_{p}, S_{q}\right)<\epsilon / 2^{k}$ for all $p, q \geqslant n_{k}$. Define a sequence of points $\left\{x_{k}\right\}_{k \in \mathbb{N}}$, where $x_{k} \in S_{n_{k}}$, as follows: let $x_{1}=x$, and $x_{k+1}$ be a point of $S_{n_{k+1}}$ such that $d\left(x_{k}, x_{k+1}\right)<\epsilon / 2^{k}$ for all $k$. Such a point can be found because $d_{H}\left(S_{n_{k}}, S_{n_{k+1}}\right)<\epsilon / 2^{k}$.

Since $\sum_{k \in \mathbb{N}} d\left(x_{k}, x_{k+1}\right)<2 \epsilon<\infty$, the sequence $\left\{x_{k}\right\}_{k \in \mathbb{N}}$ is a Cauchy sequence. Hence, it converges to a point $y \in X$ by the assumed completeness of $X$. Then,

$$
d(x, y)=\lim _{n \rightarrow \infty} d\left(x, x_{n}\right) \leqslant \sum_{k \in \mathbb{N}} d\left(x_{k}, x_{k+1}\right)<2 \epsilon
$$

Because $y \in S$ by construction, it follows that $d(x, S)<2 \epsilon$.
Theorem 4.5. If the metric space $(X, d)$ is compact, then the metric space $\left(\mathfrak{M}(X), d_{H}\right)$ is also compact.

Proof. By Proposition 4.4, $\mathfrak{M}(X)$ is complete. Therefore, it suffices to prove that $\mathfrak{M}(X)$ is totally bounded. Let $S$ be a finite $\epsilon$-net in $X$. We will show that the set of all non-empty subsets of $S$ is an $\epsilon$-net in $\mathfrak{M}(X)$.

Let $A \in \mathfrak{M}(X)$. Consider

$$
S_{A}=\{x \in S: d(x, A) \leqslant \epsilon\}
$$

Since $S$ is an $\epsilon$-net in $X$, for every $y \in A$ there exists an $x \in S$ such that $d(x, y) \leqslant \epsilon$. Because $d(x, A) \leqslant d(x, y) \leqslant \epsilon$, this point $x$ belongs to $S_{A}$. Therefore, $d\left(y, S_{A}\right) \leqslant \epsilon$ for all $y \in A$.

Since $d(x, A) \leqslant \epsilon$ for any $x \in S_{A}$, it follows that $d_{H}\left(A, S_{A}\right) \leqslant \epsilon$. Since $A$ is arbitrary, this proves that the set of subsets of $S$ is an $\epsilon$-net in $\mathfrak{M}(X)$.

### 4.2 Gromov-Hausdorff distance

In this section we follow the development in [37]. Similar treatments may be found in [80, 34].

### 4.2.1 Definition and elementary properties

Definition 4.6. Let $X$ and $Y$ be metric spaces. The Gromov-Hausdorff distance between them, denoted by $d_{\mathrm{GH}}(X, Y)$, is the infimum of the Hausdorff distances $d_{H}\left(X^{\prime}, Y^{\prime}\right)$ over all metric spaces $Z$ and subspaces $X^{\prime}$ and $Y^{\prime}$ of $Z$ that are isometric to $X$ and $Y$, respectively - see Figure 4.2.

Remark 4.7. It is not necessary to consider all possible embedding spaces $Z$. The Gromov-Hausdorff distance between two metric spaces $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$ is the infimum of those $r>0$ such that there exists a metric $d$ on the disjoint union $X \bigsqcup Y$ such that the restrictions of $d$ to $X$ and $Y$ coincide with $d_{X}$ and $d_{Y}$ and $d_{H}(X, Y)<r$ in the space $(X \bigsqcup Y, d)$.

Proposition 4.8. The distance $d_{\mathrm{GH}}$ satisfies the triangle inequality.
Proof. Given $d_{X Y}$ on $X \bigsqcup Y$ and $d_{Y Z}$ on $Y \bigsqcup Z$, define $d_{X Z}$ on $X \bigsqcup Z$ by

$$
d_{X Z}(x, z)=\inf _{y \in Y}\left\{d_{X Y}(x, y)+d_{Y Z}(y, z)\right\}
$$



Fig. 4.2. Computation of the Gromov-Hausdorff distance between metric spaces $X$ and $Y$ by embedding isometric copies $X^{\prime}$ and $Y^{\prime}$ into $Z$

### 4.2.2 Correspondences and $\epsilon$-isometries

The definition of the Gromov-Hausdorff distance $d_{\mathrm{GH}}(X, Y)$ is somewhat unwieldy, as it involves an infimum over metric spaces $Z$ and isometric embeddings of $X$ and $Y$ in $Z$. Remark 4.7 shows that it is enough to take $Z$ to be the disjoint union of $X$ and $Y$, but this still leaves the problem of finding optimal metrics on the disjoint union that extend the metrics on $X$ and $Y$. In this subsection we will give a more effective formulation of the Gromov-Hausdorff distance, as well as convenient upper and lower bounds on the distance.

Definition 4.9. Let $X$ and $Y$ be two sets. $A$ correspondence between $X$ and $Y$ is a set $\Re \subset X \times Y$ such that for every $x \in X$ there exists at least one $y \in Y$ for which $(x, y) \in \Re$, and similarly for every $y \in Y$ there exists an $x \in X$ for which $(x, y) \in \mathfrak{R}$ - see Figure 4.3.

Definition 4.10. Let $\mathfrak{R}$ be a correspondence between metric spaces $X$ and $Y$. The distortion of $\mathfrak{R}$ is defined to be

$$
\operatorname{dis} \mathfrak{R}:=\sup \left\{\left|d_{X}\left(x, x^{\prime}\right)-d_{Y}\left(y, y^{\prime}\right)\right|:(x, y),\left(x^{\prime}, y^{\prime}\right) \in \mathfrak{R}\right\},
$$

where $d_{X}$ and $d_{Y}$ are the metrics of $X$ and $Y$ respectively.


Fig. 4.3. A correspondence between two spaces

Theorem 4.11. For any two metric spaces $X$ and $Y$,

$$
d_{\mathrm{GH}}(X, Y)=\frac{1}{2} \inf _{\mathfrak{R}}(\operatorname{dis} \mathfrak{R})
$$

where the infimum is taken over all correspondences $\Re$ between $X$ and $Y$.
Proof. We first show for any $r>d_{\mathrm{GH}}(X, Y)$ that there exists a correspondence $\Re$ with dis $\Re<2 r$. Indeed, since $d_{\mathrm{GH}}(X, Y)<r$, we may assume that $X$ and $Y$ are subspaces of some metric space $Z$ and $d_{H}(X, Y)<r$ in $Z$. Define

$$
\mathfrak{R}=\{(x, y): x \in X, y \in Y, d(x, y)<r\}
$$

where $d$ is the metric of $Z$.
That $\Re$ is a correspondence follows from the fact that $d_{H}(X, Y)<r$. The estimate $\operatorname{dis} \mathfrak{R}<2 r$ follows from the triangle inequality: if $(x, y) \in \mathfrak{R}$ and $\left(x^{\prime}, y^{\prime}\right) \in \mathfrak{R}$, then

$$
\left|d\left(x, x^{\prime}\right)-d\left(y, y^{\prime}\right)\right| \leqslant d(x, y)+d\left(x^{\prime}, y^{\prime}\right)<2 r
$$

Conversely, we show that $d_{\mathrm{GH}}(X, Y) \leqslant \frac{1}{2} \operatorname{dis} \mathfrak{R}$ for any correspondence $\mathfrak{R}$. Let $\operatorname{dis} \Re=2 r$. To avoid confusion, we use the notation $d_{X}$ and $d_{Y}$ for the
metrics of $X, Y$, respectively. It suffices to show that there is a metric $d$ on the disjoint union $X \bigsqcup Y$ such that $\left.d\right|_{X \times X}=d_{X},\left.d\right|_{Y \times Y}=d_{Y}$, and $d_{H}(X, Y) \leqslant r$ in $(X \bigsqcup Y, d)$.

Given $x \in X$ and $y \in Y$, define

$$
d(x, y)=\inf \left\{d_{X}\left(x, x^{\prime}\right)+r+d_{Y}\left(y^{\prime}, y\right):\left(x^{\prime}, y^{\prime}\right) \in \mathfrak{R}\right\}
$$

(the distances within $X$ and $Y$ are already defined by $d_{X}$ and $d_{Y}$ ). Verifying the triangle inequality for $d$ and the fact that $d_{H}(X, Y) \leqslant r$ is straightforward.

Definition 4.12. Consider two metric spaces $X$ and $Y$. For $\epsilon>0$, a map $f: X \rightarrow Y$ is called an $\epsilon$-isometry if dis $f \leqslant \epsilon$ and $f(X)$ is an $\epsilon$-net in $Y$. (Here $\operatorname{dis} f:=\sup _{x, y \in X}\left|d_{X}(x, y)-d_{Y}(f(x), f(y))\right|$.)

Corollary 4.13. Consider two metric spaces $X$ and $Y$. Fix $\epsilon>0$.
(i) If $d_{\mathrm{GH}}(X, Y)<\epsilon$, then there exists a $2 \epsilon$-isometry from $X$ to $Y$.
(ii) If there exists an $\epsilon$-isometry from $X$ to $Y$, then $d_{\mathrm{GH}}(X, Y)<2 \epsilon$.

Proof. (i) Let $\mathfrak{R}$ be a correspondence between $X$ and $Y$ with dis $\Re<2 \epsilon$. For every $x \in X$, choose $f(x) \in Y$ such that $(x, f(x)) \in \mathfrak{R}$. This defines a map $f: X \rightarrow Y$. Obviously $\operatorname{dis} f \leqslant \operatorname{dis} \mathfrak{R}<2 \epsilon$. We will show that $f(X)$ is an $\epsilon$-net in $Y$.

For a $y \in Y$, consider an $x \in X$ such that $(x, y) \in \mathfrak{R}$. Since both $y$ and $f(x)$ are in correspondence with $x$, it follows that $d(y, f(x)) \leqslant d(x, x)+\operatorname{dis} \Re<2 \epsilon$. Hence, $d(y, f(X))<2 \epsilon$.
(ii) Let $f$ be an $\epsilon$-isometry. Define $\mathfrak{R} \subset X \times Y$ by

$$
\mathfrak{R}=\{(x, y) \in X \times Y: d(y, f(x)) \leqslant \epsilon\} .
$$

Then $\mathfrak{R}$ is a correspondence because $f(X)$ is an $\epsilon$-net in $Y$. If $(x, y) \in \mathfrak{R}$ and $\left(x^{\prime}, y^{\prime}\right) \in \mathfrak{R}$, then

$$
\begin{aligned}
\left|d\left(y, y^{\prime}\right)-d\left(x, x^{\prime}\right)\right| & \leqslant\left|d\left(f(x), f\left(x^{\prime}\right)\right)-d\left(x, x^{\prime}\right)\right|+d(y, f(x))+d\left(y^{\prime}, f\left(x^{\prime}\right)\right) \\
& \leqslant \operatorname{dis} f+\epsilon+\epsilon \leqslant 3 \epsilon
\end{aligned}
$$

Hence, dis $\mathbb{R}<3 \epsilon$, and Theorem 4.11 implies

$$
d_{\mathrm{GH}}(X, Y) \leqslant \frac{3}{2} \epsilon<2 \epsilon
$$

### 4.2.3 Gromov-Hausdorff distance for compact spaces

Theorem 4.14. The Gromov-Hausdorff distance is a metric on the space of isometry classes of compact metric spaces.

Proof. We already know that $d_{\mathrm{GH}}$ is a semi-metric, so only that show $d_{\mathrm{GH}}(X, Y)=0$ implies that $X$ and $Y$ are isometric.

Let $X$ and $Y$ be two compact spaces such that $d_{\mathrm{GH}}(X, Y)=0$. By Corollary 4.13 , there exists a sequence of maps $f_{n}: X \rightarrow Y$ such that $\operatorname{dis} f_{n} \rightarrow 0$.

Fix a countable dense set $S \subset X$. Using Cantor's diagonal procedure, choose a subsequence $\left\{f_{n_{k}}\right\}$ of $\left\{f_{n}\right\}$ such that for every $x \in S$ the sequence $\left\{f_{n_{k}}(x)\right\}$ converges in $Y$. By renumbering, we may assume that this holds for $\left\{f_{n}\right\}$ itself. Define a map $f: S \rightarrow Y$ as the limit of the $f_{n}$, namely, set $f(x)=\lim f_{n}(x)$ for every $x \in S$.

Because

$$
\left|d\left(f_{n}(x), f_{n}(y)\right)-d(x, y)\right| \leqslant \operatorname{dis} f_{n} \rightarrow 0
$$

we have

$$
d(f(x), f(y))=\lim d\left(f_{n}(x), f_{n}(y)\right)=d(x, y) \quad \text { for all } x, y \in S
$$

In other words, $f$ is a distance-preserving map from $S$ to $Y$. Then $f$ can be extended to a distance-preserving map from $X$ to $Y$. Now interchange the roles of $X$ and $Y$.

Proposition 4.15. Consider compact metric spaces $X$ and $\left\{X_{n}\right\}_{n \in \mathbb{N}}$. The sequence $\left\{X_{n}\right\}_{n \in \mathbb{N}}$ converges to $X$ in the Gromov-Hausdorff distance if and only if for every $\epsilon>0$ there exists a finite $\epsilon$-net $S$ in $X$ and an $\epsilon$-net $S_{n}$ in each $X_{n}$ such that $S_{n}$ converges to $S$ in the Gromov-Hausdorff distance.

Moreover these $\epsilon$-nets can be chosen so that, for all sufficiently large $n, S_{n}$ has the same cardinality as $S$.

Definition 4.16. A collection $\mathfrak{X}$ of compact metric spaces is uniformly totally bounded if for every $\epsilon>0$ there exists a natural number $N=N(\epsilon)$ such that every $X \in \mathfrak{X}$ contains an $\epsilon$-net consisting of no more than $N$ points.

Remark 4.17. Note that if the collection $\mathfrak{X}$ of compact metric spaces is uniformly totally bounded, then there is a constant $D$ such that $\operatorname{diam}(X) \leqslant D$ for all $X \in \mathfrak{X}$.

Theorem 4.18. A uniformly totally bounded class $\mathfrak{X}$ of compact metric spaces is pre-compact in the Gromov-Hausdorff topology.

Proof. Let $N(\epsilon)$ be as in Definition 4.16 and $D$ be as in Remark 4.17. Define $N_{1}=N(1)$ and $N_{k}=N_{k-1}+N(1 / k)$ for all $k \geqslant 2$. Let $\left\{X_{n}\right\}_{n \in \mathbb{N}}$ be a sequence of metric spaces from $\mathfrak{X}$.

In every space $X_{n}$, consider a union of $(1 / k)$-nets for all $k \in \mathbb{N}$. This is a countable dense collection $S_{n}=\left\{x_{i, n}\right\}_{i \in \mathbb{N}} \subset X_{n}$ such that for every $k$ the first $N_{k}$ points of $S_{n}$ form a $(1 / k)$-net in $X_{n}$. The distances $d_{X_{n}}\left(x_{i, n}, x_{j, n}\right)$ do not exceed $D$, i.e. belong to a compact interval. Therefore, using the Cantor diagonal procedure, we can extract a subsequence of $\left\{X_{n}\right\}$ in which $\left\{d_{X_{n}}\left(x_{i, n}, x_{j, n}\right)\right\}_{n \in \mathbb{N}}$ converge for all $i, j$. To simplify the notation, we assume that these sequences converge without passing to a subsequence.

We will construct the limit space $\bar{X}$ for $\left\{X_{n}\right\}_{n \in \mathbb{N}}$ as follows. First, pick an abstract countable set $X=\left\{x_{i}\right\}_{i \in \mathbb{N}}$ and define a semi-metric $d$ on $X$ by

$$
d\left(x_{i}, x_{j}\right)=\lim _{n \rightarrow \infty} d_{X_{n}}\left(x_{n, i}, x_{n, j}\right)
$$

A quotient construction gives us a metric space $X / d$. We will denote by $\bar{x}_{i}$ the point of $X / d$ obtained from $x_{i}$. Let $\bar{X}$ be the completion of $X / d$.

For $k \in \mathbb{N}$, consider the set $S^{(k)}=\left\{\bar{x}_{i}: 1 \leqslant i \leqslant N_{k}\right\} \subset \bar{X}$. Note that $S^{(k)}$ is a $(1 / k)$-net in $\bar{X}$. Indeed, every set $S_{n}^{(k)}=\left\{\bar{x}_{i, n}: 1 \leqslant i \leqslant N_{k}\right\}$ is a $(1 / k)$-net in the respective space $X_{n}$. Hence, for every $x_{i, n} \in S_{n}$ there is a $j \leqslant N_{k}$ such that $d_{X_{n}}\left(x_{i, n}, x_{j, n}\right) \leqslant 1 / k$ for infinitely many indices $n$. Passing to the limit, we see that $d\left(\bar{x}_{i}, \bar{x}_{j}\right) \leqslant 1 / k$ for this $j$. Thus, $S^{(n)}$ is a $(1 / k)$-net in $X / d$. Hence, $S^{(n)}$ is also a $(1 / k)$-net in in $\bar{X}$. Since $\bar{X}$ is complete and has a $(1 / k)$-net for any $k \in \mathbb{N}, \bar{X}$ is compact.

Furthermore, the set $S^{(k)}$ is a Gromov-Hausdorff limit of the sets $S_{n}^{(k)}$ as $n \rightarrow \infty$, because these are finite sets consisting of $N_{k}$ points (some of which may coincide) and there is a way of matching up the points of $S_{n}^{(k)}$ with those in $S^{(k)}$ so that distances converge. Thus, for every $k \in \mathbb{N}$ we have a $(1 / k)$-net in $\bar{X}$ that is a Gromov-Hausdorff limit of some $(1 / k)$-nets in the spaces $X_{n}$. By Proposition 4.15, it follows that $X_{n}$ converges to $\bar{X}$ in the Gromov-Hausdorff distance.

### 4.2.4 Gromov-Hausdorff distance for geodesic spaces

Theorem 4.19. Let $\left\{X_{n}\right\}_{n \in \mathbb{N}}$ be a sequence of geodesic spaces and $X$ a complete metric space such that $X_{n}$ converges to $X$ in the Gromov-Hausdorff distance. Then $X$ is a geodesic space.

Proof. Because $X$ is complete, it suffices to prove that for any two points $x, y \in$ $X$ there is a point $z \in X$ such that $d(x, z)=\frac{1}{2} d(x, y)$ and $d(y, z)=\frac{1}{2} d(x, y)$. Again by completeness, it further suffices to show that for any $\epsilon>0$ there is a point $z \in X$ such that $\left|d(x, z)-\frac{1}{2} d(x, y)\right|<\epsilon$ and $\left|d(y, z)-\frac{1}{2} d(x, y)\right|<\epsilon$.

Let $n$ be such that $d_{\mathrm{GH}}\left(X, X_{n}\right)<\epsilon / 4$. Then, by Theorem 4.11, there is a correspondence $\Re$ between $X$ and $X_{n}$ whose distortion is less than $\epsilon / 2$. Take points $\tilde{x}, \tilde{y} \in X_{n}$ corresponding to $x$ and $y$. Since $X_{n}$ is a geodesic space, there is a $\tilde{z} \in X_{n}$ such that $d(\tilde{x}, \tilde{z})=d(\tilde{z}, \tilde{y})=\frac{1}{2} d(\tilde{x}, \tilde{y})$. Let $z \in X$ be a point corresponding to $\tilde{z}$. Then

$$
\left|d(x, z)-\frac{1}{2} d(x, y)\right| \leqslant\left|d(\tilde{x}, \tilde{z})-\frac{1}{2} d(\tilde{x}, \tilde{y})\right|+2 \operatorname{dis} \Re<\epsilon
$$

Similarly, $\left|d(y, z)-\frac{1}{2} d(x, y)\right|<\epsilon$.
Proposition 4.20. Every compact geodesic space can be obtained as a Gromov-Hausdorff limit of a sequence of finite graphs with edge lengths.

### 4.3 Compact $\mathbb{R}$-trees and the Gromov-Hausdorff metric

### 4.3.1 Unrooted $\mathbb{R}$-trees

Definition 4.21. Let $\left(\mathbf{T}, d_{\mathrm{GH}}\right)$ be the metric space of isometry classes of compact real trees equipped with the Gromov-Hausdorff metric.

Lemma 4.22. The set $\mathbf{T}$ of compact $\mathbb{R}$-trees is a closed subset of the space of compact metric spaces equipped with the Gromov-Hausdorff distance.

Proof. It suffices to note that the limit of a sequence in $\mathbf{T}$ is a geodesic space and satisfies the four point condition.

Theorem 4.23. The metric space $\left(\mathbf{T}, d_{\mathrm{GH}}\right)$ is complete and separable.
Proof. We start by showing separability. Given a compact $\mathbb{R}$-tree, $T$, and $\varepsilon>0$, let $S_{\varepsilon}$ be a finite $\varepsilon$-net in $T$. Write $T_{\varepsilon}$ for the subtree of $T$ spanned by $S_{\varepsilon}$, that is,

$$
\begin{equation*}
T_{\varepsilon}:=\bigcup_{x, y \in S_{\varepsilon}}[x, y] \quad \text { and } \quad d_{T_{\varepsilon}}:=\left.d\right|_{T_{\varepsilon}} \tag{4.3}
\end{equation*}
$$

Obviously, $T_{\varepsilon}$ is still an $\varepsilon$-net for $T$. Hence, $d_{\mathrm{GH}}\left(T_{\varepsilon}, T\right) \leqslant d_{H}\left(T_{\varepsilon}, T\right) \leqslant \varepsilon$.
Now each $T_{\varepsilon}$ is just a "finite tree with edge lengths" and can clearly be approximated arbitrarily closely in the $d_{\mathrm{GH}}$-metric by trees with the same tree topology (that is, "shape"), and rational edge lengths. The set of isometry types of finite trees with rational edge lengths is countable, and so ( $\mathbf{T}, d_{\mathrm{GH}}$ ) is separable.

It remains to establish completeness. It suffices by Lemma 4.22 to show that any Cauchy sequence in $\mathbf{T}$ converges to some compact metric space, or, equivalently, any Cauchy sequence in $\mathbf{T}$ has a subsequence that converges to some metric space.

Let $\left(T_{n}\right)_{n \in \mathbb{N}}$ be a Cauchy sequence in $\mathbf{T}$. By Theorem 4.18, a sufficient condition for this sequence to have a subsequential limit is that for every $\varepsilon>0$ there exists a positive number $N=N(\varepsilon)$ such that every $T_{n}$ contains an $\varepsilon$-net of cardinality $N$.

Fix $\varepsilon>0$ and $n_{0}=n_{0}(\varepsilon)$ such that $d_{\mathrm{GH}}\left(T_{m}, T_{n}\right)<\varepsilon / 2$ for $m, n \geqslant n_{0}$. Let $S_{n_{0}}$ be a finite ( $\varepsilon / 2$ )-net for $T_{n_{0}}$ of cardinality $N$. Then by (4.11) for each $n \geqslant n_{0}$ there exists a correspondence $\Re_{n}$ between $T_{n_{0}}$ and $T_{n}$ such that $\operatorname{dis}\left(\Re_{n}\right)<\varepsilon$.

For each $x \in T_{n_{0}}$, choose $f_{n}(x) \in T_{n}$ such that $\left(x, f_{n}(x)\right) \in \Re_{n}$. Since for any $y \in T_{n}$ with $(x, y) \in \Re_{n}, d_{T_{n}}\left(y, f_{n}(x)\right) \leqslant \operatorname{dis}\left(\Re_{n}\right)$, for all $n \geqslant n_{0}$, the set $f_{n}\left(S_{n_{0}}\right)$ is an $\varepsilon$-net of cardinality $N$ for $T_{n}, n \geqslant n_{0}$.

### 4.3.2 Trees with four leaves

The following result is elementary, but we include the proof because it includes some formulae that will be useful later.

Lemma 4.24. The isometry class of a compact $\mathbb{R}$-tree tree $(T, d)$ with four leaves is uniquely determined by the distances between the leaves of $T$.

Proof. Let $\{a, b, c, d\}$ be the set of leaves of $T$. The tree $T$ has one of four possible shapes shown in Figure 4.4.


Fig. 4.4. The four leaf-labeled trees with four leaves

Consider case ( $I$ ), and let $e$ be the uniquely determined branch point on the tree that lies on the segments $[a, b]$ and $[a, c]$, and $f$ be the uniquely determined branch point on the tree that lies on the segments $[c, d]$ and $[a, c]$. That is,

$$
e:=Y(a, b, c)=Y(a, b, d)
$$

and

$$
f:=Y(c, d, a)=Y(c, d, b)
$$

Observe that

$$
\begin{align*}
d(a, e) & =\frac{1}{2}(d(a, b)+d(a, c)-d(b, c))=(b \cdot c)_{a}=(b \cdot d)_{a} \\
d(b, e) & =\frac{1}{2}(d(a, b)+d(b, c)-d(a, c))=(a \cdot c)_{b}=(a \cdot d)_{b} \\
d(c, f) & =\frac{1}{2}(d(c, d)+d(a, c)-d(a, d))=(d \cdot a)_{c}=(d \cdot b)_{c}  \tag{4.4}\\
d(d, f) & =\frac{1}{2}(d(c, d)+d(a, d)-d(a, c))=(c \cdot a)_{d}=(c \cdot b)_{d} \\
d(e, f) & =\frac{1}{2}(d(a, d)+d(b, c)-d(a, b)-d(c, d))=(a \cdot b)_{f}=(c \cdot d)_{e}
\end{align*}
$$

Similar observations for the other cases show that if we know the shape of the tree, then we can determine its edge lengths from leaf-to-leaf distances. Note also that

$$
\begin{align*}
& \frac{1}{2}(d(a, c)+d(b, d)-d(a, b)-d(c, d)) \\
& \quad=\left\{\begin{array}{cc}
>0 & \text { for shape (I), } \\
<0 & \text { for shape (II), } \\
=0 & \text { for shapes (III) and (IV) }
\end{array}\right. \tag{4.5}
\end{align*}
$$

This and analogous inequalities for the quantities that reconstruct the length of the "internal" edge in shapes $(I I)$ and (III), respectively, show that the shape of the tree can also be reconstructed from leaf-to-leaf distances.

### 4.3.3 Rooted $\mathbb{R}$-trees

Definition 4.25. $A$ rooted $\mathbb{R}$-tree , $(X, d, \rho)$, is a $\mathbb{R}$-tree $(X, d)$ with a distinguished point $\rho \in X$ that we call the root. It is helpful to use genealogical terminology and think of $\rho$ as a common ancestor and $h(x):=d(\rho, x)$ as the real-valued generation to which $x \in X$ belongs $(h(x)$ is also called the height of $x$ ).

We define a partial order $\leqslant$ on $X$ by declaring that

- $x \leqslant y$ if $x \in[\rho, y]$, so that $x$ is an ancestor of $y$.

Each pair $x, y \in X$ has a well-defined greatest common lower bound, $x \wedge y$, in this partial order that we think of as the most recent common ancestor of $x$ and $y$ - see Figure 4.5.

Definition 4.26. Let $\mathbf{T}^{\text {root }}$ denote the collection of all root-invariant isometry classes of rooted compact $\mathbb{R}$-trees, where we define a root-invariant isometry to be an isometry

$$
\xi:\left(X_{1}, d_{X_{1}}, \rho_{1}\right) \rightarrow\left(X_{2}, d_{X_{2}}, \rho_{2}\right) \text { with } \xi\left(\rho_{1}\right)=\rho_{2} .
$$

Define the rooted Gromov-Hausdorff distance , $d_{\mathrm{GH}}{ }^{\text {root }}\left(\left(X_{1}, \rho_{1}\right),\left(X_{2}, \rho_{2}\right)\right)$, between two rooted $\mathbb{R}$-trees $\left(X_{1}, \rho_{1}\right)$ and $\left(X_{2}, \rho_{2}\right)$


Fig. 4.5. A tree rooted at $\rho$. Here $w \leqslant x$ and $w \leqslant y$ and also $z \leqslant x$ and $z \leqslant y$. The greatest common lower bound of $x$ and $y$ is $z$.
as the infimum of $d_{\mathrm{H}}\left(X_{1}^{\prime}, X_{2}^{\prime}\right) \vee d_{Z}\left(\rho_{1}^{\prime}, \rho_{2}^{\prime}\right)$ over all rooted $\mathbb{R}$-trees $\left(X_{1}^{\prime}, \rho_{1}^{\prime}\right)$ and $\left(X_{2}^{\prime}, \rho_{2}^{\prime}\right)$ that are root-invariant isomorphic to $\left(X_{1}, \rho_{1}\right)$ and $\left(X_{2}, \rho_{2}\right)$, respectively, and that are (as unrooted trees) subspaces of a common metric space $\left(Z, d_{Z}\right)$.

Lemma 4.27. For two rooted trees $\left(X_{1}, d_{X_{1}}, \rho_{1}\right)$, and $\left(X_{2}, d_{X_{2}}, \rho_{2}\right)$,

$$
\begin{equation*}
d_{\mathrm{GH}^{\mathrm{root}}}\left(\left(X_{1}, d_{X_{1}}, \rho_{1}\right),\left(X_{2}, d_{X_{2}}, \rho_{2}\right)\right)=\frac{1}{2} \inf _{\Re^{\text {root }}} \operatorname{dis}\left(\Re^{\text {root }}\right), \tag{4.6}
\end{equation*}
$$

where now the infimum is taken over all correspondences $\Re^{\text {root }}$ between $X_{1}$ and $X_{2}$ with $\left(\rho_{1}, \rho_{2}\right) \in \Re^{\text {root }}$.

Definition 4.28. Let $\left(X_{1}, \rho_{1}\right)$ and $\left(X_{2}, \rho_{2}\right)$ be two rooted compact $\mathbb{R}$-trees, and take $\varepsilon>0$. A map $f$ is called a root-invariant $\varepsilon$-isometry from $\left(X_{1}, \rho_{1}\right)$ to $\left(X_{2}, \rho_{2}\right)$ if $f\left(\rho_{1}\right)=\rho_{2}, \operatorname{dis}(f)<\varepsilon$ and $f\left(X_{1}\right)$ is an $\varepsilon$-net for $X_{2}$.

Lemma 4.29. Let $\left(X_{1}, \rho_{1}\right)$ and $\left(X_{2}, \rho_{2}\right)$ be two rooted compact $\mathbb{R}$-trees, and take $\varepsilon>0$. Then the following hold.
(i) If $d_{\mathrm{GH}^{\text {root }}}\left(\left(X_{1}, \rho_{1}\right),\left(X_{2}, \rho_{2}\right)\right)<\varepsilon$, then there exists a root-invariant $2 \varepsilon$ isometry from $\left(X_{1}, \rho_{1}\right)$ to $\left(X_{2}, \rho_{2}\right)$.
(ii)If there exists a root-invariant $\varepsilon$-isometry from $\left(X_{1}, \rho_{1}\right)$ to $\left(X_{2}, \rho_{2}\right)$, then

$$
d_{\mathrm{GH}^{\text {root }}}\left(\left(X_{1}, \rho_{1}\right),\left(X_{2}, \rho_{2}\right)\right) \leqslant \frac{3}{2} \varepsilon .
$$

Proof. (i) Let $d_{\text {GHroot }}\left(\left(X_{1}, \rho_{1}\right),\left(X_{2}, \rho_{2}\right)\right)<\varepsilon$. By Lemma 4.27 there exists a correspondence $\Re^{\text {root }}$ between $X_{1}$ and $X_{2}$ such that $\left(\rho_{1}, \rho_{2}\right) \in \Re^{\text {root }}$ and $\operatorname{dis}\left(\Re^{\mathrm{root}}\right)<2 \varepsilon$.

Define $f: X_{1} \rightarrow X_{2}$ by setting $f\left(\rho_{1}\right)=\rho_{2}$, and choosing $f(x)$ such that $(x, f(x)) \in \Re^{\text {root }}$ for all $x \in X_{1} \backslash\left\{\rho_{1}\right\}$.

Clearly, $\operatorname{dis}(f) \leqslant \operatorname{dis}\left(\Re^{\text {root }}\right)<2 \varepsilon$.
To see that $f\left(X_{1}\right)$ is a $2 \varepsilon$-net for $X_{2}$, let $x_{2} \in X_{2}$, and choose $x_{1} \in X_{1}$ such that $\left(x_{1}, x_{2}\right) \in \Re^{\text {root }}$. Then $d_{X_{2}}\left(f\left(x_{1}\right), x_{2}\right) \leqslant d_{X_{1}}\left(x_{1}, x_{1}\right)+\operatorname{dis}\left(\Re^{\text {root }}\right)<2 \varepsilon$.
(ii) Let $f$ be a root-invariant $\varepsilon$-isometry from $\left(X_{1}, \rho_{1}\right)$ to $\left(X_{2}, \rho_{2}\right)$. Define a correspondence $\Re_{f}^{\text {root }} \subseteq X_{1} \times X_{2}$ by

$$
\begin{equation*}
\Re_{f}^{\text {root }}:=\left\{\left(x_{1}, x_{2}\right): d_{X_{2}}\left(x_{2}, f\left(x_{1}\right)\right) \leqslant \varepsilon\right\} . \tag{4.7}
\end{equation*}
$$

Then $\left(\rho_{1}, \rho_{2}\right) \in \Re_{f}^{\text {root }}$ and $\Re_{f}^{\text {root }}$ is indeed a correspondence since $f\left(X_{1}\right)$ is a $\varepsilon$-net for $X_{2}$. If $\left(x_{1}, x_{2}\right),\left(y_{1}, y_{2}\right) \in \mathbb{R}_{f}^{\text {root }}$, then

$$
\begin{align*}
\left|d_{X_{1}}\left(x_{1}, y_{1}\right)-d_{X_{2}}\left(x_{2}, y_{2}\right)\right| \leqslant & \left|d_{X_{2}}\left(f\left(x_{1}\right), f\left(y_{1}\right)\right)-d_{X_{1}}\left(x_{1}, y_{1}\right)\right| \\
& +d_{X_{2}}\left(x_{2}, f\left(x_{1}\right)\right)+d_{X_{2}}\left(f\left(x_{1}\right), y_{2}\right)  \tag{4.8}\\
& <3 \varepsilon .
\end{align*}
$$

Hence, $\operatorname{dis}\left(\Re_{f}^{\text {root }}\right)<3 \varepsilon$ and, by (4.6),

$$
d_{\mathrm{GH}}{ }^{\text {root }}\left(\left(X_{1}, \rho_{1}\right),\left(X_{2}, \rho_{2}\right)\right) \leqslant \frac{3}{2} \varepsilon .
$$

We need the following compactness criterion, that is the analogue of Theorem 4.18 and can be proved the same way, noting that the analogue of Lemma 4.22 holds for $\mathbf{T}^{\text {root }}$.

Lemma 4.30. $A$ subset $\mathcal{T} \subset \mathbf{T}^{\text {root }}$ is relatively compact if and only if for every $\varepsilon>0$ there exists a positive integer $N(\varepsilon)$ such that each $T \in \mathcal{T}$ has an $\varepsilon$-net with at most $N(\varepsilon)$ points.

Theorem 4.31. The metric space ( $\mathbf{T}^{\mathrm{root}}, d_{G H^{\mathrm{root}}}$ ) is complete and separable.
Proof. The proof follows very much the same lines as that of Theorem 4.23. The proof of separability is almost identical. The key step in establishing completeness is again to show that a Cauchy sequence in $\mathbf{T}^{\text {root }}$ has a subsequential limit. This can be shown in the same manner as in the proof of Theorem 4.23, with an appeal to Lemma 4.30 replacing one to Theorem 4.18.

### 4.3.4 Rooted subtrees and trimming

A rooted subtree of a rooted $\mathbb{R}$-tree $(T, d, \rho) \in \mathbf{T}^{\text {root }}$ is an element $\left(T^{*}, d^{*}, \rho^{*}\right) \in$ $\mathbf{T}^{\text {root }}$ that has a class representative that is a subspace of a class representative of ( $T, d, \rho$ ), with the two roots coincident. Equivalently, any class representative of $\left(T^{*}, d^{*}, \rho^{*}\right)$ can be isometrically embedded into any class representative of $(T, d, \rho)$ via an isometry that maps roots to roots. We write $T^{*} \leq^{\text {root }} T$ and note that $\leq^{\text {root }}$ is a partial order on $\mathbf{T}^{\text {root }}$.

For $\eta>0$ define $R_{\eta}: \mathbf{T}^{\text {root }} \rightarrow \mathbf{T}^{\text {root }}$ to be the map that assigns to $(T, \rho) \in \mathbf{T}^{\text {root }}$ the rooted subtree $\left(R_{\eta}(T), \rho\right)$ that consists of $\rho$ and points $a \in T$ for which the subtree

$$
S^{T, a}:=\{x \in T: a \in[\rho, x]\}
$$

(that is, the subtree above a) has height greater than or equal to $\eta$. Equivalently,

$$
R_{\eta}(T):=\left\{x \in T: \exists y \in T \text { such that } x \in[\rho, y], d_{T}(x, y) \geqslant \eta\right\} \cup\{\rho\}
$$

In particular, if $T$ has height at most $\eta$, then $R_{\eta}(T)$ is just the trivial tree consisting of the root $\rho$. See Figure 4.6 for an example of this construction.

Lemma 4.32. (i) The range of $R_{\eta}$ consists of finite rooted trees (that is, rooted compact $\mathbb{R}$-trees with finitely many leaves).
(ii) The map $R_{\eta}$ is continuous.
(iii) The family of maps $\left(R_{\eta}\right)_{\eta>0}$ is a semigroup; that is,

$$
R_{\eta^{\prime}} \circ R_{\eta^{\prime \prime}}=R_{\eta^{\prime}+\eta^{\prime \prime}} \text { for } \eta^{\prime}, \eta^{\prime \prime}>0
$$

In particular,

$$
R_{\eta^{\prime}}(T) \leq^{\text {root }} R_{\eta^{\prime \prime}}(T) \text { for } \eta^{\prime} \geqslant \eta^{\prime \prime}>0
$$

(iv) For any $(T, \rho) \in \mathbf{T}^{\text {root }}$,

$$
d_{\mathrm{GH}^{\mathrm{root}}}\left((T, \rho),\left(R_{\eta}(T), \rho\right)\right) \leqslant d_{\mathrm{H}}\left(T, R_{\eta}(T)\right) \leqslant \eta
$$

where $d_{\mathrm{H}}$ is the Hausdorff metric on compact subsets of $T$ induced by the metric $\rho$.

Lemma 4.33. Consider a sequence $\left\{T_{n}\right\}_{n \in \mathbb{N}}$ of representatives of isometry classes of rooted compact trees in $\left(\mathbf{T}, d_{\mathrm{GH}^{\mathrm{root}}}\right)$ with the following properties.

- Each set $T_{n}$ is a subset of some common set $U$.
- Each tree $T_{n}$ has the same root $\rho \in U$.
- The sequence $\left\{T_{n}\right\}_{n \in \mathbb{N}}$ is nondecreasing, that is, $T_{1} \subseteq T_{2} \subseteq \cdots \subseteq U$.
- Writing $d_{n}$ for the metric on $T_{n}$, for $m<n$ the restriction of $d_{n}$ to $T_{m}$ coincides with $d_{m}$, so that there is a well-defined metric on $T:=\bigcup_{n \in \mathbb{N}} T_{n}$ given by

$$
d(a, b)=d_{n}(a, b), \quad a, b \in T_{n}
$$



Fig. 4.6. Trimming a tree. The tree $T$ consists of both the solid and dashed edges. The $\eta$-trimming $R_{\eta}(T)$ consists of the solid edges and is composed of the points of $T$ that are distance at least $\eta$ from some leaf of $T$.

- The sequence of subsets $\left(T_{n}\right)_{n \in \mathbb{N}}$ is Cauchy in the Hausdorff distance with respect to $d$.

Then the metric completion $\bar{T}$ of $T$ is a compact $\mathbb{R}$-tree, and $d_{\mathrm{H}}\left(T_{n}, \bar{T}\right) \rightarrow 0$ as $n \rightarrow \infty$, where the Hausdorff distance is computed with respect to the extension of $d$ to $\bar{T}$. In particular,

$$
\lim _{n \rightarrow \infty} d_{\mathrm{GH}^{\mathrm{root}}}\left(\left(T_{n}, \rho\right),(\bar{T}, \rho)\right)=0
$$

### 4.3.5 Length measure on $\mathbb{R}$-trees

Fix $(T, d, \rho) \in \mathbf{T}^{\text {root }}$, and denote the Borel- $\sigma$-field on $T$ by $\mathcal{B}(T)$. Write

$$
\begin{equation*}
T^{o}:=\bigcup_{b \in T}[\rho, b[ \tag{4.9}
\end{equation*}
$$

for the skeleton of $T$.
Observe that if $T^{\prime} \subset T$ is a dense countable set, then (4.9) holds with $T$ replaced by $T^{\prime}$. In particular, $T^{o} \in \mathcal{B}(T)$ and $\left.\mathcal{B}(T)\right|_{T^{o}}=\sigma\left(\{ ] a, b\left[; a, b \in T^{\prime}\right\}\right)$, where

$$
\left.\mathcal{B}(T)\right|_{T^{o}}:=\left\{A \cap T^{o} ; A \in \mathcal{B}(T)\right\}
$$

Hence, there exists a unique $\sigma$-finite measure $\mu=\mu^{T}$ on $T$, called length measure, such that $\mu\left(T \backslash T^{o}\right)=0$ and

$$
\begin{equation*}
\mu(] a, b[)=d(a, b), \quad \forall a, b \in T . \tag{4.10}
\end{equation*}
$$

In particular, $\mu$ is the restriction to $T^{o}$ of one-dimensional Hausdorff measure on $T$.

Example 4.34. Recall from Examples 3.14 and 3.37 the construction of a rooted $\mathbb{R}$-tree ( $T_{e}, d_{T_{e}}$ ) from an excursion path $e \in U$. We can identify the length measure as follows. Given $e \in U^{\ell}$ and $a \geqslant 0$, let

$$
\mathcal{G}_{a}:=\left\{\begin{array}{c}
e(t)=a \text { and, for some } \varepsilon>0  \tag{4.11}\\
t \in[0, \ell]: \quad e(u)>a \text { for all } u \in] t, t+\varepsilon[ \\
e(t+\varepsilon)=a .
\end{array}\right\}
$$

denote the countable set of starting points of excursions of the function $e$ above the level $a$. Then $\mu^{T_{e}}$, the length measure on $T_{e}$, is just the push-forward of the measure $\int_{0}^{\infty} \mathrm{d} a \sum_{t \in \mathcal{G}_{a}} \delta_{t}$ by the quotient map. Alternatively - see Figure 4.7 write

$$
\begin{equation*}
\Gamma_{e}:=\{(s, a): s \in] 0, \ell[, a \in[0, e(s)[ \} \tag{4.12}
\end{equation*}
$$

for the region between the time axis and the graph of $e$, and for $(s, a) \in \Gamma_{e}$ denote by $\underline{s}(e, s, a):=\sup \{r<s: e(r)=a\}$ and $\bar{s}(e, s, a):=\inf \{t>s: e(t)=$ $a\}$ the start and finish of the excursion of $e$ above level $a$ that straddles time $s$. Then $\mu^{T_{e}}$ is the push-forward of the measure $\int_{\Gamma_{e}} \mathrm{~d} s \otimes \mathrm{~d} a \frac{1}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} \delta_{\underline{s}(e, s, a)}$ by the quotient map. We note that the measure $\mu^{T_{e}}$ appears in [1].

There is a simple recipe for the total length of a finite tree (that is, a tree with finitely many leaves).

Lemma 4.35. Let $(T, d, \rho) \in \mathbf{T}^{\text {root }}$ and suppose that $\left\{x_{0}, \ldots, x_{n}\right\} \subset T$ spans $T$, so that the root $\rho$ and the leaves of $T$ form a subset of $\left\{x_{0}, \ldots, x_{n}\right\}$. Then the total length of $T$ (that is, the total mass of its length measure) is given by

$$
\begin{aligned}
& d\left(x_{0}, x_{1}\right)+\sum_{k=2}^{n} \bigwedge_{0 \leqslant i<j \leqslant k-1} \frac{1}{2}\left(d\left(x_{k}, x_{i}\right)+d\left(x_{k}, x_{j}\right)-d\left(x_{i}, x_{j}\right)\right) \\
& \quad=d\left(x_{0}, x_{1}\right)+\sum_{k=2}^{n} \bigwedge_{0 \leqslant i<j \leqslant k-1}\left(x_{i} \cdot x_{j}\right)_{x_{k}}
\end{aligned}
$$

- see Figure 4.8.


Fig. 4.7. Various objects associated with an excursion $e \in U^{1}$. The set of starting points of excursions of $e$ above level $a$ is $\mathcal{G}_{a}=\{r, u\}$. The region between the graph of $e$ and the time axis is $\Gamma_{e}$. The start and finish of the excursion of $e$ above level $a$ that straddles time $s$ are $\underline{s}(e, s, a)=r$ and $\bar{s}(e, s, a)=t$.

Proof. This follows from the observation that the distance from the point $x_{k}$ to the segment $\left[x_{i}, x_{j}\right]$ is

$$
\frac{1}{2}\left(d\left(x_{k}, x_{i}\right)+d\left(x_{k}, x_{j}\right)-d\left(x_{i}, x_{j}\right)\right)=\left(x_{i} \cdot x_{j}\right)_{x_{k}}
$$

in the notation of Definition 3.7, and so the length of the segment connecting $x_{k}, 2 \leqslant k \leqslant n$, to the subtree spanned by $x_{0}, \ldots, x_{k-1}$ is

$$
\bigwedge_{0 \leqslant i<j \leqslant k-1} \frac{1}{2}\left(d\left(x_{k}, x_{i}\right)+d\left(x_{k}, x_{j}\right)-d\left(x_{i}, x_{j}\right)\right) .
$$

The formula of Lemma 4.35 can be used to establish the following result, which implies that the function that sends a tree to its total length is lower semi-continuous (and, therefore, Borel). We refer the reader to Lemma 7.3 of [63] for the proof.
Lemma 4.36. For $\eta>0$, the map $T \mapsto \mu^{T}\left(R_{\eta}\right)$ (that is, the map that takes a tree to the total length of its $\eta$-trimming) is continuous.


Fig. 4.8. The construction of Lemma 4.35. The total length of the tree is $d\left(x_{0}, x_{1}\right)+$ $d\left(x_{2}, y\right)+d\left(x_{3}, z\right)$.

The following result, when combined with the compactness criterion Lemma 4.30, gives an alternative necessary and sufficient condition for a subset of $\mathbf{T}^{\text {root }}$ to be relatively compact (Corollary 4.38 below).

Lemma 4.37. Let $T \in \mathbf{T}^{\text {root }}$ be such that $\mu^{T}(T)<\infty$. For each $\varepsilon>0$ there is an $\varepsilon$-net for $T$ of cardinality at most

$$
\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)\right]\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)+1\right] .
$$

Proof. Note that an $\frac{\varepsilon}{2}$-net for $R_{\frac{\varepsilon}{2}}(T)$ will be an $\varepsilon$-net for $T$. The set $T \backslash R_{\frac{\varepsilon}{2}}(T)^{o}$ is the union of a collection disjoint subtrees. Each leaf of $R_{\frac{\varepsilon}{2}}(T)$ belongs to a unique such subtree, and the diameter of each such subtree is at least $\frac{\varepsilon}{2}$. (There may also be other subtrees in the collection that don't contain leaves of $R_{\frac{\varepsilon}{2}}(T)$.) Thus, the number of leaves of $R_{\frac{\varepsilon}{2}}(T)$ is at most $\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)$. Enumerate the leaves of $R_{\frac{\varepsilon}{2}}(T)$ as $x_{0}, x_{1}, \ldots, x_{n}$. Each segment $\left[x_{0}, x_{i}\right], 1 \leqslant$ $i \leqslant n$, of $R_{\frac{\varepsilon}{2}}(T)$ has an $\frac{\varepsilon}{2}$-net of cardinality at $\operatorname{most}\left(\frac{\varepsilon}{2}\right)^{-1} d_{T}\left(x_{0}, x_{i}\right)+1 \leqslant$ $\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)+1$. Therefore, by taking the union of these nets, $R_{\frac{\varepsilon}{2}}(T)$ has an $\frac{\varepsilon}{2}$-net of cardinality at most $\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)\right]\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}(T)+1\right]$.

Corollary 4.38. A subset $\mathcal{T}$ of $\left(\mathbf{T}^{\mathrm{root}}, d_{\mathrm{GH}^{\mathrm{root}}}\right)$ is relatively compact if and only if for all $\varepsilon>0$,

$$
\sup \left\{\mu^{T}\left(R_{\varepsilon}(T)\right): T \in \mathcal{T}\right\}<\infty
$$

Proof. The "only if" direction follows from continuity of $T \mapsto \mu^{T}\left(R_{\varepsilon}(T)\right)$ obtained in Lemma 4.36.

Conversely, suppose that the condition of the corollary holds. Given $T \in \mathcal{T}$, an $\varepsilon$-net for $R_{\varepsilon}(T)$ is a $2 \varepsilon$-net for $T$. By Lemma $4.37, R_{\varepsilon}(T)$ has an $\varepsilon$-net of cardinality at most

$$
\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}\left(R_{\varepsilon}(T)\right)\right]\left[\left(\frac{\varepsilon}{2}\right)^{-1} \mu^{T}\left(R_{\varepsilon}(T)\right)+1\right]
$$

By assumption, the last quantity is uniformly bounded in $T \in \mathcal{T}$. Hence, the set $\mathcal{T}$ is relatively compact by Lemma 4.30 .

### 4.4 Weighted $\mathbb{R}$-trees

A weighted $\mathbb{R}$-tree is a $\mathbb{R}$-tree $(T, d)$ equipped with a probability measure $\nu$ on the Borel $\sigma$-field $\mathcal{B}(T)$. Write $\mathbf{T}^{\mathrm{wt}}$ for the space of weight-preserving isometry classes of weighted compact $\mathbb{R}$-trees, where we say that two weighted, compact $\mathbb{R}$-trees $(X, d, \nu)$ and $\left(X^{\prime}, d^{\prime}, \nu^{\prime}\right)$ are weight-preserving isometric if there exists an isometry $\phi$ between $X$ and $X^{\prime}$ such that the push-forward of $\nu$ by $\phi$ is $\nu^{\prime}$ :

$$
\begin{equation*}
\nu^{\prime}=\phi_{*} \nu:=\nu \circ \phi^{-1} . \tag{4.13}
\end{equation*}
$$

It is clear that the property of being weight-preserving isometric is an equivalence relation.

Example 4.39. Recall from Examples 3.14 and 3.37 the construction of a compact $\mathbb{R}$-tree from an excursion path $e \in U^{\ell}$. Such a $\mathbb{R}$-tree has a canonical weight, namely, the push-forward of normalized Lebesgue measure on $[0, \ell]$ by the quotient map that appears in the construction.

We want to equip $\mathbf{T}^{\mathrm{wt}}$ with a Gromov-Hausdorff type of distance that incorporates the weights on the trees.

Lemma 4.40. Let $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$ be two compact real trees such that $d_{\mathrm{GH}}\left(\left(X, d_{X}\right),\left(Y, d_{Y}\right)\right)<\varepsilon$ for some $\varepsilon>0$. Then there exists a measurable $3 \varepsilon$-isometry from $X$ to $Y$.

Proof. If $d_{\mathrm{GH}}\left(\left(X, d_{X}\right),\left(Y, d_{Y}\right)\right)<\varepsilon$, then by Theorem 4.11 there exists a correspondence $\Re$ between $X$ and $Y$ such that $\operatorname{dis}(\Re)<2 \varepsilon$. Since $\left(X, d_{X}\right)$ is compact there exists a finite $\varepsilon$-net in $X$. We claim that for each such finite $\varepsilon$-net $S^{X, \varepsilon}=\left\{x_{1}, \ldots, x_{N^{\varepsilon}}\right\} \subseteq X$, any set $S^{Y, \varepsilon}=\left\{y_{1}, \ldots, y_{N^{\varepsilon}}\right\} \subseteq Y$ such that
$\left(x_{i}, y_{i}\right) \in \Re$ for all $i \in\left\{1,2, \ldots, N^{\varepsilon}\right\}$ is a $3 \varepsilon$-net in $Y$. To see this, fix $y \in Y$. We have to show the existence of $i \in\left\{1,2, \ldots, N^{\varepsilon}\right\}$ with $d_{Y}\left(y_{i}, y\right)<3 \varepsilon$. For that choose $x \in X$ such that $(x, y) \in \Re$. Since $S^{X, \varepsilon}$ is an $\varepsilon$-net in $X$ there exists an $i \in\left\{1,2, \ldots, N^{\varepsilon}\right\}$ such that $d_{X}\left(x_{i}, x\right)<\varepsilon .\left(x_{i}, y_{i}\right) \in \Re$ implies, therefore, that $\left|d_{X}\left(x_{i}, x\right)-d_{Y}\left(y_{i}, y\right)\right| \leqslant \operatorname{dis}(\Re)<2 \varepsilon$. Hence, $d_{Y}\left(y_{i}, y\right)<3 \varepsilon$.

Furthermore, we may decompose $X$ into $N^{\varepsilon}$ possibly empty measurable disjoint subsets of $X$ by letting $X^{1, \varepsilon}:=\mathcal{B}\left(x_{1}, \varepsilon\right), X^{2, \varepsilon}:=\mathcal{B}\left(x_{2}, \varepsilon\right) \backslash X^{1, \varepsilon}$, and so on, where $\mathcal{B}(x, r)$ is the open ball $\left\{x^{\prime} \in X: d_{X}\left(x, x^{\prime}\right)<r\right\}$. Then $f$ defined by $f(x)=y_{i}$ for $x \in X^{i, \varepsilon}$ is obviously a measurable $3 \varepsilon$-isometry from $X$ to $Y$.

We also need to recall the definition of the Prohorov distance between two probability measures - see, for example, [57]. Given two probability measures $\mu$ and $\nu$ on a metric space $(X, d)$ with the corresponding collection of closed sets denoted by $\mathcal{C}$, the Prohorov distance between them is

$$
d_{\mathrm{P}}(\mu, \nu):=\inf \left\{\varepsilon>0: \mu(C) \leqslant \nu\left(C^{\varepsilon}\right)+\varepsilon \text { for all } C \in \mathcal{C}\right\}
$$

where $C^{\varepsilon}:=\left\{x \in X: \inf _{y \in C} d(x, y)<\varepsilon\right\}$. The Prohorov distance is a metric on the collection of probability measures on $X$. The following result shows that if we push measures forward with a map having a small distortion, then Prohorov distances can't increase too much.

Lemma 4.41. Suppose that $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$ are two metric spaces, $f$ : $X \rightarrow Y$ is a measurable map with $\operatorname{dis}(f) \leqslant \varepsilon$, and $\mu$ and $\nu$ are two probability measures on $X$. Then

$$
d_{\mathrm{P}}\left(f_{*} \mu, f_{*} \nu\right) \leqslant d_{\mathrm{P}}(\mu, \nu)+\varepsilon
$$

Proof. Suppose that $d_{\mathrm{P}}(\mu, \nu)<\delta$. By definition, $\mu(C) \leqslant \nu\left(C^{\delta}\right)+\delta$ for all closed sets $C \in \mathcal{C}$. If $D$ is a closed subset of $Y$, then

$$
\begin{aligned}
f_{*} \mu(D) & =\mu\left(f^{-1}(D)\right) \\
& \leqslant \mu\left(\overline{f^{-1}(D)}\right) \\
& \leqslant \nu\left(\overline{f^{-1}(D)}{ }^{\delta}\right)+\delta \\
& =\nu\left(f^{-1}(D)^{\delta}\right)+\delta
\end{aligned}
$$

Now $x^{\prime} \in f^{-1}(D)^{\delta}$ means there is $x^{\prime \prime} \in X$ such that $d_{X}\left(x^{\prime}, x^{\prime \prime}\right)<\delta$ and $f\left(x^{\prime \prime}\right) \in D$. By the assumption that $\operatorname{dis}(f) \leqslant \varepsilon$, we have $d_{Y}\left(f\left(x^{\prime}\right), f\left(x^{\prime \prime}\right)\right)<$ $\delta+\varepsilon$. Hence, $f\left(x^{\prime}\right) \in D^{\delta+\varepsilon}$. Thus,

$$
f^{-1}(D)^{\delta} \subseteq f^{-1}\left(D^{\delta+\varepsilon}\right)
$$

and we have

$$
f_{*} \mu(D) \leqslant \nu\left(f^{-1}\left(D^{\delta+\varepsilon}\right)\right)+\delta=f_{*} \nu\left(D^{\delta+\varepsilon}\right)+\delta
$$

so that $d_{\mathrm{P}}\left(f_{*} \mu, f_{*} \nu\right) \leqslant \delta+\varepsilon$, as required.

We are now in a position to define the weighted Gromov-Hausdorff distance between the two compact, weighted $\mathbb{R}$-trees $\left(X, d_{X}, \nu_{X}\right)$ and $\left(Y, d_{Y}, \nu_{Y}\right)$. For $\varepsilon>0$, set

$$
\begin{equation*}
F_{X, Y}^{\varepsilon}:=\{\text { measurable } \varepsilon \text {-isometries from } X \text { to } \mathrm{Y}\} \tag{4.14}
\end{equation*}
$$

Put

$$
\begin{align*}
& \Delta_{\mathrm{GH}^{\mathrm{wt}}}(X, Y) \\
& :=\inf \left\{\varepsilon>0: \begin{array}{c}
\text { there exist } f \in F_{X, Y}^{\varepsilon}, g \in F_{Y, X}^{\varepsilon} \text { such that } \\
d_{\mathrm{P}}\left(f_{*} \nu_{X}, \nu_{Y}\right) \leqslant \varepsilon, d_{\mathrm{P}}\left(\nu_{X}, g_{*} \nu_{Y}\right) \leqslant \varepsilon
\end{array}\right\} . \tag{4.15}
\end{align*}
$$

Note that the set on the right hand side is non-empty because $X$ and $Y$ are compact, and, therefore, bounded. It will turn out that $\Delta_{G H}{ }^{w t}$ satisfies all the properties of a metric except the triangle inequality. To rectify this, let

$$
\begin{equation*}
d_{\mathrm{GH}}{ }^{\mathrm{wt}}(X, Y):=\inf \left\{\sum_{i=1}^{n-1} \Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}}\right\} \tag{4.16}
\end{equation*}
$$

where the infimum is taken over all finite sequences of compact, weighted $\mathbb{R}$-trees $Z_{1}, \ldots Z_{n}$ with $Z_{1}=X$ and $Z_{n}=Y$.

Lemma 4.42. The map $d_{\mathrm{GH}^{\mathrm{wt}}}: \mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}} \rightarrow \mathbb{R}_{+}$is a metric on $\mathbf{T}^{\mathrm{wt}}$. Moreover,

$$
\frac{1}{2} \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}(X, Y)^{\frac{1}{4}} \leqslant d_{\mathrm{GH}}{ }^{\mathrm{wt}}(X, Y) \leqslant \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}(X, Y)^{\frac{1}{4}}
$$

for all $X, Y \in \mathbf{T}^{\mathrm{wt}}$.
Proof. It is immediate from (4.15) that the map $\Delta_{\mathrm{GH}} \mathrm{wt}$ is symmetric.
We next claim that

$$
\begin{equation*}
\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(\left(X, d_{X}, \nu_{X}\right),\left(Y, d_{Y}, \nu_{Y}\right)\right)=0 \tag{4.17}
\end{equation*}
$$

if and only if $\left(X, d_{X}, \nu_{X}\right)$ and $\left(Y, d_{Y}, \nu_{Y}\right)$ are weight-preserving isometric. The "if" direction is immediate. Note first for the converse that (4.17) implies that for all $\varepsilon>0$ there exists an $\varepsilon$-isometry from $X$ to $Y$, and, therefore, by Corollary 4.13, $d_{\mathrm{GH}}\left(\left(X, d_{X}\right),\left(Y, d_{Y}\right)\right)<2 \varepsilon$. Thus, $d_{\mathrm{GH}}\left(\left(X, d_{X}\right),\left(Y, d_{Y}\right)\right)=0$, and it follows from Theorem 4.14 that $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$ are isometric. Checking the proof of that result, we see that we can construct an isometry $f: X \rightarrow Y$ by taking any dense countable set $S \subset X$, any sequence of functions $\left(f_{n}\right)$ such that $f_{n}$ is an $\varepsilon_{n}$-isometry with $\varepsilon_{n} \rightarrow 0$ as $n \rightarrow \infty$, and letting $f$ be $\lim _{k} f_{n_{k}}$ along any subsequence such that the limit exists for all $x \in S$ (such a subsequence exists by the compactness of $Y$ ). Therefore, fix some dense subset $S \subset X$ and suppose without loss of generality that we have an isometry $f: X \rightarrow Y$ given by $f(x)=\lim _{n \rightarrow \infty} f_{n}(x), x \in S$, where $f_{n} \in F_{X, Y}^{\varepsilon_{n}}, d_{\mathrm{P}}\left(f_{n * \nu_{X}}, \nu_{Y}\right) \leqslant \varepsilon_{n}$, and $\lim _{n \rightarrow \infty} \varepsilon_{n}=0$. We will be done if we
can show that $f_{*} \nu_{X}=\nu_{Y}$. If $\mu_{X}$ is a discrete measure with atoms belonging to $S$, then

$$
\begin{gather*}
d_{\mathrm{P}}\left(f_{*} \nu_{X}, \nu_{Y}\right) \leqslant \limsup _{n}\left[d_{\mathrm{P}}\left(f_{n * \nu_{X}}, \nu_{Y}\right)+d_{\mathrm{P}}\left(f_{n *} \mu_{X}, f_{n *} \nu_{X}\right)\right. \\
\left.\quad+d_{\mathrm{P}}\left(f_{*} \mu_{X}, f_{n *} \mu_{X}\right)+d_{\mathrm{P}}\left(f_{*} \nu_{X}, f_{*} \mu_{X}\right)\right]  \tag{4.18}\\
\leqslant 2 d_{\mathrm{P}}\left(\mu_{X}, \nu_{X}\right)
\end{gather*}
$$

where we have used Lemma 4.41 and the fact that $\lim _{n \rightarrow \infty} d_{\mathrm{P}}\left(f_{*} \mu_{X}, f_{n *} \mu_{X}\right)=$ 0 because of the pointwise convergence of $f_{n}$ to $f$ on $S$. Because we can choose $\mu_{X}$ so that $d_{\mathrm{P}}\left(\mu_{X}, \nu_{X}\right)$ is arbitrarily small, we see that $f_{*} \nu_{X}=\nu_{Y}$, as required.

Now consider three spaces $\left(X, d_{X}, \nu_{X}\right),\left(Y, d_{Y}, \nu_{Y}\right)$, and $\left(Z, d_{Z}, \nu_{Z}\right)$ in $\mathbf{T}^{\mathrm{wt}}$, and constants $\varepsilon, \delta>0$, such that $\Delta_{\mathrm{GH}} \mathrm{GHt}\left(\left(X, d_{X}, \nu_{X}\right),\left(Y, d_{Y}, \nu_{Y}\right)\right)<\varepsilon$ and $\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(\left(Y, d_{Y}, \nu_{Y}\right),\left(Z, d_{Z}, \nu_{Z}\right)\right)<\delta$. Then there exist $f \in F_{X, Y}^{\varepsilon}$ and $g \in F_{Y, Z}^{\delta}$ such that $d_{\mathrm{P}}\left(f_{*} \nu_{X}, \nu_{Y}\right)<\varepsilon$ and $d_{\mathrm{P}}\left(g_{*} \nu_{Y}, \nu_{Z}\right)<\delta$. Note that $g \circ f \in F_{X, Z}^{\varepsilon+\delta}$. Moreover, by Lemma 4.41

$$
\begin{equation*}
d_{\mathrm{P}}\left((g \circ f)_{*} \nu_{X}, \nu_{Z}\right) \leqslant d_{\mathrm{P}}\left(g_{*} \nu_{Y}, \nu_{Z}\right)+d_{\mathrm{P}}\left(g_{*} f_{*} \nu_{X}, g_{*} \nu_{Y}\right)<\delta+\varepsilon+\delta . \tag{4.19}
\end{equation*}
$$

This, and a similar argument with the roles of $X$ and $Z$ interchanged, shows that

$$
\begin{equation*}
\Delta_{\mathrm{GH}^{\mathrm{wt}}}(X, Z) \leqslant 2\left[\Delta_{\mathrm{GH}^{\mathrm{wt}}}(X, Y)+\Delta_{\mathrm{GH}^{\mathrm{wt}}}(Y, Z)\right] . \tag{4.20}
\end{equation*}
$$

The second inequality in the statement of the lemma is clear. In order to see the first inequality, it suffices to show that for any $Z_{1}, \ldots Z_{n}$ we have

$$
\begin{equation*}
\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{1}, Z_{n}\right)^{\frac{1}{4}} \leqslant 2 \sum_{i=1}^{n-1} \Delta_{\mathrm{GH}}{ }_{\mathrm{Gt}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}} \tag{4.21}
\end{equation*}
$$

We will establish (4.21) by induction. The inequality certainly holds when $n=2$. Suppose it holds for $2, \ldots, n-1$. Write $S$ for the value of the sum on the right hand side of (4.21). Put

$$
\begin{equation*}
k:=\max \left\{1 \leqslant m \leqslant n-1: \sum_{i=1}^{m-1} \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}} \leqslant S / 2\right\} \tag{4.22}
\end{equation*}
$$

By the inductive hypothesis and the definition of $k$,

$$
\begin{equation*}
\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{1}, Z_{k}\right)^{\frac{1}{4}} \leqslant 2 \sum_{i=1}^{k-1} \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}} \leqslant 2(S / 2)=S . \tag{4.23}
\end{equation*}
$$

Of course,

$$
\begin{equation*}
\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{k}, Z_{k+1}\right)^{\frac{1}{4}} \leqslant S \tag{4.24}
\end{equation*}
$$

By definition of $k$,

$$
\sum_{i=1}^{k} \Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}}>S / 2
$$

so that once more by the inductive hypothesis,

$$
\begin{align*}
\Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{k+1}, Z_{n}\right)^{\frac{1}{4}} & \leqslant 2 \sum_{i=k+1}^{n-1} \Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}} \\
& =2 S-2 \sum_{i=1}^{k} \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{i}, Z_{i+1}\right)^{\frac{1}{4}}  \tag{4.25}\\
& \leqslant S
\end{align*}
$$

From (4.23), (4.24), (4.25) and two applications of (4.20) we have

$$
\begin{align*}
& \Delta_{\mathrm{GH}} \mathrm{wt} \\
&\left(Z_{1}, Z_{n}\right)^{\frac{1}{4}} \leqslant\left\{4 \left[\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(Z_{1}, Z_{k}\right)+\Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{k}, Z_{k+1}\right)\right.\right.  \tag{4.26}\\
&\left.\left.+\Delta_{\mathrm{GH}^{\mathrm{wt}}}\left(Z_{k+1}, Z_{n}\right)\right]\right\}^{\frac{1}{4}} \\
& \leqslant\left(4 \times 3 \times S^{4}\right)^{\frac{1}{4}} \\
& \leqslant 2 S
\end{align*}
$$

as required.
It is obvious by construction that $d_{\mathrm{GH}^{w t}}$ satisfies the triangle inequality. The other properties of a metric follow from the corresponding properties we have already established for $\Delta_{\mathrm{GH}} \mathrm{wt}$ and the bounds in the statement of the lemma that we have already established.

The procedure we used to construct the weighted Gromov-Hausdorff metric $d_{\mathrm{GH}}{ }^{\mathrm{wt}}$ from the semi-metric $\Delta_{\mathrm{GH}} \mathrm{wt}$ was adapted from a proof in [88] of the celebrated result of Alexandroff and Urysohn on the metrizability of uniform spaces. That proof was, in turn, adapted from earlier work of Frink and Bourbaki. The choice of the power $\frac{1}{4}$ is not particularly special, any sufficiently small power would have worked.

Proposition 4.43. A subset $\mathbf{D}$ of $\left(\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}^{\mathrm{wt}}}\right)$ is relatively compact if and only if the subset $\mathbf{E}:=\{(T, d):(T, d, \nu) \in \mathbf{D}\}$ in $\left(\mathbf{T}, d_{\mathrm{GH}}\right)$ is relatively compact.

Proof. The "only if" direction is clear. Assume for the converse that $\mathbf{E}$ is relatively compact. Suppose that $\left(\left(T_{n}, d_{T_{n}}, \nu_{T_{n}}\right)\right)_{n \in \mathbb{N}}$ is a sequence in D. By assumption, $\left(\left(T_{n}, d_{T_{n}}\right)\right)_{n \in \mathbb{N}}$ has a subsequence converging to some point $\left(T, d_{T}\right)$ of $\left(\mathbf{T}, d_{\mathrm{GH}}\right)$. For ease of notation, we will renumber and also denote this subsequence by $\left(\left(T_{n}, d_{T_{n}}\right)\right)_{n \in \mathbb{N}}$. For brevity, we will also omit specific mention of the metric on a real tree when it is clear from the context.

By Proposition 4.15, for each $\varepsilon>0$ there is a finite $\varepsilon$-net $T^{\varepsilon}$ in $T$ and for each $n \in \mathbb{N}$ a finite $\varepsilon$-net $T_{n}^{\varepsilon}:=\left\{x_{n}^{\varepsilon, 1}, \ldots, x_{n}^{\varepsilon, \# T_{n}^{\varepsilon}}\right\}$ in $T_{n}$ such that
$d_{\mathrm{GH}}\left(T_{n}^{\varepsilon}, T^{\varepsilon}\right) \rightarrow 0$ as $n \rightarrow \infty$. Without loss of generality, we may assume that $\# T_{n}^{\varepsilon}=\# T^{\varepsilon}$ for all $n \in \mathbb{N}$. We may begin with the balls of radius $\varepsilon$ around each point of $\# T_{n}^{\varepsilon}$ and decompose $T_{n}$ into $\# T_{n}^{\varepsilon}$ possibly empty, disjoint, measurable sets $\left\{T_{n}^{\varepsilon, 1}, \ldots, T_{n}^{\varepsilon, \# T^{\varepsilon}}\right\}$ of radius no greater than $\varepsilon$. Define a measurable map $f_{n}: T_{n} \rightarrow T_{n}^{\varepsilon}$ by $f_{n}^{\varepsilon}(x)=x_{n}^{\varepsilon, i}$ if $x \in T_{n}^{\varepsilon, i}$ and let $g_{n}^{\varepsilon}$ be the inclusion map from $T_{n}^{\varepsilon}$ to $T_{n}$. By construction, $f_{n}^{\varepsilon}$ and $g_{n}^{\varepsilon}$ are $\varepsilon$-isometries. Moreover, $d_{\mathrm{P}}\left(\left(g_{n}^{\varepsilon}\right)_{*}\left(f_{n}^{\varepsilon}\right)_{*} \nu_{n}, \nu_{n}\right)<\varepsilon$ and, of course, $d_{\mathrm{P}}\left(\left(f_{n}^{\varepsilon}\right)_{*} \nu_{n},\left(f_{n}^{\varepsilon}\right)_{*} \nu_{n}\right)=0$. Thus, $\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(\left(T_{n}^{\varepsilon},\left(f_{n}^{\varepsilon}\right)_{*} \nu_{n}\right),\left(T_{n}, \nu_{n}\right)\right) \leqslant \varepsilon$. By similar reasoning, if we define $h_{n}^{\varepsilon}: T_{n}^{\varepsilon} \rightarrow T^{\varepsilon}$ by $x_{n}^{\varepsilon, i} \mapsto x^{\varepsilon, i}$, then

$$
\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(\left(T_{n}^{\varepsilon},\left(f_{n}^{\varepsilon}\right)_{*} \nu_{n}\right),\left(T^{\varepsilon},\left(h_{n}^{\varepsilon}\right)_{*} \nu_{n}\right)\right) \rightarrow 0
$$

as $n \rightarrow \infty$. Since $T^{\varepsilon}$ is finite, by passing to a subsequence (and relabeling as before) we have $\lim _{n \rightarrow \infty} d_{\mathrm{P}}\left(\left(h_{n}^{\varepsilon}\right)_{*} \nu_{n}, \nu^{\varepsilon}\right)=0$ for some probability measure $\nu^{\varepsilon}$ on $T^{\varepsilon}$. Hence,

$$
\lim _{n \rightarrow \infty} \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}\left(\left(T^{\varepsilon},\left(h_{n}^{\varepsilon}\right)_{*} \nu_{n}\right),\left(T^{\varepsilon}, \nu^{\varepsilon}\right)\right)=0
$$

Therefore, by Lemma 4.42,

$$
\limsup _{n \rightarrow \infty} d_{\mathrm{GH}^{\mathrm{wt}}}\left(\left(T_{n}, \nu_{n}\right),\left(T^{\varepsilon},\left(h_{n}^{\varepsilon}\right)_{*} \nu_{n}\right)\right) \leqslant \varepsilon^{\frac{1}{4}}
$$

Now, since $\left(T, d_{T}\right)$ is compact, the family of measures $\left\{\nu^{\varepsilon}: \varepsilon>0\right\}$ is relatively compact, and so there is a probability measure $\nu$ on $T$ such that $\nu^{\varepsilon}$ converges to $\nu$ in the Prohorov distance along a subsequence $\varepsilon \downarrow 0$. Hence, by arguments similar to the above, along the same subsequence $\Delta_{\mathrm{GH}} \mathrm{Gt}\left(\left(T^{\varepsilon}, \nu^{\varepsilon}\right),(T, \nu)\right)$ converges to 0 . Again applying Lemma 4.42, we have that $d_{\mathrm{GH}^{\mathrm{wt}}}\left(\left(T^{\varepsilon}, \nu^{\varepsilon}\right),(T, \nu)\right)$ converges to 0 along this subsequence.

Combining the foregoing, we see that by passing to a suitable subsequence and relabeling, $d_{\mathrm{GH}^{\mathrm{wt}}}\left(\left(T_{n}, \nu_{n}\right),(T, \nu)\right)$ converges to 0 , as required.

Theorem 4.44. The metric space $\left(\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}^{\mathrm{wt}}}\right)$ is complete and separable.
Proof. Separability follows readily from the separability of ( $\mathbf{T}, d_{\mathrm{GH}}$ ) and the separability with respect to the Prohorov distance of the probability measures on a fixed complete, separable metric space - see, for example, [57]) - and Lemma 4.42.

It remains to establish completeness. By a standard argument, it suffices to show that any Cauchy sequence in $\mathbf{T}^{\mathrm{wt}}$ has a convergent subsequence. Let $\left(T_{n}, d_{T_{n}}, \nu_{n}\right)_{n \in \mathbb{N}}$ be a Cauchy sequence in $\mathbf{T}^{\mathrm{wt}}$. Then $\left(T_{n}, d_{T_{n}}\right)_{n \in \mathbb{N}}$ is a Cauchy sequence in $\mathbf{T}$ by Lemma 4.42. By Theorem 1 in [63] there is a $T \in \mathbf{T}$ such that $d_{\mathrm{GH}}\left(T_{n}, T\right) \rightarrow 0$, as $n \rightarrow \infty$. In particular, the sequence $\left(T_{n}, d_{T_{n}}\right)_{n \in \mathbb{N}}$ is relatively compact in $\mathbf{T}$, and, therefore, by Proposition $4.43,\left(T_{n}, d_{T_{n}}, \nu_{n}\right)_{n \in \mathbb{N}}$ is relatively compact in $\mathbf{T}^{\mathrm{wt}}$. Thus, $\left(T_{n}, d_{T_{n}}\right)_{n \in \mathbb{N}}$ has a convergent subsequence, as required.

## Root growth with re-grafting

### 5.1 Background and motivation

Recall the special case of the tree-valued Markov chain that was used in the proof of the Markov chain tree theorem, Theorem 2.1, when the underlying Markov chain is the process on $\{1,2, \ldots, n\}$ that picks a new state uniformly at each stage.

## Algorithm 5.1.

- Start with a rooted (combinatorial) tree on $n$ labeled vertices $\{1,2, \ldots, n\}$.
- Pick a vertex $v$ uniformly from
$\{1,2, \ldots, n\} \backslash\{$ current root $\}$.
- Erase the edge leading from $v$ towards the current root.
- Insert an edge from the current root to $v$ and make $v$ the new root.
- Repeat.

We know that this chain converges in distribution to the uniform distribution on rooted trees with $n$ labeled vertices.

Imagine that we do the following.

- Start with a rooted subtree (that is, one with the same root as the "big" tree).
- At each step of the chain, update the subtree by removing and adding edges as they are removed and added in the big tree and adjoining the new root of the big tree to the subtree if it isn't in the current subtree.

The subtree will evolve via two mechanisms that we might call root growth and re-grafting. Root growth occurs when the new root isn't in the current subtree, and so the new tree has an extra vertex, the new root, that is connected to the old root by a new edge. Re-grafting occurs when the new root is in the current subtree: it has the effect of severing the edge leading to a subtree of the current subtree and re-attaching it to the current root by a new edge. See Figure 5.1.


Fig. 5.1. Root growth and re-graft moves. The big tree with $n=11$ vertices consists of the solid and dashed edges in all three diagrams. In the top diagram, the current subtree has the solid edges and the vertices marked $a, b, *$. The vertices marked $c$ and \# are in the big tree but not the current subtree. The big tree and the current subtree are rooted at $a$. The bottom left diagram shows the result of a root growth move: the vertex $c$ now belongs to the new subtree, it is the root of the new big tree and the new subtree, and is connected to the old root $a$ by an edge. The vertices marked \# are not in the new subtree. The bottom right diagram shows the result of a re-graft move: the vertex $b$ is the root of the new big tree and the new subtree, and it is connected to the old root $a$ by an edge. The vertices marked $c$ and \# are not in the new subtree.

Now consider what happens as $n$ becomes large and we follow a rooted subtree that originally has $\approx \sqrt{n}$ vertices. Replace edges of length 1 with edges of length $\frac{1}{\sqrt{n}}$ and speed up time by $\sqrt{n}$.

In the limit as $n \rightarrow \infty$, it seems reasonable that we have a $\mathbb{R}$-tree-valued process with the following root growth with re-grafting dynamics.

- The edge leading to the root of the evolving tree grows at unit speed.
- Cuts rain down on the tree at unit rate per length $\times$ time, and the subtree above each cut is pruned off and re-attached at the root.

We will establish a closely related result in Section 5.4. Namely, we will show that if we have a sequence of chains following the dynamics of Algo-
rithm 5.1 such that the initial combinatorial tree of the $n^{\text {th }}$ chain re-scaled by $\sqrt{n}$ converges in the Gromov-Hausdorff distance to some compact $\mathbb{R}$-tree, then if we re-scale space and time by $\sqrt{n}$ in the $n^{\text {th }}$ chain we get weak convergence to a process with the root growth with re-grafting dynamics.

This latter result might seem counter-intuitive, because now we are working with the whole tree with $n$ vertices rather than a subtree with $\approx \sqrt{n}$ vertices. However, the assumption that the initial condition scaled by $\sqrt{n}$ converges to some compact $\mathbb{R}$-tree means that asymptotically most vertices are close to the leaves and re-arranging the subtrees above such vertices has a negligible effect in the limit.

Before we can establish such a convergence result, we need to show that the root growth with re-grafting dynamics make sense even for compact trees with infinite total length. Such trees are the sort that will typically arise in the limit when we re-scale trees with $n$ vertices by $\sqrt{n}$. This is not a trivial matter, as the set of times at which cuts appear will be dense and so the intuitive description of the dynamics does not make rigorous sense. See Theorem 5.5 for the details.

Given that the chain of Algorithm 5.1 converges at large times to the uniform rooted tree on $n$ labeled vertices and that the uniform tree on $n$ labeled vertices converges after suitable re-scaling to the Brownian continuum random tree as $n \rightarrow \infty$, it seems reasonable that the root growth with re-grafting process should converge at large times to the Brownian continuum random tree and that the Brownian continuum random tree should be the unique stationary distribution. We establish that this is indeed the case in Section 5.3. An important ingredient in the proofs of these facts will be Proposition 5.7, which says that the root growth with re-grafting process started from the trivial tree consisting of a single point is related to the Poisson line-breaking construction of the Brownian continuum random tree in Section 2.5 in the same manner that the chain of Algorithm 5.1 is related to Algorithm 2.4 for generating uniform rooted labeled trees. This is, of course, what we should expect, because the Poisson line-breaking construction arises as a limit of Algorithm 2.4 when the number of vertices goes to infinity.

### 5.2 Construction of the root growth with re-grafting process

### 5.2.1 Outline of the construction

- We want to construct a $\mathbf{T}^{\text {root }}$-valued process $X$ with the root growth and re-grafting dynamics.
- $\operatorname{Fix}(T, d, \rho) \in \mathbf{T}^{\text {root }}$. This will be $X_{0}$.
- We will construct simultaneously for each finite rooted subtree $T^{*} \leq^{\text {root }} T$ a process $X^{T^{*}}$ with $X_{0}^{T^{*}}=T^{*}$ that evolves according to the root growth with re-grafting dynamics.
- We will carry out this construction in such a way that if $T^{*}$ and $T^{* *}$ are two finite subtrees with $T^{*} \leq^{\text {root }} T^{* *}$, then $X_{t}^{T^{*}} \leq^{\text {root }} X_{t}^{T^{* *}}$ and the cut points for $X^{T^{*}}$ are those for $X^{T^{* *}}$ that happen to fall on $X_{\tau-}^{T^{*}}$ for a corresponding cut time $\tau$ of $X^{T^{* *}}$. Cut times $\tau$ for $X^{T^{* *}}$ for which the corresponding cut point does not fall on $X_{\tau-}^{T^{*}}$ are not cut times for $X^{T^{*}}$.
- The tree $(T, \rho)$ is a rooted Gromov-Hausdorff limit of finite $\mathbb{R}$-trees with root $\rho$ (indeed, any subtree of $(T, \rho)$ that is spanned by the union of a finite $\varepsilon$-net and $\{\rho\}$ is a finite $\mathbb{R}$-tree that has rooted Gromov-Hausdorff distance less than $\varepsilon$ from $(T, \rho))$.
In particular, $(T, \rho)$ is the "smallest" rooted compact $\mathbb{R}$-tree that contains all of the finite rooted subtrees of $(T, \rho)$.
- Because of the consistent projective nature of the construction, we can define $X_{t}:=X_{t}^{T}$ for $t \geqslant 0$ as the "smallest" element of $\mathbf{T}^{\text {root }}$ that contains $X_{t}^{T^{*}}$, for all finite trees $T^{*} \preceq^{\text {root }} T$.


### 5.2.2 A deterministic construction

It will be convenient to work initially in a setting where the cut times and cut points are fixed.

There are two types of cut points: those that occur at points that were present in the initial tree $T$ and those that occur at points that were added due to subsequent root growth.

Accordingly, we consider two countable subsets $\pi_{0} \subset \mathbb{R}^{++} \times T^{o}$ and $\pi \subset$ $\left\{(t, x) \in \mathbb{R}^{++} \times \mathbb{R}^{++}: x \leqslant t\right\}$. See Figure 5.2.

Assumption 5.2. Suppose that the sets $\pi_{0}$ and $\pi$ have the following properties.
(a) For all $t_{0}>0$, each of the sets $\pi_{0} \cap\left(\left\{t_{0}\right\} \times T^{o}\right)$ and $\left.\left.\pi \cap\left(\left\{t_{0}\right\} \times\right] 0, t_{0}\right]\right)$ has at most one point and at least one of these sets is empty.
(b) For all $t_{0}>0$ and all finite subtrees $T^{\prime} \subseteq T$, the set $\left.\left.\pi_{0} \cap(] 0, t_{0}\right] \times T^{\prime}\right)$ is finite.
(c) For all $t_{0}>0$, the set $\pi \cap\left\{(t, x) \in \mathbb{R}^{++} \times \mathbb{R}^{++}: x \leqslant t \leqslant t_{0}\right\}$ is finite.

Remark 5.3. Conditions (a)-(c) of Assumption 5.2 will hold almost surely if $\pi_{0}$ and $\pi$ are realizations of Poisson point processes with respective intensities $\lambda \otimes \mu$ and $\lambda \otimes \lambda$ (where $\lambda$ is Lebesgue measure), and it is this random mechanism that we will introduce later to produce a stochastic process having the root growth with re-grafting dynamics.

Consider a finite rooted subtree $T^{*} \leq^{\text {root }} T$. It will avoid annoying circumlocutions about equivalence via root-invariant isometries if we work with particular class representatives for $T^{*}$ and $T$, and, moreover, suppose that $T^{*}$ is embedded in $T$.

Put $\tau_{0}^{*}:=0$, and let $0<\tau_{1}^{*}<\tau_{2}^{*}<\ldots$ (the cut times for $X^{T^{*}}$ ) be the points of $\left\{t>0: \pi_{0}\left(\{t\} \times T^{*}\right)>0\right\} \cup\left\{t>0: \pi\left(\{t\} \times \mathbb{R}^{++}\right)>0\right\}$.


Fig. 5.2. The sets of points $\pi_{0}$ and $\pi$

Step 1 (Root growth). At any time $t \geqslant 0, X_{t}^{T^{*}}$ as a set is given by the disjoint union $\left.\left.T^{*} \amalg\right] 0, t\right]$. For $t>0$, the root of $X_{t}^{T^{*}}$ is the point $\left.\left.\rho_{t}:=t \in\right] 0, t\right]$. The metric $d_{t}^{T^{*}}$ on $X_{t}^{T^{*}}$ is defined inductively as follows.

Set $d_{0}^{T^{*}}$ to be the metric on $X_{0}^{T^{*}}=T^{*}$; that is, $d_{0}^{T^{*}}$ is the restriction of $d$ to $T^{*}$. Suppose that $d_{t}^{T^{*}}$ has been defined for $0 \leqslant t \leqslant \tau_{n}^{*}$. Define $d_{t}^{T^{*}}$ for $\tau_{n}^{*}<t<\tau_{n+1}^{*}$ by

$$
d_{t}^{T^{*}}(a, b):= \begin{cases}d_{\tau_{n}^{*}}(a, b), & \text { if } a, b \in X_{\tau_{n}^{*}}^{T^{*}},  \tag{5.1}\\ |b-a|, & \text { if } \left.a, b \in] \tau_{n}^{*}, t\right], \\ \left|a-\tau_{n}^{*}\right|+d_{\tau_{n}^{*}}\left(\rho_{\tau_{n}^{*}}, b\right), & \text { if } \left.a \in] \tau_{n}^{*}, t\right], b \in X_{\tau_{n}^{*}}^{T^{*}} .\end{cases}
$$

Step 2 (Re-Grafting). Note that the left-limit $X_{\tau_{n+1}^{*}-}^{T^{*}}$ exists in the rooted Gromov-Hausdorff metric. As a set this left-limit is the disjoint union

$$
\left.\left.\left.\left.X_{\tau_{n}^{*}}^{T^{*}} \amalg\right] \tau_{n}^{*}, \tau_{n+1}^{*}\right]=T^{*} \amalg\right] 0, \tau_{n+1}^{*}\right],
$$

and the corresponding metric $d_{\tau_{n+1}^{*}-}$ is given by a prescription similar to (5.1).
Define the $(n+1)^{\text {st }}$ cut point for $X^{T^{*}}$ by

5 Root growth with re-grafting

$$
p_{n+1}^{*}:= \begin{cases}a \in T^{*}, & \text { if } \pi_{0}\left(\left\{\left(\tau_{n+1}^{*}, a\right)\right\}\right)>0 \\ \left.x \in] 0, \tau_{n+1}^{*}\right], & \text { if } \pi\left(\left\{\left(\tau_{n+1}^{*}, x\right)\right\}\right)>0\end{cases}
$$

Let $S_{n+1}^{*}$ be the subtree above $p_{n+1}^{*}$ in $X_{\tau_{n+1}^{*}-}^{T^{*}}$, that is,

$$
\begin{equation*}
S_{n+1}^{*}:=\left\{b \in X_{\tau_{n+1}^{*}-}^{T^{*}}: p_{n+1}^{*} \in\left[\rho_{\tau_{n+1}^{*}-}^{*}, b[ \}\right.\right. \tag{5.2}
\end{equation*}
$$

Define the metric $d_{\tau_{n+1}^{*}}$ by

$$
\begin{aligned}
& d_{\tau_{n+1}^{*}}(a, b) \\
& := \begin{cases}d_{\tau_{n+1}^{*}-}(a, b), & \text { if } a, b \in S_{n+1}^{*}, \\
d_{\tau_{n+1}^{*}-}^{*}(a, b), & \text { if } a, b \in X_{\tau_{n+1}^{*}}^{T^{*}} \backslash S_{n+1}^{*}, \\
d_{\tau_{n+1}^{*}-}\left(a, \rho_{\tau_{n+1}^{*}}^{*}\right)+d_{\tau_{n+1}^{*}-}\left(p_{n+1}^{*}, b\right), & \text { if } a \in X_{\tau_{n+1}^{*}}^{T^{*} \backslash S_{n+1}^{*}}, b \in S_{n+1}^{*}\end{cases}
\end{aligned}
$$

In other words $X_{\tau_{n+1}^{*}}^{T^{*}}$ is obtained from $X_{\tau_{n+1}^{*}-}^{T^{*}}$ by pruning off the subtree $S_{n+1}^{*}$ and re-attaching it to the root. See Figure 5.3.


Fig. 5.3. Pruning off the subtree $S$ and regrafting it at the root $\rho$

Now consider two other finite, rooted subtrees $\left(T^{* *}, \rho\right)$ and $\left(T^{* * *}, \rho\right)$ of $T$ such that $T^{*} \cup T^{* *} \subseteq T^{* * *}$ (with induced metrics).

Build $X^{T^{* *}}$ and $X^{T^{* * *}}$ from $\pi_{0}$ and $\pi$ in the same manner as $X^{T^{*}}$ (but starting at $T^{* *}$ and $\left.T^{* * *}\right)$. It is clear from the construction that:

- $X_{t}^{T^{*}}$ and $X_{t}^{T^{* *}}$ are rooted subtrees of $X_{t}^{T^{* * *}}$ for all $t \geqslant 0$,
- the Hausdorff distance between $X_{t}^{T^{*}}$ and $X_{t}^{T^{* *}}$ as subsets of $X_{t}^{T^{* * *}}$ does not depend on $T^{* * *}$,
- the Hausdorff distance is constant between jumps of $X^{T^{*}}$ and $X^{T^{* *}}$ (when only root growth is occurring in both processes).
The following lemma shows that the Hausdorff distance between $X_{t}^{T^{*}}$ and $X_{t}^{T^{* *}}$ as subsets of $X_{t}^{T^{* * *}}$ does not increase at jump times.
Lemma 5.4. Let $T$ be a finite rooted tree with root $\rho$ and metric $d$, and let $T^{\prime}$ and $T^{\prime \prime}$ be two rooted subtrees of $T$ (both with the induced metrics and root $\rho$ ). Fix $p \in T$, and let $S$ be the subtree in $T$ above $p$ (recall (5.2)). Define a new metric $\hat{d}$ on $T$ by putting

$$
\hat{d}(a, b):= \begin{cases}d(a, b), & \text { if } a, b \in S \\ d(a, b), & \text { if } a, b \in T \backslash S \\ d(a, p)+d(\rho, b), & \text { if } a \in S, b \in T \backslash S\end{cases}
$$

Then the sets $T^{\prime}$ and $T^{\prime \prime}$ are also subtrees of $T$ equipped with the induced metric $\hat{d}$, and the Hausdorff distance between $T^{\prime}$ and $T^{\prime \prime}$ with respect to $\hat{d}$ is not greater than that with respect to $d$.

Proof. Suppose that the Hausdorff distance between $T^{\prime}$ and $T^{\prime \prime}$ under $d$ is less than some given $\varepsilon>0$. Given $a \in T^{\prime}$, there then exists $b \in T^{\prime \prime}$ such that $d(a, b)<\varepsilon$. Because $d(a, a \wedge b) \leqslant d(a, b)$ and $a \wedge b \in T^{\prime \prime}$, we may suppose (by replacing $b$ by $a \wedge b$ if necessary) that $b \leqslant a$.

We claim that $\hat{d}(a, c)<\varepsilon$ for some $c \in T^{\prime \prime}$. This and the analogous result with the roles of $T^{\prime}$ and $T^{\prime \prime}$ interchanged will establish the result.

If $a, b \in S$ or $a, b \in T \backslash S$, then $\hat{d}(a, b)=d(a, b)<\varepsilon$. The only other possibility is that $a \in S$ and $b \in T \backslash S$, in which case $p \in[b, a]$ (for $T$ equipped with $d$ ). Then $\hat{d}(a, \rho)=d(a, p) \leqslant d(a, b)<\varepsilon$, as required (because $\left.\rho \in T^{\prime \prime}\right)$.

Now let $T_{1} \subseteq T_{2} \subseteq \cdots$ be an increasing sequence of finite subtrees of $T$ such that $\bigcup_{n \in \mathbb{N}} T_{n}$ is dense in $T$. Thus, $\lim _{n \rightarrow \infty} d_{\mathrm{H}}\left(T_{n}, T\right)=0$.

Let $X^{1}, X^{2}, \ldots$ be constructed from $\pi_{0}$ and $\pi$ starting with $T_{1}, T_{2}, \ldots$. Applying Lemma 5.4 yields

$$
\lim _{m, n \rightarrow \infty} \sup _{t \geqslant 0} d_{\mathrm{GH}^{\mathrm{root}}}\left(X_{t}^{m}, X_{t}^{n}\right)=0 .
$$

Hence, by completeness of $\mathbf{T}^{\text {root }}$, there exists a càdlàg $\mathbf{T}^{\text {root }}$-valued process $X$ such that $X_{0}=T$ and

$$
\lim _{m \rightarrow \infty} \sup _{t \geqslant 0} d_{\mathrm{GH}^{\text {root }}}\left(X_{t}^{m}, X_{t}\right)=0 .
$$

A priori, the process $X$ could depend on the choice of the approximating sequence of trees $\left\{T_{n}\right\}_{n \in \mathbb{N}}$. To see that this is not so, consider two approximating sequences $T_{1}^{1} \subseteq T_{2}^{1} \subseteq \cdots$ and $T_{1}^{2} \subseteq T_{2}^{2} \subseteq \cdots$.

For $k \in \mathbb{N}$, write $T_{n}^{3}$ for the smallest rooted subtree of $T$ that contains both $T_{n}^{1}$ and $T_{n}^{2}$. As a set, $T_{n}^{3}=T_{n}^{1} \cup T_{n}^{2}$. Now let $\left\{\left(X_{t}^{n, i}\right\}_{t \geqslant 0}\right)_{n \in \mathbb{N}}$ for $i=1,2,3$ be the corresponding sequences of finite tree-value processes and let $\left(X_{t}^{\infty, i}\right)_{t \geqslant 0}$ for $i=1,2,3$ be the corresponding limit processes. By Lemma 5.4,

$$
\begin{align*}
d_{\mathrm{GH}^{\mathrm{root}}}\left(X_{t}^{n, 1}, X_{t}^{n, 2}\right) & \leqslant d_{\mathrm{GH}^{\mathrm{root}}}\left(X_{t}^{n, 1}, X_{t}^{n, 3}\right)+d_{\mathrm{GH}^{\mathrm{root}}}\left(X_{t}^{n, 2}, X_{t}^{n, 3}\right) \\
& \leqslant d_{\mathrm{H}}\left(X_{t}^{n, 1}, X_{t}^{n, 3}\right)+d_{\mathrm{H}}\left(X_{t}^{n, 2}, X_{t}^{n, 3}\right) \\
& \leqslant d_{\mathrm{H}}\left(T_{n}^{1}, T_{n}^{3}\right)+d_{\mathrm{H}}\left(T_{n}^{2}, T_{n}^{3}\right)  \tag{5.3}\\
& \leqslant d_{\mathrm{H}}\left(T_{n}^{1}, T\right)+d_{\mathrm{H}}\left(T_{n}^{2}, T\right) \rightarrow 0
\end{align*}
$$

as $n \rightarrow \infty$.
Thus, for each $t \geqslant 0$ the sequences $\left\{X_{t}^{n, 1}\right\}_{n \in \mathbb{N}}$ and $\left\{X_{t}^{n, 2}\right\}_{n \in \mathbb{N}}$ do indeed have the same rooted Gromov-Hausdorff limit and the process $X$ does not depend on the choice of approximating sequence for the initial tree $T$.

### 5.2.3 Putting randomness into the construction

We constructed a $\mathbf{T}^{\text {root }}$-valued function $t \mapsto X_{t}$ starting with a fixed triple ( $T, \pi_{0}, \pi$ ), where $T \in \mathbf{T}^{\text {root }}$ and $\pi_{0}, \pi$ satisfy the conditions of Assumption 5.2. We now want to think of $X$ as a function of time and such triples.

Let $\Omega^{*}$ be the set of triples $\left(T, \pi_{0}, \pi\right)$, where $T$ is a rooted compact $\mathbb{R}$ tree (that is, a class representative of an element of $\mathbf{T}^{\text {root }}$ ) and $\pi_{0}, \pi$ satisfy Assumption 5.2.

The root invariant isometry equivalence relation on rooted compact $\mathbb{R}$ trees extends naturally to an equivalence relation on $\Omega^{*}$ by declaring that two triples $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right)$ and ( $\left.T^{\prime \prime}, \pi_{0}^{\prime \prime}, \pi^{\prime \prime}\right)$, where $\pi_{0}^{\prime}=\left\{\left(\sigma_{i}^{\prime}, x_{i}^{\prime}\right): i \in \mathbb{N}\right\}$ and $\pi_{0}^{\prime \prime}=\left\{\left(\sigma_{i}^{\prime \prime}, x_{i}^{\prime \prime}\right): i \in \mathbb{N}\right\}$, are equivalent if there is a root invariant isometry $f$ mapping $T^{\prime}$ to $T^{\prime \prime}$ and a permutation $\gamma$ of $\mathbb{N}$ such that $\sigma_{i}^{\prime \prime}=\sigma_{\gamma(i)}^{\prime}$ and $x_{i}^{\prime \prime}=$ $f\left(x_{\gamma(i)}^{\prime}\right)$ for all $i \in \mathbb{N}$. Write $\Omega$ for the resulting quotient space of equivalence classes. There is a natural measurable structure on $\Omega$ : we refer to [63] for the details.

Given $T \in \mathbf{T}^{\text {root }}$, let $\mathbf{P}^{T}$ be the probability measure on $\Omega$ defined by the following requirements.

- The measure $\mathbf{P}^{T}$ assigns all of its mass to the set $\left\{\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \in \Omega: T^{\prime}=\right.$ $T\}$.
- Under $\mathbf{P}^{T}$, the random variable $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \mapsto \pi_{0}^{\prime}$ is a Poisson point process on the set $\mathbb{R}^{++} \times T^{o}$ with intensity $\lambda \otimes \mu$, where $\mu$ is the length measure on $T$.
- Under $\mathbf{P}^{T}$, the random variable $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \mapsto \pi^{\prime}$ is a Poisson point process on the set $\left\{(t, x) \in \mathbb{R}^{++} \times \mathbb{R}^{++}: x \leqslant t\right\}$ with intensity $\lambda \otimes \lambda$ restricted to this set.
- The random variables $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \mapsto \pi_{0}^{\prime}$ and $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \mapsto \pi^{\prime}$ are independent under $\mathbf{P}^{T}$.

Of course, the random variable $\left(T^{\prime}, \pi_{0}^{\prime}, \pi^{\prime}\right) \mapsto \pi_{0}^{\prime}$ takes values in a space of equivalence classes of countable sets rather than a space of sets per se, so, more formally, this random variable has the law of the image of a Poisson process on an arbitrary class representative under the appropriate quotient map.

For $t \geqslant 0, g$ a bounded Borel function on $\mathbf{T}^{\text {root }}$, and $T \in \mathbf{T}^{\text {root }}$, set

$$
\begin{equation*}
P_{t} g(T):=\mathbf{P}^{T}\left[g\left(X_{t}\right)\right] \tag{5.4}
\end{equation*}
$$

With a slight abuse of notation, let $\tilde{R}_{\eta}$ for $\eta>0$ also denote the map from $\Omega$ into $\Omega$ that sends $\left(T, \pi_{0}, \pi\right)$ to $\left(R_{\eta}(T), \pi_{0} \cap\left(\mathbb{R}^{++} \times\left(R_{\eta}(T)\right)^{o}\right), \pi\right)$.

Theorem 5.5. (i) If $T \in \mathbf{T}^{\text {root }}$ is finite, then $\left(X_{t}\right)_{t \geqslant 0}$ under $\mathbf{P}^{T}$ is a Markov process that evolves via the root growth with re-grafting dynamics on finite trees.
(ii) For all $\eta>0$ and $T \in \mathbf{T}^{\text {root }}$, the law of $\left(X_{t} \circ \tilde{R}_{\eta}\right)_{t \geqslant 0}$ under $\mathbf{P}^{T}$ coincides with the law of $\left(X_{t}\right)_{t \geqslant 0}$ under $\mathbf{P}^{R_{\eta}(T)}$.
(iii) For all $T \in \mathbf{T}^{\text {root }}$, the law of $\left(X_{t}\right)_{t \geqslant 0}$ under $\mathbf{P}^{R_{\eta}(T)}$ converges as $\eta \downarrow 0$ to that of $\left(X_{t}\right)_{t \geqslant 0}$ under $\mathbf{P}^{T}$ (in the sense of convergence of laws on the space of càdlàg $\mathbf{T}^{\mathrm{root}}$-valued paths equipped with the Skorohod topology).
(iv) For $g \in \operatorname{bB}\left(\mathbf{T}^{\text {root }}\right)$, the map $(t, T) \mapsto P_{t} g(T)$ is $\mathcal{B}\left(\mathbb{R}^{+}\right) \times \mathcal{B}\left(\mathbf{T}^{\text {root }}\right)$ measurable.
(v) The process $\left(X_{t}, \mathbf{P}^{T}\right)$ is strong Markov and has transition semigroup $\left(P_{t}\right)_{t \geqslant 0}$.

Proof. (i) This is clear from the definition of the root growth and re-grafting dynamics.
(ii) It is enough to check that the push-forward of the probability measure $\mathbf{P}^{T}$ under the map $R_{\eta}: \Omega \rightarrow \Omega$ is the measure $\mathbf{P}^{R_{\eta}(T)}$.

This, however, follows from the observation that the restriction of length measure on a tree to a subtree is just length measure on the subtree.
(iii) This is immediate from part (ii) and part (iv) of Lemma 4.32. Indeed, we have that

$$
\sup _{t \geqslant 0} d_{\mathrm{GH}} \mathrm{root}\left(X_{t}, X_{t} \circ \tilde{R}_{\eta}\right) \leqslant d_{\mathrm{H}}\left(T, R_{\eta}(T)\right) \leqslant \eta
$$

(iv) By a monotone class argument, it is enough to consider the case where the test function $g$ is continuous. It follows from part (iii) that $P_{t} g\left(R_{\eta}(T)\right)$ converges pointwise to $P_{t} g(T)$ as $\eta \downarrow 0$, and it is not difficult to show using Lemma 4.32 and part (i) that $(t, T) \mapsto P_{t} g\left(R_{\eta}(T)\right)$ is $\mathcal{B}\left(\mathbb{R}^{+}\right) \times \mathcal{B}\left(\mathbf{T}^{\text {root }}\right)$ measurable, but we omit the details.
(v) By construction and Lemma 4.33, we have for $t \geqslant 0$ and $\left(T, \pi_{0}, \pi\right) \in \Omega$ that, as a set, $X_{t}^{o}\left(T, \pi_{0}, \pi\right)$ is the disjoint union $\left.\left.T^{o} \amalg\right] 0, t\right]$.

Put

$$
\begin{aligned}
\theta_{t}(T, & \left.\pi_{0}, \pi\right) \\
:= & \left(X_{t}\left(T, \pi_{0}, \pi\right),\left\{(s, x) \in \mathbb{R}^{++} \times T^{o}:(t+s, x) \in \pi_{0}\right\}\right. \\
& \left.\left\{(s, x) \in \mathbb{R}^{++} \times \mathbb{R}^{++}:(t+s, t+x) \in \pi\right\}\right) \\
= & \left(X_{t}\left(T, \pi_{0}, \pi\right),\left\{(s, x) \in \mathbb{R}^{++} \times X_{t}^{o}\left(T, \pi_{0}, \pi\right):(t+s, x) \in \pi_{0}\right\}\right. \\
& \left.\left\{(s, x) \in \mathbb{R}^{++} \times \mathbb{R}^{++}:(t+s, t+x) \in \pi\right\}\right) .
\end{aligned}
$$

Thus, $\theta_{t}$ maps $\Omega$ into $\Omega$. Note that $X_{s} \circ \theta_{t}=X_{s+t}$ and that $\theta_{s} \circ \theta_{t}=\theta_{s+t}$, that is, the family $\left(\theta_{t}\right)_{t \geqslant 0}$ is a semigroup.

Fix $t \geqslant 0$ and $\left(T, \pi_{0}, \pi\right) \in \Omega$. Write $\mu^{\prime}$ for the measure on $\left.\left.T^{o} \amalg\right] 0, t\right]$ that restricts to length measure on $T^{o}$ and to Lebesgue measure on $\left.] 0, t\right]$. Write $\mu^{\prime \prime}$ for the length measure on $X_{t}^{o}\left(T, \pi_{0}, \pi\right)$.

The strong Markov property will follow from a standard strong Markov property for Poisson processes if we can show that $\mu^{\prime}=\mu^{\prime \prime}$.

This equality is clear from the construction if $T$ is finite: the tree $X_{t}\left(T, \pi_{0}, \pi\right)$ is produced from the tree $T$ and the set $\left.] 0, t\right]$ by a finite number of dissections and rearrangements.

The equality for general $T$ follows from the construction and Lemma 4.33.

### 5.2.4 Feller property

The proof of Theorem 5.5 depended on an argument that showed that if we have two finite subtrees of a given tree that are close in the GromovHausdorff distance, then the resulting root growth with re-grafting processes can be coupled together on the same probability space so that they stay close together. It is believable that if we start the root growth with re-grafting process with any two trees that are close together (whether or not they are finite or subtrees of of a common tree), then the resulting processes will be close in some sense. The following result, which implies that the measure induced by the root growth with re-grafting process on path space is weakly continuous in the starting state with respect to the Skorohod topology on path space can be established by a considerably more intricate coupling argument: we refer to [63] for the details.

Proposition 5.6. If the function $f: \mathbf{T}^{\mathrm{root}} \rightarrow \mathbb{R}$ is continuous and bounded, then the function $P_{t} f$ is also continuous and bounded for each $t \geqslant 0$.

### 5.3 Ergodicity, recurrence, and uniqueness

### 5.3.1 Brownian CRT and root growth with re-grafting

Recall that Algorithm 2.4 for generating uniform rooted tree on $n$ labeled vertices was derived from Algorithm 5.1, the tree-valued Markov chain appearing in the proof of the Markov chain tree theorem that has the uniform rooted tree on $n$ labeled vertices as its stationary distribution. Recall also that the Poisson line-breaking construction of the Brownian continuum random tree in Section 2.5 is an asymptotic version of Algorithm 2.4, whilst the root growth with re-grafting process was motivated as an asymptotic version of Algorithm 5.1. Therefore, it seems reasonable that there should be a connection between the Poisson line-breaking construction and the root growth with re-grafting process. We establish the connection in this subsection.

Let us first present the Poisson line-breaking construction in a more "dynamic" way that will make the comparison with the root growth with regrafting process a little more transparent.

- Write $\tau_{1}, \tau_{2}, \ldots$ for the successive arrival times of an inhomogeneous Poisson process with arrival rate $t$ at time $t \geqslant 0$. Call $\tau_{n}$ the $n^{\text {th }}$ cut time
- Start at time 0 with the 1 -tree (that is a line segment with two ends), $\mathcal{R}_{0}$, of length zero ( $\mathcal{R}_{0}$ is "really" the trivial tree that consists of one point only, but thinking this way helps visualize the dynamics more clearly for this semi-formal description). Identify one end of $\mathcal{R}_{0}$ as the root.
- Let this line segment grow at unit speed until the first cut time $\tau_{1}$.
- At time $\tau_{1}$ pick a point uniformly on the segment that has been grown so far. Call this point the first cut point.
- Between time $\tau_{1}$ and time $\tau_{2}$, evolve a tree with 3 ends by letting a new branch growing away from the first cut point at unit speed.
- Proceed inductively: Given the $n$-tree (that is, a tree with $n+1$ ends), $\mathcal{R}_{\tau_{n}-}$, pick the $n$-th cut point uniformly on $\mathcal{R}_{\tau_{n}-}$ to give an $n+1$-tree, $\mathcal{R}_{\tau_{n}}$, with one edge of length zero, and for $t \in\left[\tau_{n}, \tau_{n+1}\left[\right.\right.$, let $\mathcal{R}_{t}$ be the tree obtained from $\mathcal{R}_{\tau_{n}}$ by letting a branch grow away from the $n^{\text {th }}$ cut point with unit speed.
The tree $\mathcal{R}_{\tau_{n}-}$ is $n^{\text {th }}$ step of the Poisson line-breaking construction, and the Brownian CRT is the limit of the increasing family of rooted finite trees $\left(\mathcal{R}_{t}\right)_{t \geqslant 0}$.

We will now use the ingredients appearing in the construction of $\mathcal{R}$ to construct a version of the root growth with re-grafting process started at the trivial tree.

- Let $\tau_{1}, \tau_{2}, \ldots$ be as in the construction of the $\mathcal{R}$.
- Start with the 1-tree (with one end identified as the root and the other as a leaf), $\mathcal{T}_{0}$, of length zero.
- Let this segment grow at unit speed on the time interval $\left[0, \tau_{1}[\right.$, and for $t \in\left[0, \tau_{1}\left[\right.\right.$ let $\mathcal{T}_{t}$ be the rooted 1 -tree that has its points labeled by the interval $[0, t]$ in such a way that the root is $t$ and the leaf is 0 .
- At time $\tau_{1}$ sample the first cut point uniformly along the tree $\mathcal{T}_{\tau_{1}-}$, prune off the piece of $\mathcal{I}_{\tau_{1}-}$ that is above the cut point (that is, prune off the interval of points that are further away from the root $t$ than the first cut point).
- Re-graft the pruned segment such that its cut end and the root are glued together. Just as we thought of $\mathcal{T}_{0}$ as a tree with two points, (a leaf and a root) connected by an edge of length zero, we take $\mathcal{I}_{\tau_{1}}$ to be the the rooted 2 -tree obtained by "ramifying" the root $\mathcal{I}_{\tau_{1}-}$ into two points (one of which we keep as the root) that are joined by an edge of length zero.
- Proceed inductively: Given the labeled and rooted $n$-tree, $\mathcal{T}_{\tau_{n-1}}$, for $t \in$ $\left[\tau_{n-1}, \tau_{n}\left[\right.\right.$, let $\mathcal{T}_{t}$ be obtained by letting the edge containing the root grow at unit speed so that the points in $\mathcal{I}_{t}$ correspond to the points in the interval $[0, t]$ with $t$ as the root. At time $\tau_{n}$, the $n^{\text {th }}$ cut point is sampled randomly along the edges of the $n$-tree, $\mathcal{T}_{\tau_{n}-}$, and the subtree above the cut point (that is the subtree of points further away from the root than the cut point) is pruned off and re-grafted so that its cut end and the root are glued together. The root is then "ramified" as above to give an edge of length zero leading from the root to the rest of the tree.

Let $\left(\mathcal{R}_{t}\right)_{t \geqslant 0},\left(\mathcal{T}_{t}\right)_{t \geqslant 0}$, and $\left\{\tau_{n}\right\}_{n \in \mathbb{N}}$ be as above. Note that $\left(\mathcal{T}_{t}\right)_{t \geqslant 0}$ has the same law as $\left(X_{t}\right)_{t \geqslant 0}$ under $\mathbf{P}^{T_{0}}$, where $T_{0}$ is the trivial tree.

Proposition 5.7. The two random finite rooted trees $\mathcal{R}_{\tau_{n}-}$ and $\mathcal{T}_{\tau_{n}-}$ have the same distribution for all $n \in \mathbb{N}$.

Proof. Let $R_{n}$ denote the object obtained by taking the rooted finite tree with edge lengths $\mathcal{R}_{\tau_{n}-}$ and labeling the leaves with $1, \ldots, n$, in the order that they are added in Aldous's construction. Let $T_{n}$ be derived similarly from the rooted finite tree with edge lengths $\mathcal{T}_{\tau_{n}-}$, by labeling the leaves with $1, \ldots, n$ in the order that they appear in the root growth with re-grafting construction. It will suffice to show that $R_{n}$ and $T_{n}$ have the same distribution. Note that both $R_{n}$ and $T_{n}$ are rooted bifurcating trees with $n$ labeled leaves and edge lengths. Such a tree $S_{n}$ is uniquely specified by its shape, denoted shape $\left(S_{n}\right)$, that is a rooted, bifurcating, leaf-labeled combinatorial tree, and by the list of its $(2 n-1)$ edge lengths in a canonical order determined by its shape, say

$$
\operatorname{lengths}\left(S_{n}\right):=\left(\operatorname{length}\left(S_{n}, 1\right), \ldots, \text { length }\left(S_{n}, 2 n-1\right)\right)
$$

where the edge lengths are listed in order of traversal of edges by first working along the path from the root to leaf 1, then along the path joining that path to leaf 2 , and so on.

Recall that $\tau_{n}$ is the $n$th point of a Poisson process on $\mathbb{R}^{++}$with rate $t d t$. We construct $R_{n}$ and $T_{n}$ on the same probability space using cuts at
points $U_{i} \tau_{i}, 1 \leqslant i \leqslant n-1$, where $U_{1}, U_{2}, \ldots$ is a sequence of independent random variables uniformly distributed on the interval ]0,1] and independent of the sequence $\left\{\tau_{n}\right\}_{n \in \mathbb{N}}$. Then, by construction, the common collection of edge lengths of $R_{n}$ and of $T_{n}$ is the collection of lengths of the $2 n-1$ subintervals of $] 0, \tau_{n}$ ] obtained by cutting this interval at the $2 n-2$ points

$$
\left\{X_{i}^{(n)}: 1 \leqslant i \leqslant 2 n-2\right\}:=\bigcup_{i=1}^{n-1}\left\{U_{i} \tau_{i}, \tau_{i}\right\}
$$

where the $X_{i}^{(n)}$ are indexed to increase in $i$ for each fixed $n$. Let $X_{0}^{(n)}:=0$ and $X_{2 n-1}^{(n)}:=\tau_{n}$. Then

$$
\begin{gather*}
\operatorname{length}\left(R_{n}, i\right)=X_{i}^{(n)}-X_{i-1}^{(n)}, \quad 1 \leqslant i \leqslant 2 n-1  \tag{5.5}\\
\operatorname{length}\left(T_{n}, i\right)=\operatorname{length}\left(R_{n}, \sigma_{n, i}\right), \quad 1 \leqslant i \leqslant 2 n-1 \tag{5.6}
\end{gather*}
$$

for some almost surely unique random indices $\sigma_{n, i} \in\{1, \ldots 2 n-1\}$ such that $i \mapsto \sigma_{n, i}$ is almost surely a permutation of $\{1, \ldots 2 n-1\}$. According to [10, Lemma 21], the distribution of $R_{n}$ may be characterized as follows:
(i) the sequence lengths $\left(R_{n}\right)$ is exchangeable, with the same distribution as the sequence of lengths of subintervals obtained by cutting $\left.] 0, \tau_{n}\right]$ at $2 n-2$ uniformly chosen points $\left\{U_{i} \tau_{n}: 1 \leqslant i \leqslant 2 n-2\right\}$;
(ii) shape $\left(R_{n}\right)$ is uniformly distributed on the set of all $1 \times 3 \times 5 \times \cdots \times(2 n-3)$ possible shapes;
(iii) lengths $\left(R_{n}\right)$ and $\operatorname{shape}\left(R_{n}\right)$ are independent.

In view of this characterization and (5.6), to show that $T_{n}$ has the same distribution as $R_{n}$ it is enough to show that
(a) the random permutation $\left\{i \mapsto \sigma_{n, i}: 1 \leqslant i \leqslant 2 n-1\right\}$ is a function of shape $\left(T_{n}\right)$;
(b) $\operatorname{shape}\left(T_{n}\right)=\Psi_{n}\left(\operatorname{shape}\left(R_{n}\right)\right)$ for some bijective map $\Psi_{n}$ from the set of all possible shapes to itself.

This is trivial for $n=1$, so we assume below that $n \geqslant 2$. Before proving (a) and (b), we recall that (ii) above involves a natural bijection

$$
\begin{equation*}
\left(I_{1}, \ldots, I_{n-1}\right) \leftrightarrow \operatorname{shape}\left(R_{n}\right) \tag{5.7}
\end{equation*}
$$

where $I_{n-1} \in\{1, \ldots, 2 n-3\}$ is the unique $i$ such that

$$
U_{n-1} \tau_{n-1} \in\left(X_{i-1}^{(n-1)}, X_{i}^{(n-1)}\right)
$$

Hence, $I_{n-1}$ is the index in the canonical ordering of edges of $R_{n-1}$ of the edge that is cut in the transformation from $R_{n-1}$ to $R_{n}$ by attachment of an additional edge, of length $\tau_{n}-\tau_{n-1}$, connecting the cut-point to leaf $n$. Thus, (ii) and (iii) above correspond via (5.7) to the facts that $I_{1}, \ldots, I_{n-1}$
are independent and uniformly distributed over their ranges, and independent of lengths $\left(R_{n}\right)$. These facts can be checked directly from the construction of $\left\{R_{n}\right\}_{n \in \mathbb{N}}$ from $\left\{\tau_{n}\right\}_{n \in \mathbb{N}}$ and $\left\{U_{n}\right\}_{n \in \mathbb{N}}$ using standard facts about uniform order statistics.

Now (a) and (b) follow from (5.7) and another bijection

$$
\begin{equation*}
\left(I_{1}, \ldots, I_{n-1}\right) \leftrightarrow \operatorname{shape}\left(T_{n}\right) \tag{5.8}
\end{equation*}
$$

where each possible value $i$ of $I_{m}$ is identified with edge $\sigma_{m, i}$ in the canonical ordering of edges of $T_{m}$. This is the edge of $T_{m}$ whose length equals length $\left(R_{m}, i\right)$. The bijection (5.8), and the fact that $\sigma_{n, i}$ depends only on $\operatorname{shape}\left(T_{n}\right)$, will now be established by induction on $n \geqslant 2$. For $n=2$ the claim is obvious. Suppose for some $n \geqslant 3$ that the correspondence between $\left(I_{1}, \ldots, I_{n-2}\right)$ and shape $\left(T_{n-1}\right)$ has been established, and that the length of edge $\sigma_{n-1, i}$ in the canonical ordering of edges of $T_{n-1}$ is equals the length of the $i$ th edge in the canonical ordering of edges of $R_{n-1}$, for some $\sigma_{n-1, i}$ that is a function of $i$ and shape $\left(T_{n-1}\right)$. According to the construction of $T_{n}$, if $I_{n-1}=i$ then $T_{n}$ is derived from $T_{n-1}$ by splitting $T_{n-1}$ into two branches at some point along edge $\sigma_{n-1, i}$ in the canonical ordering of the edges of $T_{n-1}$, and forming a new tree from the two branches and an extra segment of length $\tau_{n}-\tau_{n-1}$. Clearly, shape $\left(T_{n}\right)$ is determined by shape $\left(T_{n-1}\right)$ and $I_{n-1}$, and in the canonical ordering of the edge lengths of $T_{n}$ the length of the $i$ th edge equals the length of the edge $\sigma_{n, i}$ of $R_{n}$, for some $\sigma_{n, i}$ that is a function of shape $\left(T_{n-1}\right)$ and $I_{n-1}$, and, therefore, a function of shape $\left(T_{n}\right)$. To complete the proof, it is enough by the inductive hypothesis to show that the map

$$
\left(\operatorname{shape}\left(T_{n-1}\right), I_{n-1}\right) \rightarrow \operatorname{shape}\left(T_{n}\right)
$$

just described is invertible. But shape $\left(T_{n-1}\right)$ and $I_{n-1}$ can be recovered from shape $\left(T_{n}\right)$ by the following sequence of moves:

- delete the edge attached to the root of $\operatorname{shape}\left(T_{n}\right)$
- split the remaining tree into its two branches leading away from the internal node to which the deleted edge was attached;
- re-attach the bottom end of the branch not containing leaf $n$ to leaf $n$ on the other branch, joining the two incident edges to form a single edge;
- the resulting shape is shape $\left(T_{n-1}\right)$, and $I_{n-1}$ is the index such that the joined edge in shape $\left(T_{n-1}\right)$ is the edge $\sigma_{n-1, I_{n-1}}$ in the canonical ordering of edges on shape $\left(T_{n-1}\right)$.


### 5.3.2 Coupling

Lemma 5.8. For any $(T, d, \rho) \in \mathbf{T}^{\text {root }}$ we can build on the same probability space two $\mathbf{T}^{\text {root }}$-valued processes $X^{\prime}$ and $X^{\prime \prime}$ such that:

- $X^{\prime}$ has the law of $X$ under $\mathbf{P}^{T_{0}}$, where $T_{0}$ is the trivial tree consisting of just the root,
- $X^{\prime \prime}$ has the law of $X$ under $\mathbf{P}^{T}$,
- for all $t \geqslant 0$,

$$
\begin{equation*}
d_{\mathrm{GH}^{\mathrm{root}}}\left(X_{t}^{\prime}, X_{t}^{\prime \prime}\right) \leqslant d_{\mathrm{GH}^{\mathrm{root}}}\left(T_{0}, T\right)=\sup \{d(\rho, x): x \in T\} \tag{5.9}
\end{equation*}
$$

$$
\begin{equation*}
\lim _{t \rightarrow \infty} d_{\mathrm{GH}^{\text {root }}}\left(X_{t}^{\prime}, X_{t}^{\prime \prime}\right)=0, \quad \text { almost surely } \tag{5.10}
\end{equation*}
$$

Proof. The proof follows almost immediately from construction of $X$ and Lemma 5.4. The only point requiring some comment is (5.10).

For that it will be enough to show for any $\varepsilon>0$ that for $\mathbf{P}^{T}$-a.e. $\left(T, \pi_{0}, \pi\right) \in$ $\Omega$ there exists $t>0$ such that the projection of $\left.\left.\pi_{0} \cap(] 0, t\right] \times T^{o}\right)$ onto $T$ is an $\varepsilon$-net for $T$.

Note that the projection of $\left.\left.\pi_{0} \cap(] 0, t\right] \times T^{o}\right)$ onto $T$ is a Poisson process under $\mathbf{P}^{T}$ with intensity $t \mu$, where $\mu$ is the length measure on $T$. Moreover, $T$ can be covered by a finite collection of $\varepsilon$-balls, each with positive $\mu$-measure.

Therefore, the $\mathbf{P}^{T}$-probability of the set of $\left(T, \pi_{0}, \pi\right) \in \Omega$ such that the projection of $\left.\left.\pi_{0} \cap(] 0, t\right] \times T^{o}\right)$ onto $T$ is an $\varepsilon$-net for $T$ increases as $t \rightarrow \infty$ to 1.

### 5.3.3 Convergence to equilibrium

Proposition 5.9. For any $T \in \mathbf{T}^{\text {root }}$, the law of $X_{t}$ under $\mathbf{P}^{T}$ converges weakly to that of the Brownian CRT as $t \rightarrow \infty$.

Proof. It suffices by Lemma 5.8 to consider the case where $T$ is the trivial tree.

We saw in the Proposition 5.7 that, in the notation of that result, $\mathcal{T}_{\tau_{n}-}$ has the same distribution as $\mathcal{R}_{\tau_{n}-\text {. }}$.

Moreover, $\mathcal{R}_{t}$ converges in distribution to the continuum random tree as $t \rightarrow \infty$ if we use Aldous's metric on trees that comes from thinking of them as closed subsets of $\ell^{1}$ with the root at the origin and equipped with the Hausdorff distance.

By construction, $\left(\mathcal{T}_{t}\right)_{t \geqslant 0}$ has the root growth with re-grafting dynamics started at the trivial tree. Clearly, the rooted Gromov-Hausdorff distance between $\mathcal{T}_{t}$ and $\mathcal{T}_{\tau_{n+1}-}$ is at most $\tau_{n+1}-\tau_{n}$ for $\tau_{n} \leqslant t<\tau_{n+1}$.

It remains to observe that $\tau_{n+1}-\tau_{n} \rightarrow 0$ in probability as $n \rightarrow \infty$.

### 5.3.4 Recurrence

Proposition 5.10. Consider a non-empty open set $U \subseteq \mathbf{T}^{\text {root }}$. For each $T \in$ $\mathbf{T}^{\text {root }}$,

$$
\begin{equation*}
\mathbf{P}^{T}\left\{\text { for all } s \geqslant 0, \text { there exists } t>s \text { such that } X_{t} \in U\right\}=1 \tag{5.11}
\end{equation*}
$$

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Proof. It is straightforward, but notationally rather tedious, to show that if $B^{\prime} \subseteq \mathbf{T}^{\text {root }}$ is any ball and $T_{0}$ is the trivial tree, then

$$
\begin{equation*}
\mathbf{P}^{T_{0}}\left\{X_{t} \in B^{\prime}\right\}>0 \tag{5.12}
\end{equation*}
$$

for all $t$ sufficiently large.
Thus, for any ball $B^{\prime} \subseteq \mathbf{T}^{\text {root }}$ there is, by Lemma 5.8 , a ball $B^{\prime \prime} \subseteq \mathbf{T}^{\text {root }}$ containing the trivial tree such that

$$
\begin{equation*}
\inf _{T \in B^{\prime \prime}} \mathbf{P}^{T}\left\{X_{t} \in B^{\prime}\right\}>0 \tag{5.13}
\end{equation*}
$$

for each $t$ sufficiently large.
By a standard application of the Markov property, it therefore suffices to show for each $T \in \mathbf{T}^{\text {root }}$ and each ball $B^{\prime \prime}$ around the trivial tree that

$$
\begin{equation*}
\mathbf{P}^{T}\left\{\text { there exists } t>0 \text { such that } X_{t} \in B^{\prime \prime}\right\}=1 \tag{5.14}
\end{equation*}
$$

By another standard application of the Markov property, equation (5.14) will follow if we can show that there is a constant $p>0$ depending on $B^{\prime \prime}$ such that for any $T \in \mathbf{T}^{\text {root }}$

$$
\liminf _{t \rightarrow \infty} \mathbf{P}^{T}\left\{X_{t} \in B^{\prime \prime}\right\}>p
$$

This, however, follows from Proposition 5.9 and the observation that for any $\varepsilon>0$ the law of the Brownian CRT assigns positive mass to the set of trees with height less than $\varepsilon$ : this is just the observation that the law of the Brownian excursion assigns positive mass to the set of excursion paths with maximum less that $\varepsilon / 2$.

### 5.3.5 Uniqueness of the stationary distribution

Proposition 5.11. The law of the Brownian CRT is the unique stationary distribution for $X$. That is, if $\xi$ is the law of the CRT, then

$$
\int \xi(d T) P_{t} f(T)=\int \xi(d T) f(T)
$$

for all $t \geqslant 0$ and $f \in \mathrm{~b} \mathcal{B}\left(\mathbf{T}^{\mathrm{root}}\right)$, and $\xi$ is the unique probability measure on $\mathbf{T}^{\text {root }}$ with this property.

Proof. This is a standard argument given Proposition 5.9 and the Feller property for the semigroup $\left(P_{t}\right)_{t \geqslant 0}$ established in Proposition 5.6, but we include the details for completeness.

Consider a test function $f: \mathbf{T}^{\text {root }} \rightarrow \mathbb{R}$ that is continuous and bounded. By Proposition 5.6, the function $P_{t} f$ is also continuous and bounded for each $t \geqslant 0$.

Therefore, by Proposition 5.9,

$$
\begin{align*}
\int \xi(d T) f(T) & =\lim _{s \rightarrow \infty} \int \xi(d T) P_{s} f(T)=\lim _{s \rightarrow \infty} \int \xi(d T) P_{s+t} f(T)  \tag{5.15}\\
& =\lim _{s \rightarrow \infty} \int \xi(d T) P_{s}\left(P_{t} f\right)(T)=\int \xi(d T) P_{t} f(T)
\end{align*}
$$

for each $t \geqslant 0$. Hence, $\xi$ is stationary.
Moreover, if $\zeta$ is a stationary measure, then

$$
\begin{align*}
\int \zeta(d T) f(T) & =\int \zeta(d T) P_{t} f(T)  \tag{5.16}\\
& \rightarrow \int \zeta(d T)\left(\int \xi(d T) f(T)\right)=\int \xi(d T) f(T)
\end{align*}
$$

and $\zeta=\xi$, as claimed.

### 5.4 Convergence of the Markov chain tree algorithm

We would like to show that Algorithm 5.1 converges to a process having the root growth with re-grafting dynamics after suitable re-scaling of time and edge lengths of the evolving tree. It will be more convenient for us to work with the continuous time version of the algorithm in which the transitions are made at the arrival times of an independent Poisson process with rate 1.

The continuous time version of Algorithm 5.1 involves a labeled combinatorial tree, but, by symmetry, if we don't record the labeling and associate rooted labeled combinatorial trees with rooted compact real trees having edges that are line segments with length 1 , then the resulting process will still be Markovian.

It will be convenient to use the following notation for re-scaling the distances in a $\mathbb{R}$-tree: $T=(T, d, \rho)$ is a rooted compact real tree and $c>0$, we write $c T$ for the tree $(T, c d, \rho)$ (that is, $c T=T$ as sets and the roots are the same, but the metric is re-scaled by $c$ ).

Proposition 5.12. Let $Y^{n}=\left(Y_{t}^{n}\right)_{t \geqslant 0}$ be a sequence of Markov processes that take values in the space of rooted compact real trees with integer edge lengths and evolve according to the dynamics associated with the continuoustime version of Algorithm 5.1. Suppose that each tree $Y_{0}^{n}$ is non-random with total branch length $N_{n}$, that $N_{n}$ converges to infinity as $n \rightarrow \infty$, and that $N_{n}^{-1 / 2} Y_{0}^{n}$ converges in the rooted Gromov-Hausdorff metric to some rooted compact real tree $T$ as $n \rightarrow \infty$. Then, in the sense of weak convergence of processes on the space of càdlàg paths equipped with the Skorohod topology, $\left(N_{n}^{-1 / 2} Y^{n}\left(N_{n}^{1 / 2} t\right)\right)_{t \geqslant 0}$ converges as $n \rightarrow \infty$ to the root growth with re-grafting process $X$ under $\mathbf{P}^{T}$.
Proof. Define $Z^{n}=\left(Z_{t}^{n}\right)_{t \geqslant 0}$ by

$$
Z_{t}^{n}:=N_{n}^{-1 / 2} Y^{n}\left(N_{n}^{1 / 2} t\right)
$$

For $\eta>0$, let $Z^{\eta, n}$ be the $\mathbf{T}^{\text {root }}$-valued process constructed as follows.

- Set $Z_{0}^{\eta, n}=R_{\eta_{n}}\left(Z_{0}^{n}\right)$, where $\eta_{n}:=N_{n}^{-1 / 2}\left\lfloor N_{n}^{1 / 2} \eta\right\rfloor$.
- The value of $Z^{\eta, n}$ is unchanged between jump times of $\left(Z_{t}^{n}\right)_{t \geqslant 0}$.
- At a jump time $\tau$ for $\left(Z_{t}^{n}\right)_{t \geqslant 0}$, the tree $Z_{\tau}^{\eta, n}$ is the subtree of $Z_{\tau}^{n}$ spanned by $Z_{\tau-}^{\eta, n}$ and the root of $Z_{\tau}^{n}$.

An argument similar to that in the proof of Lemma 5.4 shows that

$$
\sup _{t \geqslant 0} d_{\mathrm{H}}\left(Z_{t}^{n}, Z_{t}^{\eta, n}\right) \leqslant \eta_{n}
$$

and so it suffices to show that $Z^{\eta, n}$ converges weakly as $n \rightarrow \infty$ to $X$ under $\mathbf{P}^{R_{\eta}(T)}$.

Note that $Z_{0}^{\eta, n}$ converges to $R_{\eta}(T)$ as $n \rightarrow \infty$. Moreover, if $\Lambda$ is the map that sends a tree to its total length (that is, the total mass of its length measure), then $\lim _{n \rightarrow \infty} \Lambda\left(Z_{0}^{\eta, n}\right)=\Lambda \circ R_{\eta}(T)<\infty$ by Lemma 4.36 below.

The pure jump process $Z^{\eta, n}$ is clearly Markovian. If it is in a state $\left(T^{\prime}, \rho^{\prime}\right)$, then it jumps with the following rates.

- With rate $N_{n}^{1 / 2}\left(N_{n}^{1 / 2} \Lambda\left(T^{\prime}\right)\right) / N_{n}=\Lambda\left(T^{\prime}\right)$, one of the $N_{n}^{1 / 2} \Lambda\left(T^{\prime}\right)$ points in $T^{\prime}$ that are at distance a positive integer multiple of $N_{n}^{-1 / 2}$ from the root $\rho^{\prime}$ is chosen uniformly at random and the subtree above this point is joined to $\rho^{\prime}$ by an edge of length $N_{n}^{-1 / 2}$. The chosen point becomes the new root and a segment of length $N_{n}^{-1 / 2}$ that previously led from the new root toward $\rho^{\prime}$ is erased. Such a transition results in a tree with the same total length as $T^{\prime}$.
- With rate $N_{n}^{1 / 2}-\Lambda\left(T^{\prime}\right)$, a new root not present in $T^{\prime}$ is attached to $\rho^{\prime}$ by an edge of length $N_{n}^{-1 / 2}$. This results in a tree with total length $\Lambda\left(T^{\prime}\right)+N_{n}^{-1 / 2}$.

It is clear that these dynamics converge to those of the root growth with regrafting process, with the first class of transitions leading to re-graftings in the limit and the second class leading to root growth.

## The wild chain and other bipartite chains

### 6.1 Background

The wild chain was introduced informally in Chapter 1. We will now describe it more precisely.

The state space of the wild chain is the set $\mathbf{T}^{*}$ consisting of rooted $\mathbb{R}$-trees such that each edge has length 1, each vertex has finite degree, and if the tree is infinite there is a single infinite length path from the root. Let $\mu$ denote the PGW(1) measure (that is, the distribution of the Galton-Watson tree with mean 1 Poisson offspring distribution) on the set $\mathbf{T}_{<\infty}$ of finite trees in $\mathbf{T}^{*}$, and let $\nu$ denote the distribution of a PGW(1) tree "conditioned to be infinite". It is well-known that $\nu$ is concentrated on the set $\mathbf{T}_{\infty}^{*}:=\mathbf{T}^{*} \backslash \mathbf{T}_{<\infty}$ consisting of infinite trees with a single infinite path from the root. A realization of $\nu$ may be constructed by taking a semi-infinite path, thought of as infinitely many vertices connected by edges of length 1 and appending independent realizations of $\mu$ at each vertex. When started in a finite tree from $\mathbf{T}_{<\infty}$, at rate one for each vertex the wild chain attaches that vertex by an edge to the root of a realization of $\nu$. Conversely (and somewhat heuristically), when started in an infinite tree from $\mathbf{T}_{\infty}^{*}$, at rate one for each vertex the wild chain prunes off and discards the infinite subtree above that vertex, leaving a finite tree.

The set of times when the state of the wild chain is an infinite tree has Lebesgue measure zero, but it is the uncountable set of points of increase of a continuous additive functional (so that it looks qualitatively like the zero set of a Brownian motion).

The aim of this chapter is to use Dirichlet form methods to construct and study a general class of symmetric Markov processes on a generic totally disconnected state space. Specializing this construction leads to a class of processes that we call bipartite chains. This class contains the wild chain as a special case.

In general, we take the state space of the processes we construct to be a Lusin space $E$ such that there exists a countable algebra $\mathcal{R}$ of simultaneously
closed and open subsets of $E$ that is a base for the topology of $E$. Note that $E$ is indeed totally disconnected - see Theorem 33.B of [129]. Conversely, if $E$ is any totally disconnected compact metric space, then there exists a collection $\mathcal{R}$ with the required properties - see Theorem 2.94 of [85].

The following are two instances of such spaces. More examples, including an arbitrary local field and the compactification of an infinite tree, are described in Section 6.2.

Example 6.1. Let $E$ be $\overline{\mathbb{N}}:=\mathbb{N} \cup\{\infty\}$, the usual one-point compactification of the positive integers $\mathbb{N}:=\{1,2, \ldots\}$. Equip $E$ with the usual total order and let $\mathcal{R}$ be the algebra generated by sets of the form $\{y: x \leqslant y\}, x \in \mathbb{N}$. That is, $\mathcal{R}$ consists of finite subsets of $\mathbb{N}$ and sets that contain a subset of the form $\{z, z+1, z+2, \ldots, \infty\}$ for $z \in \mathbb{N}$.

Example 6.2. Let $E$ be the collection $\mathbf{T}_{\leqslant \infty}$ of rooted trees with every vertex having finite degree. Write $\mathbf{T}_{\leqslant n}$ for the subset of $\mathbf{T}_{\leqslant \infty}$ consisting of trees with height at most $n$. For $m>n$, there is a natural projection map from $\rho_{m n}: \mathbf{T}_{\leqslant m} \rightarrow \mathbf{T}_{\leqslant n}$ that throws away vertices of height greater than $n$ and the edges leading to them. We can identify $\mathbf{T}_{\leqslant \infty}$ with the projective limit of this projective system and give it the corresponding projective limit topology (each $\mathbf{T}_{\leqslant n}$ is given the discrete topology), so that $\mathbf{T}_{\leqslant \infty}$ is Polish. Equip $\mathbf{T}_{\leqslant \infty}$ with the inclusion partial order (that is, $x \leqslant y$ if $x$ is a sub-tree of $y$ ). Let $\mathcal{R}$ be the algebra generated by sets of the form $\{y: x \leqslant y\}, x \in \mathbf{T}_{<\infty}:=\bigcup_{n} \mathbf{T}_{\leqslant n}$. Equivalently, if $\rho_{n}: \mathbf{T}_{\leqslant \infty} \rightarrow \mathbf{T}_{\leqslant n}$ is the projection map that throws away vertices of height greater than $n$ and the edges leading to them, then $\mathcal{R}$ is the collection of sets of the form $\rho_{n}^{-1}(B)$ for finite or co-finite $B \subseteq \mathbf{T}_{\leqslant n}$, as $n$ ranges over $\mathbb{N}$.

Our main existence result is the following. We prove it in Section 6.3. Appendix A contains a summary of the relevant Dirichlet form theory.

Notation 6.3. Denote by $\mathcal{C}$ the subalgebra of $b C(E)(:=$ continuous bounded functions on $E$ ) generated by the indicator functions of sets in $\mathcal{R}$.

Theorem 6.4. Consider two probability measures $\mu$ and $\nu$ on $E$ and a nonnegative Borel function $\kappa$ on $E \times E$. Define a $\sigma$-finite measure $\Lambda$ on $E \times E$ by $\Lambda(d x, d y):=\kappa(x, y) \mu(d x) \nu(d y)$. Suppose that the following hold:
(a) the closed support of the measure $\mu$ is $E$;
(b) $\Lambda([(E \backslash R) \times R] \cup[R \times(E \backslash R)])<\infty$ for all $R \in \mathcal{R}$;
(c) $\int \kappa(x, y) \mu(d x)=\infty$ for $\nu_{s}$-a.e. $y$, where $\nu_{s}$ is the singular component in the Lebesgue decomposition of $\nu$ with respect to $\mu$;
(d) there exists a sequence $\left(R_{n}\right)_{n \in \mathbb{N}}$ of sets in $\mathcal{R}$ such that $\bigcap_{m=n}^{\infty} R_{m}$ is compact for all $n, \sum_{n \in \mathbb{N}} \mu\left(E \backslash R_{n}\right)<\infty$, and

$$
\sum_{n \in \mathbb{N}} \Lambda\left(\left[\left(E \backslash R_{n}\right) \times R_{n}\right] \cup\left[R_{n} \times\left(E \backslash R_{n}\right)\right]\right)<\infty
$$

Then there is a recurrent $\mu$-symmetric Hunt process $\mathbf{X}=\left(X_{t}, \mathbb{P}^{x}\right)$ on $E$ whose Dirichlet form is the closure of the form $\mathcal{E}$ on $\mathcal{C}$ defined by

$$
\mathcal{E}(f, g)=\iint(f(y)-f(x))(g(y)-g(x)) \Lambda(d x, d y), f, g \in \mathcal{C}
$$

Our standing assumption throughout this chapter is that the conditions of Theorem 6.4 hold.

In order to produce processes that are reminiscent of the wild chain, we need to assume a little more structure on $E$. Say that $E$ is bipartite if there is a countable, dense subset $E^{o} \subseteq E$ such that each point of $E^{o}$ is isolated. In particular, $E^{o}$ is open. In Example 6.1 we can take $E^{o}=\mathbb{N}$. In Example 6.2 we can take $E^{o}=T_{<\infty}$. We will see more examples in Section 6.2. Put $E^{*}=E \backslash E^{o}$. Note that $E^{*}$ is the boundary of the open set $E^{o}$.

Definition 6.5. We will call the process $\mathbf{X}$ described in Theorem 6.4 a bipartite Markov chain if the space $E$ is bipartite and, in the notation of Theorem 6.4:

- $\mu$ is concentrated on $E^{o}$,
- $\quad \nu$ is concentrated on $E^{*}$.

Remark 6.6. For bipartite chains, the measures $\mu$ and $\nu$ are mutually singular and $\nu_{s}=\nu$ in the notation of Theorem 6.4. The reference measure $\mu$ is invariant for $\mathbf{X}$, that is, $\mathbb{P}^{\mu}\left\{X_{t} \in \cdot\right\}=\mu$ for each $t \geqslant 0$. Thus, for any $x \in E^{o}$ we have $\mathbb{P}^{x}\left\{X_{t} \in E^{o}\right\}=1$ for each $t \geqslant 0$, and so $\mathbf{X}$ is Markov chain on the countable set $E^{o}$ in the same sense that the Feller-McKean chain is a Markov chain on the rationals (the Feller-McKean chain is one-dimensional Brownian motion time-changed by a continuous additive functional that has as its Revuz measure a purely atomic probability measure that assigns positive mass to each rational).

We establish in Proposition 6.14 that the sample-paths of $\mathbf{X}$ bounce backwards and forwards between $E^{o}$ and $E^{*}$ in the same manner that the sample paths of the wild chain bounce backwards and forwards between the finite and infinite trees. Also, we show in Proposition 6.16 that under suitable conditions $\mu$ is the unique invariant distribution for $\mathbf{X}$ that assigns all of its mass to $E^{o}$, and, moreover, for any probability measure $\gamma$ concentrated on $E^{o}$ the law of $X_{t}$ under $\mathbb{P}^{\gamma}$ converges in total variation to $\mu$ as $t \rightarrow \infty$.

In Section 6.6 we prove that, in the general setting of Theorem 6.4, the measure $\nu$ is the Revuz measure of a positive continuous additive functional (PCAF). We can, therefore, time-change $\mathbf{X}$ using the inverse of this PCAF. When this procedure is applied to a bipartite chain, it produces a Markov process with state space that is a subset of $E^{*}$. In particular, we observe in Example 6.24 that instances of this time-change construction lead to "spherically symmetric" Lévy processes on local fields.

A useful tool for proving the last fact is a result from Section 6.5. There we consider a certain type of equivalence relation on $E$ with associated map
$\pi$ onto the corresponding quotient space. We give conditions on the Dirichlet form $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ that are sufficient for the process $\pi \circ \mathbf{X}$ to be a symmetric Hunt process.

Notation 6.7. Write $(\cdot, \cdot)_{\mu}$ for the $L^{2}(E, \mu)$ inner product and $\left(T_{t}\right)_{t \geqslant 0}$ for the semigroup on $L^{2}(E, \mu)$ associated with the form $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$.

### 6.2 More examples of state spaces

Example 6.8. Let $E$ be the usual path-space of a discrete-time Markov chain with countable state-space $S$ augmented by a distinguished cemetery state $\partial$ to form $S^{\partial}=S \cup\{\partial\}$. That is, $E$ is the subset of the space of sequences $\left(S^{\partial}\right)^{\mathbb{N}_{0}}$ (where $\left.\mathbb{N}_{0}:=\{0,1,2, \ldots\}\right)$ consisting of sequences $\left\{x_{n}\right\}_{n \in \mathbb{N}_{0}}$ such that if $x_{n}=\partial$ for some $n$, then $x_{m}=\partial$ for all $m>n$. Give $E$ the subspace topology inherited from the product topology on $\left(S^{\partial}\right)^{\mathbb{N}_{0}}$ (where each factor has the discrete topology), so that $E$ is Polish. Given $x \in E$, write $\zeta(x):=$ $\inf \left\{n: x_{n}=\partial\right\} \in \mathbb{N}_{0} \cup\{\infty\}$ for the death-time of $x$. Define a partial order on $E$ by declaring that $x \leqslant y$ if $\zeta(x) \leqslant \zeta(y)$ and $x_{n}=y_{n}$ for $0 \leqslant n<\zeta(x)$. (In particular, if $x$ and $y$ are such that $\zeta(x)=\zeta(y)=\infty$, then $x \leqslant y$ if and only if $x=y$.) Let $\mathcal{R}$ be the algebra generated by sets of the form $\{y: x \leqslant y\}$, $\zeta(x)<\infty$. When $\# S=k<\infty$, we can think of $E$ as the regular $k$-ary rooted tree along with its set of ends. In particular, when $k=1$ we recover Example 6.1. This example is bipartite with $E^{o}=\{x: \zeta(x)<\infty\}$,

Example 6.9. A local field $\mathbb{K}$ is a locally compact, non-discrete, totally disconnected, topological field. We refer the reader to [135] or [123] for a full discusion of these objects and for proofs of the facts outlined below. More extensive summaries and references to the literature on probability in a local field setting can be found in [58] and [62].

There is a real-valued mapping on $\mathbb{K}$ that we denote by $x \mapsto|x|$. This map, called the valuation takes the values $\left\{q^{k}: k \in \mathbb{Z}\right\} \cup\{0\}$, where $q=p^{c}$ for some prime $p$ and positive integer $c$ and has the properties

$$
\begin{gathered}
|x|=0 \Leftrightarrow x=0 \\
|x y|=|x||y| \\
|x+y| \leqslant|x| \vee|y| .
\end{gathered}
$$

The mapping $(x, y) \mapsto|x-y|$ on $\mathbb{K} \times \mathbb{K}$ is a metric on $\mathbb{K}$ that gives the topology of $K$.

Put $\mathbb{D}=\{x:|x| \leqslant 1\}$. The set $\mathbb{D}$ is a ring (the so-called ring of integers of $K)$. If we choose $\rho \in K$ so that $|\rho|=q^{-1}$, then

$$
\rho^{k} \mathbb{D}=\left\{x:|x| \leqslant q^{-k}\right\}=\left\{x:|x|<q^{-(k-1)}\right\} .
$$

Every ball is of the form $x+\rho^{k} \mathbb{D}$ for some $x \in \mathbb{K}$ and $k \in \mathbb{Z}$, and, in particular, all balls are both closed and open. For $\ell<k$ the additive quotient group
$\rho^{\ell} \mathbb{D} / \rho^{k} \mathbb{D}$ has order $q^{k-\ell}$. Consequently, $\mathbb{D}$ is the union of $q$ disjoint translates of $\rho \mathbb{D}$. Each of these components is, in turn, the union of $q$ disjoint translates of $\rho^{2} \mathbb{D}$, and so on. Thus, we can think of the collection of balls contained in $\mathbb{D}$ as being arranged in an infinite rooted $q$-ary tree: the root is $\mathbb{D}$ itself, the nodes at level $k$ are the balls of radius $q^{-k}\left(=\operatorname{cosets}\right.$ of $\left.\rho^{k} \mathbb{D}\right)$, and the $q$ "children" of such a ball are the $q$ cosets of $\rho^{k+1} \mathbb{D}$ that it contains. We can uniquely associate each point in $\mathbb{D}$ with the sequence of balls that contain it, and so we can think of the points in $\mathbb{D}$ as the ends this tree - see Figure 6.1.


Fig. 6.1. Schematic drawing of the ring of integers $\mathbb{D}$ when $q=p=7$

This tree picture alone does not capture all the algebraic structure of $\mathbb{D}$; the rings of integers for the $p$-adic numbers and the $p$-series field (that is, the field of formal Laurent series with coefficients drawn from the finite field with $p$ elements) are both represented by a $p$-ary tree, even though the $p$-adic field has characteristic 0 whereas the $p$-series field has characteristic $p$. (As an aside, a locally compact, non-discrete, topological field that is not totally disconnected is necessarily either the real or the complex numbers. Every local field is either a finite algebraic extension of the $p$-adic number field for some prime $p$ or a finite algebraic extension of the $p$-series field.)

We can take either $E=\mathbb{K}$ or $E=\mathbb{D}$, with $\mathcal{R}$ the algebra generated by the balls. The same comment applies to Banach spaces over local fields defined as in [123], and we leave the details to the reader.

Example 6.10. In the notation of Example 6.2, let $\mathbf{T}_{\infty}^{*}$ be the subset of $\mathbf{T}_{\leqslant \infty}$ consisting of infinite trees through which there is a unique infinite path starting at the root, that is, trees with only one end. Put $\mathbf{T}^{*}=\mathbf{T}_{<\infty} \cup \mathbf{T}_{\infty}^{*}$. It is not hard to see that $E=\mathbf{T}^{*}$ satisfies our hypothesis, with $\mathcal{R}$ the trace on $\mathbf{T}^{*}$ of the algebra of subsets of $\mathbf{T}_{\leqslant \infty}$ described in Example 6.2.

Example 6.11. Suppose that the pairs $\left(E_{1}, \mathcal{R}_{1}\right), \ldots,\left(E_{N}, \mathcal{R}_{N}\right)$ each satisfy our hypotheses. Put $E:=\prod_{i} E_{i}$, equip $E$ with the product topology, and set $\mathcal{R}$ to be the algebra generated by subsets of $E$ of the form $\prod_{i} R_{i}$ with $R_{i} \in \mathcal{R}_{i}$. If each of the factors $E_{i}$ is bipartite with corresponding countable dense sets of isolated point $E_{i}^{o}$, then $E$ is also bipartite with countable dense set of isolated points $\prod_{i} E_{i}^{o}$. Similar observations holds for sums rather than products, and we leave the details to the reader.

### 6.3 Proof of Theorem 6.4

We first check that $\mathcal{E}$ is well-defined on $\mathcal{C}$. Any $f \in \mathcal{C}$ can be written $f=$ $\sum_{i=1}^{N} a_{i} \mathbf{1}_{R_{i}}$ for suitable $R_{i} \in \mathcal{R}$ and constants $a_{i}$, and condition (b) is just the condition that $\mathcal{E}\left(\mathbf{1}_{R}, \mathbf{1}_{R}\right)<\infty$ for all $R \in \mathcal{R}$. It is clear that $\mathcal{E}$ is a symmetric, non-negative, bilinear form on $\mathcal{C}$.

We next check that $\mathcal{E}$ defined on $\mathcal{C}$ is closable (as a form on $L^{2}(E, \mu)$ ). Let $\left(f_{n}\right)_{n \in \mathbb{N}}$ be a sequence in $\mathcal{C}$ such that

$$
\begin{equation*}
\lim _{n \rightarrow \infty}\left(f_{n}, f_{n}\right)_{\mu}=0 \tag{6.1}
\end{equation*}
$$

and

$$
\begin{equation*}
\lim _{m, n \rightarrow \infty} \mathcal{E}\left(f_{m}-f_{n}, f_{m}-f_{n}\right)=0 \tag{6.2}
\end{equation*}
$$

We need to show that

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \mathcal{E}\left(f_{n}, f_{n}\right)=0 \tag{6.3}
\end{equation*}
$$

Put $\Lambda_{s}(d x, d y)=\kappa(x, y) \mu(d x) \nu_{s}(d y)$. For $M>0$ put $\Lambda^{M}(d x, d y)=$ $[\kappa(x, y) \wedge M] \mu(d x) \nu(d y)$ and $\Lambda_{s}^{M}(d x, d y)=[\kappa(x, y) \wedge M] \mu(d x) \nu_{s}(d y)$. From (6.1) we have

$$
\lim _{m, n \rightarrow \infty} \iint\left(f_{m}(x)-f_{n}(x)\right)^{2} \Lambda_{s}^{M}(d x, d y)=0, \forall M>0
$$

and from (6.2) we have

$$
\begin{equation*}
\lim _{m, n \rightarrow \infty} \iint\left(\left\{f_{m}(y)-f_{n}(y)\right\}-\left\{f_{m}(x)-f_{n}(x)\right\}\right)^{2} \Lambda^{M}(d x, d y)=0, \forall M>0 \tag{6.4}
\end{equation*}
$$

So, by Minkowski's inequality,

$$
\begin{equation*}
\lim _{m, n \rightarrow \infty} \iint\left(f_{m}(y)-f_{n}(y)\right)^{2} \Lambda_{s}^{M}(d x, d y)=0, \forall M>0 \tag{6.5}
\end{equation*}
$$

Thus, by (6.1), (6.5) and (c), there exists a Borel function $f$ and a sequence $\left(n_{k}\right)_{k \in \mathbb{N}}$ such that $\lim _{k \rightarrow \infty} f_{n_{k}}=0, \mu$-a.e. (and, therefore, $\nu_{a}$-a.e., where $\nu_{a}=$ $\nu-\nu_{s}$ is the absolutely continuous component in the Lebesgue decomposition of $\nu$ with respect to $\mu$ ), and $\lim _{k \rightarrow \infty} f_{n_{k}}=f, \nu_{s}$-a.e.

Now, by Fatou, (6.2) and Minkowski's inequality,

$$
\begin{aligned}
\iint f^{2}(y) \Lambda_{s}(d x, d y) & =\iint \lim _{k \rightarrow \infty}\left(f_{n_{k}}(y)-f_{n_{k}}(x)\right)^{2} \Lambda_{s}(d x, d y) \\
& \leqslant \liminf _{k \rightarrow \infty} \iint\left(f_{n_{k}}(y)-f_{n_{k}}(x)\right)^{2} \Lambda_{s}(d x, d y) \\
& <\infty
\end{aligned}
$$

and so, by (c), $f=0, \nu_{s}$-a.e. Finally, by Fatou and (6.2),

$$
\begin{aligned}
& \lim _{m \rightarrow \infty} \iint\left(f_{m}(y)-f_{m}(x)\right)^{2} \Lambda(d x, d y) \\
& \quad=\lim _{m \rightarrow \infty} \iint \lim _{k \rightarrow \infty}\left(\left\{f_{m}(y)-f_{n_{k}}(y)\right\}-\left\{f_{m}(x)-f_{n_{k}}(x)\right\}\right)^{2} \Lambda(d x, d y) \\
& \quad \leqslant \lim _{m \rightarrow \infty} \liminf _{k \rightarrow \infty} \iint\left(\left\{f_{m}(y)-f_{n_{k}}(y)\right\}-\left\{f_{m}(x)-f_{n_{k}}(x)\right\}\right)^{2} \Lambda(d x, d y) \\
& \quad=0
\end{aligned}
$$

as required.
Write $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ for the closure of the form $(\mathcal{E}, \mathcal{C})$. To complete the proof that $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ is a Dirichlet form, it only remains to show that this form is Markov. By Theorem A.7, this will be accomplished if we can show that the unit contraction acts on $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$. That is, we have to show for any $f \in \mathcal{C}$ that

$$
\begin{equation*}
(f \vee 0) \wedge 1 \in \mathcal{C} \tag{6.6}
\end{equation*}
$$

and

$$
\begin{equation*}
\mathcal{E}((f \vee 0) \wedge 1,(f \vee 0) \wedge 1) \leqslant \mathcal{E}(f, f) \tag{6.7}
\end{equation*}
$$

Considering claim (6.6), first observe that $f \in \mathcal{C}$ if and only if there exist pairwise disjoint $R_{1}, \ldots, R_{N}$ and constants $a_{1}, \ldots, a_{N}$ such that $f=\sum_{i} a_{i} \mathbf{1}_{R_{i}}$. Thus,

$$
(f \wedge 0) \vee 1=\sum_{i}\left(\left(a_{i} \vee 0\right) \wedge 1\right) \mathbf{1}_{R_{i}} \in \mathcal{C}
$$

The claim (6.7) is immediate from the definition of $\mathcal{E}$ on $\mathcal{C}$.
We will appeal to Theorem A. 8 to establish that $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ is the Dirichlet form of a $\mu$-symmetric Hunt process, $\mathbf{X}$. It is immediate that conditions (a)(c) of that result hold for $\mathcal{C}$, so it remains to check the tightness condition (d). Take $K_{n}=\bigcap_{m=n}^{\infty} R_{m}$. Then

$$
\begin{aligned}
\operatorname{Cap}\left(E \backslash K_{n}\right) & \leqslant \sum_{m=n}^{\infty} \operatorname{Cap}\left(E \backslash R_{m}\right) \\
& \leqslant \sum_{m=n}^{\infty}\left(\mathcal{E}\left(\mathbf{1}_{E \backslash R_{m}}, \mathbf{1}_{E \backslash R_{m}}\right)+\left(\mathbf{1}_{E \backslash R_{m}}, \mathbf{1}_{E \backslash R_{m}}\right)_{\mu}\right) \\
& =\sum_{m=n}^{\infty}\left(\Lambda\left(\left[\left(E \backslash R_{m}\right) \times R_{m}\right] \cup\left[R_{m} \times\left(E \backslash R_{m}\right)\right]\right)+\mu\left(E \backslash R_{m}\right)\right)
\end{aligned}
$$

The rightmost sum is finite by (d), and so we certainly have

$$
\lim _{n \rightarrow \infty} \operatorname{Cap}\left(E \backslash K_{n}\right)=0
$$

Finally, because constants belong to $\mathcal{D}(\mathcal{E})$, it follows from Theorem 1.6.3 of [72] that $\mathbf{X}$ is recurrent.

Remark 6.12. (i) Note that Example A. 2 doesn't apply to give the closability of $\mathcal{E}$ unless $\nu_{s}=0$.
(ii) Suppose that $\mathcal{S} \subseteq \mathcal{R}$ generates $\mathcal{R}$, then it suffices to check condition (b) just for $R \in \mathcal{S}$, as the following argument shows. We remarked in the proof that condition (b) was just the statement that $\mathcal{E}\left(\mathbf{1}_{R}, \mathbf{1}_{R}\right)<\infty$ for all $R \in \mathcal{R}$. Note that $\mathbf{1}_{R}$ for $R \in \mathcal{R}$ is a finite linear combination of functions of the form $f=\prod_{i=1}^{N} \mathbf{1}_{S_{i}}$ for $S_{1}, \ldots, S_{N} \in \mathcal{S}$, and so it suffices to show that $\mathcal{E}(f, f)<\infty$ for such $f$. Observe that if $a_{1}, \ldots, a_{N} \in \mathbb{R}$ and $b_{1}, \ldots, b_{N} \in \mathbb{R}$ satisfy $\left|a_{i}\right| \leqslant 1$ and $\left|b_{i}\right| \leqslant 1$ for $1 \leqslant i \leqslant N$, then

$$
\left|\prod_{i=1}^{N} a_{i}-\prod_{i=1}^{N} b_{i}\right|=\left|\sum_{i=1}^{N}\left(\prod_{j=1}^{i-1} a_{j}\right)\left(a_{i}-b_{i}\right)\left(\prod_{k=i+1}^{N} b_{k}\right)\right| \leqslant \sum_{i=1}^{N}\left|a_{i}-b_{i}\right| .
$$

Therefore,

$$
\begin{aligned}
(f(y)-f(x))^{2} & =|f(y)-f(x)| \\
& \leqslant \sum_{i=1}^{N}\left(\mathbf{1}_{\left(E \backslash S_{i}\right) \times S_{i}}(x, y)+\mathbf{1}_{S_{i} \times\left(E \backslash S_{i}\right)}(x, y)\right),
\end{aligned}
$$

and applying the assumption that (b) holds for all $R \in \mathcal{S}$ gives the result.
(iii) We emphasize that the elements of $\mathcal{D}(\mathcal{E})$ are elements of $L^{2}(E, \mu)$ and are thus equivalence classes of functions. It is clear from the above proof that if $f, g \in \mathcal{D}(\mathcal{E})$, then there are representatives $\hat{f}$ and $\hat{g}$ of the $L^{2}(E, \mu)$ equivalence classes of $f$ and $g$ such that

$$
\mathcal{E}(f, g)=\iint(\hat{f}(y)-\hat{f}(x))(\hat{g}(y)-\hat{g}(x)) \Lambda(d x, d y)
$$

Some care must be exercised here: it is clear that if $\nu_{s} \neq 0$, then we cannot substitute an arbitrary choice of representatives into the right-hand side to compute $\mathcal{E}(f, g)$.
(iv) The above proof appealed to Theorem A.8, which is Theorem 7.3.1 of [72]. Although our state-space $E$ is, in general, not locally compact, much of the theory developed in [72] for the locally compact setting still applies - see Remark A.9.

We present several examples of set-ups satisfying the conditions of the Theorem 6.4 at the end of Section 6.4.

### 6.4 Bipartite chains

Assume for this section that $\mathbf{X}$ is a bipartite chain.
Notation 6.13. For a Borel set $B \subseteq E$, put $\sigma_{B}=\inf \left\{t>0: X_{t} \in B\right\}$ and $\tau_{B}=\inf \left\{t>0: X_{t} \notin B\right\}$.

## Proposition 6.14.

(i) Consider $x \in E^{o}$. If $\int \kappa(x, z) \nu(d z)=0$, then $\mathbb{P}^{x}\left\{\tau_{\{x\}}<\infty\right\}=0$. Otherwise,

$$
\mathbb{P}^{x}\left\{\tau_{\{x\}}>t, X_{\tau_{\{x\}}} \in d y\right\}=\exp \left(-t \int \kappa(x, z) \nu(d z)\right) \frac{\kappa(x, y) \nu(d y)}{\int \kappa(x, z) \nu(d z)}
$$

and, in particular, $\mathbb{P}^{x}\left\{X_{\tau_{\{x\}}} \in E^{*}\right\}=1$.
(ii) For q.e. $x \in E^{*}, \mathbb{P}^{x}\left\{X_{t} \in E^{o}\right\}=1$ for Lebesgue almost all $t \geqslant 0$. In particular, $\mathbb{P}^{x}\left\{\sigma_{E^{o}}=0\right\}=1$ for $q . e . x \in E^{*}$.

Proof. (i) Because each $x \in E^{o}$ is isolated, it follows from standard considerations that $\mathbb{P}\left\{\tau_{\{x\}}>t\right\}=\exp (-\alpha t)$, where

$$
\begin{aligned}
\mu(\{x\}) \alpha & =-\lim _{t \downarrow 0}\left(\frac{1}{t}\left(T_{t}-I\right) \mathbf{1}_{x}, \mathbf{1}_{x}\right)_{\mu} \\
& =\mathcal{E}\left(\mathbf{1}_{x}, \mathbf{1}_{x}\right)=\mu(\{x\}) \int \kappa(x, z) \nu(d z)
\end{aligned}
$$

Observe for $f, g \in \mathcal{C}$ that $\mathcal{E}(f, g)=\iint(f(y)-f(x))(g(y)-g(x)) J(d x, d y)$, where $J(d x, d y)=(1 / 2)[\Lambda(d x, d y)+\Lambda(d y, d x)]$ is the symmetrization of $\Lambda$. Note that $J$ is a symmetric measure that assigns no mass to the diagonal of $E \times E$. This representation of $\mathcal{E}$ is the one familiar from the Beurling-Deny formula. The result now follows from Lemma 4.5.5 of [72].
(ii) This is immediate from the Markov property, Fubini and the observation $\mathbb{P}^{\mu}\left\{X_{t} \notin E^{o}\right\}=\mu\left(E^{*}\right)=0$ for all $t \geqslant 0$.

Definition 6.15. Define a subprobability kernel $\xi$ on $E$ by $\xi(x, B)=\mu \otimes$ $\nu\left(\left\{\left(x^{\prime}, y\right): \kappa(x, y)>0, \kappa\left(x^{\prime}, y\right)>0, x^{\prime} \in B\right\}\right)$. Note that $\xi(x, \cdot) \leqslant \mu$. Say that $\mathbf{X}$ is graphically irreducible if there exists $x_{0} \in E^{o}$ such that for all $x \in E^{o}$ there exists $n \in \mathbb{N}$ for which $\xi^{n}\left(x_{0},\{x\}\right)>0$.

Recall that a measure $\eta$ is invariant for $\mathbf{X}$ if $\mathbb{P}^{\eta}\left\{X_{t} \in \cdot\right\}=\eta$ for all $t \geqslant 0$.
Proposition 6.16. Suppose that $\mathbf{X}$ is graphically irreducible. Then $\mu$ is the unique invariant probability measure for $\mathbf{X}$ such that $\mu\left(E^{o}\right)=1$. If $\gamma$ is any other probability measure such that $\gamma\left(E^{o}\right)=1$, then

$$
\lim _{t \rightarrow \infty} \sup _{B}\left|\mathbb{P}^{\gamma}\left\{X_{t} \in B\right\}-\mu(B)\right|=0
$$

Proof. By standard coupling arguments, both claims will hold if we can show

$$
\begin{equation*}
\mathbb{P}^{x}\left\{\sigma_{\{y\}}<\infty\right\}=1, \text { for all } x, y \in E^{o} . \tag{6.8}
\end{equation*}
$$

For (6.8) it suffices by Theorem 4.6 .6 of [72] to check that the recurrent form $\mathcal{E}$ is irreducible in the sense of Section 1.6 of [72]. Furthermore, applying Theorem 1.6.1 of [72] (and the fact that $1 \in \mathcal{D}(\mathcal{E})$ with $\mathcal{E}(1,1)=0$ ), it is certainly enough to establish that if $B$ is any Borel set with $\mathbf{1}_{B} \in \mathcal{D}(\mathcal{E})$ and

$$
\begin{equation*}
0=\mathcal{E}\left(\mathbf{1}_{B}, \mathbf{1}_{B}\right)+\mathcal{E}\left(\mathbf{1}_{E \backslash B}, \mathbf{1}_{E \backslash B}\right)=2 \mathcal{E}\left(\mathbf{1}_{B}, \mathbf{1}_{B}\right) \tag{6.9}
\end{equation*}
$$

then $\mu(B)$ is either 0 or 1 .
Suppose that (6.9) holds. By Remark 6.12(iii), there is a Borel function $\hat{f}$ with $\hat{f}=\mathbf{1}_{B}, \mu$-a.e., such that

$$
\begin{align*}
0 & =\mathcal{E}\left(\mathbf{1}_{B}, \mathbf{1}_{B}\right) \\
& =\iint(\hat{f}(y)-\hat{f}(x))^{2} \Lambda(d x, d y)  \tag{6.10}\\
& =\iint\left(\hat{f}(y)-\mathbf{1}_{B}(x)\right)^{2} \Lambda(d x, d y)
\end{align*}
$$

Suppose first that $x_{0} \in B$, where $x_{0}$ is as in Definition 6.15. From (6.10),

$$
\int(\hat{f}(y)-1)^{2} \kappa\left(x_{0}, y\right) \nu(d y)=0
$$

and so $\nu\left(\left\{y: \hat{f} \neq 1, \kappa\left(x_{0}, y\right)>0\right\}\right)=0$. Therefore, again from (6.10), $\xi\left(x_{0},\{x\right.$ : $\left.\left.\mathbf{1}_{B}(x) \neq 1\right\}\right)=0$. That is, if $\xi\left(x_{0},\{x\}\right)>0$, then $x \in B$. Continuing in this way, we get that if $x \in E^{o}$ is such that $\xi^{n}\left(x_{0},\{x\}\right)>0$ for some $n$, then $x \in B$. Thus, $E^{o} \subseteq B$ and $\mu(B)=1$. A similar argument shows that if $x_{0} \notin B$, then $\mu(B)=0$.

Example 6.17. Suppose that we are in the setting of Example 6.1 with $E^{o}=\mathbb{N}$. Let $\mu$ be an arbitrary fully supported probability measure on $\mathbb{N}$ and put $\nu=\delta_{\infty}$. In order that the conditions of Theorem 6.4 hold we only need $\kappa$ to satisfy $\sum_{x \in \mathbb{N}} \kappa(x, \infty) \mu(\{x\})=\infty$. The conditions of Proposition 6.16 will hold if and only if $\kappa(x, \infty)>0$ for all $x \in \mathbb{N}$.

Example 6.18. We recall the Dirichlet form for the wild chain. Here $E=\mathbf{T}^{*}$ from Example 6.10, $\mu$ is the PGW(1) distribution and $\nu$ is the distribution of a PGW(1) tree "conditioned to be infinite". A more concrete description of $\nu$ is the following. Each $y \in \mathbf{T}_{\infty}^{*}$ has a unique path $\left(u_{0}, u_{1}, u_{2}, \ldots\right)$ starting at the root. There is a bijection between $\mathbf{T}_{\infty}^{*}$ and $\mathbf{T}_{<\infty} \times \mathbf{T}_{<\infty} \times \ldots$ that is given by identifying $y \in \mathbf{T}_{\infty}^{*}$ with the sequence of finite trees $\left(y_{0}, y_{1}, y_{2}, \ldots\right)$, where $y_{i}$ is the tree rooted at $u_{i}$ in the forest obtained by deleting the edges of the path $\left(u_{0}, u_{1}, u_{2}, \ldots\right)$ - see Figure 6.2.


Fig. 6.2. The bijection between $\mathbf{T}_{\infty}^{*}$ and $\mathbf{T}_{<\infty} \times \mathbf{T}_{<\infty} \times \ldots$

The probability measure $\nu$ on $\mathbf{T}_{\infty}^{*}$ is the push-forward by this bijection of the probability measure $\mu \times \mu \times \ldots$ on $\mathbf{T}_{<\infty} \times \mathbf{T}_{<\infty} \times \ldots$

Rather than describe $\kappa(x, y)$ explicitly, it is more convenient (and equally satisfactory for our purposes) to describe the measures

$$
q^{\uparrow}(x, d y):=\kappa(x, y) \nu(d y)
$$

for each $x$ and

$$
q^{\downarrow}(y, d x):=\kappa(x, y) \mu(d x)
$$

for each $y$. Given $x \in \mathbf{T}_{<\infty}, y \in \mathbf{T}_{\infty}^{*}$, and a vertex $u$ of $x$, let $(x / u / y) \in \mathbf{T}_{\infty}^{*}$ denote the tree rooted at the root of $x$ that is obtained by inserting a new
edge from $u$ to the root of $y$. Then

$$
\begin{equation*}
q^{\uparrow}(x, f):=\sum_{u \in x} \int f((x / u / y)) \nu(d y) \tag{6.11}
\end{equation*}
$$

for $f$ a non-negative Borel function on $\mathbf{T}^{*}$.
For $y \in \mathbf{T}_{\infty}^{*}$ with infinite path from the root $\left(u_{0}, u_{1}, u_{2}, \ldots\right)$ and $i \in \mathbb{N}_{0}$, removing the edge $\left(u_{i}, u_{i+1}\right)$ produces two trees, one finite rooted at $u_{0}$ and one infinite rooted at $u_{i+1}$. Let $k_{i}(y) \in \mathbf{T}_{<\infty}$ denote the finite tree. Then (6.11) is equivalent to

$$
\begin{equation*}
q^{\downarrow}(y, f)=\sum_{i \in \mathbb{N}_{0}} f\left(k_{i}(y)\right) \tag{6.12}
\end{equation*}
$$

for $f$ a non-negative Borel function on $\mathbf{T}^{*}$.
Let us now check the conditions of Theorem 6.4. Condition (a) is obvious. Turning to condition (b), recall that any $R \in \mathcal{R}$ is of the form $\left\{x: \rho_{n}(x) \in B\right\}$ for some $n \in \mathbb{N}$ and finite or co-finite $B \subseteq \mathbf{T}_{\leqslant n}$, where $\rho_{n}$ is defined in Example 6.2. Note that $\left[\left(\mathbf{T}^{*} \backslash R\right) \times R\right] \cup\left[R \times\left(\mathbf{T}^{*} \backslash R\right)\right] \subseteq\left\{(x, y): \rho_{n}(x) \neq\right.$ $\left.\rho_{n}(y)\right\}$. Moreover, if $y \in \mathbf{T}_{\infty}^{*}$ is of the form $\left(x / u / y^{\prime}\right)$ for some $u \in x$ and $y^{\prime} \in \mathbf{T}_{\infty}^{*}$, then $\rho_{n}(x) \neq \rho_{n}(y)$ if and only if $u$ has height less than $n$. Therefore, by (6.11),

$$
\Lambda\left(\left[\left(\mathbf{T}^{*} \backslash R\right) \times R\right] \cup\left[R \times\left(\mathbf{T}^{*} \backslash R\right)\right]\right) \leqslant \int \#\left(\rho_{n-1}(x)\right) \mu(d x)=n
$$

where we recall that the expected size of the $k^{\text {th }}$ generation in a critical Galton-Watson branching process is 1 .

It is immediate from (6.12) that

$$
\int \kappa(x, y) \mu(d x)=q^{\downarrow}(y, 1)=\infty
$$

for $\nu=\nu_{s}$ almost every $y$, and so condition (c) holds.
Finally, consider condition (d). Put $S_{n, c}:=\left\{x: \#\left(\rho_{n}(x)\right) \leqslant c\right\}$. We will take $R_{n}=S_{n, c_{n}}$ for some sequence of constants $\left(c_{n}\right)_{n \in \mathbb{N}}$. Note that $\bigcap_{m=n}^{\infty} S_{m, c_{m}}$ is compact for all $n$, whatever the choice of $\left(c_{n}\right)_{n \in \mathbb{N}}$. By choosing $c_{n}$ large enough, we can certainly make $\mu\left(\mathbf{T}^{*} \backslash S_{n, c_{n}}\right) \leqslant 2^{-n}$. From the argument for part (b) we know that $\left[\left(\mathbf{T}^{*} \backslash S_{n, c}\right) \times S_{n, c}\right] \cup\left[S_{n, c} \times\left(\mathbf{T}^{*} \backslash S_{n, c}\right)\right]=$ $S_{n, c} \times\left(\mathbf{T}^{*} \backslash S_{n, c}\right)$ is contained in the set $\left\{(x, y): \rho_{n}(x) \neq \rho_{n}(y)\right\}$ that has finite $\Lambda$ measure. Of course, $\lim _{c \rightarrow \infty} \mathbf{T}^{*} \backslash S_{n, c}=\varnothing$. Therefore, by dominated convergence, $\lim _{c \rightarrow \infty} \Lambda\left(\left[\left(\mathbf{T}^{*} \backslash S_{n, c}\right) \times S_{n, c}\right] \cup\left[S_{n, c} \times\left(\mathbf{T}^{*} \backslash S_{n, c}\right)\right]\right)=0$, and by choosing $c_{n}$ large enough we can make $\Lambda\left(\left[\left(\mathbf{T}^{*} \backslash S_{n, c_{n}}\right) \times S_{n, c_{n}}\right] \cup\left[S_{n, c_{n}} \times\left(\mathbf{T}^{*} \backslash S_{n, c_{n}}\right)\right]\right) \leqslant$ $2^{-n}$.

It is obvious that the extra bipartite chain conditions hold with $E^{o}=\mathbf{T}_{<\infty}$. The condition of Proposition 6.16 also holds. More specifically, we can take $x_{0}$ in Definition 6.15 to be the trivial tree consisting of only a root. By (6.11) and (6.12), the measure $\xi^{n}\left(x_{0}, \cdot\right)$ assigns positive mass to every tree $x \in \mathbf{T}_{<\infty}$ with at most $n$ children in the first generation (that is, $x \in \mathbf{T}_{<\infty}$ such that $\left.\#\left(\rho_{1}(x)\right) \leqslant n+1\right)$, and so $\mathbf{X}$ is indeed graphically irreducible.

Example 6.19. Suppose that we are in the setting of Example 6.8 with $\# S<\infty$ (so that $E$ is compact) and $E^{o}$ the set $\{x: \zeta(x)<\infty\}$, as above. Note that $E^{*}=S^{\mathbb{N}_{0}}$. Fix a probability measure $P$ on $S$ with full support, an $S \times S$ stochastic matrix $Q$ with positive entries and a probability measure $R$ on $\mathbb{N}_{0}$. Define a probability measure $\mu$ on $E^{o}$ by $\mu\left(\left\{x: \zeta(x)=n, x_{0}=s_{0}, \ldots, x_{n-1}=\right.\right.$ $\left.\left.s_{n-1}\right\}\right)=R(n) P\left(s_{0}\right) Q\left(s_{0}, s_{1}\right) \ldots Q\left(s_{n-2}, s_{n-1}\right)$. In other words, $\mu$ is the law of a Markov chain with initial distribution $P$ and transition matrix $Q$ killed at an independent time with distribution $R$. Define $\nu$ on $E^{*}$ by $\nu\left(\left\{s_{0}\right\} \times \cdots \times\right.$ $\left.\left\{s_{n}\right\} \times S \times S \times \ldots\right)=P\left(s_{0}\right) Q\left(s_{0}, s_{1}\right) \ldots Q\left(s_{n-1}, s_{n}\right)$. Thus, $\nu$ is the law of the unkilled chain with initial distribution $P$ and transition matrix $Q$. Define $\kappa(x, y)$ for $x \in E^{o}$ and $y \in E^{*}$ by $\kappa(x, y)=K(\zeta(x)) \mathbf{1}_{x \leqslant y}$ for some sequence of non-negative constants $K(n), n \in \mathbb{N}_{0}$.

In order that the conditions of Theorem 6.4 hold, we only need $K$ to satisfy $\sum_{x \leqslant y} K(\zeta(x)) \mu(\{x\})=\infty$ for $\nu$-a.e. $y \in E^{*}$. For example, if $q_{*}=\min _{s, s^{\prime}} Q\left(s, s^{\prime}\right)$, then it suffices that $\sum_{n \in \mathbb{N}_{0}} K(n) R(n) q_{*}^{n}=\infty$. In particular, if $\nu$ is the law of a sequence of i.i.d. uniform draws from $S$ (so that $P(s)=S\left(s, s^{\prime}\right)=(\# S)^{-1}$ for all $s, s^{\prime} \in S$ ), then we require $\sum_{n \in \mathbb{N}_{0}} K(n) R(n)(\# S)^{-n}=\infty$.

In general, $\mathbf{X}$ will be graphically irreducible with $x_{0}=(\partial, \partial, \ldots)$ (and, therefore, the condition of Proposition 6.16 holds) if $K(n)>0$ for all $n \in \mathbb{N}_{0}$.

### 6.5 Quotient processes

Return to the general set-up of Theorem 6.4. Suppose that $\mathcal{R}^{\prime}$ is a subalgebra of $\mathcal{R}$ and write $\mathcal{C}^{\prime}$ for the subalgebra of $\mathcal{C}$ generated by the indicator functions of sets in $\mathcal{R}^{\prime}$. We can define an equivalence relation on $E$ by declaring that $x$ and $y$ are equivalent if $f(x)=f(y)$ for all $f \in \mathcal{C}^{\prime}$. Let $\bar{E}$ denote the corresponding quotient space equipped with the quotient topology and denote by $\pi: E \rightarrow \bar{E}$ the quotient map. It is not hard to check that $\bar{E}$ is a Lusin space and that the algebra $\overline{\mathcal{R}}:=\left\{\pi R: R \in \mathcal{R}^{\prime}\right\}$ consists of simultaneously closed and open sets and is a base for the topology of $\bar{E}$. Write $\overline{\mathcal{C}}$ for the algebra generated by the indicator functions of sets in $\overline{\mathcal{R}}$. Note that $\mathcal{C}^{\prime}=\{\bar{f} \circ \pi: \bar{f} \in \overline{\mathcal{C}}\}$.

Proposition 6.20. Suppose that the following hold:
(a) $\mu=\nu$;
(b) there exists a Borel function $\bar{\kappa}: \bar{E} \times \bar{E} \rightarrow \mathbb{R}_{+}$such that $\kappa(x, y)=\bar{\kappa}(\pi x, \pi y)$ for $\pi x \neq \pi y$;
(c) $\bar{E}$ is compact;
(d) $\mu_{\mathcal{R}^{\prime}}[f]:=\mu\left[f \mid \sigma\left(\mathcal{R}^{\prime}\right)\right]=\mu[f \mid \sigma(\pi)]$ has a version in $\mathcal{C}^{\prime}$ for all $f \in \mathcal{C}$.

Then the hypotheses of Theorem 6.4 hold with $E, \mathcal{R}, \mathcal{C}, \mu, \nu, \kappa$ replaced by $\bar{E}, \overline{\mathcal{R}}, \overline{\mathcal{C}}, \bar{\mu}, \bar{\nu}, \bar{\kappa}$, where $\bar{\mu}=\bar{\nu}$ is the push-forward of $\mu=\nu$ by $\pi$. Moreover, if $(\overline{\mathcal{E}}, \mathcal{D}(\overline{\mathcal{E}}))$ denotes the resulting Dirichlet form, then $\pi \circ \mathbf{X}$ is a $\bar{\mu}$-symmetric Hunt process with Dirichlet form $(\overline{\mathcal{E}}, \mathcal{D}(\overline{\mathcal{E}}))$.

Proof. It is clear that the hypotheses of Theorem 6.4 hold with $E, \mathcal{R}, \mathcal{C}, \mu, \nu, \kappa$ replaced by $\bar{E}, \overline{\mathcal{R}}, \overline{\mathcal{C}}, \bar{\mu}, \bar{\nu}, \bar{\kappa}$.

Let $\left(\bar{T}_{t}\right)_{t \geqslant 0}$ denote the semigroup on $L^{2}(\bar{E}, \bar{\mu})$ corresponding to $\overline{\mathcal{E}}$. The proof $\pi \circ \mathbf{X}$ is a $\bar{\mu}$-symmetric Hunt process with $\operatorname{Dirichlet~form~}(\overline{\mathcal{E}}, \mathcal{D}(\overline{\mathcal{E}}))$ will be fairly straightforward once we establish that $T_{t}(\bar{f} \circ \pi)=\left(\bar{T}_{t} \bar{f}\right) \circ \pi$ for all $t \geqslant 0$ and $\bar{f} \in L^{2}(\bar{E}, \bar{\mu})$ (see Theorem 13.5 of [128] for a proof that this suffices for $\pi \circ \mathbf{X}$ to be a Hunt process - the proof that $\pi \circ \mathbf{X}$ is $\bar{\mu}$-symmetric and the identification of the associated Dirichlet form are then easy). Equivalently, writing $\left(G_{\alpha}\right)_{\alpha>0}$ and $\left(\bar{G}_{\alpha}\right)_{\alpha>0}$ for the resolvents corresponding to $\left(T_{t}\right)_{t \geqslant 0}$ and $\left(\bar{T}_{t}\right)_{t \geqslant 0}$, we need to establish that $G_{\alpha}(\bar{f} \circ \pi)=\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi$ for all $\alpha>0$ and $\bar{f} \in L^{2}(\bar{E}, \bar{\mu})$. This is further equivalent to establishing that $\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi \in \mathcal{D}(\mathcal{E})$ and $\mathcal{E}\left(\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi, g\right)+\alpha\left(\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi, g\right)_{\mu}=(\bar{f} \circ \pi, g)_{\mu}$ for all $g \in \mathcal{C}-$ see Equation (1.3.7) of [72].

Fix $\bar{f} \in L^{2}(\bar{E}, \bar{\mu})$ and $g \in \mathcal{C}$. By assumption, $\mu_{\mathcal{R}^{\prime}}[g]=\bar{g} \circ \pi$ for some $\bar{g} \in \overline{\mathcal{C}}$. Also, it is fairly immediate from the definition of $\bar{E}$ that $\bar{h} \in \mathcal{D}(\overline{\mathcal{E}})$ if and only if $\bar{h} \circ \pi \in \mathcal{D}(\mathcal{E})$, and that $\overline{\mathcal{E}}(\bar{h}, \bar{h})=\mathcal{E}(\bar{h} \circ \pi, \bar{h} \circ \pi)$. Hence, by Remark 6.12 (iii),

$$
\begin{aligned}
& \mathcal{E}(\bar{h} \circ \pi, g)=\iint(\bar{h} \circ \pi(y)-\bar{h} \circ \pi(x))(g(y)-g(x)) \Lambda(d x, d y) \\
& =\iint_{\{(x, y): \pi x \neq \pi y\}}(\bar{h} \circ \pi(y)-\bar{h} \circ \pi(x))(g(y)-g(x)) \Lambda(d x, d y) \\
& =\iint(\bar{h} \circ \pi(y)-\bar{h} \circ \pi(x))(g(y)-g(x)) \bar{\kappa}(\pi x, \pi y) \mu(d x) \mu(d y) \\
& =\iint(\bar{h} \circ \pi(y)-\bar{h} \circ \pi(x))\left(\mu_{\mathcal{R}^{\prime}}[g](y)-\mu_{\mathcal{R}^{\prime}}[g](x)\right) \bar{\kappa}(\pi x, \pi y) \mu(d x) \mu(d y) \\
& =\iint(\bar{h} \circ \pi(y)-\bar{h} \circ \pi(x))(\bar{g} \circ \pi(y)-\bar{g} \circ \pi(x)) \bar{\kappa}(\pi x, \pi y) \mu(d x) \mu(d y) \\
& =\iint(\bar{h}(w)-\bar{h}(v))(\bar{g}(w)-\bar{g}(v)) \bar{\kappa}(v, w) \bar{\mu}(d v) \bar{\mu}(d w) \\
& =\overline{\mathcal{E}}(\bar{h}, \bar{g}) .
\end{aligned}
$$

Of course,

$$
(\bar{h} \circ \pi, g)_{\mu}=(\bar{h} \circ \pi, \bar{g} \circ \pi)_{\mu}=(\bar{h}, \bar{g})_{\bar{\mu}}
$$

Therefore,

$$
\begin{aligned}
& \mathcal{E}\left(\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi, g\right)+\alpha\left(\left(\bar{G}_{\alpha} \bar{f}\right) \circ \pi, g\right)_{\mu}=\overline{\mathcal{E}}\left(\bar{G}_{\alpha} \bar{f}, \bar{g}\right)+\alpha\left(\bar{G}_{\alpha} \bar{f}, \bar{g}\right)_{\bar{\mu}} \\
& \quad=(\bar{f}, \bar{g})_{\bar{\mu}}=(\bar{f} \circ \pi, \bar{g} \circ \pi)_{\mu}=(\bar{f} \circ \pi, g)_{\mu},
\end{aligned}
$$

as required.
We will see an application of Proposition 6.20 at the end of Section 6.7.

### 6.6 Additive functionals

We are still in the general setting of Theorem 6.4.

Proposition 6.21. The probability measure $\nu$ assigns no mass to sets of zero capacity, and there is a positive continuous additive functional $\left(A_{t}\right)_{t \geqslant 0}$ with Revuz measure $\nu$.

Proof. The reference measure $\mu$ assigns no mass to sets of zero capacity, so it suffices to show that $\nu_{s}$ assigns no mass to sets of zero capacity. For $M>0$ put $G_{M}:=\left\{y: \int[\kappa(x, y) \wedge M] \mu(d x) \geqslant 1\right\}$ and define a subprobability measure $\nu_{s}^{M}$ by $\nu_{s}^{M}:=\nu_{s}\left(\cdot \cap G_{M}\right)$. By (c) of Theorem $6.4, \nu_{s}\left(E \backslash \bigcup_{M} G_{M}\right)=0$, and so it suffices to show for each $M$ that $\nu_{s}^{M}$ assigns no mass to sets of zero capacity.

Observe for $f \in \mathcal{C}$ that

$$
\begin{aligned}
& \left(\int|f(y)| \nu_{s}^{M}(d y)\right)^{2} \leqslant \int f^{2}(y) \nu_{s}^{M}(d y) \leqslant \iint f^{2}(y) \Lambda^{M}(d x, d y) \\
& \quad \leqslant 2\left(\iint(f(y)-f(x))^{2} \Lambda^{M}(d x, d y)+\iint f^{2}(x) \Lambda^{M}(d x, d y)\right) \\
& \quad \leqslant 2(1 \vee M)\left(\mathcal{E}(f, f)+(f, f)_{\mu}\right)
\end{aligned}
$$

The development leading to Lemma 2.2.3 of [72] can now be followed to show that for all Borel sets $B$ we have $\nu_{s}^{M}(B) \leqslant C_{M} \operatorname{Cap}(B)^{1 / 2}$ for a suitable constant $C_{M}$ (the argument in [72] is in a locally compact setting, but it carries over without difficulty to our context).

The existence and uniqueness of $\left(A_{t}\right)_{t \geqslant 0}$ follows from Theorem 5.1.4 of [72].

Remark 6.22. In the bipartite chain case, the distribution under $\mathbb{P}^{\mu}$ of $X_{\zeta}$, where $\zeta:=\tau_{\left\{X_{0}\right\}}$, is mutually absolutely continuous with respect to $\nu$, and Proposition 6.21 is obvious.

### 6.7 Bipartite chains on the boundary

Return to the bipartite chain setting. Following the construction in Section 6.2 of [72], let $\mathbf{Y}$ denote the process $\mathbf{X}$ time-changed according to the positive continuous additive functional $A$. That is, $Y_{t}=X_{\gamma_{t}}$ where $\gamma_{t}=\inf \{s>0$ : $\left.A_{s}>t\right\}$. Write $\tilde{E}$ for the support of $A$. We have $\tilde{E} \subseteq \breve{E}:=\operatorname{supp} \nu \subseteq E^{*}$ and $\nu(\breve{E} \backslash \tilde{E})=0$.

Let $\breve{\mathcal{R}}=\{R \cap \breve{E}: R \in \mathcal{R}\}$ and put $\breve{\mathcal{C}}=\left\{f_{\mid \breve{E}}: f \in \mathcal{C}\right\}$. Note that $\breve{\mathcal{C}}$ is also the algebra generated by $\breve{\mathcal{R}}$.

Theorem 6.23. The process $\mathbf{Y}$ is a recurrent $\nu$-symmetric Hunt process with state-space $\breve{E}$ and Dirichlet form given by the closure of the form $\breve{\mathcal{E}}$ on $\breve{\mathcal{C}}$ defined by

$$
\breve{\mathcal{E}}(f, g)=\iint(f(y)-f(z))(g(y)-g(z)) \breve{\kappa}(y, z) \nu(d y) \nu(d z), f, g \in \breve{\mathcal{C}}
$$

where

$$
\breve{\kappa}(y, z)=\int \kappa(x, y) \frac{\kappa(x, z)}{\int \kappa(x, w) \nu(d w)} \mu(d x)
$$

(with the convention $0 / 0=0$ ).
Proof. By Theorem A.2.6 and Theorem 4.1.3 of [72],

$$
\mathbb{P}^{y}\left\{\sigma_{\breve{E}}=0\right\}=1 \text { for q.e. } y \in \breve{E}
$$

Hence, for q.e. $y \in \breve{E}$ we have $\lim _{\epsilon \downarrow 0} \inf \left\{t>\epsilon: X_{t} \in \breve{E}\right\}=0$, $\mathbb{P}^{y}$-a.s. Moreover, it follows from parts (i) and (ii) of Proposition 6.14 and the observation $\nu(\breve{E} \backslash \tilde{E})=0$ that for q.e. $y \in \breve{E}$ we have $\inf \left\{t>\epsilon: X_{t} \in \breve{E}\right\}=\inf \left\{t>\epsilon: X_{t} \in\right.$ $\tilde{E}\}$ for all $\epsilon>0, \mathbb{P}^{y}$-a.s. Combining this with Proposition 6.21 gives

$$
\mathbb{P}^{y}\left\{\sigma_{\tilde{E}}=0\right\}=1 \text { for q.e. and } \nu \text {-a.e. } y \in \breve{E}
$$

Define $H_{\tilde{E}} f(x):=\mathbb{P}^{x}\left[f\left(X_{\sigma_{\tilde{E}}}\right)\right]$ for $f$ a bounded Borel function on $E$. It follows from part (i) of Proposition 6.14 and what we have just observed that

$$
H_{\tilde{E}} f(x)=\frac{\int f(y) \kappa(x, y) \nu(d y)}{\int \kappa(x, y) \nu(d y)}, \text { for } \mu \text {-a.e. } \mathrm{x}
$$

and

$$
H_{\tilde{E}} f(x)=f(x), \text { for } \nu \text {-a.e. x. }
$$

The result now follows by applying Theorem 6.2.1 of [72].
Example 6.24. Suppose that we are in the setting of Example 6.19. For $y, z \in E^{*}=S^{\mathbb{N}_{0}}, y \neq z$, define $\delta(y, z)=\inf \left\{n: y_{n} \neq z_{n}\right\}$. Note that $\int \kappa(x, w) \nu(d w)=K(\zeta(x)) \nu(\{w: x \leqslant w\})=K(\zeta(x)) \mu(\{x\}) / R(\zeta(x))$ for $x \in E^{o}$ and so

$$
\begin{equation*}
\breve{\kappa}(y, z)=\sum_{n \leqslant \delta(y, z)} K(n) R(n) . \tag{6.13}
\end{equation*}
$$

We will now apply the results of Section 6.5 with $E, \mathbf{X}, \mu, \mathcal{E}$ replaced by $\breve{E}=E^{*}=S^{\mathbb{N}_{0}}, \mathbf{Y}, \nu, \breve{\mathcal{E}}$. Fix $N \in \mathbb{N}_{0}$ and let $\mathcal{R}^{\prime}$ be the algebra of subsets of $S^{\mathbb{N}_{0}}$ of the form $B_{0} \times \cdots \times B_{N} \times S \times S \times \ldots$.. We can identify the quotient space $\bar{E}$ with $S^{N+1}$ and the quotient map $\pi$ with the map $\left(y_{0}, y_{1}, \ldots\right) \mapsto\left(y_{0}, \ldots y_{N}\right)$. Then we can identify $\bar{\mu}$, which we emphasise is now the push-forward $\nu$ by $\pi$, with the measure that assigns mass $P\left(s_{0}\right) Q\left(s_{0}, s_{1}\right) \ldots Q\left(s_{N-1}, s_{N}\right)$ to $\left(s_{0}, \ldots s_{N}\right)$. Note that $\pi y \neq \pi z$ for $y, z \in S^{\mathbb{N}_{0}}$ is equivalent to $\delta(y, z) \leqslant N$, and it is immediate from (6.13) that Proposition 6.20 applies and $\pi \circ \mathbf{Y}$ is a $\bar{\mu}$-symmetric Markov chain on the finite state-space $S^{N+1}$. In terms of jump rates, $\pi \circ \mathbf{Y}$ jumps from $\bar{y}$ to $\bar{z} \neq \bar{y}$ at rate $\left(\sum_{n \leqslant \delta(\bar{y}, \bar{z})} K(n) R(n)\right) \bar{\mu}(\{\bar{z}\})$, where $\delta(\bar{y}, \bar{z})$ is defined in the obvious way.

As a particular example of this construction, consider the case when $\# S=p^{c}$ for some prime $p$ and integer $c \geqslant 1$. We can identify $S^{\mathbb{N}_{0}}$ (as
a set) with the ring of integers $\mathbb{D}$ of a local field $\mathbb{K}$ as in Example 6.9. If we take $P(s)=Q\left(s, s^{\prime}\right)=p^{-c}$ for all $s, s^{\prime} \in S$, then we can identify $\nu$ with the normalised Haar measure on $\mathbb{D}$. It is clear that $\mathbf{Y}$ is a Lévy process on $\mathbb{D}$ with "spherically symmetric" Lévy measure $\phi(|y|) \nu(d y)$, where $\phi\left(p^{-c n}\right)=\sum_{\ell=0}^{n} K(\ell) R(\ell)$. The condition $\sum_{n \in \mathbb{N}_{0}} K(n) R(n) p^{-c n}=\infty$ of Example 6.19 is equivalent to $\int_{\mathbb{D}} \phi(|y|) \nu(d y)=\infty$. Conversely, any Lévy process on $\mathbb{D}$ with Lévy measure of the form $\psi(|y|) \nu(d y)$ with $\psi$ non-increasing and $\int_{\mathbb{D}} \psi(|y|) \nu(d y)=\infty$ can be produced by this construction (Lévy processes on $\mathbb{D}$ are completely characterised by their Lévy measures - there is no analogue of the drift or Gaussian components of the Euclidean case, see [59]). The latter condition is equivalent to the paths of the process almost surely not being step-functions, that is, to the times at which jumps occur being almost surely dense. When $\psi(|y|)=a|y|^{-(\alpha+1)}$ for some $a>0$ and $0<\alpha<\infty$, the resultant process is analogous to a symmetric stable process. Lévy processes on local fields and totally disconnected Abelian groups in general are considered in [59] and the special case of the $p$-adic numbers has been considered by a number of authors - see Chapter 1 for a discussion.

## Diffusions on a $\mathbb{R}$-tree without leaves: snakes and spiders

### 7.1 Background

Let $(T, d)$ be a $\mathbb{R}$-tree without ends as in Section 3.4. Suppose that that there is a $\sigma$-finite Borel measure $\mu$ on the set on $\mathbf{E}_{+}$of ends at $+\infty$ such that $0<\mu(B)<\infty$ for every ball $B$ in the metric $\delta$. In particular, the support of $\mu$ is all of $E_{+}$.

The existence of such a measure $\mu$ is a more restrictive assumption on $T$ than it might first appear. Let $\bar{\mu}$ be a finite measure on $E_{+}$that is equivalent to $\mu$. Recall from (3.4) that $T_{t}, t \in \mathbb{R}$, is the set of points in $T$ with height $t$. As we remarked in Section 3.4.2, the set $\left\{\zeta \in E_{+}: \zeta \mid t=x\right\}$ is a ball in $E_{+}$for each $x \in T_{t}$ and two such balls are disjoint. Because the $\bar{\mu}$ measure of each such ball is non-zero, the set $T_{t}$ is necessarily countable. Hence, by observations made in Section 3.4.2, both the complete metric spaces $T$ and $E_{+}$are separable, and, therefore, Lusin.

We will be interested in the $T$-valued process $X$ that evolves in the following manner. The real-valued process $H$, where $H_{t}=h\left(X_{t}\right)$, evolves as a standard Brownian motion. For small $\epsilon>0$ the conditional probability of the event $\left\{X_{t+\epsilon} \in C\right\}$ given $X_{t}$ and $H$ is approximately

$$
\frac{\mu\left\{y: y\left|H_{t+\epsilon} \in C, y\right| H_{t}=X_{t}\right\}}{\mu\left\{y: y \mid H_{t}=X_{t}\right\}} .
$$

In particular, if $H_{t+\epsilon}<H_{t}$, then $X_{t+\epsilon}$ is approximately $X_{t} \mid H_{t+\epsilon}$. An intuitive description of these dynamics is given in Figure 7.1.

This evolution is reminiscent of Le Gall's Brownian snake process - see, for example, $[97,98,99,100]$ - with the difference that the "height" process $H$ is a Brownian motion here rather than a reflected Brownian motion and the role of Wiener measure on $C\left(\mathbb{R}_{+}, \mathbb{R}^{d}\right)$ in the snake construction is played here by $\mu$.


Fig. 7.1. A heuristic description of the dynamics of $X$. When $X$ is at position $x$ it makes an infinitesimal move up or down with equal probability. Conditional on $X$ moving down, it takes the branch leading to the set of ends $A$ with probability $\mu(A) /(\mu(A)+\mu(B))$ and the branch leading to the set of ends $B$ with probability $\mu(B) /(\mu(A)+\mu(B))$.

### 7.2 Construction of the diffusion process

For $x \in T$ and real numbers $b<c$ with $b<h(x)$, define a probability measure $\mu(x, b, c ; \cdot)$ on $T$ by

$$
\mu(x, b, c ; A):=\frac{\mu\left\{\xi \in E_{+}: \xi|c \in A, \xi| b=x \mid b\right\}}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}}
$$

- see Figure 7.2.

Let $\left(B_{t}, P^{a}\right)$ be a standard (real-valued) Brownian motion. Write $m_{t}:=$ $\inf _{0 \leqslant s \leqslant t} B_{s}$. Recall that the pair $\left(m_{t}, B_{t}\right)$ has joint density

$$
\phi_{a, t}(b, c):=\sqrt{\frac{2}{\pi}} \frac{c-2 b+a}{t^{3 / 2}} \exp \left(-\frac{(c-2 b+a)^{2}}{2 t}\right), b<a \wedge c
$$

under $P^{a}$ - see, for example, Corollary 30 in Chapter 1 of [70].
Theorem 7.1. There is a Markov semigroup $\left(P_{t}\right)_{t \geqslant 0}$ on $T$ defined by


Fig. 7.2. The measure $\mu(x, b, c ; \cdot)$ is supported on the set $\{y \in T: h(y)=c, y \mid b=$ $x \mid b\}$, and the mass it assigns to the set $A$ is the normalized $\mu$ mass of the shaded subset of $E_{+}$.

$$
P_{t} f(x):=P^{h(x)}\left[\mu\left(x, m_{t}, B_{t} ; f\right)\right] .
$$

Furthermore, there is a strong Markov process $\left(X_{t}, \mathbb{P}^{x}\right)$ on $T$ with continuous sample paths and semigroup $\left(P_{t}\right)_{t \geqslant 0}$.

Proof. The proof of the semigroup property of $\left(P_{t}\right)_{t \geqslant 0}$ is immediate from the Markov property of Brownian motion and the readily checked observation that for $x, x^{\prime} \in T, b<c, b<h(x)$, and $b^{\prime}<c \wedge c^{\prime}$ we have

$$
\int \mu\left(x^{\prime}, b^{\prime}, c^{\prime} ; A\right) \mu\left(x, b, c ; d x^{\prime}\right)=\mu\left(x, b \wedge b^{\prime}, c^{\prime} ; A\right)
$$

By Kolmogorov's extension theorem, there is a Markov process $\left(X_{t}, \mathbb{P}^{x}\right)$ on $T$ with semigroup $\left(P_{t}\right)_{t \geqslant 0}$. In order to show that a version of $X$ can be chosen with continuous sample paths, it suffices because ( $T, d$ ) is complete and separable to check Kolmogorov's continuity criterion. Because of the Markov property of $X$, it further suffices to observe for $\alpha>0$ that, by definition of $\left(P_{t}\right)_{t \geqslant 0}$,

$$
\begin{aligned}
\mathbb{P}^{x} & {\left[d\left(x, X_{t}\right)^{\alpha}\right] } \\
& =P^{h(x)}\left[\frac{\int\left[h(x)+h\left(\xi \mid B_{t}\right)-2 h\left(x \wedge\left(\xi \mid B_{t}\right)\right)\right]^{\alpha} \mathbf{1}\left\{\xi\left|m_{t}=x\right| m_{t}\right\} \mu(d \xi)}{\mu\left\{\xi \in E_{+}: \xi\left|m_{t}=x\right| m_{t}\right\}}\right] \\
& \leqslant P^{h(x)}\left[\frac{\int\left[h(x)+B_{t}-2 m_{t}\right]^{\alpha} \mathbf{1}\left\{\xi\left|m_{t}=x\right| m_{t}\right\} \mu(d \xi)}{\mu\left\{\xi \in E_{+}: \xi\left|m_{t}=x\right| m_{t}\right\}}\right] \\
& \leqslant C P^{h(x)}\left[\left|h(x)-m_{t}\right|^{\alpha}+\left|m_{t}-B_{t}\right|^{\alpha}\right] \\
& \leqslant C^{\prime} t^{\alpha / 2}
\end{aligned}
$$

for some constants $C, C^{\prime}$ that depend on $\alpha$ but not on $x \in T$.
The claim that $X$ is strong Markov will follow if we can show that $P_{t}$ maps $b C(T)$ into itself - see, for example, Sections III.8, III. 9 of [120]. It is assumed there that the underlying space is locally compact and the semigroup maps the space of continuous functions that vanish at infinity into itself, but this stronger assumption is only needed to establish the existence of a process with càdlàg sample paths and plays no role in the proof of the strong Markov property). By definition, for $f \in b \mathcal{B}(T)$ and $t>0$

$$
\begin{aligned}
P_{t} f(x)= & \int_{-\infty}^{h(x)} \int_{b}^{\infty} \frac{\int f(\xi \mid c) \mathbf{1}\{\xi|b=x| b\} \mu(d \xi)}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}} \\
& \times \sqrt{\frac{2}{\pi}} \frac{c-2 b+h(x)}{t^{3 / 2}} \exp \left(-\frac{(c-2 b+h(x))^{2}}{2 t}\right) d c d b
\end{aligned}
$$

for $t>0$. The right-hand side can be written as $\int_{-\infty}^{\infty} \int_{-\infty}^{\infty} F_{f, x}(b, c) d c d b$ for a certain function $F_{f, x}$. Recall from (3.2) that $\left|h(x)-h\left(x^{\prime}\right)\right| \leqslant d\left(x, x^{\prime}\right)$. Also, if $b<h(x)$, then $x^{\prime}|b=x| b$ for $x^{\prime}$ such that $d\left(x, x^{\prime}\right) \leqslant h(x)-b$. Therefore, $\lim _{x^{\prime} \rightarrow x} F_{f, x^{\prime}}(b, c)=F_{f, x}(b, c)$ except possibly at $b=h(x)$. Moreover, if $\sup _{x}|f(x)| \leqslant C$, then $\left|F_{f, x}(b, c)\right| \leqslant C F_{1, x}(b, c)$. Because

$$
\lim _{x^{\prime} \rightarrow x} \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} F_{1, x^{\prime}}(b, c) d c d b=\lim _{x^{\prime} \rightarrow x} 1=1=\int_{-\infty}^{\infty} \int_{-\infty}^{\infty} F_{1, x}(b, c) d c d b
$$

a standard generalization of the dominated convergence theorem - see, for example, Proposition 18 in Chapter 11 of [121] - shows that if $f \in b \mathcal{B}(T)$, then $P_{t} f \in b C(T)$ for $t>0$.

### 7.3 Symmetry and the Dirichlet form

Write $\lambda$ for Lebesgue measure on $\mathbb{R}$. Consider the measure $\nu$ that is obtained by pushing forward the measure $\mu \otimes \lambda$ on $E_{+} \times \mathbb{R}$ with the map $(\xi, a) \mapsto \xi \mid a$ - see Figure 7.3. Note that for $x \in T$ with $h(x)=h^{*}$ and $\epsilon>0$ we have

$$
\begin{aligned}
& \nu\{y \in T: d(x, y) \leqslant \epsilon\} \\
& \quad \leqslant \nu\left\{y \in T: y\left|\left(h^{*}-\epsilon\right)=x\right|\left(h^{*}-\epsilon\right), h^{*}-\epsilon \leqslant h(y) \leqslant h^{*}+\epsilon\right\} \\
& \quad \leqslant 2 \epsilon \mu\left\{\xi \in E_{+}: \xi\left|\left(h^{*}-\epsilon\right)=x\right|\left(h^{*}-\epsilon\right)\right\}
\end{aligned}
$$



Fig. 7.3. The definition of the measure $\nu$ on $T$ in terms of the measure $\mu$ on $E_{+}$.

That is, $\nu$ assigns finite mass to balls in $T$ and, in particular, is Radon.
We begin by showing that each operator $P_{t}, t>0$, can be continuously extended from $b \mathcal{B}(T) \cap L^{2}(T, \nu)$ to $L^{2}(T, \nu)$ and that the resulting semigroup is a strongly continuous, self-adjoint, Markovian semigroup on $L^{2}(T, \nu)$.

Observe that if $f \in b \mathcal{B}(T)$, then

$$
\begin{aligned}
& P_{t} f(x)=\int_{E_{+}} \int_{\mathbb{R}} \int_{-\infty}^{h(x) \wedge c} \frac{f(\xi \mid c) \mathbf{1}\{\xi|b=x| b\}}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}} \phi_{h(x), t}(b, c) d b d c \mu(d \xi) \\
& =\int_{T} f(y) \int_{-\infty}^{h(x) \wedge h(y)} \frac{\mathbf{1}\{x|b=y| b\}}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}} \\
& \quad \times \sqrt{\frac{2}{\pi}} \frac{h(x)+h(y)-2 b}{t^{3 / 2}} \exp \left(-\frac{(h(x)+h(y)-2 b)^{2}}{2 t}\right) d b \nu(d y) \\
& =\int_{T} f(y) \int_{-\infty}^{h(x \wedge y)} \frac{1}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}} \\
& \quad \times \sqrt{\frac{2}{\pi}} \frac{h(x)+h(y)-2 b}{t^{3 / 2}} \exp \left(-\frac{(h(x)+h(y)-2 b)^{2}}{2 t}\right) d b \nu(d y)
\end{aligned}
$$

for $t>0$. Consequently, $P_{t} f(x)=\int_{T} p_{t}(x, y) f(y) \nu(d y)$ for the jointly continuous, everywhere positive transition density

$$
\begin{align*}
p_{t}(x, y):= & \int_{-\infty}^{h(x \wedge y)} \frac{1}{\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}}  \tag{7.1}\\
& \times \sqrt{\frac{2}{\pi}} \frac{h(x)+h(y)-2 b}{t^{3 / 2}} \exp \left(-\frac{(h(x)+h(y)-2 b)^{2}}{2 t}\right) d b
\end{align*}
$$

Moreover, because $\mu\left\{\xi \in E_{+}: \xi|b=x| b\right\}=\mu\left\{\xi \in E_{+}: \xi|b=y| b\right\}$ when $b \leqslant h(x \wedge y)$ (equivalently, when $x|b=y| b)$, we have $p_{t}(x, y)=p_{t}(y, x)$. Therefore, there exists a self-adjoint, Markovian semigroup on $L^{2}(T, \nu)$ that coincides with $\left(P_{t}\right)_{t \geqslant 0}$ on $b \mathcal{B}(T) \cap L^{2}(T, \nu)$ (cf. Section1.4 of [72]). With the usual abuse of notation, we also denote this semigroup by $\left(P_{t}\right)_{t \geqslant 0}$.

Because $\nu$ is Radon, $b C(T) \cap L^{1}(T, \nu)$ is dense in $L^{2}(T, \nu)$. It is immediate from the definition of $\left(P_{t}\right)_{t \geqslant 0}$ that $\lim _{t \downarrow 0} P_{t} f(x)=f(x)$ for all $f \in b C(T)$ and $x \in T$. Therefore, by Lemma 1.4.3 of [72], the semigroup $\left(P_{t}\right)_{t \geqslant 0}$ is strongly continuous on $L^{2}(T, \nu)$.

We now proceed to identify the Dirichlet form corresponding to $\left(P_{t}\right)_{t \geqslant 0}$.
Definition 7.2. Let $\mathcal{A}$ denote the class of functions $f \in b C(T)$ such that there exists $g \in \mathcal{B}(T)$ with the property that

$$
\begin{equation*}
f(\xi \mid b)-f(\xi \mid a)=\int_{a}^{b} g(\xi \mid u) d u, \xi \in E_{+},-\infty<a<b<\infty \tag{7.2}
\end{equation*}
$$

Note for $\xi \in E_{+}$that if $A \in \mathcal{B}(T)$ with $A \subseteq[a, b]$, where $-\infty<a<b<\infty$, then

$$
\begin{aligned}
\mu\left\{\zeta \in E_{+}: \zeta|b=\xi| b\right\} \lambda(A) & \leqslant \nu\{\xi \mid u: u \in A\} \\
& \leqslant \mu\left\{\zeta \in E_{+}: \zeta|a=\xi| a\right\} \lambda(A)
\end{aligned}
$$

Therefore, the function $g$ in (7.2) is unique up to $\nu$-null sets, and (with the usual convention of using function notation to denote equivalence classes of functions) we denote $g$ by $\nabla f$.

Definition 7.3. Write $\mathcal{D}$ for the class of functions $f \in \mathcal{A} \cap L^{2}(T, \nu)$ such that $\nabla f \in L^{2}(T, \nu)$.

Remark 7.4. By the observations made in Definition 7.2, the integral

$$
\int_{a}^{b} \bar{g}(\xi \mid u) d u
$$

is well-defined for any $\xi \in E_{+}$and $\bar{g} \in L^{2}(T, \nu)$.
Theorem 7.5. The Dirichlet form $\mathcal{E}$ corresponding to the strongly continuous, self-adjoint, Markovian semigroup $\left(P_{t}\right)_{t \geqslant 0}$ on $L^{2}(T, \nu)$ has domain $\mathcal{D}$ and is given by

$$
\begin{equation*}
\mathcal{E}(f, g)=\frac{1}{2} \int_{T} \nabla f(x) \nabla g(x) \nu(d x), f, g \in \mathcal{D} \tag{7.3}
\end{equation*}
$$

Proof. A virtual reprise of the argument in Example A. 1 shows that the form $\mathcal{E}^{\prime}$ given by the right-hand side of $(7.3)$ is a Dirichlet form.

Let $\left(G_{\alpha}\right)_{\alpha>0}$ denote the resolvent corresponding to $\left(P_{t}\right)_{t \geqslant 0}$ : that is, $G_{\alpha} f=$ $\int_{0}^{\infty} e^{-\alpha t} P_{t} f d t$ for $f \in L^{2}(T, \nu)$. In order to show that $\mathcal{E}=\mathcal{E}^{\prime}$, it suffices to show that $G_{\alpha}\left(L^{2}(T, \nu)\right) \subseteq \mathcal{D}$ and $\mathcal{E}_{\alpha}^{\prime}\left(G_{\alpha} f, g\right):=\mathcal{E}^{\prime}\left(G_{\alpha} f, g\right)+\alpha(f, g)=$ $(f, g)$ for $f \in L^{2}(T, \nu)$ and $g \in \mathcal{D}$, where we write $(\cdot, \cdot)$ for the $L^{2}(T, \nu)$ inner product. By a simple approximation argument, it further suffices to check that $G_{\alpha}\left(b \mathcal{B}(T) \cap L^{2}(T, \nu)\right) \subseteq \mathcal{D}$ and $\mathcal{E}_{\alpha}^{\prime}\left(G_{\alpha} f, g\right)=(f, g)$ for $f \in b \mathcal{B}(T) \cap L^{2}(T, \nu)$ and $g \in \mathcal{D}$

Observe that

$$
\int_{0}^{\infty} e^{-\alpha t} \phi_{a, t}(b, c) d t=2 \exp (-\sqrt{2 \alpha}(c-2 b+a)), b<a \wedge c
$$

- see Equations 3.71.13 and 6.23.15 of [143]. Therefore, for $f \in b \mathcal{B}(T) \cap L^{2}(T, \nu)$ we have

$$
\begin{equation*}
G_{\alpha} f(x)=2 \int_{-\infty}^{h(x)} \int_{b}^{\infty} \mu(x, b, c ; f) e^{-\sqrt{2 \alpha}(c-2 b+h(x))} d c d b \tag{7.4}
\end{equation*}
$$

Thus, $G_{\alpha} f \in \mathcal{A}$ with

$$
\begin{align*}
\nabla\left(G_{\alpha} f\right)(x)= & 2 \int_{h(x)}^{\infty} \mu(x, h(x), c ; f) e^{-\sqrt{2 \alpha}(c-h(x))} d c  \tag{7.5}\\
& -\sqrt{2 \alpha} G_{\alpha} f(x)
\end{align*}
$$

In order to show that $G_{\alpha} f \in \mathcal{D}$ is remains to show that the first term on the righ-hand side of $(7.5)$ is in $L^{2}(T, \nu)$. By the Cauchy-Schwarz inequality and recalling the definition of $T_{t}$ from (3.4),

$$
\begin{aligned}
& \int_{T} {\left[\int_{h(x)}^{\infty} \mu(x, h(x), c ; f) e^{-\sqrt{2 \alpha}(c-h(x))} d c\right]^{2} \nu(d x) } \\
&= \int_{-\infty}^{\infty} \sum_{x \in T_{a}}\left[\int_{a}^{\infty} \frac{\int_{E_{+}} f(\xi \mid c) \mathbf{1}\{\xi \mid a=x\} \mu(d \xi)}{\mu\{\xi: \xi \mid a=x\}} e^{-\sqrt{2 \alpha}(c-a)} d c\right]^{2} \\
& \times \mu\{\xi: \xi \mid a=x\} d a \\
& \leqslant \frac{1}{2 \sqrt{2 \alpha}} \int_{-\infty}^{\infty} \sum_{x \in T_{a}}\left[\int_{a}^{\infty}\left[\frac{\int_{E_{+}} f(\xi \mid c) \mathbf{1}\{\xi \mid a=x\} \mu(d \xi)}{\mu\{\xi: \xi \mid a=x\}}\right]^{2} e^{-\sqrt{2 \alpha}(c-a)} d c\right] \\
& \times \mu\{\xi: \xi \mid a=x\} d a \\
& \leqslant \frac{1}{2 \sqrt{2 \alpha}} \int_{-\infty}^{\infty} \sum_{x \in T_{a}}\left[\int_{a}^{\infty} \frac{\int_{E_{+}} f^{2}(\xi \mid c) \mathbf{1}\{\xi \mid a=x\} \mu(d \xi)}{\mu\{\xi: \xi \mid a=x\}} e^{-\sqrt{2 \alpha}(c-a)} d c\right] \\
& \quad \times \mu\{\xi: \xi \mid a=x\} d a \\
&=\frac{1}{2 \sqrt{2 \alpha}} \int_{-\infty}^{\infty} \int_{a}^{\infty}\left[\int_{E_{+}}\left[f^{2}(\xi \mid c) \mu(d \xi)\right] e^{-\sqrt{2 \alpha}(c-a)} d c d a\right. \\
&=\frac{1}{4 \alpha} \int_{-\infty}^{\infty} \int_{E_{+}} f^{2}(\xi \mid c) d c \mu(d \xi)=\frac{1}{4 \alpha} \int_{T} f^{2}(x) \nu(d x)<\infty
\end{aligned}
$$

as required.
From (7.5) we have for $g \in \mathcal{D}$ that

$$
\begin{align*}
& \mathcal{E}^{\prime}\left(G_{\alpha} f, g\right) \\
& =\int_{-\infty}^{\infty} \int_{E+}\left[\int_{a}^{\infty} \mu(\xi \mid a, a, c ; f) e^{-\sqrt{2 \alpha}(c-a)} d c\right] \nabla g(\xi \mid a) \mu(d \xi) d a  \tag{7.6}\\
& \quad-\frac{1}{2} \sqrt{2 \alpha} \int_{-\infty}^{\infty} \int_{E+} G_{\alpha} f(x) \nabla g(\xi \mid a), \mu(d \xi) d a .
\end{align*}
$$

Consider the first term on the right-hand side of (7.6). Note that it can be written as

$$
\begin{align*}
& \int_{-\infty}^{\infty} \quad \sum_{x \in T_{a}}\left[\int_{a}^{\infty} \frac{\int_{E_{+}} f(\xi \mid c) 1\{\xi \mid a=x\} \mu(d \xi)}{\mu\{\xi: \xi \mid a=x\}} e^{-\sqrt{2 \alpha}(c-a)} d c\right] \\
& \quad \times \nabla g(x) \mu\{\xi: \xi \mid a=x\} d a  \tag{7.7}\\
& \quad=\int_{-\infty}^{\infty} \int_{E+}\left[\int_{a}^{\infty} f(x \mid c) e^{-\sqrt{2 \alpha}(c-a)} d c\right] \nabla g(\xi \mid a) \mu(d \xi) d a .
\end{align*}
$$

Substitute (7.7) into (7.6), integrate by parts, and use (7.5) to get that

$$
\begin{aligned}
& \mathcal{E}^{\prime}\left(G_{\alpha} f, g\right)=\int_{E_{+}} \int_{-\infty}^{\infty} f(\xi \mid a) g(\xi \mid a) d a \mu(d x) \\
& \quad-\sqrt{2 \alpha} \int_{E_{+}} \int_{-\infty}^{\infty}\left[\int_{a}^{\infty} f(x \mid c) e^{-\sqrt{2 \alpha}(c-a)} d c\right] g(\xi \mid a) d a \mu(d \xi) \\
& \quad+\sqrt{2 \alpha} \int_{E_{+}} \int_{-\infty}^{\infty}\left[\int_{a}^{\infty} \mu(\xi \mid a, a, c ; f) e^{-\sqrt{2 \alpha}(c-a)} d c\right] g(\xi \mid a) d a \mu(d \xi) \\
& \quad-\alpha \int_{E_{+}} \int_{-\infty}^{\infty} G_{\alpha} f(\xi \mid a) g(\xi \mid a) d a \mu(d x) .
\end{aligned}
$$

Argue as in (7.7) to see that the second and third terms on the right-hand side cancel and so

$$
\mathcal{E}^{\prime}\left(G_{\alpha} f, g\right)=(f, g)-\alpha\left(G_{\alpha} f, g\right),
$$

as required.
Remark 7.6. We wish to apply to $X$ the theory of symmetric processes and their associated Dirichlet forms developed in [72]. Because $T$ is not generally locally compact, we need to to check that the conditions of Theorem A. 8 hold - see Remark A.9.

We first show that conditions (a)-(c) of Theorem A. 8 hold. That is, that there is a countably generated subalgebra $\mathcal{C} \subseteq b C(T) \cap \mathcal{D}$ such that $\mathcal{C}$ is $\mathcal{E}_{1-}$ dense in $\mathcal{D}, \mathcal{C}$ separates points of $T$, and for each $x \in T$ there exists $f \in \mathcal{C}$ with $f(x)>0$. Let $\mathcal{C}_{0}$ be a countable subset of $b C(T) \cap L^{2}(T, \nu)$ that separates points of $T$ and is such that for every $x \in T$ there exists $f \in \mathcal{C}_{0}$ with $f(x)>0$. Let $\mathcal{C}$ be the algebra generated by the countable collection $\bigcup_{\alpha} G_{\alpha} \mathcal{C}_{0}$, where the union is over the positive rationals. It is clear that $\mathcal{C}$ is $\mathcal{E}_{1}$-dense in $\mathcal{D}$. We observed in the proof of Theorem 7.1 that $P_{t}: b C(T) \rightarrow b C(T)$ for all $t \geqslant 0$ and $\lim _{t \downarrow 0} P_{t} f(x)=f(x)$ for all $f \in b C(T)$. Thus, $G_{\alpha}: b C(T) \rightarrow b C(T)$ for all $\alpha>0$ and $\lim _{\alpha \rightarrow \infty} \alpha G_{\alpha} f(x)=f(x)$ for all $f \in b C(T)$. Therefore, $\mathcal{C}$ separates points of $T$ and for every $x \in T$ there exists $f \in \mathcal{C}$ with $f(x)>0$.

It remains to check that the tightness condition (d) of Theorem A. 8 holds. That is, for all $\epsilon>0$ there exists a compact set $K$ such that $\operatorname{Cap}(T \backslash K)<\epsilon$ where Cap denotes the capacity associated with $\mathcal{E}_{1}$. However, it follows from the sample path continuity of $X$ and Theorem IV.1.15 of [106] that, in the terminology of that result, the process $X$ is $\nu$-tight. Conditions IV.3.1 (i) - (iii) of [106] then hold by Theorem IV.5.1 of [106], and this suffices by Theorem III.2.11 of [106] to establish condition that (d) of Theorem A. 8 holds.

### 7.4 Recurrence, transience, and regularity of points

The Green operator $G$ associated with the semigroup $\left(P_{t}\right)_{t \geqslant 0}$ is defined by $G f(x):=\int_{0}^{\infty} P_{t} f(x) d t=\sup _{\alpha>0} G_{\alpha} f(x)$ for $f \in p \mathcal{B}(T)$. In the terminology of [72], we say that $X$ is transient is $G f<\infty$, $\nu$-a.e., for any $f \in L_{+}^{1}(T, \nu)$, whereas $X$ is recurrent if $G f \in\{0, \infty\}, \nu$-a.e., for any $f \in L_{+}^{1}(T, \nu)$.

As we observed in Section 7.3, $X$ has symmetric transition densities $p_{t}(x, y)$ with respect to $\nu$ such that $p_{t}(x, y)>0$ for all $x, y \in T$. Consequently, in the terminology of [72], $X$ is irreducible. Therefore, by Lemma 1.6.4 of [72], $X$ is either transient or recurrent, and if $X$ is recurrent, then $G f=\infty$ for any $f \in L_{+}^{1}(T, \nu)$ that is not $\nu$-a.e. 0 .

Taking limits as $\alpha \downarrow 0$ in (7.4), we see that

$$
G f(x)=\int_{T} g(x, y) f(y) \nu(d y)
$$

where

$$
\begin{align*}
g(x, y) & :=2 \int_{-\infty}^{h(x \wedge y)} \frac{1}{\mu\{\xi: \xi|b=x| b\}} d b  \tag{7.8}\\
& =2 \int_{-\infty}^{h(x \wedge y)} \frac{1}{\mu\{\xi: \xi|b=y| b\}} d b .
\end{align*}
$$

Note that the integrals

$$
\begin{equation*}
\int_{-\infty}^{a} \frac{1}{\mu\{\xi: \xi|b=\zeta| b\}} d b, a \in \mathbb{R}, \zeta \in E_{+} \tag{7.9}
\end{equation*}
$$

are either simultaneously finite or infinite. The following is now obvious.
Theorem 7.7. If the integrals in (7.9) are finite (resp. infinite), then $g(x, y)<$ $\infty$ (resp. $g(x, y)=\infty)$ for all $x, y \in T$ and $X$ is transient (resp. recurrent).

Remark 7.8. For $B \in \mathcal{B}(T)$ write $\sigma_{B}:=\inf \left\{t>0: X_{t} \in B\right\}$. We note from Theorem 4.6.6 and Problem 4.6.3 of [72] that if $\mathbb{P}^{x}\left\{\sigma_{B}<\infty\right\}>0$ for some $x \in T$, then $\mathbb{P}^{x}\left\{\sigma_{B}<\infty\right\}>0$ for all $x \in T$. Moreover, if $X$ is recurrent, then $\mathbb{P}^{x}\left\{\sigma_{B}<\infty\right\}>0$ for some $x \in T$ implies that $\mathbb{P}^{x}\left\{\forall N \in \mathbb{N}, \exists t>N: X_{t} \in B\right\}=$ 1 for all $x \in T$.

Given $y \in T$, write $\sigma_{y}$ for $\sigma_{\{y\}}$. Set $C=\{z \in T: y \leqslant z\}$. Pick $x \leqslant y$ with $x \neq y$. By definition of $\left(P_{t}\right)_{t \geqslant 0}, \mathbb{P}^{x}\left\{X_{t} \in C\right\}>0$ for all $t>0$. In particular, $\mathbb{P}^{x}\left\{\sigma_{C}<\infty\right\}>0$. It follows from Axioms I and II that if $\gamma: \mathbb{R}_{+} \mapsto T$ is any continuous map with $\{x, z\} \subset \gamma\left(\mathbb{R}_{+}\right)$for some $z \in C$, then $y \in \gamma\left(\mathbb{R}_{+}\right)$ also. Therefore, by the sample path continuity of $X, \mathbb{P}^{x}\left\{\sigma_{y}<\infty\right\}>0$ for this particular choice of $x$. However, Remark 7.8 then gives that $\mathbb{P}^{x}\left\{\sigma_{y}<\infty\right\}>0$ for all $x \in T$. By Theorem 4.1.3 of [72] we have that points are regular for themselves. That is, $\mathbb{P}^{x}\left\{\sigma_{x}=0\right\}=1$ for all $x \in T$.

### 7.5 Examples

Recall the the family of $\mathbb{R}$-tree without ends $(T, d)$ construction in Subsection 3.4.3 for a prime number $p$ and constants $r_{-}, r_{+} \geqslant 1$.

In the notation of Subsection 3.4.3, define a Borel measure $\mu$ on $E_{+}$as follows. Write $\ldots \leqslant w_{-1} \leqslant w_{0}=1 \leqslant w_{1} \leqslant w_{2} \leqslant \ldots$ for the possible values of $w(\cdot, \cdot)$. That is, $w_{k}=\sum_{i=0}^{k} r_{+}^{i}$ if $k \geqslant 0$, whereas $w_{k}=1-\sum_{i=0}^{-k} r_{-}^{i}$ if $k<0$. By construction, closed balls in $E_{+}$all have diameters of the form $2^{-w_{k}}$ for some $k \in \mathbb{Z}$ and such a ball is the disjoint union of $p$ balls of diameter $2^{-w_{k+1}}$. We can, therefore, uniquely define $\mu$ by requiring that each closed ball of diameter $2^{-w_{k}}$ has mass $p^{-k}$. The measure $\mu$ is nothing but the (unique up to constants) Haar measure on the locally compact Abelian group $E_{+}$.

Applying Theorem 7.7, we see that $X$ will be transient if and only if $\sum_{k \in \mathbb{N}_{0}} p^{-k} r_{-}^{k}<\infty$, that is, if and only if $r_{-}<p$. As we might have expected, transience and recurrence are unaffected by the value of $r_{+}$: Theorem 7.7 shows that transience and recurrence are features of the structure of $T$ "near" $\dagger$, whereas $r_{+}$only dictates the structure of the $T$ "near" points of $E_{+}$.

### 7.6 Triviality of the tail $\sigma$-field

Theorem 7.9. For all $x \in T$ the tail $\sigma$-field $\bigcap_{s \geqslant 0} \sigma\left\{X_{t}: t \geqslant s\right\}$ is $\mathbb{P}^{x}$-trivial (that is, consists of sets with $\mathbb{P}^{x}$-measure 0 or 1 ).

Proof. Fix $x \in T$. By the continuity of the sample paths of $X, \sigma_{x \mid a}=\inf \{t>$ $\left.0: h\left(X_{t}\right)=a\right\}$. Because $h(X)$ is a Brownian motion, this stopping time is $\mathbb{P}^{x}$-a.s. finite. Put $T_{0}:=0$ and $T_{k}:=\sigma_{x \mid(h(x)-k)}$ for $k=1,2, \ldots$ By the strong Markov property we get that $\mathbb{P}^{x}\left\{T_{1}<T_{2}<\cdots<\infty\right\}=1$. Set $X_{k}(t):=$ $X\left(\left(T_{k}+t\right) \wedge T_{k+1}\right)$ for $k=0,1, \ldots$ Note that the tail $\sigma$-field in the statement of the result can also be written as $\bigcap_{k \geqslant 0} \sigma\left\{\left(T_{\ell}, X_{\ell}\right): \ell \geqslant k\right\}$.

By the strong Markov property, the pairs $\left(\left(T_{k+1}-T_{k}, X_{k}\right)\right)_{k \in \mathbb{N}_{0}}$ are independent. Moreover, by the spatial homogeneity of Brownian motion, the random variables $\left(T_{k+1}-T_{k}\right)_{k \in \mathbb{N}_{0}}$ are identically distributed. The result now follows from Lemma 7.10 below.

Lemma 7.10. Let $\left\{\left(Y_{n}, Z_{n}\right)\right\}_{n \in \mathbb{N}}$ be a sequence of independent $\mathbb{R} \times \mathbf{U}$-valued random variables, where $(\mathbf{U}, \mathcal{U})$ is a measurable space. Suppose further that that the random variables $Y_{n}, n \in \mathbb{N}$, have a common distribution. Put $W_{n}:=$ $Y_{1}+\ldots+Y_{n}$. Then the tail $\sigma$-field $\bigcap_{m \in \mathbb{N}} \sigma\left\{\left(W_{n}, Z_{n}\right): n \geqslant m\right\}$ is trivial.

Proof. Consider a real-valued random variable $V$ that is measurable with respect to the tail $\sigma$-field in the statement. For each $m \in \mathbb{N}$ we have by conditioning on $\sigma\left\{W_{n}: n \geqslant m\right\}$ and using Kolmogorov's zero-one law that there is a $\sigma\left\{W_{n}: n \geqslant m\right\}$-measurable random variable $V_{m}^{\prime}$ such that $V_{m}^{\prime}=V$ almost surely. Consequently, there is a random variable $V^{\prime}$ measurable with respect to $\bigcap_{m \in \mathbb{N}} \sigma\left\{W_{n}: n \geqslant m\right\}$ such that $V^{\prime}=V$ almost surely, and the proof is completed by an application of the Hewitt-Savage zero-one law.
Definition 7.11. A function $f \in \mathcal{B}\left(T \times \mathbb{R}_{+}\right.$) (resp. $f \in \mathcal{B}(T)$ ) is said to be space-time harmonic (resp. harmonic ) if $0 \leqslant f<\infty$ and $P_{s} f(\cdot, t)=f(\cdot, s+t)$ (resp. $P_{s} f=f$ ) for all $s, t \geqslant 0$.

Remark 7.12. There does not seem to be a generally agreed upon convention for the use of the term "harmonic". It is often used for the analogous definition without the requirement that the function is non-negative, and $P_{t} f(x)=$ $\mathbb{P}^{x}\left[f\left(X_{t}\right)\right]$ is sometimes replaced by $\mathbb{P}^{x}\left[f\left(X_{\tau}\right)\right]$ for suitable stopping times $\tau$. Also, the terms invariant and regular are sometimes used.

The following is a standard consequence of the triviality of the tail $\sigma$-field and irreducibility of the process, but we include a proof for completeness.

Corollary 7.13. There are no non-constant bounded space-time harmonic functions (and, a fortiori, no non-constant bounded harmonic functions).

Proof. Suppose that $f$ is a bounded space-time harmonic function. For each $x \in T$ and $s \geqslant 0$ the process $\left(f\left(X_{t}, s+t\right)\right)_{t \geqslant 0}$ is a bounded $\mathbb{P}^{x}$-martingale. Therefore $\lim _{t \rightarrow \infty} f\left(X_{t}, s+t\right)$ exists $\mathbb{P}^{x}$-a.s. and $f(x, s)=\mathbb{P}^{x}\left[\lim _{t \rightarrow \infty} f\left(X_{t}, s+\right.\right.$ $t)]=\lim _{t \rightarrow \infty} f\left(X_{t}, s+t\right), \mathbb{P}^{x}$-a.s., by the triviality of the tail. By the Markov property and the fact that $X$ has everywhere positive transition densities with respect to $\nu$ we get that $f(s, x)=f(t, y)$ for $\nu$-a.e. $y$ for each $t>s$, and it is clear from this that $f$ is a constant.

Remark 7.14. The conclusion of Corollary 7.13 for harmonic functions has the following alternative probabilistic proof. By the arguments in the proof of Theorem 7.9 we have that if $n \in \mathbb{Z}$ is such that $n<h(x)$, then $\mathbb{P}^{x}\left\{\sigma_{\left.x\right|_{n}}<\right.$ $\left.\sigma_{x \mid(n-1)}<\sigma_{x \mid(n-2)}<\cdots<\infty\right\}=1$. Suppose that $f$ is a bounded harmonic function. Then $f(x)=\mathbb{P}^{x}\left[\lim _{t \rightarrow \infty} f\left(X_{t}\right)\right]=\lim _{k \rightarrow \infty} f(x \mid(-k))$. Now note for each pair $x, y \in T$ that $x|(-k)=y|(-k)$ for $k \in \mathbb{N}$ sufficiently large.

### 7.7 Martin compactification and excessive functions

Suppose in this section that $X$ is transient. Recall that $f \in \mathcal{B}(T)$ is excessive for $\left(P_{t}\right)_{t \geqslant 0}$ if $0 \leqslant f<\infty, P_{t} f \leqslant f$, and $\lim _{t \downarrow 0} P_{t} f=f$ pointwise. Recall the definition of harmonic function from Section 7.6. In this section we will obtain an integral representation for the excessive and harmonic functions.

Fix $x_{0} \in T$ and define $k: T \times T \rightarrow \mathbb{R}$, the corresponding Martin kernel, by

$$
\begin{align*}
k(x, y) & :=\frac{g(x, y)}{g\left(x_{0}, y\right)}=\frac{\int_{-\infty}^{h(x \wedge y)} \mu\{\xi: \xi|b=y| b\}^{-1} d b}{\int_{-\infty}^{h\left(x_{0} \wedge y\right)} \mu\{\xi: \xi|b=y| b\}^{-1} d b}  \tag{7.10}\\
& =\frac{\int_{-\infty}^{h(x \wedge y)} \mu\{\xi: \xi|b=x| b\}^{-1} d b}{\int_{-\infty}^{h\left(x_{0} \wedge y\right)} \mu\left\{\xi: \xi\left|b=x_{0}\right| b\right\}^{-1} d b}
\end{align*}
$$

Note that the function $k$ is continuous in both arguments and

$$
0<\mathbb{P}^{x}\left\{\sigma_{x_{0}}<\infty\right\} \leqslant k(x, y)=\frac{\mathbb{P}^{x}\left\{\sigma_{y}<\infty\right\}}{\mathbb{P}^{x_{0}}\left\{\sigma_{y}<\infty\right\}} \leqslant \mathbb{P}^{x_{0}}\left\{\sigma_{x}<\infty\right\}^{-1}<\infty
$$

We can follow the standard approach to constructing a Martin compactification when there are well-behaved potential kernel densities (e.g. [94, 108]). That is, we choose a countable, dense subset $S \subset T$ and compactify $T$ using the sort of Stone-C̆ech-like procedure described in Section 3.4.2 to obtain a metrizable compactification $T^{M}$ such that a sequence $\left\{y_{n}\right\}_{n \in \mathbb{N}} \subset T$ converges if and only if $\lim _{n} k\left(x, y_{n}\right)$ exists for all $x \in T$. We discuss the analytic interpretation of the Martin compactification later in this section. We investigate the probabilistic features of the compactification and the connection with Doob $h$-transforms in Section 7.8. We first show that $T^{M}$ coincides with the compactification $\bar{T}$ of Section 3.4.2.

Proposition 7.15. The compact metric spaces $\bar{T}$ and $T^{M}$ are homeomorphic, so that $T^{M}$ can be identified with $T \cup E$. If we define

$$
k(x, \eta):=\frac{\int_{-\infty}^{h(x \wedge \eta)} \mu\{\xi: \xi|b=\eta| b\}^{-1} d b}{\int_{-\infty}^{h\left(x_{0} \wedge \eta\right)} \mu\{\xi: \xi|b=\eta| b\}^{-1} d b}, x \in T, \eta \in T \cup E_{+},
$$

and $k(x, \dagger)=1$, then $k(x, \cdot)$ is continuous on $T \cup E$. Moreover,

$$
\sup _{x \in B} \sup _{\eta \in T \cup E} k(x, \eta)<\infty
$$

for all balls $B \subset T$.
Proof. The rest of the proof will be almost immediate once we show for a sequence $\left\{y_{n}\right\}_{n \in \mathbb{N}} \subset T$ that $\lim _{n} k\left(x, y_{n}\right)$ exists for all $x \in T$ if and only if $\lim _{n} h\left(x \wedge y_{n}\right)$ exists (in the extended sense) for all $x \in T$.

It is clear that if $\lim _{n} h\left(x \wedge y_{n}\right)$ exists for all $x \in T$, then $\lim _{n} k\left(x, y_{n}\right)$ exists for all $x \in T$.

Suppose, on the other hand, that $\lim _{n} k\left(x, y_{n}\right)$ exists for all $x \in T$ but $\lim _{n} h\left(x^{\prime} \wedge y_{n}\right)$ does not exist for some $x^{\prime} \in T$. Then we can find $\epsilon>0$ and $a<h\left(x^{\prime}\right)-\epsilon$ such that $x^{\prime \prime}:=x^{\prime} \mid a \in T, \liminf _{n} h\left(x^{\prime} \wedge y_{n}\right) \leqslant a-\epsilon$, and $\lim \sup _{n} h\left(x^{\prime} \wedge y_{n}\right) \geqslant a+\epsilon$. This implies that for any $N \in \mathbb{N}$ there exists $p, q \geqslant$ $N$ such that $h\left(x^{\prime \prime} \wedge y_{p}\right)=h\left(x^{\prime} \wedge y_{p}\right)$ and $h\left(x^{\prime \prime} \wedge y_{q}\right)=a<a+\epsilon / 2<h\left(x^{\prime} \wedge y_{q}\right)$. Thus, we obtain the contradiction

$$
\liminf _{n} \frac{k\left(x^{\prime}, y_{n}\right)}{k\left(x^{\prime \prime}, y_{n}\right)}=\liminf _{n} \frac{g\left(x^{\prime}, y_{n}\right)}{g\left(x^{\prime \prime}, y_{n}\right)}=1
$$

while

$$
\begin{aligned}
\limsup _{n} \frac{k\left(x^{\prime}, y_{n}\right)}{k\left(x^{\prime \prime}, y_{n}\right)} & =\limsup _{n} \frac{g\left(x^{\prime}, y_{n}\right)}{g\left(x^{\prime \prime}, y_{n}\right)} \\
& \geqslant \frac{\int_{-\infty}^{a+\epsilon / 2} \mu\left\{\xi: \xi\left|b=x^{\prime}\right| b\right\}^{-1} d b}{\int_{-\infty}^{a} \mu\left\{\xi: \xi\left|b=x^{\prime}\right| b\right\}^{-1} d b}>1
\end{aligned}
$$

The following theorem essentially follows from results in [108], with most of the work that is particular to our setting being the argument that the points of $E_{+}$are, in the terminology of [108], . Unfortunately, the standing assumption in [108] is that the state-space is locally compact. The requirement for this hypothesis can be circumvented using the special features of our process, but checking this requires a fairly close reading of much of [108]. Later, more probabilistic or measure-theoretic, approaches to the Martin boundary such as $[51,74,73,86]$ do not require local compactness, but are rather less concrete and less pleasant to compute with. Therefore, we sketch the relevant arguments.

Definition 7.16. An excessive function $f$ is said to be a potential if

$$
\lim _{t \rightarrow \infty} P_{t} f=0
$$

(The term purely excessive function is also sometimes used.)
Theorem 7.17. If $u$ is an excessive function, then there is a unique finite measure $\gamma$ on $\bar{T}=T \cup E$ such that $u(x)=\int_{T \cup E} k(x, \eta) \gamma(d \eta), x \in T$. Furthermore, $u$ is harmonic (resp. a potential) if and only if $\gamma(T)=0$ (resp. $\gamma(E)=0)$.

Proof. From Theorem XII. 17 in [43] there exists a sequence $\left\{f_{n}\right\}_{n \in \mathbb{N}}$ of bounded non-negative functions such that $G f_{n}$ is bounded for all $n$ and $G f_{1}(x) \leqslant G f_{2}(x) \leqslant \ldots \leqslant G f_{n}(x) \uparrow u(x)$ as $n \rightarrow \infty$ for all $x \in T$. Define a measure $\gamma_{n}$ by $\gamma_{n}(d y):=g\left(x_{0}, y\right) f_{n}(y) \nu(d y)$, so that $G f_{n}(x)=\int_{T} k(x, y) \gamma_{n}(d y)$. Note that $\gamma_{n}(T)=G f_{n}\left(x_{0}\right) \leqslant u\left(x_{0}\right)<\infty$. We can think of $\left\{\gamma_{n}\right\}_{n \in \mathbb{N}}$ as a sequence of finite measures on the compact space $\bar{T}$ with bounded total mass. Therefore, there exists a subsequence $\left(n_{\ell}\right)_{\ell \in \mathbb{N}}$ such that $\gamma=\lim _{\ell} \gamma_{n_{\ell}}$ exists in the topology of weak convergence of finite measures on $\bar{T}$. By Proposition 7.15, each of the functions $k(x, \cdot)$ is bounded and continuous, and so

$$
\begin{aligned}
\int_{T \cup E} k(x, \eta) \gamma(d \eta) & =\lim _{\ell} \int_{T \cup E} k(x, \eta) \gamma_{n_{\ell}}(d \eta) \\
& =\lim _{\ell} \int_{T} k(x, y) \gamma_{n_{\ell}}(d y) \\
& =\lim _{\ell} G f_{n_{\ell}}(x)=u(x)
\end{aligned}
$$

This completes the proof of existence. We next consider the the uniqueness claim.

Note first of all that the set of excessive functions is a cone; that is, it is closed under addition and multiplication by non-negative constants. This cone has an associated strong order: we say that $f \ll g$ for two excessive functions if $g=f+h$ for some excessive function $h$. As remarked in XII. 34 of [43], for any two excessive functions $f$ and $g$ there is a greatest lower bound excessive function $h$ such that $h \ll f, h \ll g$ and $h^{\prime} \ll h$ for any other excessive
function $h^{\prime}$ with $h^{\prime} \ll f$ and $h^{\prime} \ll g$. There is a similarly defined least upper bound. Moreover, if $h$ and $k$ are respectively the greatest lower bounds and least upper bounds of two excessive functions $f$ and $g$, then $f+g=h+k$. Thus, the cone of excessive functions is a lattice in the strong order.

From Proposition 7.15 , all excessive functions are bounded on balls and a fortiori $\nu$-integrable on balls. Thus, the excessive functions are a subset of the separable, locally convex, topological vector space $L_{\mathrm{loc}}^{1}(T, \nu)$ of locally $\nu$-integrable functions equipped with the metrizable topology of $L^{1}(T, \nu)$ convergence on balls.

Consider the convex set of excessive functions $u$ such that $u\left(x_{0}\right)=1$. Any measure appearing in the representation of such a function $u$ is necessarily a probability measure. Given a sequence $\left\{u_{n}\right\}_{n \in \mathbb{N}}$ of such functions, we can, by the weak compactness argument described above, find a subsequence $\left(u_{n_{\ell}}\right)_{\ell \in \mathbb{N}}$ that converges bounded pointwise, and, therefore, also in $L_{\text {loc }}^{1}(T, \nu)$, to some limit $u$. Thus, the set of excessive functions $u$ such that $u\left(x_{0}\right)=1$ is convex, compact and metrizable.

An arbitrary excessive function is a non-negative multiple of an excessive function $u$ with $u\left(x_{0}\right)=1$. Consequently, the cone of excessive functions is a cone in a locally convex, separable, topological vector space with a compact and metrizable base and this cone is a lattice in the associated strong order. The Choquet uniqueness theorem - see Theorem X. 64 of [43] - guarantees that every excessive function $u$ with $u\left(x_{0}\right)=1$ can be represented uniquely as an integral over the extreme points of the compact convex set of such functions.

Write $k_{\eta}$ for the excessive function $k(\cdot, \eta), \eta \in T \cup E$. The uniqueness claim will follow provided we can show for all $\eta \in T \cup E$ that the function $k_{\eta}$ is an extreme point. That is, we must show that if $k_{\eta}=\int_{T \cup E} k_{\eta^{\prime}} \gamma\left(d \eta^{\prime}\right)$ for some probability measure $\gamma$, then $\gamma$ is necessarily the point mass at $\eta$.

Each of the functions $k_{y}, y \in T$, is clearly a potential. A direct calculation using (7.4), which we omit, shows that if $\xi \in E$, then $\alpha G_{\alpha} k_{\xi}=k_{\xi}$ for all $\alpha>0$, and this implies that $k_{\xi}$ is harmonic. Thus, $\lim _{t \rightarrow \infty} P_{t} k_{\eta}$ is either 0 or $k_{\eta}$ depending on whether $\eta \in T$ or $\eta \in E$. In particular, if $k_{\eta}=\int_{T \cup E} k_{\eta^{\prime}} \gamma\left(d \eta^{\prime}\right)$, then

$$
\lim _{t \rightarrow \infty} P_{t} k_{\eta}=\int_{T \cup E} \lim _{t \rightarrow \infty} P_{t} k_{\eta^{\prime}} \gamma\left(d \eta^{\prime}\right)=\int_{E} k_{\eta^{\prime}} \gamma\left(d \eta^{\prime}\right)
$$

Thus, $\gamma$ must be concentrated on $T$ if $\eta \in T$ and on $E$ if $\eta \in E$.
Suppose now for $y \in T$ that

$$
k_{y}(x)=\int_{T} k_{y^{\prime}}(x) \gamma\left(d y^{\prime}\right)
$$

or, equivalently, that

$$
\frac{g(x, y)}{g\left(x_{0}, y\right)}=\int_{T} \frac{g\left(x, y^{\prime}\right)}{g\left(x_{0}, y^{\prime}\right)} \gamma\left(d y^{\prime}\right)
$$

Thus, we have

$$
\int_{T} g\left(x, y^{\prime}\right) \pi\left(d y^{\prime}\right)=\int_{T} g\left(x, y^{\prime}\right) \rho\left(d y^{\prime}\right)
$$

where $\pi$ is the measure $\delta_{y} / g\left(x_{0}, y\right)$ and $\rho$ is the measure $\gamma / g\left(x_{0}, \cdot\right)$. Let $g_{\alpha}$ be the kernel corresponding to the operator $G$; that is,

$$
G f(x)=\int_{T} g_{\alpha}(x, y) f(y) \nu(d y)
$$

It is straightforward to check that $\alpha G_{\alpha} G=G-G_{\alpha}$ (this is just a special instance of the resolvent equation). Thus

$$
\int_{T} g_{\alpha}\left(x, y^{\prime}\right) \pi\left(d y^{\prime}\right)=\int_{T} g_{\alpha}\left(x, y^{\prime}\right) \rho\left(d y^{\prime}\right)
$$

and

$$
\int_{T} \int_{T} f(x) g_{\alpha}\left(x, y^{\prime}\right) \pi\left(d y^{\prime}\right) \nu(d x)=\int_{T} \int_{T} f(x) g_{\alpha}\left(x, y^{\prime}\right) \rho\left(d y^{\prime}\right) \nu(d x)
$$

for any bounded continuous function $f$. Since $g_{\alpha}$ is symmetric,

$$
\int_{T} f(x) g_{\alpha}\left(x, y^{\prime}\right) \nu(d x)=\int_{T} g_{\alpha}\left(y^{\prime}, x\right) f(x) \nu(d x)
$$

Moreover,

$$
\alpha \int_{T} g_{\alpha}\left(y^{\prime}, x\right) f(x) \nu(d x) \leqslant \sup _{x \in T}|f(x)|
$$

and

$$
\lim _{\alpha \rightarrow \infty} \alpha \int_{T} g_{\alpha}\left(y^{\prime}, x\right) f(x) \nu(d x)=f\left(y^{\prime}\right)
$$

for all $y^{\prime} \in T$. Thus, $\int_{T} f\left(y^{\prime}\right) \pi\left(d y^{\prime}\right)=\int_{T} f\left(y^{\prime}\right) \rho\left(d y^{\prime}\right)$ for any bounded continuous function $f$, and $\pi=\rho$ as required. The argument we have just given is essentially a special case of the principle of masses - see, for example, Proposition 1.1 of [75].

Similarly, suppose for some $\xi \in E_{+}$that $k_{\xi}(x)=\int_{E} k_{\xi^{\prime}}(x) \gamma(d \xi)$. For $x \in T$ and $a>h(x \wedge \xi)$

$$
\begin{aligned}
k_{\xi}(x) \geqslant & \mathbb{P}^{x}\left[k_{\xi}\left(X_{\sigma_{\xi \mid a}}\right)\right] \\
= & \frac{g(x, \xi \mid a)}{g(\xi|a, \xi| a)} k(\xi \mid a, \xi) \\
= & \frac{\int_{-\infty}^{h(x \wedge(\xi \mid a))} \mu\{\zeta: \zeta|b=(\xi \mid a)| b\}^{-1} d b}{\int_{-\infty}^{h(\xi \mid a)} \mu\{\zeta: \zeta|b=(\xi \mid a)| b\}^{-1} d b} \\
& \times \frac{\int_{-\infty}^{h((\xi \mid a) \wedge \xi)} \mu\{\zeta: \zeta=\xi \mid b\}^{-1} d b}{\int_{-\infty}^{h\left(x_{0} \wedge \xi\right)} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b} \\
= & \frac{\int_{-\infty}^{h(x \wedge \xi)} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b}{\int_{-\infty}^{a} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b} \\
& \times \frac{\int_{-\infty}^{a} \mu\{\zeta: \zeta=\xi \mid b\}^{-1} d b}{\int_{-\infty}^{h\left(x_{0} \wedge \xi\right)} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b} \\
= & k_{\xi}(x)
\end{aligned}
$$

Thus, $k_{\xi}(x)=\mathbb{P}^{x}\left[k_{\xi}\left(X_{\sigma_{\xi \mid a}}\right)\right]$ for all $a$ sufficiently large. On the other hand, a similar argument shows for $\xi^{\prime} \in E_{+} \backslash\{\xi\}$ that

$$
k_{\xi^{\prime}}(x) \geqslant \mathbb{P}^{x}\left[k_{\xi^{\prime}}\left(X_{\sigma_{\xi \mid a}}\right)\right]
$$

and

$$
\mathbb{P}^{x}\left[k_{\xi^{\prime}}\left(X_{\sigma_{\xi \mid a}}\right)\right]=\frac{\int_{-\infty}^{h\left(\xi \wedge \xi^{\prime}\right)} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b}{\int_{-\infty}^{a} \mu\{\zeta: \zeta|b=\xi| b\}^{-1} d b} k_{\xi^{\prime}}(x),
$$

for sufficiently large $a$, where the right-hand side converges to 0 as $a \rightarrow 0$. Similarly, $\lim _{a \rightarrow \infty} \mathbb{P}^{x}\left[k_{\dagger}\left(X_{\sigma_{\xi \mid \alpha}}\right)\right]=0$. This clearly shows that if $k_{\xi}=\int_{E} k_{\xi^{\prime}} \gamma\left(d \xi^{\prime}\right)$, then $\gamma$ cannot assign any mass to $E \backslash\{\xi\}$. Uniqueness for the representation of $k_{\dagger}$ is handled similarly and the proof of the uniqueness claim is complete.

Lastly, the claim regarding representation of harmonic functions and potentials is immediate from what we have already shown.

Remark 7.18. Theorem 7.17 can be used as follows to give an analytic proof (in the transient case) of the conclusion of Corollary 7.13 that bounded harmonic functions are necessarily constant.

First extend the definition of the Green kernel $g$ to $T \cup E$ by setting

$$
\begin{aligned}
g(\eta, \rho) & :=2 \int_{-\infty}^{h(\eta \wedge \rho)} \mu\{\zeta: \zeta|b=\eta| b\}^{-1} d b \\
& =2 \int_{-\infty}^{h(\eta \wedge \rho)} \mu\{\zeta: \zeta|b=\rho| b\}^{-1} d b
\end{aligned}
$$

By Theorem 7.17, non-constant bounded harmonic functions exist if and only if there is a non-trivial finite measure $\gamma$ concentrated on $E_{+}$such that

$$
\begin{equation*}
\sup _{x \in T} \int_{E_{+}} k(x, \zeta) \gamma(d \zeta)<\infty \tag{7.11}
\end{equation*}
$$

Note that for any ball $B \subset E_{+}$of the form $B=\left\{\zeta \in E_{+}: \zeta \mid h\left(x^{*}\right)=x^{*}\right\}$ for $h\left(x^{*}\right) \geqslant h\left(x_{0}\right)$ we have $g\left(x_{0}, \zeta\right)=g\left(x_{0}, x^{*}\right)$. Thus, by possibly replacing the measure $\gamma$ in (7.11) by its trace on a ball, we have that non-constant bounded harmonic functions exist if and only if there is a probability measure (that we also denote by $\gamma$ ) concentrated on a ball $B \subset E_{+}$such that

$$
\begin{equation*}
\sup _{x \in T} \int_{B} g(x, \zeta) \gamma(d \zeta)<\infty \tag{7.12}
\end{equation*}
$$

Observe that $g(\xi \mid t, \zeta)$ increases monotonically to $g(\xi, \zeta)$ as $t \rightarrow \infty$ and so, by monotone convergence, (7.12) holds if and only if

$$
\begin{equation*}
\sup _{\xi \in E_{+}} \int_{B} g(\xi, \zeta) \gamma(d \zeta)<\infty \tag{7.13}
\end{equation*}
$$

It is further clear that if (7.13) holds, then

$$
\begin{equation*}
\int_{B} \int_{B} g(\xi, \zeta) \gamma(d \xi) \gamma(d \zeta)<\infty \tag{7.14}
\end{equation*}
$$

Suppose that (7.14) holds. For $b \in \mathbb{R}$ write $T_{b}^{\gamma}$ for the subset of $T_{b}$ consisting of $x \in T_{b}$ such that $\gamma\{\xi \in B: \eta \mid b=x\}>0$. In other words, $T_{b}^{\gamma}$ is the collection of points of the form $\eta \mid b$ for some $\eta$ in the closed support of $\gamma$. Note that $\sum_{x \in T_{b}^{\gamma}} \mu\{\eta: \eta \mid b=x\} \leqslant \mu(B)$ if $2^{-b}$ is at most the diameter of $B$. Applying Jensen's inequality, we obtain the contradiction

$$
\begin{aligned}
& \int_{B} \int_{B} g(\xi, \zeta) \gamma(d \xi) \gamma(d \zeta) \\
& \quad=2 \int_{-\infty}^{\infty} \int_{B} \int_{B} \frac{\mathbf{1}\{\xi|b=\zeta| b\}}{\mu\{\eta: \eta|b=\xi| b\}} \gamma(d \xi) \gamma(d \zeta) d b \\
& \quad=2 \int_{-\infty}^{\infty} \int_{B} \frac{\gamma\{\eta: \eta|b=\xi| b\}}{\mu\{\eta: \eta|b=\xi| b\}} \gamma(d \xi) d b \\
& \quad \geqslant 2 \int_{-\infty}^{\infty}\left[\int_{B} \frac{\mu\{\eta: \eta|b=\xi| b\}}{\gamma\{\eta: \eta|b=\xi| b\}} \gamma(d \xi)\right]^{-1} d b \\
& \quad=2 \int_{-\infty}^{\infty}\left[\sum_{x \in T_{b}^{\gamma}} \frac{\mu\{\eta: \eta \mid b=x\}}{\gamma\{\eta: \eta \mid b=x\}} \gamma\{\eta: \eta \mid b=x\}\right]^{-1} d b \\
& \quad=\infty
\end{aligned}
$$

### 7.8 Probabilistic interpretation of the Martin compactification

Suppose that $X$ is transient and consider the harmonic functions $k_{\xi}=k(\cdot, \xi)$, $\xi \in E_{+}$, introduced in Section 7.7 and the corresponding Doob h-transform
laws $\mathbb{P}_{k_{\xi}}^{x}, x \in T$. That is, $\mathbb{P}_{k_{\xi}}^{x}, x \in T$, is the collection of laws of a Markov process $X^{\xi}$ such that $\mathbb{P}_{k_{\xi}}^{x}\left[f\left(X_{t}^{\xi}\right)\right]=k_{\xi}(x)^{-1} \mathbb{P}^{x}\left[k_{\xi}\left(X_{t}\right) f\left(X_{t}\right)\right], f \in b \mathcal{B}(T)$. The following result says that the process $X^{\xi}$ can be thought of as " $X$ conditioned to converge to $\xi$."
Theorem 7.19. For all $x \in T, \mathbb{P}_{k_{\xi}}^{x}\left\{\lim _{t \rightarrow \infty} X_{t}^{\xi}=\xi\right\}=1$.
Proof. Note that $X^{\xi}$ has Green kernel $k_{\xi}(x)^{-1} g(x, y) k_{\xi}(y)<\infty$. Thus, $X^{\xi}$ is transient.

Now observe that $\lim _{t \rightarrow \infty} X_{t}^{\xi}$ exists. This is so because, by compactness, the limit exists along a subsequence and if two subsequences had different limits then there would be a ball in $T$ that was visited infinitely often contradicting transience.

Thus, it suffices to show that if $a>h(x \wedge \xi)$, then $\mathbb{P}_{k_{\xi}}^{x}\left\{\sigma_{\xi \mid a}<\infty\right\}=1$. However, after some algebra,

$$
\begin{aligned}
\mathbb{P}_{k_{\xi}}^{x}\left\{\sigma_{\xi \mid a}<\infty\right\} & =k_{\xi}(x)^{-1} \mathbb{P}^{x}\left[k_{\xi}\left(X_{\xi \mid a}\right) \mathbf{1}\left\{\sigma_{\xi \mid a}<\infty\right\}\right] \\
& =\frac{1}{k(x, \xi)} \frac{g(x, \xi \mid a)}{g(\xi|a, \xi| a)} k(\xi \mid a, \xi)=1
\end{aligned}
$$

Remark 7.20. Recall that $\left(h\left(X_{t}\right)\right)_{t \geqslant 0}$ is a standard Brownian motion under $\mathbb{P}^{x}$. We can ask what $\left(h\left(X_{t}^{\xi}\right)\right)_{t \geqslant 0}$ looks like under $\mathbb{P}_{k_{\varepsilon}}^{x}$. Arguing as in the proof of Theorem 7.24 below and using Girsanov's theorem, we have under $\mathbb{P}_{k_{\xi}}^{x}$ that

$$
h\left(X_{t}^{\xi}\right)=h\left(X_{0}^{\xi}\right)+W_{t}+D_{t}
$$

where $W$ is a standard Brownian motion and

$$
D_{t}=\int_{0}^{t}\left[\frac{1\left\{X_{s} \leqslant \xi\right\}}{\mu\left\{\zeta: X_{s} \leqslant \zeta\right\}}\right] /\left[\int_{-\infty}^{h\left(X_{s}\right)} \frac{1}{\mu\left\{\zeta: X_{s} \mid b \leqslant \zeta\right\}} d b\right] d s
$$

In other words, when $X_{t}^{\xi}$ is not on the ray $R_{\xi}$ the height process $h\left(X_{t}^{\xi}\right)$ evolves as a standard Brownian motion, but when $X_{t}^{\xi}$ is on the ray $R_{\xi}:=\{x \in T$ : $x \leqslant \xi\}$ the height experiences an added positive drift toward $\xi$.

### 7.9 Entrance laws

A probability entrance law for the semigroup $\left(P_{t}\right)_{t \geqslant 0}$ is a family $\left(\gamma_{t}\right)_{t>0}$ of probability measures on $T$ such that $\gamma_{s} P_{t}=\gamma_{s+t}$ for all $s, t>0$. Given such a probability entrance law, we can construct on some probability space $(\Omega, \mathcal{F}, \mathbb{P})$ a continuous process that, with a slight abuse of notation, we denote $X=\left(X_{t}\right)_{t>0}$ such that $X_{t}$ has law $\gamma_{t}$ and $X$ is a time-homogeneous Markov process with transition semigroup $\left(P_{t}\right)_{t \geqslant 0}$.

In this section we show that the only probability entrance laws are the trivial ones (that is, there is no way to start the process "from infinity" in some sense).
Theorem 7.21. If $\left(\gamma_{t}\right)_{t>0}$ is a probability entrance law for $\left(P_{t}\right)_{t \geqslant 0}$, then $\gamma_{t}=$ $\gamma_{0} P_{t}, t>0$, for some probability measure $\gamma_{0}$ on $T$.
Proof. Construct a Ray-Knight compactification ( $T^{R}, \rho$ ), say, as in Section17 of [128]. Write $\left(\bar{P}_{t}\right)_{t \geqslant 0}$ and $\left(\bar{G}_{\alpha}\right)_{\alpha>0}$ for the corresponding extended semigroup and resolvent.

Construct $X$ with one-dimensional distributions $\left(\gamma_{t}\right)_{t>0}$ and semigroup $\left(P_{t}\right)_{t \geqslant 0}$ as described above. By Theorem 40.4 of [128], $\lim _{t \downarrow 0} X_{t}$ exists in the Ray topology, and if $\gamma_{0}$ denotes the law of this limit, then $\gamma_{0} \bar{P}_{t}$ is concentrated on $T$ for all $t>0$ and $\gamma_{t}$ is the restriction of $\gamma_{0} \bar{P}_{t}$ to $T$. We need, therefore, to establish that $\gamma_{0}$ is concentrated on $T$. Moreover, it suffices to consider the case when $\gamma_{0}$ is a point mass at some $x_{0} \in T^{R}$, so that $\lim _{t \downarrow 0} X_{t}=x_{0}$ in the Ray topology. Note by Theorem 4.10 of [128] that the germ $\sigma$-field $\mathcal{F}_{0+}:=\bigcap_{\epsilon} \sigma\left\{X_{t}: 0 \leqslant t \leqslant \epsilon\right\}$ is trivial under $\mathbb{P}$ in this case.

By construction of $\left(P_{t}\right)_{t \geqslant 0}$, the family obtained by pushing forward each $\gamma_{t}$ by the map $h$ is an entrance law for standard Brownian motion on $\mathbb{R}$. Because Brownian motion is a Feller-Dynkin process, the only entrance laws for it are the trivial ones $\left(\rho Q_{t}\right)_{t>0}$, where $\left(Q_{t}\right)_{t \geqslant 0}$ is the semigroup of Brownian motion and $\rho$ is a probability measure on $\mathbb{R}$. Thus, by the triviality $\mathcal{F}_{0+}$, there is a constant $h_{0} \in \mathbb{R}$ such that $\lim _{t \downarrow 0} h\left(X_{t}\right)=h_{0}$, $\mathbb{P}$-a.s.

As usual, regard functions on $T$ as functions on $T^{R}$ by extending them to be 0 on $T^{R} \backslash T$. For every $f \in b \mathcal{B}(T)$ we have by Theorem 40.4 of [128] that $\lim _{t \downarrow 0} G_{\alpha} f\left(X_{t}\right)=\lim _{t \downarrow 0} \bar{G}_{\alpha} f\left(X_{t}\right)=\bar{G}_{\alpha} f(x)$.

From (7.4),

$$
G_{\alpha} f(x)=\int_{T} g_{\alpha}(x, y) f(y) \nu(d y)
$$

where

$$
\begin{align*}
g_{\alpha}(x, y) & :=2 \int_{-\infty}^{h(x \wedge y)} \frac{\exp (-\sqrt{2 \alpha}(h(x)+h(y)-2 b))}{\mu\{\xi: \xi|b=x| b\}} d b  \tag{7.15}\\
& =2 \int_{-\infty}^{h(x \wedge y)} \frac{\exp (-\sqrt{2 \alpha}(h(x)+h(y)-2 b))}{\mu\{\xi: \xi|b=y| b\}} d b
\end{align*}
$$

It follows straightforwardly that $\lim _{t \downarrow 0} h\left(X_{t} \wedge y\right)$ exists for all $y \in T$, $\mathbb{P}$-a.s., and so, by the discussion in Section 3.4.2 and the triviality of $\mathcal{F}_{0+}$, there exists $\eta \in T \cup E$ such that $h(\eta) \leqslant h_{0}$ and $\lim _{t \downarrow 0} h\left(X_{t} \wedge y\right)=h(\eta \wedge y), \mathbb{P}$-a.s. Note, in particular, that we actually have $\eta \in T \cup\{\dagger\}$ because $h(\eta)<\infty$. Moreover, we conclude that

$$
\begin{aligned}
& \int_{0}^{\infty} e^{-\alpha t} \gamma_{t}(f) d t=\bar{G}_{\alpha} f\left(x_{0}\right) \\
& \quad=2 \int_{T}\left[\int_{-\infty}^{h(\eta \wedge y)} \frac{\exp \left(-\sqrt{2 \alpha}\left(h_{0}+h(y)-2 b\right)\right)}{\mu\{\xi: \xi|b=y| b\}} d b\right] \nu(d y)
\end{aligned}
$$

for all $f \in b \mathcal{B}(T)$.
We cannot have $\eta=\dagger$, because this would imply that $\gamma_{t}$ is the null measure for all $t>0$. If $\eta \in T$ and $h_{0}=h(\eta)$, then we have $\gamma_{t}=\delta_{\eta} P_{t}$.

We need, therefore, only rule out the possibility that $\eta \in T$ but $h(\eta)<h_{0}$. In this case we have

$$
\int_{0}^{\infty} e^{-\alpha t} \gamma_{t}(f) d t=\exp \left(-\sqrt{2 \alpha}\left(h_{0}-h(\eta)\right)\right) \int_{0}^{\infty} e^{-\alpha t} \delta_{\eta} P_{t}(f) d t
$$

and so, by comparison of Laplace transforms, $\gamma_{t}=\int_{0}^{t} \delta_{\eta} P_{t-s} \kappa(d s)$, where $\kappa$ is a certain stable- $\frac{1}{2}$ distribution. In particular, $\gamma_{t}$ has total mass $\kappa([0, t])<1$ and is not a probability distribution.

### 7.10 Local times and semimartingale decompositions

Our aim in this section is to give a semimartingale decomposition for the process $H_{\xi}(t):=h\left(X_{t} \wedge \xi\right), t \geqslant 0$, for $\xi \in E_{+}$.

This result will be analogous to the classical Tanaka's formula for a standard Brownian motion $B$ that says

$$
B(t)_{+}=B(0)_{+}+\int_{0}^{t} \mathbf{1}\{B(s)>0\} d B(s)+\frac{1}{2} \ell(t)
$$

where $\ell$ is the local time of the Brownian motion at 0 . In other words, $B_{+}$ is constant (at 0 ) over time intervals when $B<0$ and during time intervals when $B \geqslant 0$ it evolves like a standard Brownian motion except at 0 when it gets an additive positive "kick" from the local time.

From the intuitive description of $X$ in the Section 7.1 , we similarly expect $H_{\xi}$ to remain constant over time intervals when $X_{t}$ is not in the ray $R_{\xi}:=$ $\{x \in T: x \leqslant \xi\}$. During time intervals when $X_{t}$ is in $R_{\xi}$ we expect $H_{\xi}$ to evolve as a standard Brownian motion except at branch points of $T$ where it receives negative "kicks" from a local time additive functional. Here the magnitude of the kicks will be related to how much $\mu$-mass is being lost to the rays that are branching off from $R_{\xi}$.

To make this description precise, we first need to introduce the appropriate local time processes and then use Fukushima's stochastic calculus for Dirichlet processes (in much the same way that Tanaka's formula follows from the standard Itô's formula for Brownian motion). Unfortunately, this involves appealing to quite a large body of material from [72], but it would have required lengthening this section considerably to state in detail the results that we use.

We showed in Section 7.4 that $\mathbb{P}^{x}\left\{\sigma_{y}<\infty\right\}$ for any $x, y \in T$. By Theorems 4.2.1 and 2.2 .3 of [72], the point mass $\delta_{y}$ at any $y \in T$ belongs to the set of measures $S_{00}$. (See (2.2.10) of [72] for a definition of $S_{00}$. Another way of seeing that $\delta_{y}$ is in $S_{00}$ is just to observe that $\sup _{x} g_{\alpha}(x, y)<\infty$ for all $\alpha>0$.)

By Theorem 5.1.6 of [72] there exists for each $y \in T$ a strict sense positive continuous additive functional $L^{y}$ with Revuz measure $\delta_{y}$. As usual, we call $L^{y}$ the local time at $y$.

Definition 7.22. Given $\xi \in E_{+}$, write $m_{\xi}$ for the Radon measure on $T$ that is supported on the ray $R_{\xi}$ and for each $a \in \mathbb{R}$ assigns mass $\mu\left\{\zeta \in E_{+}: \zeta|a=\xi| a\right\}$ to the set $\{\xi \mid b: b \geqslant a\}=\left\{x \in R_{\xi}: h(x) \geqslant a\right\}$.

Remark 7.23. Note that $m_{\xi}$ is a discrete measure that is concentrated on the countable set of points of the form $\xi \wedge \zeta$ for some $\zeta \in E_{+} \backslash\{\xi\}$ (that is, on the points where other rays branch from $R_{\xi}$ ).

Theorem 7.24. For each $\xi \in E_{+}$and $x \in T$ the process $H_{\xi}$ has a semimartingale decomposition

$$
H_{\xi}(t)=H_{\xi}(0)+M_{\xi}(t)-\frac{1}{2} \int_{R_{\xi}} L^{y}(t) m_{\xi}(d y), t \geqslant 0
$$

under $\mathbb{P}^{x}$, where $M_{\xi}$ is a continuous, square-integrable martingale with quadratic variation

$$
\left\langle M_{\xi}\right\rangle(t)=\int_{0}^{t} \mathbf{1}\{X(s) \leqslant \xi\} d s, t \geqslant 0
$$

Moreover, the martingales $M_{\xi}$ and $M_{\xi^{\prime}}$ for $\xi, \xi^{\prime} \in E_{+}$have covariation

$$
\left\langle M_{\xi}, M_{\xi^{\prime}}\right\rangle_{t}=\int_{0}^{t} 1\left\{X(s) \leqslant \xi \wedge \xi^{\prime}\right\} d s, t \geqslant 0
$$

Proof. For $\xi \in E_{+}, x \in T$, and $A \in \mathbb{N}$, set $h_{\xi}(x)=h(x \wedge \xi)$ and $h_{\xi}^{A}(x)=$ $(-A) \vee(h(x \wedge \xi) \wedge A)$.

It is clear that $h_{\xi}^{A}$ is in the domain $\mathcal{D}$ of the Dirichlet form $\mathcal{E}$, with $\nabla h_{\xi}^{A}(x)=1\{\xi|(-A) \leqslant x \leqslant \xi| A\}$. Given $f \in \mathcal{D}$, it follows from the product rule that

$$
2 \mathcal{E}\left(h_{\xi}^{A} f, h_{\xi}^{A} f\right)-\mathcal{E}\left(\left(h_{\xi}^{A}\right)^{2}, f\right)=\int_{T} f(x) \mathbf{1}\{\xi|(-A) \leqslant x \leqslant \xi| A\} \nu(d x)
$$

In the terminology of Section 3.2 of [72], the energy measure corresponding to $h_{\xi}^{A}$ is $\nu_{\xi}^{A}(d x):=\mathbf{1}\{\xi|(-A) \leqslant x \leqslant \xi| A\} \nu(d x)$. A similar calculation shows that the joint energy measure corresponding to a pair of functions $h_{\xi}^{A}$ and $h_{\xi^{\prime}}^{A^{\prime}}$ is $1\left[\{\xi|(-A) \leqslant x \leqslant \xi| A\} \cap\left\{\xi^{\prime}\left|\left(-A^{\prime}\right) \leqslant x \leqslant \xi^{\prime}\right| A^{\prime}\right\}\right] \nu(d x)=\left(\nu_{\xi}^{A} \wedge \nu_{\xi^{\prime}}^{A^{\prime}}\right)(d x)$ in the usual lattice structure on measures.

An integration by parts establishes that for any $f \in \mathcal{D}$ we have

$$
\mathcal{E}\left(h_{\xi}^{A}, f\right)=\frac{1}{2} \int_{T} f(x) \tilde{m}_{\xi}^{A}(d x)
$$

where

$$
\tilde{m}_{\xi}^{A}:=m_{\xi}^{A}-\mu\{\zeta: \zeta|(-A)=\xi|(-A)\} \delta_{\xi \mid(-A)}+\mu\{\zeta: \zeta|A=\xi| A\} \delta_{\xi \mid A}
$$

with

$$
m_{\xi}^{A}(d x):=1\{\xi|(-A) \leqslant x \leqslant \xi| A\} m_{\xi}(d x)
$$

Now $\nu_{\xi}^{A}$ is the Revuz measure of the strict sense positive continuous additive functional $\int_{0}^{t} \mathbf{1}\{\xi|(-A) \leqslant X(s) \leqslant \xi| A\} d s$ and $\nu_{\xi}^{A} \wedge \nu_{\xi^{\prime}}^{A^{\prime}}$ is the Revuz measure of the strict sense positive continuous additive functional $\int_{0}^{t} \mathbf{1}\left[\{\xi|(-A) \leqslant X(s) \leqslant \xi| A\} \cap\left\{\xi^{\prime}\left|\left(-A^{\prime}\right) \leqslant X(s) \leqslant \xi^{\prime}\right| A^{\prime}\right\}\right] d s$. A straightforward calculation shows that $\sup _{x} \int g_{\alpha}(x, y) m_{\xi}^{A}(d y)<\infty$, and so $m_{\xi}^{A} \in S_{00}$ is the Revuz measure of the strict sense positive continuous additive functional $\int_{R_{\xi}} L^{y}(t) m_{\xi}^{A}(d y)$ (because the integral is just a sum, we do not need to address the measurability of $\left.y \mapsto L^{y}(t)\right)$.

Put $H_{\xi}^{A}(t):=h_{\xi}^{A}(X(t)), t \geqslant 0$. Theorem 5.2.5 of [72] applies to give that

$$
H_{\xi}^{A}(t)=H_{\xi}^{A}(0)+M_{\xi}^{A}(t)-\frac{1}{2} \int_{R_{\xi}} L^{y}(t) \tilde{m}_{\xi}^{A}(d y), t \geqslant 0
$$

under $\mathbb{P}^{x}$ for each $x \in T$, where $M_{\xi}^{A}$ is a continuous, square-integrable martingale with quadratic variation

$$
\left\langle M_{\xi}^{A}\right\rangle(t)=\int_{0}^{t} \mathbf{1}\{\xi|(-A) \leqslant X(s) \leqslant \xi| A\} d s
$$

Moreover, the martingales $M_{\xi}^{A}$ and $M_{\xi^{\prime}}^{A^{\prime}}$ for $\xi, \xi^{\prime} \in E_{+}$have covariation

$$
\begin{aligned}
& \left\langle M_{\xi}^{A}, M_{\xi^{\prime}}^{A^{\prime}}\right\rangle(t) \\
& \quad=\int_{0}^{t} \mathbf{1}\left[\{\xi|(-A) \leqslant X(s) \leqslant \xi| A\} \cap\left\{\xi^{\prime}\left|\left(-A^{\prime}\right) \leqslant X(s) \leqslant \xi^{\prime}\right| A^{\prime}\right\}\right] d s
\end{aligned}
$$

In particular,

$$
\begin{align*}
& \left\langle M_{\xi}^{B}-M_{\xi}^{A}\right\rangle(t) \\
& \quad=\int_{0}^{t} \mathbf{1}[\{\xi|(-B) \leqslant X(s) \leqslant \xi| B\} \backslash\{\xi|(-A) \leqslant X(s) \leqslant \xi| A\}] d s \tag{7.16}
\end{align*}
$$

for $A<B$.
For each $t \geqslant 0$ we have that $H_{\xi}^{A}(s)=H_{\xi}(s)$ and $\int_{R_{\xi}} L^{y}(s) \tilde{m}_{\xi}^{A}(d y)=$ $\int_{R_{\xi}} L^{y}(s) m_{\xi}(d y)$ for all $0 \leqslant s \leqslant t$ when $A>\sup \left\{\left|H_{\xi}(s)\right|: 0 \leqslant s \leqslant t\right\}$, $\mathbb{P}^{x}$-a.s. Therefore, there exists a continuous process $M_{\xi}$ such that $M_{\xi}^{A}(s)=M_{\xi}(s)$ for all $0 \leqslant s \leqslant t$ when $A>\sup \left\{\left|H_{\xi}(s)\right|: 0 \leqslant s \leqslant t\right\}, \mathbb{P}^{x}$-a.s. It follows from (7.16) that $\lim _{A \rightarrow \infty} \mathbb{P}^{x}\left[\sup _{0 \leqslant s \leqslant t}\left|M_{\xi}^{A}(s)-M_{\xi}(s)\right|^{2}\right]=0$. By standard arguments, the processes $M_{\xi}$ are continuous, square-integrable martingales with the stated quadratic variation and covariation properties.

Remark 7.25. There is more that can be said about the process $H_{\xi}$. For instance, given $x \in T$ and $\xi \in E_{+}$with $x \in R_{\xi}$ and $a>h(x)$, we can explicitly calculate the Laplace transform of $\inf \left\{t>0: H_{\xi}(t)=a\right\}=\sigma_{\xi \mid a}$ under $\mathbb{P}^{x}$. We have

$$
\mathbb{P}^{x}\left[\exp \left(-\alpha \sigma_{\xi \mid a}\right)\right]=g_{\alpha}(x, \xi \mid a) / g_{\alpha}(\xi|a, \xi| a)
$$

where $g_{\alpha}$ is given explicitly by (7.15). When $X$ is transient, the distribution of $\sigma_{\xi \mid a}$ has an atom at $\infty$ and we have

$$
\mathbb{P}^{x}\left\{\sup _{0 \leqslant t<\infty} H_{\xi}(t) \geqslant a\right\}=\mathbb{P}^{x}\left\{\sigma_{\xi \mid a}<\infty\right\}=g(x, \xi \mid a) / g(\xi|a, \xi| a)
$$

By the strong Markov property, the càdlàg process $\left(\sigma_{\xi \mid a}\right)_{a \geqslant h(x)}$ has independent (although, of course, non-stationary) increments under $\mathbb{P}^{x}$, with the usual appropriate definition of this notion for non-decreasing $\mathbb{R} \cup\{+\infty\}$-valued processes.

## $\mathbb{R}$-trees from coalescing particle systems

### 8.1 Kingman's coalescent

Here is a quick description of Kingman's coalescent (which we will hereafter simply refer to as the coalescent). Let $\mathcal{P}$ denote the collection of partitions of $\mathbb{N}$. For $n \in \mathbb{N}$ let $\mathcal{P}_{n}$ denote the collection of partitions of $\mathbb{N}_{\leqslant n}:=\{1,2, \ldots, n\}$. Write $\rho_{n}$ for the natural restriction map from $\mathcal{P}$ onto $\mathcal{P}_{n}$. Kingman [90] showed that there was a (unique in law) $\mathcal{P}$-valued Markov process $\Pi$ such that for all $n \in \mathbb{N}$ the restricted process $\Pi_{n}:=\rho_{n} \circ \Pi$ is a $\mathcal{P}_{n}$-valued, time-homogeneous Markov chain with initial state $\Pi_{n}(0)$ the trivial partition $\{\{1\}, \ldots,\{n\}\}$ and the following transition rates: if $\Pi_{n}$ is in a state with $k$ blocks, then

- a jump occurs at rate $\binom{k}{2}$,
- the new state is one of the $\binom{k}{2}$ partitions that can be obtained by merging two blocks of the current state,
- and all such possibilities are equally likely.

Let $N(t)$ denote the number of blocks of the partition $\Pi(t)$. It was shown in [90] that almost surely, $N(t)<\infty$ for all $t>0$ and the process $N$ is a pure-death Markov chain that jumps from $k$ to $k-1$ at rate $\binom{k}{2}$ for $k>1$ (the state 1 is a trap). Therefore, the construction in Example 3.41 applies to construct a compact $\mathbb{R}$-tree from $\Pi$. Let ( $\mathbb{S}, \delta$ ) denote the corresponding (random) ultrametric space that arises from looking at the closure of the leaves (that is, $\mathbb{N}$ ) in that tree, as in Example 3.41. We note that some properties of the space ( $\mathbb{N}, \delta$ ) were considered explicitly in Section 4 of [10]. We will apply Proposition B. 3 to show that the Hausdorff and packing dimensions of $\mathbb{S}$ are both 1 and that, in the terminology of [112] - see, also, [27, 113, 114] - the space $\mathbb{S}$ is a.s. capacity-equivalent to the unit interval $[0,1]$.

Theorem 8.1. Almost surely, the Hausdorff and packing dimensions of the random compact metric space $\mathbb{S}$ are both 1 . There exist random variables $C^{*}, C^{* *}$ such that almost surely $0<C^{*} \leqslant C^{* *}<\infty$ and for every gauge $f$

$$
C^{*} \operatorname{Cap}_{f}([0,1]) \leqslant \operatorname{Cap}_{f}(\mathbb{S}) \leqslant C^{* *} \operatorname{Cap}_{f}([0,1])
$$

Proof. We will apply Proposition B.3.
Note that $\sigma_{n}:=\inf \{t>0: N(t)=n\}$ is of the form $\tau_{n+1}+\tau_{n+2}+\ldots$, where the $\tau_{k}$ are independent and $\tau_{k}$ is exponential with rate $\binom{k}{2}$. Thus

$$
\begin{equation*}
\mathbb{P}\left[\sigma_{n}\right]=\frac{2}{(n+1) n}+\frac{2}{(n+2)(n+1)}+\cdots=\frac{2}{n} \tag{8.1}
\end{equation*}
$$

It is easy to check that

$$
\lim _{t \downarrow 0} t N(t)=\lim _{n \rightarrow \infty} \sigma_{n} N\left(\sigma_{n}\right)=\lim _{n \rightarrow \infty} \sigma_{n} n=2, \text { a.s. }
$$

- see, for example, the arguments that lead to Equation (35) in [18].

It was shown in [90] that almost surely for all $t>0$ the asymptotic block frequencies

$$
F_{i}(t):=\lim _{n \rightarrow \infty} n^{-1}\left|\left\{j \in \mathbb{N}_{\leqslant n}: j \sim_{\Pi(t)} I_{i}(t)\right\}\right|, 1 \leqslant i \leqslant N(t)
$$

exist and

$$
F_{1}(t)+\cdots+F_{N(t)}(t)=1
$$

We claim that

$$
\begin{equation*}
\lim _{t \downarrow 0} t^{-1} \sum_{i=1}^{N(t)} F_{i}(t)^{2}=1, \text { a.s. } \tag{8.2}
\end{equation*}
$$

To see this, set $X_{n, i}:=F_{i}\left(\sigma_{n}\right)$ for $n \in \mathbb{N}$ and $1 \leqslant i \leqslant n$, and observe from (8.1) that it suffices to establish

$$
\begin{equation*}
\lim _{n \rightarrow \infty} n \sum_{i=1}^{n} X_{n, i}^{2}=2, \text { a.s. } \tag{8.3}
\end{equation*}
$$

By the "paintbox" construction in Section 5 of [90] the random variable $\sum_{i=1}^{n} X_{n, i}^{2}$ has the same law as $U_{(1)}^{2}+\left(U_{(2)}-U_{(1)}\right)^{2}+\cdots+\left(U_{(n-1)}-U_{(n-2)}\right)^{2}+$ $\left(1-U_{(n-1)}\right)^{2}$, where $U_{(1)} \leqslant \ldots \leqslant U_{(n-1)}$ are the order statistics corresponding to i.i.d. random variables $U_{1}, \ldots, U_{n-1}$ that are uniformly distributed on $[0,1]$ - see Figure 8.1 and Section 4.2 of [18] for an exposition from which essentially this figure was taken with permission. By a classical result on the spacings between order statistics of i.i.d. uniform random variables - see, for example, Section III.3.(e) of [66] - the law of $\sum_{i=1}^{n} X_{n, i}^{2}$ is the same as that of $\left(\sum_{i=1}^{n} T_{i}^{2}\right) /\left(\sum_{i=1}^{n} T_{i}\right)^{2}$, where $T_{1}, \ldots, T_{n}$ are i.i.d. mean one exponential random variables.

Now for any $0<\varepsilon<1$ we have, recalling $\mathbb{P}\left[T_{i}^{2}\right]=2$,

$$
\begin{aligned}
& \mathbb{P}\left\{\left(\sum_{i=1}^{n} T_{i}^{2}\right) /\left(\sum_{i=1}^{n} T_{i}\right)^{2}>(1+\varepsilon)(1-\varepsilon)^{-2} 2 n^{-1}\right\} \\
& \quad \leqslant \mathbb{P}\left\{\sum_{i=1}^{n}\left(T_{i}^{2}-\mathbb{P}\left[T_{i}^{2}\right]\right)>2 \varepsilon n\right\}+\mathbb{P}\left\{\sum_{i=1}^{n}\left(T_{i}-\mathbb{P}\left[T_{i}\right]\right)<-\varepsilon n\right\}
\end{aligned}
$$



Fig. 8.1. Kingman's description of the block frequencies in the coalescent. Let $V_{1}, V_{2}, \ldots$ be independent random variables uniformly distributed on [0, 1]. For $\sigma_{n} \leqslant t<\sigma_{n-1}$ put $Y_{1}(t)=V_{(1)}, Y_{2}(t)=V_{(2)}-V_{(1)}, \ldots, Y_{n}(t)=1-V_{(n-1)}$, where $V_{(1)}, \ldots, V_{(n-1)}$ are the order statistics of $V_{1}, \ldots V_{n-1}$. Then, as set valued processes, the block proportions $\left(\left\{F_{1}(t), \ldots F_{N(t)}(t)\right\}\right)_{t \geqslant 0}$ and the spacings $\left(\left\{Y_{1}(t), \ldots Y_{N(t)}(t)\right\}\right)_{t \geqslant 0}$ have the same distribution.

A fourth moment computation and Markov's inequality show that both terms on the right-hand side are bounded above by $c(\varepsilon) n^{-2}$ for a suitable constant $c(\epsilon)$. A similar bound holds for

$$
\mathbb{P}\left\{\left(\sum_{i=1}^{n} T_{i}^{2}\right) /\left(\sum_{i=1}^{n} T_{i}\right)^{2}<(1-\varepsilon)(1+\varepsilon)^{-2} 2 n^{-1}\right\}
$$

The claim (8.3) and, hence, (8.2) now follows by an application of the BorelCantelli Lemma.

The proof is finished by an appeal to Proposition B. 3 and the observation there exist constants $0<c^{\#} \leqslant c^{\# \#}<\infty$ such that

$$
c^{\#}\left(\int_{0}^{1} f(t) d t\right)^{-1} \leqslant \operatorname{Cap}_{f}([0,1]) \leqslant c^{\# \#}\left(\int_{0}^{1} f(t) d t\right)^{-1}
$$

(this is described as "classical" in [113] and follows by arguments similar to those used in Section 3 of that paper to prove a higher dimensional analogue of this fact).

### 8.2 Coalescing Brownian motions

Let $\mathbb{T}$ denote the circle of circumference $2 \pi$. It is possible to construct a stochastic process $\mathbf{Z}=\left(Z_{1}(t), Z_{2}(t), \ldots\right)$ such that:

- each coordinate process $Z_{i}$ evolves as a Brownian motion on $\mathbb{T}$ with uniformly distributed starting point,
- until they collide, different coordinate processes evolve independently,
- after they collide, two coordinate processes follow the same evolution
- see, for example, [44]. We can then define a coalescing partition valued process $\Pi$ be declaring that $i \sim_{\Pi(t)} j$ if $Z_{i}(t)=Z_{j}(t)$ (that is, $i$ and $j$ are in the same block of $\Pi(t)$ if the particles $i$ and $j$ have coalesced by times $t$ ). Let $N(t)$ denote the number of blocks of $\Pi(t)$. We will show below that almost surely $N(t)<\infty$ for all $t>0$, and the procedure in Example 3.41 gives a $\mathbb{R}$-tree with leaves corresponding to $\mathbb{N}$ and a compactification of $\mathbb{N}$ that we will denote by $(\mathbb{S}, \delta)$.

Our main result is the following.
Theorem 8.2. Amost surely, the random compact metric space ( $\mathbb{S}, \delta$ ) has Hausdorff and packing dimensions both equal to $\frac{1}{2}$. There exist random variables $K^{*}, K^{* *}$ such that $0<K^{*} \leqslant K^{* *}<\infty$ and for every gauge $f$

$$
K^{*} \operatorname{Cap}_{f}\left(C_{\frac{1}{2}}\right) \leqslant \operatorname{Cap}_{f}(\mathbb{S}) \leqslant K^{* *} \operatorname{Cap}_{f}\left(C_{\frac{1}{2}}\right)
$$

where $C_{\frac{1}{2}}$ is the middle- $\frac{1}{2}$ Cantor set.
Remark 8.3. One of the assertions of the following result is that $\mathbb{S}$ is a.s. capacity-equivalent to $C_{\frac{1}{2}}$. Hence, by the results of [113], $\mathbb{S}$ is also a.s. capacity-equivalent to the zero set of (one-dimensional) Brownian motion.

Before proving Theorem 8.2, we will need to do some preliminary computations to enable us to check the conditions of Proposition B.3.

Given a finite non-empty set $A \subseteq \mathbb{T}$, let $W^{A}$ be a process taking values in the space of finite subsets of $\mathbb{T}$ that describes the evolution of a finite set of indistinguishable Brownian particles with the features that $W^{A}(0)=A$ and that particles evolve independently between collisions but when two particles collide they coalesce into a single particle.

Write $\mathcal{O}$ for the collection of open subsets of $\mathbb{T}$ that are either empty or consist of a finite union of open intervals with distinct end-points. Given $B \in \mathcal{O}$, define on some probability space $(\Sigma, \mathcal{G}, \mathbb{Q})$ an $\mathcal{O}$-valued process $V^{B}$, the annihilating circular Brownian motion as follows. The end-points of the
constituent intervals execute independent Brownian motions on $\mathbb{T}$ until they collide, at which point they annihilate each other. If the two colliding endpoints are from different intervals, then those two intervals merge into one interval. If the two colliding end-points are from the same interval, then that interval vanishes (unless the interval was arbitrarily close to $\mathbb{T}$ just before the collision, in which case the process takes the value $\mathbb{T}$ ). The process is stopped when it hits the empty set or $\mathbb{T}$.

We have the following duality relation between $W^{A}$ and $V^{B}$. An analogous result for the coalescing Brownian flow on $\mathbb{R}$ is on p18 of [22].

Proposition 8.4. For all finite, non-empty subsets $A \subseteq \mathbb{T}$, all sets $B \in \mathcal{O}$, and all $t \geqslant 0$,

$$
\mathbb{P}\left\{W^{A}(t) \subseteq B\right\}=\mathbb{Q}\left\{A \subseteq V^{B}(t)\right\}
$$

Proof. For $N \in \mathbb{N}$, let $\mathbb{Z}_{N}:=\{0,1, \ldots N-1\}$ denote the integers modulo $N$. Let $\mathbb{Z}_{N}^{\frac{1}{2}}:=\left\{\frac{1}{2}, \frac{3}{2}, \ldots, \frac{2 N-1}{2}\right\}$ denote the half-integers modulo $N$. A nonempty subset $D$ of $\mathbb{Z}_{N}$ can be (uniquely) decomposed into "intervals": an interval of $D$ is an equivalence class for the equivalence relation on the points of $D$ defined by $x \sim y$ if and only if $x=y,\{x, x+1, \ldots, y-1, y\} \subseteq D$, or $\{y, y+1, \ldots, x-1, x\} \subseteq D$ (with all arithmetic modulo $N$ ). Any interval other than $\mathbb{Z}_{N}$ itself has an associated pair of (distinct) "end-points" in $\mathbb{Z}_{N}^{\frac{1}{2}}$ : if the interval is $\{a, a+1, \ldots, b-1, b\}$, then the corresponding end-points are $a-\frac{1}{2}$ and $b+\frac{1}{2}$ (with all arithmetic modulo $N$ ). Note that the end-points of different intervals of $D$ are distinct.

For $C \subseteq \mathbb{Z}_{N}$, let $W_{N}^{C}$ be a process on some probability space $\left(\Omega^{\prime}, \mathcal{F}^{\prime}, \mathbb{P}^{\prime}\right)$ taking values in the collection of non-empty subsets of $\mathbb{Z}_{N}$ that is defined in the same manner as $W^{A}$, with Brownian motion on $\mathbb{T}$ replaced by simple, symmetric (continuous time) random walk on $\mathbb{Z}_{N}$ (that is, by the continuous time Markov chain on $\mathbb{Z}_{N}$ that only makes jumps from $x$ to $x+1$ or $x$ to $x-1$ at a common rate $\lambda>0$ for all $\left.x \in \mathbb{Z}_{N}\right)$. For $D \subseteq \mathbb{Z}_{N}$, let $V_{N}^{D}$ be a process taking values in the collection of subsets of $\mathbb{Z}_{N}$ that is defined on some probability space $\left(\Sigma^{\prime}, \mathcal{G}^{\prime}, \mathbb{Q}^{\prime}\right)$ in the same manner as $V^{B}$, with Brownian motion on $\mathbb{T}$ replaced by simple, symmetric (continuous time) random walk on $\mathbb{Z}_{N}^{\frac{1}{2}}$ (with the same jump rate $\lambda$ as in the definition of $W_{N}^{C}$ ). That is, end-points of intervals evolve as annihilating random walks on $\mathbb{Z}_{N}^{\frac{1}{2}}$.

The proposition will follow by a straightforward weak limit argument if we can show the following duality relationship between the coalescing "circular" random walk $W_{N}^{C}$ and the annihilating "circular" random walk $V_{N}^{D}$ :

$$
\begin{equation*}
\mathbb{P}^{\prime}\left\{W_{N}^{C}(t) \subseteq D\right\}=\mathbb{Q}^{\prime}\left\{C \subseteq V_{N}^{D}(t)\right\} \tag{8.4}
\end{equation*}
$$

for all non-empty subsets of $C \subseteq \mathbb{Z}_{N}$, all subsets of $D \subseteq \mathbb{Z}_{N}$, and all $t \geqslant 0$.
It is simple, but somewhat tedious, to establish (8.4) by a generator calculation using the usual generator criterion for duality - see, for example, Corollary 4.4.13 of [56]. However, as Tom Liggett pointed out to us, there
is an easier route. A little thought shows that $V_{N}^{D}$ is nothing other than the (simple, symmetric) voter model on $\mathbb{Z}_{N}$. The analogous relationship between the annihilating random walk and the voter model on $\mathbb{Z}$ due to [124] is usually called the border equation - see Section 2 of [32] for a discussion and further references. The relationship (8.4) is then just the analogue of the usual duality between the voter model and coalescing random walk on $\mathbb{Z}$ and it can be established in a similar manner by Harris's graphical method (again see Section 2 of [32] for a discussion and references and Figure 8.2 for an illustration).


Fig. 8.2. The graphical construction of the (symmetric, nearest neighbor) voter model on $\mathbb{Z}_{16}$. Time proceeds up the page. The initial configuration is at the bottom of the diagram. Horizontal arrows issue from each site at rate $\lambda$, and are equally likely to point left or right. The state of the site at the head of an arrow is changed to the current state of the site at the tail. Arrows wrap around modulo 16. Going forwards in time, the boundaries between blocks of 0 s and blocks of 1 s execute a family of continuous time annihilating simple random walks. By reversing the direction of the vertical and horizontal arrows, it is possible to trace back from some location in space and time to the ultimate origin at time 0 of the state at that location. The resulting history is a continuous time simple random walk. Any two such histories evolve independently until they collide, after which they coalesce.

Define a set-valued processes $W^{[n]}, n \in \mathbb{N}$, and $W$ by

$$
W^{[n]}(t):=\left\{Z_{1}(t), Z_{2}(t), \ldots, Z_{n}(t)\right\} \subseteq \mathbb{T}, t \geqslant 0
$$

and

$$
W(t):=\left\{Z_{1}(t), Z_{2}(t), \ldots\right\} \subseteq \mathbb{T}, t \geqslant 0
$$

Thus, $W^{[1]}(t) \subseteq W^{[2]}(t) \subseteq \ldots, \bigcup_{n \in \mathbb{N}} W^{[n]}(t)=W(t)$, and the cardinality of $W(t)$ is $N(t)$, the number of blocks in the partition $\Pi(t)$.

Corollary 8.5. For $t>0$,

$$
\mathbb{P}[N(t)]=1+2 \sum_{n \in \mathbb{N}} \exp \left(-\left(\frac{n}{2}\right)^{2} t\right)<\infty
$$

and

$$
\lim _{t \downarrow 0} t^{\frac{1}{2}} \mathbb{P}[N(t)]=2 \sqrt{\pi}
$$

Proof. Note that if $B$ is a single open interval (so that for all $t \geqslant 0$ the set $V^{B}(t)$ is either an interval or empty) and we let $L(t)$ denote the length of $V^{B}(t)$, then $L$ is a Brownian motion on [0, 2 $\pi$ ] with infinitesimal variance 2 that is stopped at the first time it hits $\{0,2 \pi\}$.

Now, for $M \in \mathbb{N}$ and $0 \leqslant i \leqslant M-1$ we have from the translation invariance of $Z$ and Proposition 8.4 that

$$
\begin{aligned}
& \mathbb{P}\left\{W^{[n]}(t) \cap[2 \pi i / M, 2 \pi(i+1) / M] \neq \varnothing\right\} \\
&=1-\mathbb{P}\left\{W^{[n]}(t) \subseteq\right] 0,2 \pi(M-1) / M[ \} \\
&=1-\mathbb{P}\left\{W^{[n]}(0) \subseteq V^{00,2 \pi(M-1) / M[ }(t)\right\}
\end{aligned}
$$

where we take the annihilating process $V^{] 0,2 \pi(M-1) / M[ }$ to be defined on the same probability space $(\Omega, \mathcal{F}, \mathbb{P})$ as the process $\mathbf{Z}$ that was used to construct $W^{[n]}$ and $W$, and we further take the processes $V^{0,2 \pi(M-1) / M[ }$ and $\mathbf{Z}$ to be independent. Thus,

$$
\begin{aligned}
& \mathbb{P}\{ W(t) \cap[2 \pi i / M, 2 \pi(i+1) / M] \neq \varnothing\} \\
& \quad=1-\mathbb{P}\left\{V^{00,2 \pi(M-1) / M[ }(t)=\mathbb{T}\right\} \\
& \quad=1-\tilde{\mathbb{P}}\{\tilde{\tau} \leqslant 2 t, \tilde{B}(\tilde{\tau})=2 \pi \mid \tilde{B}(0)=2 \pi(M-1) / M\},
\end{aligned}
$$

where $\tilde{B}$ is a standard one-dimensional Brownian motion on some probability space $(\tilde{\Omega}, \tilde{\mathcal{F}}, \tilde{\mathbb{P}})$ and $\tilde{\tau}=\inf \{s \geqslant 0: \tilde{B}(s) \in\{0,2 \pi\}\}$.

By Theorem 4.1.1 of [91] we have

$$
\begin{aligned}
\mathbb{P} & {[|W(t)|] } \\
& =\lim _{M \rightarrow \infty} \mathbb{P}\left[\sum_{i=0}^{M-1} \mathbf{1}\{W(t) \cap[2 \pi i / M, 2 \pi(i+1) / M] \neq \varnothing\}\right] \\
& =\lim _{M \rightarrow \infty} M(1-\tilde{\mathbb{P}}\{\tilde{\tau} \leqslant 2 t, \tilde{B}(\tilde{\tau})=2 \pi \mid \tilde{B}(0)=2 \pi(M-1) / M\}) \\
& =1-\lim _{M \rightarrow \infty} M \frac{2}{\pi} \sum_{n \in \mathbb{N}} \frac{(-1)^{n}}{n} \sin \left(n \pi\left(\frac{M-1}{M}\right)\right) \exp \left(-\left(\frac{n}{2}\right)^{2} t\right) \\
& =1+2 \sum_{n \in \mathbb{N}} \exp \left(-\left(\frac{n}{2}\right)^{2} t\right) \\
& =\theta\left(\frac{t}{4 \pi}\right)<\infty,
\end{aligned}
$$

where

$$
\begin{equation*}
\theta(u):=\sum_{n=-\infty}^{\infty} \exp \left(-\pi n^{2} u\right) \tag{8.5}
\end{equation*}
$$

is the Jacobi theta function (we refer the reader to [31] for a survey of many of the other probabilistic interpretations of the theta function). The proof is completed by recalling that $\theta$ satisfies the functional equation $\theta(u)=u^{-\frac{1}{2}} \theta\left(u^{-1}\right)$ and noting that $\lim _{u \rightarrow \infty} \theta(u)=1$.

For $t>0$ the random partition $\Pi(t)$ is exchangeable with a finite number of blocks. Let $1=I_{1}(t)<I_{2}(t)<\ldots<I_{N(t)}(t)$ be the list in increasing order of the minimal elements of the blocks of $\Pi(t)$. Results of Kingman see Section 11 of [11] for a unified account - and the fact that $\Pi$ evolves by pairwise coalescence of blocks give that $\mathbb{P}$-a.s. for all $t>0$ the asymptotic frequencies

$$
F_{i}(t)=\lim _{n \rightarrow \infty} n^{-1}\left|\left\{j \in \mathbb{N}_{\leqslant n}: j \sim_{\Pi(t)} I_{i}(t)\right\}\right|
$$

exist for $1 \leqslant i \leqslant N(t)$ and $F_{1}(t)+\cdots+F_{N(t)}(t)=1$.
Lemma 8.6. Almost surely,

$$
\lim _{t \downarrow 0} t^{-\frac{1}{2}} \sum_{i=1}^{N(t)} F_{i}(t)^{2}=\frac{2}{\pi^{3 / 2}}
$$

Proof. Put $T_{i j}:=\inf \left\{t \geqslant 0: Z_{i}(t)=Z_{j}(t)\right\}$ for $i \neq j$. Observe that

$$
\begin{aligned}
\mathbb{P}\left[\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right] & =\mathbb{P}\left[\lim _{n \rightarrow \infty} \frac{1}{n^{2}} \sum_{i=1}^{n} \sum_{k=1}^{n} \mathbf{1}\left\{j \sim_{\Pi(t)} k\right\}\right] \\
& =\mathbb{P}\left\{1 \sim_{\Pi(t)} 2\right\} \\
& =\mathbb{P}\left\{T_{12} \leqslant t\right\} .
\end{aligned}
$$

From Theorem 4.1.1 of [91] we have

$$
\begin{aligned}
\mathbb{P} & \left\{T_{12} \leqslant t\right\} \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} 1-\frac{4}{\pi} \sum_{n \in \mathbb{N}} \sin \left(\frac{(2 n-1) x}{2}\right) \frac{1}{2 n-1} \exp \left(-\left(\frac{2 n-1}{2}\right)^{2} t\right) d x \\
& =\frac{8}{\pi^{2}} \sum_{n \in \mathbb{N}} \frac{1}{(2 n-1)^{2}}\left\{1-\exp \left(-\left(\frac{2 n-1}{2}\right)^{2} t\right)\right\} \\
& =\frac{2}{\pi^{2}} \int_{0}^{t} \sum_{n \in \mathbb{N}} \exp \left(-\left(\frac{2 n-1}{2}\right)^{2} s\right) d s \\
& =\frac{2}{\pi^{2}} \int_{0}^{t} \frac{1}{2}\left\{\sum_{n=-\infty}^{\infty} \exp \left(-n^{2} \frac{s}{4}\right)-\sum_{n=-\infty}^{\infty} \exp \left(-n^{2} s\right)\right\} d s \\
& =\frac{1}{\pi^{2}} \int_{0}^{t}\left\{\theta\left(\frac{s}{4 \pi}\right)-\theta\left(\frac{s}{\pi}\right)\right\} d s,
\end{aligned}
$$

where $\theta$ is again the Jacobi theta function defined in (8.5). By the properties of $\theta$ recalled after (8.5),

$$
\begin{equation*}
\lim _{t \downarrow 0} t^{-\frac{1}{2}} \mathbb{P}\left[\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right]=\lim _{t \downarrow 0} t^{-\frac{1}{2}} \mathbb{P}\left\{T_{12} \leqslant t\right\}=\frac{2}{\pi^{3 / 2}} \tag{8.6}
\end{equation*}
$$

Now

$$
\begin{aligned}
& \mathbb{P}\left[\left(\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right)^{2}\right] \\
& \quad=\mathbb{P}\left[\lim _{n \rightarrow \infty} \frac{1}{n^{4}} \sum_{i_{1}=1}^{n} \sum_{i_{2}=1}^{n} \sum_{i_{3}=1}^{n} \sum_{i_{4}=1}^{n} \mathbf{1}\left\{i_{1} \sim_{\Pi(t)} i_{2}, i_{3} \sim_{\Pi(t)} i_{4}\right\}\right] \\
& \quad=\mathbb{P}\left\{1 \sim_{\Pi(t)} 2,3 \sim_{\Pi(t)} 4\right\},
\end{aligned}
$$

and so

$$
\begin{align*}
\operatorname{Var}\left(\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right) & =\mathbb{P}\left\{1 \sim_{\Pi(t)} 2,3 \sim_{\Pi(t)} 4\right\}-\mathbb{P}\left\{T_{12} \leqslant t\right\}^{2}  \tag{8.7}\\
& =\mathbb{P}\left\{1 \sim_{\Pi(t)} 2,3 \sim_{\Pi(t)} 4\right\}-\mathbb{P}\left\{T_{12} \leqslant t, T_{23} \leqslant t\right\}
\end{align*}
$$

Observe that

$$
\begin{aligned}
& \mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t, T_{13}>t, T_{14}>t, T_{23}>t, T_{24}>t\right\} \\
& \quad \leqslant \mathbb{P}\left\{1 \sim_{\Pi(t)} 2,3 \sim_{\Pi(t)} 4,\left|W^{[4]}(t)\right| \neq 1\right\} \\
& \quad \leqslant \mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t\right\}
\end{aligned}
$$

and

$$
\begin{aligned}
& \mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t\right\}-\mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t, T_{13}>t, T_{14}>t, T_{23}>t, T_{24}>t\right\} \\
& \quad \leqslant \sum_{i=1,2} \sum_{j=3,4} \mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t, T_{i j} \leqslant t\right\}
\end{aligned}
$$

Thus

$$
\begin{align*}
\operatorname{Var}\left(\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right) \leqslant & \mathbb{P}\left\{1 \sim_{\Pi(t)} 2 \sim_{\Pi(t)} 3 \sim_{\Pi(t)} 4\right\}  \tag{8.8}\\
& +\sum_{i=1,2} \sum_{j=3,4} \mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t, T_{i j} \leqslant t\right\}
\end{align*}
$$

Put $D_{i j}:=\left|Z_{i}(0)-Z_{j}(0)\right|$. We have

$$
\begin{align*}
\mathbb{P}\{1 & \left.\sim_{\Pi(t)} 2 \sim_{\Pi(t)} 3 \sim_{\Pi(t)} 4\right\} \\
= & \mathbb{P}\left\{T_{12} \leqslant t, T_{13} \wedge T_{23} \leqslant t, T_{14} \wedge T_{24} \wedge T_{34} \leqslant t\right\} \\
= & \mathbb{P}\left(\left\{T_{12} \leqslant t, T_{13} \wedge T_{23} \leqslant t, T_{14} \wedge T_{24} \wedge T_{34} \leqslant t\right\}\right. \\
& \left.\backslash\left\{D_{12} \leqslant t^{\frac{2}{5}},\left(D_{13} \wedge D_{23}\right) \leqslant t^{\frac{2}{5}},\left(D_{14} \wedge D_{24} \wedge D_{34}\right) \leqslant t^{\frac{2}{5}}\right\}\right)  \tag{8.9}\\
& +\mathbb{P}\left\{D_{12} \leqslant t^{\frac{2}{5}},\left(D_{13} \wedge D_{23}\right) \leqslant t^{\frac{2}{5}},\left(D_{14} \wedge D_{24} \wedge D_{34}\right) \leqslant t^{\frac{2}{5}}\right\} \\
\leqslant & \sum_{1 \leqslant i<j \leqslant 4} \mathbb{P}\left\{T_{i j} \leqslant t, D_{i j}>t^{\frac{2}{5}}\right\}+\mathbb{P}\left\{\max _{1 \leqslant i<j \leqslant 4} D_{i j} \leqslant 3 t^{\frac{2}{5}}\right\}
\end{align*}
$$

where we have appealed to the triangle inequality in the last step. Because $\frac{2}{5}<\frac{1}{2}$, an application of the reflection principle and Brownian scaling certainly gives that the probability $\mathbb{P}\left\{T_{i j} \leqslant t, D_{i j}>t^{\frac{2}{5}}\right\}$ is $o\left(t^{\alpha}\right)$ as $t \downarrow 0$ for any $\alpha>0$. Moreover, by the translation invariance of $m$ (the common distribution of the $\left.Z_{i}(0)\right)$, the second term in the rightmost member of (8.9) is at most

$$
\begin{aligned}
& \mathbb{P}\left\{\left|Z_{2}(0)-Z_{1}(0)\right| \leqslant 3 t^{\frac{2}{5}},\left|Z_{3}(0)-Z_{1}(0)\right| \leqslant 3 t^{\frac{2}{5}},\left|Z_{4}(0)-Z_{1}(0)\right| \leqslant 3 t^{\frac{2}{5}}\right\} \\
& \quad=\mathbb{P}\left\{\left|Z_{2}(0)\right| \leqslant 3 t^{\frac{2}{5}},\left|Z_{3}(0)\right| \leqslant 3 t^{\frac{2}{5}},\left|Z_{4}(0)\right| \leqslant 3 t^{\frac{2}{5}}\right\} \\
& \quad=c t^{\frac{6}{5}}
\end{aligned}
$$

for a suitable constant $c$ when $t$ is sufficiently small. Therefore,

$$
\begin{align*}
& \mathbb{P}\left\{1 \sim_{\Pi(t)} 2 \sim_{\Pi(t)} 3 \sim_{\Pi(t)} 4\right\} \\
& \quad=\mathbb{P}\left\{\left\{T_{12} \leqslant t, T_{13} \wedge T_{23} \leqslant t, T_{14} \wedge T_{24} \wedge T_{34} \leqslant t\right\}\right.  \tag{8.10}\\
& \quad=O\left(t^{\frac{6}{5}}\right), \quad \text { as } t \downarrow 0
\end{align*}
$$

A similar argument establishes that

$$
\begin{equation*}
\mathbb{P}\left\{T_{12} \leqslant t, T_{34} \leqslant t, T_{i j} \leqslant t\right\}=O\left(t^{\frac{6}{5}}\right), \quad \text { as } t \downarrow 0, \tag{8.11}
\end{equation*}
$$

for $i=1,2$ and $j=3,4$.
Substituting (8.10) and (8.11) into (8.8) gives

$$
\operatorname{Var}\left(\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right)=O\left(t^{\frac{6}{5}}\right), \quad \text { as } t \downarrow 0
$$

This establishes the desired result when combined with the expectation calculation (8.6), Chebyshev's inequality, a standard Borel-Cantelli argument, and the monotonicity of $\sum_{i=1}^{N(t)} F_{i}(t)^{2}$.

We may suppose that on our probability space $(\Omega, \mathcal{F}, \mathbb{P})$ there is a sequence $B_{1}, B_{2}, \ldots$ of i.i.d. one-dimensional standard Brownian motions with initial distribution the uniform distribution on $[0,2 \pi]$ and that $Z_{i}$ is defined by setting $Z_{i}(t)$ to be the image of $B_{i}(t)$ under the usual homomorphism from $\mathbb{R}$ onto $\mathbb{T}$. For $n \in \mathbb{N}$ and $0 \leqslant j \leqslant 2^{n}-1$, let $I_{1}^{n, j} \leqslant I_{2}^{n, j} \leqslant \ldots$ be a list in increasing order of the set of indices $\left\{i \in \mathbb{N}: B_{i}(0) \in\left[2 \pi j / 2^{n}, 2 \pi(j+1) / 2^{n}[ \}\right.\right.$. Put $B_{i}^{n, j}:=B_{I_{i}^{n, j}}$ and $Z_{i}^{n, j}:=Z_{I_{i}^{n, j}}$. Thus, $\left(B_{i}^{n, j}\right)_{i \in \mathbb{N}}$ is an i.i.d. sequence of standard $\mathbb{R}$-valued Brownian motions and $\left(Z_{i}^{n, j}\right)_{i \in \mathbb{N}}$ is an i.i.d. sequence of standard $\mathbb{T}$-valued Brownian motions. In each case the corresponding initial distribution is uniform on $\left[2 \pi j / 2^{n}, 2 \pi(j+1) / 2^{n}[\right.$. Moreover, for $n \in \mathbb{N}$ fixed the sequences $\left(B_{i}^{n, j}\right)_{i \in \mathbb{N}}$ are independent as $j$ varies and the same is true of the sequences $\left(Z_{i}^{n, j}\right)_{i \in \mathbb{N}}$.

Let $\underline{W}$ (resp. $\underline{W}^{n, j}, W^{n, j}$ ) be the coalescing system defined in terms of $\left(B_{i}\right)_{i \in \mathbb{N}}\left(\right.$ resp. $\left.\left(B_{i}^{n, j}\right)_{i \in \mathbb{N}},\left(Z_{i}^{n, j}\right)_{i \in \mathbb{N}}\right)$ in the same manner that $W$ is defined in terms of $\left(Z_{i}\right)_{i \in \mathbb{N}}$.

It is clear by construction that

$$
\begin{equation*}
N(t)=|W(t)| \leqslant \sum_{i=0}^{2^{n}-1}\left|W^{n, i}(t)\right| \leqslant \sum_{i=0}^{2^{n}-1}\left|\underline{W}^{n, i}(t)\right|, \quad t>0, n \in \mathbb{N} \tag{8.12}
\end{equation*}
$$

Lemma 8.7. The expectation $\mathbb{P}[|\underline{W}(1)|]$ is finite.
Proof. There is an obvious analogue of the duality relation Proposition 8.4 for systems of coalescing and annihilating one-dimensional Brownian motions. Using this duality and arguing as in the proof of Corollary 8.5, it is easy to see that, letting $\bar{L}$ and $\bar{U}$ be two independent, standard, real-valued Brownian motions on some probability space $(\bar{\Omega}, \overline{\mathcal{F}}, \overline{\mathbb{P}})$ with $\bar{L}(0)=\bar{U}(0)=0$,
$\mathbb{P}[|\underline{W}(1)|]$

$$
\begin{aligned}
= & \lim _{M \rightarrow \infty} \sum_{i=-\infty}^{\infty} \mathbb{P}\{\underline{W}(1) \cap[2 \pi i / M, 2 \pi(i+1) / M] \neq \varnothing\} \\
= & \lim _{M \rightarrow \infty} \sum_{i=-\infty}^{\infty} \overline{\mathbb{P}}\left\{\min _{0 \leqslant t \leqslant 1}((\bar{U}(t)+2 \pi(i+1) / M)-(\bar{L}(t)+2 \pi i / M))>0\right. \\
& {[\bar{L}(1)+2 \pi i / M, \bar{U}(1)+2 \pi(i+1) / M] \cap[0,2 \pi] \neq \varnothing\} } \\
\leqslant & \limsup _{M \rightarrow \infty} c^{\prime} M \overline{\mathbb{P}}\left[\mathbf{1}\left\{\min _{0 \leqslant t \leqslant 1}(\bar{U}(t)-\bar{L}(t))>-2 \pi / M\right\}\left(\bar{U}(1)-\bar{L}(1)+c^{\prime \prime}\right)\right]
\end{aligned}
$$

for suitable constants $c^{\prime}$ and $c^{\prime \prime}$. Noting that $(\bar{U}-\bar{L}) / \sqrt{2}$ is a standard Brownian motion, the result follows from a straightforward calculation with the joint distribution of the minimum up to time 1 and value at time 1 of such a process - see, for example, Corollary 30 in Section 1.3 of [70].

Proposition 8.8. Almost surely,

$$
0<\liminf _{t \downarrow 0} t^{\frac{1}{2}} N(t) \leqslant \limsup _{t \downarrow 0} t^{\frac{1}{2}} N(t)<\infty
$$

Proof. By the Cauchy-Schwarz inequality,

$$
1=\left(\sum_{i=1}^{N(t)} F_{i}(t)\right)^{2} \leqslant N(t) \sum_{i=1}^{N(t)} F_{i}(t)^{2}
$$

Hence, by Lemma 8.6,

$$
\liminf _{t \downarrow} t^{\frac{1}{2}} N(t) \geqslant \frac{\pi^{\frac{3}{2}}}{2}, \quad \mathbb{P}-\text { a.s. }
$$

On the other hand, for each $n \in \mathbb{N},\left|\underline{W}^{n, i}\left(2^{-2 n}\right)\right|, i=0, \ldots, 2^{n}-1$, are i.i.d. random variables that, by Brownian scaling, have the same distribution as $|\underline{W}(1)|$. By (8.12),

$$
t^{\frac{1}{2}} N(t) \leqslant \frac{1}{2^{n-1}} \sum_{i=0}^{2^{n}-1}\left|\underline{W}^{n, i}\left(2^{-2 n}\right)\right|
$$

for $2^{-2 n}<t \leqslant 2^{-2(n-1)}$. An application of Lemma 8.7 and the following strong law of large numbers for triangular arrays completes the proof.

Lemma 8.9. Consider a triangular array $\left\{X_{n, i}: 1 \leqslant i \leqslant 2^{n}, n \in \mathbb{N}\right\}$ of identically distributed, real-valued, mean zero, random variables on some probability space $(\Omega, \mathcal{F}, \mathbb{P})$ such that the collection $\left\{X_{n, i}: 1 \leqslant i \leqslant 2^{n}\right\}$ is independent for each $n \in \mathbb{N}$. Then

$$
\lim _{n \rightarrow \infty} 2^{-n}\left(X_{n, 1}+\cdots+X_{n, 2^{n}}\right)=0, \mathbb{P}-\text { a.s. }
$$

Proof. This sort of result appears to be known in the theory of complete convergence. For example, it follows from the much more general Theorem A in [23] by taking $N_{n}=2^{n}$ and $\psi(t)=2^{t}$ in the notation of that result - see also the Example following that result. For the sake of completeness, we give a short proof that was pointed out to us by Michael Klass.

Let $\left\{Y_{n}: n \in \mathbb{N}\right\}$ be an independent identically distributed sequence with the same common distribution as the $X_{n, i}$. By the strong law of large numbers, for any $\varepsilon>0$ the probability that $\left|Y_{1}+\cdots+Y_{2^{n}}\right|>\varepsilon 2^{n}$ infinitely often is 0 . Therefore, by the triangle inequality, for any $\varepsilon>0$ the probability that $\left|Y_{2^{n}+1}+\cdots+Y_{2^{n+1}}\right|>\varepsilon 2^{n}$ infinitely often is 0 ; and so, by the Borel-Cantelli lemma for sequences of independent events,

$$
\sum_{n} \mathbb{P}\left\{\left|Y_{2^{n}+1}+\cdots+Y_{2^{n+1}}\right|>\varepsilon 2^{n}\right\}<\infty
$$

for all $\varepsilon>0$. The last sum is also

$$
\sum_{n} \mathbb{P}\left\{\left|X_{n, 1}+\cdots+X_{n, 2^{n}}\right|>\varepsilon 2^{n}\right\},
$$

and an application of the "other half" of the Borel-Cantelli lemma for possibly dependent events establishes that for all $\varepsilon>0$ the probability of $\mid X_{n, 1}+\cdots+$ $X_{n, 2^{n}} \mid>\varepsilon 2^{n}$ infinitely often is 0 , as required.

We can now give the proof of Theorem 8.2. Proposition 8.8 and Lemma 8.6 verify the conditions of Proposition B.3. The proof is then completed using Equation (10) of [113] that gives upper and lower bounds on the capacity of $C_{\frac{1}{2}}$ in an arbitrary gauge.

## Subtree prune and re-graft

### 9.1 Background

As we mentioned in Chapter 1, Markov chains that move through a space of finite trees are an important ingredient in several algorithms in phylogenetic analysis, and one standard set of moves that is implemented in several phylogenetic software packages is the set of subtree prune and re-graft (SPR) moves.

In an SPR move, a binary tree $T$ (that is, a tree in which all non-leaf vertices have degree three) is cut "in the middle of an edge" to give two subtrees, say $T^{\prime}$ and $T^{\prime \prime}$. Another edge is chosen in $T^{\prime}$, a new vertex is created "in the middle" of that edge, and the cut edge in $T^{\prime \prime}$ is attached to this new vertex. Lastly, the "pendant" cut edge in $T^{\prime}$ is removed along with the vertex it was attached to in order to produce a new binary tree that has the same number of vertices as $T$ - see Figure 9.1.

In this chapter we investigate the asymptotics of the simplest possible treevalued Markov chain based on the SPR moves, namely the chain in which the two edges that are chosen for cutting and for re-attaching are chosen uniformly (without replacement) from the edges in the current tree. Intuitively, the continuous time Markov process we discuss arises as limit when the number of vertices in the tree goes to infinity, the edge lengths are re-scaled by a constant factor so that initial tree converges in a suitable sense to a continuous analogue of a combinatorial tree (more specifically, a compact real tree), and the time scale of the Markov chain is sped up by an appropriate factor. We do not, in fact, prove such a limit theorem. Rather, we use Dirichlet form techniques to establish the existence of a process that has the dynamics we would expect from such a limit.

The process we construct has as its state space the set of pairs $(T, \nu)$, where $T$ is a compact real tree and $\nu$ is a probability measure on $T$. Let $\mu$ be the length measure associated with $T$. Our process jumps away from $T$ by first choosing a pair of points $(u, v) \in T \times T$ according to the rate measure $\mu \otimes \nu$ and then transforming $T$ into a new tree by cutting off the


Fig. 9.1. A subtree prune and re-graft operation
subtree rooted at $u$ that does not contain $v$ and re-attaching this subtree at $v$. This jump kernel (which typically has infinite total mass - so that jumps are occurring on a dense countable set) is precisely what we would expect for a limit (as the number of vertices goes to infinity) of the particular SPR Markov chain on finite trees described above in which the edges for cutting and re-attachment are chosen uniformly at each stage. The limit process is reversible with respect to the distribution of Brownian CRT weighted with the probability measure that comes from the push-forward of Lebesgue measure on $[0,1]$ as in Example 4.39.

For $\mathbb{R}$-trees arising from an excursion path, the counterpart of an SPR move is the excision and re-insertion of a sub-excursion. Figure 9.2 illustrates such an operation.

We follow the development of [65] in this chapter.

### 9.2 The weighted Brownian CRT

Consider the Itô excursion measure for excursions of standard Brownian motion away from 0 . This $\sigma$-finite measure is defined subject to a normalization of Brownian local time at 0 , and we take the usual normalization of local


Fig. 9.2. A subtree prune and re-graft operation on an excursion path: the excursion starting at time $u$ in the top picture is excised and inserted at time $v$, and the resulting gap between the two points marked \# is closed up. The two points marked \# (resp. *) in the top (resp. bottom) picture correspond to a single point in the associated real tree.
times at each level that makes the local time process an occupation density in the spatial variable for each fixed value of the time variable. The excursion measure is the sum of two measures, one that is concentrated on non-negative excursions and one that is concentrated on non-positive excursions. Let $\mathbb{N}$ be the part that is concentrated on non-negative excursions. Thus, in the notation of Example 3.14, $\mathbb{N}$ is a $\sigma$-finite measure on the space of excursion paths $U$, where we equip $U$ with the $\sigma$-field $\mathcal{U}$ generated by the coordinate maps.

Define a map $v: U \rightarrow U^{1}$ by $e \mapsto \frac{e(\zeta(e) \cdot)}{\sqrt{\zeta(e)}}$. Then

$$
\mathbb{P}(\Gamma):=\frac{\mathbb{N}\left\{v^{-1}(\Gamma) \cap\{e \in U: \zeta(e) \geqslant c\}\right\}}{\mathbb{N}\{e \in U: \zeta(e) \geqslant c\}}, \quad \Gamma \in \mathcal{U}
$$

does not depend on $c>0$ - see, for example, Exercise 12.2.13.2 in [117]. The probability measure $\mathbb{P}$ is called the law of normalized non-negative Brownian excursion. We have

$$
\begin{equation*}
\mathbb{N}\{e \in U: \zeta(e) \in \mathrm{d} c\}=\frac{\mathrm{d} c}{2 \sqrt{2 \pi c^{3}}} \tag{9.1}
\end{equation*}
$$

and, defining $\mathcal{S}_{c}: U^{1} \rightarrow U^{c}$ by

$$
\begin{equation*}
\mathcal{S}_{c} e:=\sqrt{c} e(\cdot / c) \tag{9.2}
\end{equation*}
$$

we have

$$
\begin{equation*}
\int \mathbb{N}(\mathrm{d} e) G(e)=\int_{0}^{\infty} \frac{\mathrm{d} c}{2 \sqrt{2 \pi c^{3}}} \int_{U^{1}} \mathbb{P}(\mathrm{~d} e) G\left(\mathcal{S}_{c} e\right) \tag{9.3}
\end{equation*}
$$

for a non-negative measurable function $G: U \rightarrow \mathbb{R}$.
Recall from Example 4.39 how each $e \in U^{1}$ is associated with a weighted compact $\mathbb{R}$-tree $\left(T_{e}, d_{T_{e}}, \nu_{T_{e}}\right)$. Let $\mathbf{P}$ be the probability measure on ( $\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}^{\mathrm{wt}}}$ ) that is the push-forward of the normalized excursion measure by the map $e \mapsto\left(T_{2 e}, d_{T_{2 e}}, \nu_{T_{2 e}}\right)$, where $2 e \in U^{1}$ is just the excursion path $t \mapsto 2 e(t)$.

Thus, the probability measure $\mathbf{P}$ is the distribution of an object consisting of the Brownian CRT equipped with its natural weight. Recall that the Brownian continuum random tree arises as the limit of a uniform random tree on $n$ vertices when $n \rightarrow \infty$ and edge lengths are rescaled by a factor of $1 / \sqrt{n}$. The associated weight on each realization of the continuum random tree is the probability measure that arises in this limiting construction by taking the uniform probability measure on realizations of the approximating finite trees. Therefore, the probability measure $\mathbf{P}$ can be viewed informally as the "uniform distribution" on ( $\left.\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}}{ }^{\mathrm{wt}}\right)$.

### 9.3 Campbell measure facts

For the purposes of constructing the Markov process that is of interest to us, we need to understand picking a random weighted tree $\left(T, d_{T}, \nu_{T}\right)$ according to the continuum random tree distribution $\mathbf{P}$, picking a point $u$ according to the length measure $\mu^{T}$ and another point $v$ according to the weight $\nu_{T}$, and then decomposing $T$ into two subtrees rooted at $u$ - one that contains $v$ and one that does not (we are being a little imprecise here, because $\mu^{T}$ will be an infinite measure, $\mathbf{P}$ almost surely).

In order to understand this decomposition, we must understand the corresponding decomposition of excursion paths under normalized excursion measure. Because subtrees correspond to sub-excursions and because of our observation in Example 4.34 that for an excursion $e$ the length measure $\mu^{T_{e}}$ on the corresponding tree is the push-forward of the measure $\int_{\Gamma_{e}} \mathrm{~d} s \otimes \mathrm{~d} a \frac{1}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} \delta_{\underline{s}(e, s, a)}$ by the quotient map, we need to understand the decomposition of the excursion $e$ into the excursion above $a$ that straddles $s$ and the "remaining" excursion when when $e$ is chosen according to the standard Brownian excursion distribution $\mathbb{P}$ and $(s, a)$ is chosen according to the $\sigma$-finite measure $\mathrm{d} s \otimes \mathrm{~d} a \frac{1}{\bar{s}(e, s, a)-\underline{s}(e, s, a)}$ on $\Gamma_{e}$ - see Figure 9.3.

Given an excursion $e \in U$ and a level $a \geqslant 0$ write:

- $\zeta(e):=\inf \{t>0: e(t)=0\}$ for the "length" of $e$,


Fig. 9.3. The decomposition of the excursion $e$ in the top picture into the excursion $\hat{e}^{s, a}$ above level $a$ that straddles time $s$ in the middle picture and the "remaining" excursion $\check{e}^{s, a}$ in the bottom picture.

- $\quad \ell_{t}^{a}(e)$ for the local time of $e$ at level $a$ up to time $t$,
- $\quad e^{\downarrow a}$ for $e$ time-changed by the inverse of $t \mapsto \int_{0}^{t} \mathrm{~d} s 1\{e(s) \leqslant a\}$ (that is, $e^{\downarrow a}$ is $e$ with the sub-excursions above level $a$ excised and the gaps closed up),
- $\quad \ell_{t}^{a}\left(e^{\downarrow a}\right)$ for the local time of $e^{\downarrow a}$ at the level $a$ up to time $t$,
- $U^{\uparrow a}(e)$ for the set of sub-excursion intervals of $e$ above $a$ (that is, an element of $U^{\uparrow a}(e)$ is an interval $I=\left[g_{I}, d_{I}\right]$ such that $e\left(g_{I}\right)=e\left(d_{I}\right)=a$ and $e(t)>a$ for $\left.g_{I}<t<d_{I}\right)$,
- $\mathcal{N}^{\uparrow a}(e)$ for the counting measure that puts a unit mass at each point $\left(s^{\prime}, e^{\prime}\right)$, where, for some $I \in U^{\uparrow a}(e), s^{\prime}:=\ell_{g_{I}}^{a}(e)$ is the amount of local time of $e$ at level $a$ accumulated up to the beginning of the sub-excursion $I$ and $e^{\prime} \in U$ is given by

$$
e^{\prime}(t)= \begin{cases}e\left(g_{I}+t\right)-a, & 0 \leqslant t \leqslant d_{I}-g_{I} \\ 0, & t>d_{I}-g_{I}\end{cases}
$$

is the corresponding piece of the path $e$ shifted to become an excursion above the level 0 starting at time 0 ,

- $\hat{e}^{s, a} \in U$ and $\check{e}^{s, a} \in U$, for the subexcursion "above" $(s, a) \in \Gamma_{e}$, that is,

$$
\hat{e}^{s, a}(t):=\left\{\begin{array}{cc}
e(\underline{s}(e, s, a)+t)-a, 0 \leqslant t \leqslant \bar{s}(e, s, a)-\underline{s}(e, s, a) \\
0, & t>\bar{s}(e, s, a)-\underline{s}(e, s, a)
\end{array}\right.
$$

respectively"below" $(s, a) \in \Gamma_{e}$, that is,

$$
\check{e}^{s, a}(t):=\left\{\begin{array}{cl}
e(t), & 0 \leqslant t \leqslant \underline{s}(e, s, a), \\
e(t+\bar{s}(e, s, a)-\underline{s}(e, s, a)), & t>\underline{s}(e, s, a) .
\end{array}\right.
$$

- $\sigma_{s}^{a}(e):=\inf \left\{t \geqslant 0: \ell_{t}^{a}(e) \geqslant s\right\}$ and $\tau_{s}^{a}(e):=\inf \left\{t \geqslant 0: \ell_{t}^{a}(e)>s\right\}$,
- $\tilde{e}^{s, a} \in U$ for $e$ with the interval ] $\sigma_{s}^{a}(e), \tau_{s}^{a}(e)$ [containing an excursion above level $a$ excised, that is,

$$
\tilde{e}^{s, a}(t):= \begin{cases}e(t), & 0 \leqslant t \leqslant \sigma_{s}^{a}(e) \\ e\left(t+\tau_{s}^{a}(e)-\sigma_{s}^{a}(e)\right), & t>\sigma_{s}^{a}(e)\end{cases}
$$

The following path decomposition result under the $\sigma$-finite measure $\mathbb{N}$ is preparatory to a decomposition under the probability measure $\mathbb{P}$, Corollary 9.2 , that has a simpler intuitive interpretation.

Proposition 9.1. For non-negative measurable functions $F$ on $\mathbb{R}_{+}$and $G, H$ on $U$,

$$
\begin{aligned}
& \int \mathbb{N}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\overline{\bar{s}}(e, s, a)-\underline{s}(e, s, a)} F(\underline{s}(e, s, a)) G\left(\hat{e}^{s, a}\right) H\left(\check{e}^{s, a}\right) \\
& \quad=\int \mathbb{N}(\mathrm{d} e) \int_{0}^{\infty} \mathrm{d} a \int \mathcal{N}^{\uparrow a}(e)\left(\mathrm{d}\left(s^{\prime}, \mathrm{e}^{\prime}\right)\right) F\left(\sigma_{s^{\prime}}^{a}(e)\right) G\left(e^{\prime}\right) H\left(\tilde{e}^{s^{s}, a}\right) \\
& \quad=\mathbb{N}[G] \mathbb{N}\left[H \int_{0}^{\zeta} \mathrm{d} s F(s)\right] .
\end{aligned}
$$

Proof. The first equality is just a change in the order of integration and has already been remarked upon in Example 4.34.

Standard excursion theory - see, for example, [119, 117, 29] - says that under $\mathbb{N}$, the random measure $e \mapsto \mathcal{N}^{\uparrow a}(e)$ conditional on $e \mapsto e^{\downarrow a}$ is a Poisson random measure with intensity measure $\lambda^{\downarrow a}(e) \otimes \mathbb{N}$, where $\lambda^{\downarrow a}(e)$ is Lebesgue measure restricted to the interval $\left[0, \ell_{\infty}^{a}(e)\right]=\left[0,2 \ell_{\infty}^{a}\left(e^{\downarrow a}\right)\right]$.

Note that $\tilde{e}^{s^{\prime}, a}$ is constructed from $e^{\downarrow a}$ and $\mathcal{N}^{\uparrow a}(e)-\delta_{\left(s^{\prime}, e^{\prime}\right)}$ in the same way that $e$ is constructed from $e^{\downarrow a}$ and $\mathcal{N}^{\uparrow a}(e)$. Also, $\sigma_{s^{\prime}}^{a}\left(\tilde{e}^{e^{\prime}, a}\right)=\sigma_{s^{\prime}}^{a}(e)$. Therefore, by the Campbell-Palm formula for Poisson random measures see, for example, Section 12.1 of [41] -

$$
\begin{aligned}
& \int \mathbb{N}(\mathrm{d} e) \int_{0}^{\infty} \mathrm{d} a \int \mathcal{N}^{\uparrow a}(e)\left(\mathrm{d}\left(s^{\prime}, \mathrm{e}^{\prime}\right)\right) F\left(\sigma_{s^{\prime}}^{a}(e)\right) G\left(e^{\prime}\right) H\left(\tilde{e}^{s^{\prime}, a}\right) \\
& \quad=\int \mathbb{N}(\mathrm{d} e) \int_{0}^{\infty} \mathrm{d} a \mathbb{N}\left[\int \mathcal{N}^{\uparrow a}(e)\left(\mathrm{d}\left(s^{\prime}, \mathrm{e}^{\prime}\right)\right) F\left(\sigma_{s^{\prime}}^{a}(e)\right) G\left(e^{\prime}\right) H\left(\tilde{e}^{s^{\prime}, a}\right) \mid e^{\downarrow a}\right] \\
& =\int \mathbb{N}(\mathrm{d} e) \int_{0}^{\infty} \mathrm{d} a \mathbb{N}[G] \mathbb{N}\left[\left\{\int_{0}^{\ell_{\infty}^{a}(e)} \mathrm{d} s^{\prime} F\left(\sigma_{s^{\prime}}^{a}(e)\right)\right\} H \mid e^{\downarrow a}\right] \\
& =\mathbb{N}[G] \int_{0}^{\infty} \mathrm{d} a \int \mathbb{N}(\mathrm{~d} e)\left(\left\{\int \mathrm{d} \ell_{s}^{a}(e) F(s)\right\} H(e)\right) \\
& =\mathbb{N}[G] \int \mathbb{N}(\mathrm{d} e)\left(\left\{\int_{0}^{\infty} \mathrm{d} a \int \mathrm{~d} \ell_{s}^{a}(e) F(s)\right\} H(e)\right) \\
& =\mathbb{N}[G] \mathbb{N}\left[H \int_{0}^{\zeta} \mathrm{d} s F(s)\right] .
\end{aligned}
$$

The next result says that if we pick an excursion $e$ according to the standard excursion distribution $\mathbb{P}$ and then pick a point $(s, a) \in \Gamma_{e}$ according to the $\sigma$-finite length measure corresponding to the length measure $\mu^{T_{e}}$ on the associated tree $T_{e}$, then the following objects are independent:
(a) the length of the excursion above level $a$ that straddles time $s$,
(b) the excursion obtained by taking the excursion above level $a$ that straddles time $s$, turning it (by a shift of axes) into an excursion $\hat{e}^{s, a}$ above level zero starting at time zero, and then Brownian re-scaling $\hat{e}^{s, a}$ to produce an excursion of unit length,
(c) the excursion obtained by taking the excursion $\check{e}^{s, a}$ that comes from excising $\hat{e}^{s, a}$ and closing up the gap, and then Brownian re-scaling $\check{e}^{s, a}$ to produce an excursion of unit length,
(d) the starting time $\underline{s}(e, s, a)$ of the excursion above level $a$ that straddles time $s$ rescaled by the length of $\check{e}^{s, a}$ to give a time in the interval $[0,1]$.
Moreover, the length in (a) is "distributed" according to the $\sigma$-finite measure

$$
\frac{1}{2 \sqrt{2 \pi}} \frac{\mathrm{~d} \rho}{\sqrt{(1-\rho) \rho^{3}}}, \quad 0 \leqslant \rho \leqslant 1
$$

the unit length excursions in (b) and (c) are both distributed as standard Brownian excursions (that is, according to $\mathbb{P}$ ), and the time in (d) is uniformly distributed on the interval $[0,1]$.

Corollary 9.2. For non-negative measurable functions $F$ on $\mathbb{R}_{+}$and $K$ on $U \times U$,

$$
\begin{aligned}
& \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} F\left(\frac{\underline{s}(e, s, a)}{\zeta\left(\check{e}^{s, a}\right)}\right) K\left(\hat{e}^{s, a}, e^{s, a}\right) \\
& =\left\{\int_{0}^{1} \mathrm{~d} u F(u)\right\} \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} K\left(\hat{e}^{s, a}, \check{e}^{s, a}\right) \\
& =\left\{\int_{0}^{1} \mathrm{~d} u F(u)\right\} \frac{1}{2 \sqrt{2 \pi}} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{(1-\rho) \rho^{3}}} \int \mathbb{P}\left(\mathrm{~d} e^{\prime}\right) \otimes \mathbb{P}\left(\mathrm{d} e^{\prime \prime}\right) K\left(\mathcal{S}_{\rho} e^{\prime}, \mathcal{S}_{1-\rho} e^{\prime \prime}\right) .
\end{aligned}
$$

Proof. For a non-negative measurable function $L$ on $U \times U$, it follows straightforwardly from Proposition 9.1 that

$$
\begin{align*}
& \int \mathbb{N}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} F\left(\frac{\underline{s}(e, s, a)}{\zeta\left(e^{s, a}\right)}\right) L\left(\hat{e}^{s, a}, e^{s, a}\right)  \tag{9.4}\\
& =\left\{\int_{0}^{1} \mathrm{~d} u F(u)\right\} \int \mathbb{N}\left(\mathrm{de}^{\prime}\right) \otimes \mathbb{N}\left(\mathrm{de}^{\prime \prime}\right) L\left(e^{\prime}, e^{\prime \prime}\right) \zeta\left(e^{\prime \prime}\right)
\end{align*}
$$

The left-hand side of equation (9.4) is, by (9.3),

$$
\begin{equation*}
\left.\int_{0}^{\infty} \frac{\mathrm{d} c}{2 \sqrt{2 \pi c^{3}}} \int \mathbb{P}(\mathrm{~d} e) \int_{\Gamma_{\mathcal{S}_{c} e}} \mathrm{~d} s \otimes \mathrm{~d} a \frac{F\left(\frac{\underline{s}\left(\mathcal{S}_{c} e, s, a\right)}{\zeta\left(\mathcal{S}_{c} e\right.}\right) L\left(\widehat{\mathcal{S}}_{c} e, a\right.}{s, a},{\widehat{\mathcal{S}_{c}}}^{s, a}\right) \tag{9.5}
\end{equation*}
$$

If we change variables to $t=s / c$ and $b=a / \sqrt{c}$, then the integral for $(s, a)$ over $\Gamma_{\mathcal{S}_{c} e}$ becomes an integral for $(t, b)$ over $\Gamma_{e}$. Also,

$$
\begin{aligned}
\underline{s}\left(\mathcal{S}_{c} e, c t, \sqrt{c} b\right) & =\sup \left\{r<c t: \sqrt{c} e\left(\frac{r}{c}\right)<\sqrt{c} b\right\} \\
& =c \sup \{r<t: e(r)<b\} \\
& =c \underline{s}(e, t, b),
\end{aligned}
$$

and, by similar reasoning,

$$
\bar{s}\left(\mathcal{S}_{c} e, c t, \sqrt{c} b\right)=c \bar{s}(e, t, b)
$$

and

$$
\zeta\left({\widetilde{\mathcal{S}_{c}}}^{c t, \sqrt{c} b}\right)=c \zeta\left(\bar{e}^{t, b}\right)
$$

Thus, (9.5) is

$$
\begin{equation*}
\int_{0}^{\infty} \frac{\mathrm{d} c}{2 \sqrt{2 \pi c^{3}}} \int \mathbb{P}(\mathrm{~d} e) \sqrt{c} \int_{\Gamma_{e}} \mathrm{~d} t \otimes \mathrm{~d} b \frac{F\left(\frac{\underline{s}(e, t, b)}{\bar{\zeta}\left(\tilde{e}^{t, b}\right)}\right) L\left({\widehat{\mathcal{S}} e^{c} e}^{c t, \sqrt{c} b}, \widehat{\mathcal{S}}_{c} e^{c t, \sqrt{c b} b}\right)}{\bar{s}(e, t, b)-\underline{s}(e, t, b)} \tag{9.6}
\end{equation*}
$$

Now suppose that $L$ is of the form

$$
L\left(e^{\prime}, e^{\prime \prime}\right)=K\left(\mathcal{R}_{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)} e^{\prime}, \mathcal{R}_{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)} e^{\prime \prime}\right) \frac{M\left(\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)\right)}{\sqrt{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)}}
$$

where, for ease of notation, we put for $e \in U$, and $c>0$,

$$
\mathcal{R}_{c} e:=\mathcal{S}_{c^{-1}} e=\frac{1}{\sqrt{c}} e(c \cdot) .
$$

Then (9.6) becomes

$$
\begin{equation*}
\left.\int_{0}^{\infty} \frac{\mathrm{d} c}{2 \sqrt{2 \pi c^{3}}} \int \mathbb{P}(\mathrm{~d} e) \int_{\Gamma_{e}} \mathrm{~d} t \otimes \mathrm{~d} b \frac{F\left(\frac{s(e, t, b)}{\zeta\left(e^{t} t, b\right.}\right)}{}\right) K\left(\hat{e}^{t, b}, e^{t, b}\right) M(c) . \tag{9.7}
\end{equation*}
$$

Since (9.7) was shown to be equivalent to the left hand side of (9.4), it follows from (9.3) that

$$
\begin{align*}
\int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} & \frac{\mathrm{~d} t \otimes \mathrm{~d} b}{\bar{s}(e, t, b)-\underline{s}(e, t, b)} F\left(\frac{\underline{s}(e, t, b)}{\zeta\left(e^{t, b}\right)}\right) K\left(\hat{e}^{t, b}, e^{t, b}\right) \\
& =\frac{\int_{0}^{1} \mathrm{~d} u F(u)}{\mathbb{N}[M]} \int \mathbb{N}\left(\mathrm{de}^{\prime}\right) \otimes \mathbb{N}\left(\mathrm{de}^{\prime \prime}\right) L\left(e^{\prime}, e^{\prime \prime}\right) \zeta\left(e^{\prime \prime}\right) \tag{9.8}
\end{align*}
$$

and the first equality of the statement follows.
We have from the identity (9.8) that, for any $C>0$,

$$
\begin{aligned}
& \mathbb{N}\{\zeta(e)>C\} \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} K\left(e^{s, a}, e^{s, a}\right) \\
& =\int \mathbb{N}\left(\mathrm{de}^{\prime}\right) \otimes \mathbb{N}\left(\mathrm{de}^{\prime \prime}\right) K\left(\mathcal{R}_{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)} e^{\prime}, \mathcal{R}_{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)} e^{\prime \prime}\right) \frac{1\left\{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)>C\right\}}{\sqrt{\zeta\left(e^{\prime}\right)+\zeta\left(e^{\prime \prime}\right)}} \zeta\left(e^{\prime \prime}\right) \\
& =\int_{0}^{\infty} \frac{\mathrm{d} c^{\prime}}{2 \sqrt{2 \pi c^{\prime 3}}} \int_{0}^{\infty} \frac{\mathrm{d} c^{\prime \prime}}{2 \sqrt{2 \pi c^{\prime \prime}}} \\
& \quad \int \mathbb{P}\left(\mathrm{d} e^{\prime}\right) \otimes \mathbb{P}\left(\mathrm{d} e^{\prime \prime}\right) K\left(\mathcal{R}_{c^{\prime}+c^{\prime \prime}} \mathcal{S}_{c^{\prime}} e^{\prime}, \mathcal{R}_{c^{\prime}+c^{\prime \prime}} \mathcal{S}_{c^{\prime \prime}} e^{\prime \prime}\right) \frac{1\left\{c^{\prime}+c^{\prime \prime}>C\right\}}{\sqrt{c^{\prime}+c^{\prime \prime}}} .
\end{aligned}
$$

Make the change of variables $\rho=\frac{c^{\prime}}{c^{\prime}+c^{\prime \prime}}$ and $\xi=c^{\prime}+c^{\prime \prime}$ (with corresponding Jacobian factor $\xi$ ) to get

$$
\begin{aligned}
& \int_{0}^{\infty} \frac{\mathrm{d} c^{\prime}}{2 \sqrt{2 \pi c^{\prime 3}}} \int_{0}^{\infty} \frac{\mathrm{d} c^{\prime \prime}}{2 \sqrt{2 \pi c^{\prime \prime}}} \\
& =\left(\frac{1}{2 \sqrt{2 \pi}}\right)^{2} \int_{0}^{\infty} d \xi \int_{0}^{1} \frac{\mathrm{~d} \rho \xi}{\sqrt{\rho^{3}(1-\rho) \xi^{4}}} \frac{1\{\xi>C\}}{\sqrt{\xi}} \\
& \int \mathbb{P}\left(\mathrm{d} e^{\prime \prime}\right) K\left(\mathcal{R}_{c^{\prime}+c^{\prime \prime}} \mathcal{S}_{c^{\prime}} e^{\prime}, \mathcal{R}_{c^{\prime}+c^{\prime \prime}} \mathcal{S}_{c^{\prime \prime}} e^{\prime \prime}\right) \frac{1\left\{c^{\prime}+c^{\prime \prime}>C\right\}}{\sqrt{c^{\prime}+c^{\prime \prime}}} \\
& =\left(\frac{1}{2 \sqrt{2 \pi}}\right)^{2}\left\{\int_{C}^{\infty} \frac{d \xi}{\sqrt{\xi^{3}}}\right\} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{\rho^{3}(1-\rho)}} \\
& \int \mathbb{P}\left(\mathrm{d} e^{\prime}\right) \otimes \mathbb{P}\left(\mathrm{S} e_{\rho} e^{\prime \prime}, \mathcal{S}_{1-\rho} e^{\prime \prime}\right) K\left(\mathcal{S}_{\rho} e^{\prime}, \mathcal{S}_{1-\rho} e^{\prime \prime}\right),
\end{aligned}
$$

and the corollary follows upon recalling (9.1).

Corollary 9.3. (i) For $x>0$,

$$
\begin{aligned}
& \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} 1\left\{\max _{0 \leqslant t \leqslant \zeta\left(\hat{e}^{s, a}\right)} \hat{e}^{s, a}>x\right\} \\
& \quad=2 \sum_{n \in \mathbb{N}} n x \exp \left(-2 n^{2} x^{2}\right)
\end{aligned}
$$

(ii) For $0<p \leqslant 1$,

$$
\int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} 1\left\{\zeta\left(\hat{e}^{s, a}\right)>p\right\}=\sqrt{\frac{1-p}{2 \pi p}} .
$$

Proof. (i) Recall first of all from Theorem 5.2.10 in [92] that

$$
\mathbb{P}\left\{e \in U^{1}: \max _{0 \leqslant t \leqslant 1} e(t)>x\right\}=2 \sum_{n \in \mathbb{N}}\left(4 n^{2} x^{2}-1\right) \exp \left(-2 n^{2} x^{2}\right)
$$

By Corollary 9.2 applied to $K\left(e^{\prime}, e^{\prime \prime}\right):=1\left\{\max _{t \in\left[0, \zeta\left(e^{\prime}\right)\right]} e^{\prime}(t) \geqslant x\right\}$ and $F \equiv 1$,

$$
\begin{aligned}
& \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} 1\left\{\max _{0 \leqslant t \leqslant \zeta\left(\hat{e}^{s, a}\right)} \hat{e}^{s, a}>x\right\} \\
& \quad=\frac{1}{2 \sqrt{2 \pi}} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{\rho^{3}(1-\rho)}} \mathbb{P}\left\{\max _{t \in[0, \rho]} \sqrt{\rho} e(t / \rho)>x\right\} \\
& \quad=\frac{1}{2 \sqrt{2 \pi}} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{\rho^{3}(1-\rho)}} \mathbb{P}\left\{\max _{t \in[0,1]} e(t)>\frac{x}{\sqrt{\rho}}\right\} \\
& \quad=\frac{1}{2 \sqrt{2 \pi}} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{\rho^{3}(1-\rho)}} 2 \sum_{n \in \mathbb{N}}\left(4 n^{2} \frac{x^{2}}{\rho}-1\right) \exp \left(-2 n^{2} \frac{x^{2}}{\rho}\right) \\
& \quad=2 \sum_{n \in \mathbb{N}} n x \exp \left(-2 n^{2} x^{2}\right),
\end{aligned}
$$

as claimed.
(ii) Corollary 9.2 applied to $K\left(e^{\prime}, e^{\prime \prime}\right):=1\left\{\zeta\left(e^{\prime}\right) \geqslant p\right\}$ and $F \equiv 1$ immediately yields

$$
\begin{aligned}
& \int \mathbb{P}(\mathrm{d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} 1\left\{\zeta\left(\hat{e}^{s, a}\right)>p\right\} \\
& \quad=\frac{1}{2 \sqrt{2 \pi}} \int_{p}^{1} \frac{\mathrm{~d} \rho}{\sqrt{\rho^{3}(1-\rho)}}=\sqrt{\frac{1-p}{2 \pi p}} .
\end{aligned}
$$

We conclude this section by calculating the expectations of some functionals with respect to $\mathbf{P}$ (the the "uniform distribution" on ( $\left.\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}}{ }^{\mathrm{wt}}\right)$ as introduced in the end of Section 9.2).

For $\varepsilon>0, T \in \mathbf{T}$, and $\rho \in T$, write $R_{\varepsilon}(T, \rho)$ for the $\varepsilon$-trimming of the rooted $\mathbb{R}$-tree obtained by rooting $T$ at $\rho$ (recall Subsection 4.3.4). With a slight abuse of notation, set

$$
R_{\varepsilon}(T):=\left\{\begin{array}{cc}
\bigcap_{\rho \in T} R_{\varepsilon}(T, \rho), & \operatorname{diam}(T)>\varepsilon  \tag{9.9}\\
\text { singleton }, & \operatorname{diam}(T) \leqslant \varepsilon
\end{array}\right.
$$

For $T \in \mathbf{T}^{\mathrm{wt}}$ recall the length measure $\mu^{T}$ from (4.10). Given $(T, d) \in \mathbf{T}^{\mathrm{wt}}$ and $u, v \in T$, let

$$
\begin{equation*}
S^{T, u, v}:=\{x \in T: u \in] v, x[ \} \tag{9.10}
\end{equation*}
$$

denote the subtree of $T$ that differs from its closure by the point $u$, which can be thought of as its root, and consists of points that are on the "other side" of $u$ from $v$ (recall ] $v, x[$ is the open segment in $T$ between $v$ and $x$ ).
Lemma 9.4. (i) For $x>0$,

$$
\begin{aligned}
\mathbf{P} & {\left[\mu^{T} \otimes \nu_{T}\left\{(u, v) \in T \times T: \operatorname{height}\left(S^{T, u, v}\right)>x\right\}\right] } \\
& =\mathbf{P}\left[\int_{T} \nu_{T}(\mathrm{~d} v) \mu^{T}\left(R_{x}(T, v)\right)\right] \\
& =2 \sum_{n \in \mathbb{N}} n x \exp \left(-n^{2} x^{2} / 2\right)
\end{aligned}
$$

(ii) For $1<\alpha<\infty$,

$$
\begin{aligned}
\mathbf{P} & {\left[\int_{T} \nu_{T}(\mathrm{~d} v) \int_{T} \mu^{T}(\mathrm{~d} u)\left(\operatorname{height}\left(S^{T, u, v}\right)\right)^{\alpha}\right] } \\
& =2^{-\frac{1}{2}} \alpha \Gamma\left(\frac{\alpha}{2}+\frac{1}{2}\right) \zeta(\alpha)
\end{aligned}
$$

where, as usual, $\zeta(\alpha):=\sum_{n \geqslant 1} n^{-\alpha}$.
(iii) For $0<p \leqslant 1$,

$$
\mathbf{P}\left[\mu^{T} \otimes \nu_{T}\left\{(u, v) \in T \times T: \nu_{T}\left(S^{T, u, v}\right)>p\right\}\right]=\sqrt{\frac{2(1-p)}{\pi p}}
$$

(iv) For $\frac{1}{2}<\beta<\infty$,

$$
\mathbf{P}\left[\int_{T} \nu_{T}(\mathrm{~d} v) \int_{T} \mu^{T}(\mathrm{~d} u)\left(\nu_{T}\left(S^{T, u, v}\right)\right)^{\beta}\right]=2^{-\frac{1}{2}} \frac{\Gamma\left(\beta-\frac{1}{2}\right)}{\Gamma(\beta)}
$$

Proof. (i) The first equality is clear from the definition of $R_{x}(T, v)$ and Fubini's theorem.

Turning to the equality of the first and last terms, first recall that $\mathbf{P}$ is the push-forward on ( $\mathbf{T}^{\mathrm{wt}}, d_{\mathrm{GH}^{\mathrm{wt}}}$ ) of the normalized excursion measure $\mathbb{P}$ by the map $e \mapsto\left(T_{2 e}, d_{T_{2 e}}, \nu_{T_{2 e}}\right)$, where $2 e \in U^{1}$ is just the excursion path $t \mapsto$ $2 e(t)$. In particular, $T_{2 e}$ is the quotient of the interval [ 0,1$]$ by the equivalence relation defined by $2 e$. By the invariance of the standard Brownian excursion under random re-rooting - see Section 2.7 of [13] - the point in $T_{2 e}$ that corresponds to the equivalence class of $0 \in[0,1]$ is distributed according to
$\nu_{T_{2 e}}$ when $e$ is chosen according to $\mathbb{P}$. Moreover, recall from Example 4.34 that for $e \in U^{1}$, the length measure $\mu^{T_{e}}$ is the push-forward of the measure $\mathrm{d} s \otimes \mathrm{~d} a \frac{1}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} \delta_{\underline{s}(e, s, a)}$ on the sub-graph $\Gamma_{e}$ by the quotient map defined in (3.14).

It follows that if we pick $T$ according to $\mathbf{P}$ and then pick $(u, v) \in T \times T$ according to $\mu^{T} \otimes \nu_{T}$, then the subtree $S^{T, u, v}$ that arises has the same $\sigma$-finite law as the tree associated with the excursion $2 \hat{e}^{s, a}$ when $e$ is chosen according to $\mathbb{P}$ and $(s, a)$ is chosen according to the measure $\mathrm{d} s \otimes \mathrm{~d} a \frac{1}{\overline{s(e, s, a)-\underline{s}(e, s, a)}} \delta_{\underline{s}(e, s, a)}$ on the sub-graph $\Gamma_{e}$.

Therefore, by part (i) of Corollary 9.3,

$$
\begin{aligned}
\mathbf{P} & {\left[\int_{T} \nu_{T}(\mathrm{~d} v) \int_{T} \mu^{T}(\mathrm{~d} u) 1\left\{\operatorname{height}\left(S^{T, u, v}\right)>x\right\}\right] } \\
& =2 \int \mathbb{P}(\mathrm{~d} e) \int_{\Gamma_{e}} \frac{\mathrm{~d} s \otimes \mathrm{~d} a}{\bar{s}(e, s, a)-\underline{s}(e, s, a)} 1\left\{\max _{0 \leqslant t \leqslant \zeta\left(\hat{e}^{s, a}\right)} \hat{e}^{s, a}>\frac{x}{2}\right\} \\
& =2 \sum_{n \in \mathbb{N}} n x \exp \left(-n^{2} x^{2} / 2\right) .
\end{aligned}
$$

Part (ii) is a consequence of part (i) and some straightforward calculus.
Part (iii) follows immediately from part(ii) of Corollary 9.3.
Part (iv) is a consequence of part (iii) and some more straightforward calculus.

### 9.4 A symmetric jump measure

In this section we will construct and study a measure on $\mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}$ that is related to the decomposition discussed at the beginning of Section 9.3.

Define a map $\Theta$ from $\{((T, d), u, v): T \in \mathbf{T}, u \in T, v \in T\}$ into $\mathbf{T}$ by setting $\Theta((T, d), u, v):=\left(T, d^{(u, v)}\right)$ where letting

$$
d^{(u, v)}(x, y):=\left\{\begin{array}{cc}
d(x, y), & \text { if } x, y \in S^{T, u, v} \\
d(x, y), & \text { if } x, y \in T \backslash S^{T, u, v} \\
d(x, u)+d(v, y), & \text { if } x \in S^{T, u, v}, y \in T \backslash S^{T, u, v} \\
d(y, u)+d(v, x), & \text { if } y \in S^{T, u, v}, x \in T \backslash S^{T, u, v}
\end{array}\right.
$$

That is, $\Theta((T, d), u, v)$ is just $T$ as a set, but the metric has been changed so that the subtree $S^{T, u, v}$ with root $u$ is now pruned and re-grafted so as to have root $v$.

If $(T, d, \nu) \in \mathbf{T}^{\mathrm{wt}}$ and $(u, v) \in T \times T$, then we can think of $\nu$ as a weight on $\left(T, d^{(u, v)}\right)$, because the Borel structures induces by $d$ and $d^{(u, v)}$ are the same. With a slight misuse of notation we will, therefore, write $\Theta((T, d, \nu), u, v)$ for $\left(T, d^{(u, v)}, \nu\right) \in \mathbf{T}^{\mathrm{wt}}$. Intuitively, the mass contained in $S^{T, u, v}$ is transported along with the subtree.

Define a kernel $\kappa$ on $\mathbf{T}^{\mathrm{wt}}$ by

$$
\kappa\left(\left(T, d_{T}, \nu_{T}\right), \mathbf{B}\right):=\mu^{T} \otimes \nu_{T}\{(u, v) \in T \times T: \Theta(T, u, v) \in \mathbf{B}\}
$$

for $\mathbf{B} \in \mathcal{B}\left(\mathbf{T}^{\mathrm{wt}}\right)$. Thus, $\kappa\left(\left(T, d_{T}, \nu_{T}\right), \cdot\right)$ is the jump kernel described informally in Section 9.1.

We show in part (i) of Lemma 9.5 below that the kernel $\kappa$ is reversible with respect to the probability measure $\mathbf{P}$. More precisely, we show that if we define a measure $J$ on $\mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}$ by

$$
J(\mathbf{A} \times \mathbf{B}):=\int_{\mathbf{A}} \mathbf{P}(\mathrm{d} T) \kappa(T, \mathbf{B})
$$

for $\mathbf{A}, \mathbf{B} \in \mathcal{B}\left(\mathbf{T}^{\mathrm{wt}}\right)$, then $J$ is symmetric.

## Lemma 9.5. Then

(i) The measure $J$ is symmetric.
(ii) For each compact subset $\mathbf{K} \subset \mathbf{T}^{\mathbf{w t}}$ and open subset $\mathbf{U}$ such that $\mathbf{K} \subset$ $\mathbf{U} \subseteq \mathbf{T}$,

$$
J\left(\mathbf{K}, \mathbf{T}^{\mathrm{wt}} \backslash \mathbf{U}\right)<\infty
$$

(iii)

$$
\int_{\mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}} J(\mathrm{~d} T, \mathrm{~d} S) \Delta_{\mathrm{GH}^{\mathrm{wt}}}^{2}(T, S)<\infty .
$$

Proof. (i) Given $e^{\prime}, e^{\prime \prime} \in U^{1}, 0 \leqslant u \leqslant 1$, and $0<\rho \leqslant 1$, define $e^{\circ}\left(\cdot ; e^{\prime}, e^{\prime \prime}, u, \rho\right) \in$ $U^{1}$ by

$$
\begin{aligned}
& e^{\circ}\left(t ; e^{\prime}, e^{\prime \prime}, u, \rho\right) \\
& \quad:= \begin{cases}\mathcal{S}_{1-\rho} e^{\prime \prime}(t), & 0 \leqslant t \leqslant(1-\rho) u \\
\mathcal{S}_{1-\rho} e^{\prime \prime}((1-\rho) u)+\mathcal{S}_{\rho} e^{\prime}(t-(1-\rho) u), & (1-\rho) u \leqslant t \leqslant(1-\rho) u+\rho, \\
\mathcal{S}_{1-\rho} e^{\prime \prime}(t-\rho), & (1-\rho) u+\rho \leqslant t \leqslant 1\end{cases}
\end{aligned}
$$

That is, $e^{\circ}\left(\cdot ; e^{\prime}, e^{\prime \prime}, u, \rho\right)$ is the excursion that arises from Brownian re-scaling $e^{\prime}$ and $e^{\prime \prime}$ to have lengths $\rho$ and $1-\rho$, respectively, and then inserting the re-scaled version of $e^{\prime}$ into the re-scaled version of $e^{\prime \prime}$ at a position that is a fraction $u$ of the total length of the re-scaled version of $e^{\prime \prime}$.

Define a measure $\mathbb{J}$ on $U^{1} \times U^{1}$ by

$$
\begin{aligned}
& \int_{U^{1} \times U^{1}} \mathbb{J}\left(\mathrm{~d} e^{*}, \mathrm{~d} e^{* *}\right) K\left(e^{*}, e^{* *}\right) \\
& :=\int_{[0,1]^{2}} \mathrm{~d} u \otimes \mathrm{~d} v \frac{1}{2 \sqrt{2 \pi}} \int_{0}^{1} \frac{\mathrm{~d} \rho}{\sqrt{(1-\rho) \rho^{3}}} \int \mathbb{P}\left(\mathrm{~d} e^{\prime}\right) \otimes \mathbb{P}\left(\mathrm{d} e^{\prime \prime}\right) \\
& \quad \times K\left(e^{\circ}\left(\cdot ; e^{\prime}, e^{\prime \prime}, u, \rho\right), e^{\circ}\left(\cdot ; e^{\prime}, e^{\prime \prime}, v, \rho\right)\right)
\end{aligned}
$$

Clearly, the measure $\mathbb{J}$ is symmetric. It follows from the discussion at the beginning of the proof of part (i) of Lemma 9.4 and Corollary 9.2 that the
measure $J$ is the push-forward of the symmetric measure $2 \mathbb{J}$ by the map that sends the pair $\left(e^{*}, e^{* *}\right) \in U^{1} \times U^{1}$ to the pair

$$
\left(\left(T_{2 e^{*}}, d_{T_{2 e^{*}}}, \nu_{T_{2 e^{*}}}\right),\left(T_{2 e^{* *}}, d_{T_{2 e^{* *}}}, \nu_{T_{2 e^{* *}}}\right)\right) .
$$

Hence, $J$ is also symmetric.
(ii) The result is trivial if $\mathbf{K}=\varnothing$, so we assume that $\mathbf{K} \neq \varnothing$. Since $\mathbf{T}^{\mathrm{wt}} \backslash \mathbf{U}$ and $\mathbf{K}$ are disjoint closed sets and $\mathbf{K}$ is compact, we have that

$$
c:=\inf _{T \in \mathbf{K}, S \in \mathbf{U}} \Delta_{\mathrm{GH}^{\mathrm{wt}}}(T, S)>0
$$

Fix $T \in \mathbf{K}$. If $(u, v) \in T \times T$ is such that $\Delta_{\mathrm{GH}}(T, \Theta(T, u, v)) \geqslant c$, then either

- $u \in R_{c}(T)$, or
- there exists $\rho \in T^{o}$ such that $u \notin R_{c}(T, \rho)$ and $\nu_{T}\left(S^{T, u, \rho}\right) \geqslant c$ (recall that $R_{c}(T)$ is the $c$-trimming of $T$, that $R_{c}(T, \rho)$ is the $c$-trimming of $T$ rooted at $\rho$, and that $S^{T, u, \rho}$ is the subtree of $T$ consisting of points that are on the other side of $u$ to $\rho$ ).

Hence, we have

$$
\begin{aligned}
& J\left(\mathbf{K}, \mathbf{T}^{\mathrm{wt}} \backslash \mathbf{U}\right) \\
& \leqslant \int_{\mathbf{K}} \mathbf{P}\{\mathrm{d} T\} \kappa\left(T,\left\{S: \Delta_{\mathrm{GH}^{\mathrm{wt}}}(T, S)>c\right\}\right) \\
& \leqslant \int_{\mathbf{K}} \mathbf{P}(\mathrm{d} T) \mu^{T}\left(R_{c}(T)\right) \\
&+\int_{\mathbf{K}} \mathbf{P}(\mathrm{d} T) \int_{T} \nu_{T}(\mathrm{~d} v) \mu^{T}\left\{u \in T: \nu_{T}\left(S^{T, u, v}\right)>c\right\} \\
&<\infty
\end{aligned}
$$

where we have used Lemma 9.4 and the observation that

$$
\mu^{T}\left(R_{c}(T)\right) \leqslant \int_{T} \nu_{T}(\mathrm{~d} v) \mu^{T}\left(R_{c}(T, v)\right)
$$

because $R_{c}(T) \subseteq R_{c}(T, v)$ for all $v \in T$.
(iii) Similar reasoning yields that

$$
\begin{aligned}
& \int_{\mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}} J(\mathrm{~d} T, \mathrm{~d} S) \Delta_{\mathrm{GH}^{\mathrm{wt}}}^{2}(T, S) \\
& \quad=\int_{\mathbf{T}^{\mathrm{wt}}} \mathbf{P}\{\mathrm{~d} T\} \int_{0}^{\infty} \mathrm{d} t 2 t \kappa\left(T,\left\{S: \Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}(T, S)>t\right\}\right) \\
& \leqslant \\
& \leqslant \int_{\mathbf{T}^{\mathrm{wt}}} \mathbf{P}(\mathrm{~d} T) \int_{0}^{\infty} \mathrm{d} t 2 t \mu^{T}\left(R_{t}(T)\right) \\
& \quad+\int_{\mathbf{T}^{\mathrm{wt}}} \mathbf{P}(\mathrm{~d} T) \int_{0}^{\infty} \mathrm{d} t 2 t \int_{T} \nu_{T}(\mathrm{~d} v) \mu^{T}\left\{u \in T: \nu_{T}\left\{S^{T, u, v}\right\}>t\right\} \\
& \leqslant \\
& \leqslant \int_{0}^{\infty} \mathrm{d} t 2 t \int_{\mathbf{T}^{\mathrm{wt}}} \mathbf{P}(\mathrm{~d} T) \mu^{T}\left(R_{t}(T)\right) \\
& \quad+\int_{\mathbf{T}^{\mathrm{wt}}} \mathbf{P}(\mathrm{~d} T) \int_{T} \nu_{T}(\mathrm{~d} v) \int_{T} \mu^{T}(\mathrm{~d} u) \nu_{T}^{2}\left(S^{T, u, v}\right) \\
& \quad<\infty,
\end{aligned}
$$

where we have applied Lemma 9.4 once more.

### 9.5 The Dirichlet form

Consider the bilinear form

$$
\mathcal{E}(f, g):=\int_{\mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}} J(\mathrm{~d} T, \mathrm{~d} S)(f(S)-f(T))(g(S)-g(T)),
$$

for $f, g$ in the domain

$$
\mathcal{D}^{*}(\mathcal{E}):=\left\{f \in L^{2}\left(\mathbf{T}^{\mathrm{wt}}, \mathbf{P}\right): f \text { is measurable, and } \mathcal{E}(f, f)<\infty\right\},
$$

(here as usual, $L^{2}\left(\mathbf{T}^{\mathbf{w t}}, \mathbf{P}\right)$ is equipped with the inner product $(f, g)_{\mathbf{P}}:=$ $\left.\int \mathbf{P}(\mathrm{d} x) f(x) g(x)\right)$. By the argument in Example 1.2.1 in [72] and Lemma 9.5, $\left(\mathcal{E}, \mathcal{D}^{*}(\mathcal{E})\right)$ is well-defined, symmetric and Markovian.

Lemma 9.6. The form $\left(\mathcal{E}, \mathcal{D}^{*}(\mathcal{E})\right)$ is closed. That is, if $\left(f_{n}\right)_{n \in \mathbb{N}}$ be a sequence in $\mathcal{D}^{*}(\mathcal{E})$ such that

$$
\lim _{m, n \rightarrow \infty}\left(\mathcal{E}\left(f_{n}-f_{m}, f_{n}-f_{m}\right)+\left(f_{n}-f_{m}, f_{n}-f_{m}\right)_{\mathbf{P}}\right)=0
$$

then there exists $f \in \mathcal{D}^{*}(\mathcal{E})$ such that

$$
\lim _{n \rightarrow \infty}\left(\mathcal{E}\left(f_{n}-f, f_{n}-f\right)+\left(f_{n}-f, f_{n}-f\right)_{\mathbf{P}}\right)=0 .
$$

Proof. Let $\left(f_{n}\right)_{n \in \mathbb{N}}$ be a sequence such that $\lim _{m, n \rightarrow \infty} \mathcal{E}\left(f_{n}-f_{m}, f_{n}-f_{m}\right)+$ $\left(f_{n}-f_{m}, f_{n}-f_{m}\right)_{\mathbf{P}}=0$ (that is, $\left(f_{n}\right)_{n \in \mathbb{N}}$ is Cauchy with respect to $\mathcal{E}(\cdot, \cdot)+$ $\left.(\cdot, \cdot)_{\mathbf{P}}\right)$. There exists a subsequence $\left(n_{k}\right)_{k \in \mathbb{N}}$ and $f \in L_{2}\left(\mathbf{T}^{\mathbf{w t}}, \mathbf{P}\right)$ such that $\lim _{k \rightarrow \infty} f_{n_{k}}=f, \mathbf{P}$-a.s, and $\lim _{k \rightarrow \infty}\left(f_{n_{k}}-f, f_{n_{k}}-f\right)_{\mathbf{P}}=0$. By Fatou's Lemma,

$$
\int J(\mathrm{~d} T, \mathrm{~d} S)\left((f(S)-f(T))^{2} \leqslant \liminf _{k \rightarrow \infty} \mathcal{E}\left(f_{n_{k}}, f_{n_{k}}\right)<\infty\right.
$$

and so $f \in \mathcal{D}^{*}(\mathcal{E})$. Similarly,

$$
\begin{aligned}
\mathcal{E} & \left(f_{n}-f, f_{n}-f\right) \\
& =\int J(\mathrm{~d} T, \mathrm{~d} S) \lim _{k \rightarrow \infty}\left(\left(f_{n}-f_{n_{k}}\right)(S)-\left(f_{n}-f_{n_{k}}\right)(T)\right)^{2} \\
& \leqslant \liminf _{k \rightarrow \infty} \mathcal{E}\left(f_{n}-f_{n_{k}}, f_{n}-f_{n_{k}}\right) \rightarrow 0
\end{aligned}
$$

as $n \rightarrow \infty$. Thus, $\left\{f_{n}\right\}_{n \in \mathbb{N}}$ has a subsequence that converges to $f$ with respect to $\mathcal{E}(\cdot, \cdot)+(\cdot, \cdot)_{\mathbf{P}}$, but, by the Cauchy property, this implies that $\left\{f_{n}\right\}_{n \in \mathbb{N}}$ itself converges to $f$.

Let $\mathcal{L}$ denote the collection of functions $f: \mathbf{T}^{\mathrm{wt}} \rightarrow \mathbb{R}$ such that

$$
\begin{equation*}
\sup _{T \in \mathbf{T}^{\mathrm{wt}}}|f(T)|<\infty \tag{9.11}
\end{equation*}
$$

and

$$
\begin{equation*}
\sup _{S, T \in \mathbf{T}^{\mathrm{wt}}, S \neq T} \frac{|f(S)-f(T)|}{\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}(S, T)}<\infty \tag{9.12}
\end{equation*}
$$

Note that $\mathcal{L}$ consists of continuous functions and contains the constants. It follows from (4.20) that $\mathcal{L}$ is both a vector lattice and an algebra. By Lemma 9.7 below, $\mathcal{L} \subseteq \mathcal{D}^{*}(\mathcal{E})$. Therefore, the closure of $(\mathcal{E}, \mathcal{L})$ is a Dirichlet form that we will denote by $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$.

Lemma 9.7. Suppose that $\left\{f_{n}\right\}_{n \in \mathbb{N}}$ is a sequence of functions from $\mathbf{T}^{\mathrm{wt}}$ into $\mathbb{R}$ such that

$$
\begin{gathered}
\sup _{n \in \mathbb{N}} \sup _{T \in \mathbf{T}^{\mathrm{wt}}}\left|f_{n}(T)\right|<\infty, \\
\sup _{n \in \mathbb{N}} \sup _{S, T \in \mathbf{T}^{\mathrm{wt}}, S \neq T} \frac{\left|f_{n}(S)-f_{n}(T)\right|}{\Delta_{\mathrm{GH}}{ }^{\mathrm{wt}}(S, T)}<\infty,
\end{gathered}
$$

and

$$
\lim _{n \rightarrow \infty} f_{n}=f, \quad \text { P-a.s. }
$$

for some $f: \mathbf{T}^{\mathrm{wt}} \rightarrow \mathbb{R}$. Then $\left\{f_{n}\right\}_{n \in \mathbb{N}} \subset \mathcal{D}^{*}(\mathcal{E})$, $f \in \mathcal{D}^{*}(\mathcal{E})$, and

$$
\lim _{n \rightarrow \infty}\left(\mathcal{E}\left(f_{n}-f, f_{n}-f\right)+\left(f_{n}-f, f_{n}-f\right)_{\mathbf{P}}\right)=0
$$

Proof. By the definition of the measure $J$, see (9.4), and the symmetry of $J$ (Lemma 9.5(i)), we have that $f_{n}(x)-f_{n}(y) \rightarrow f(x)-f(y)$ for $J$-almost every pair $(x, y)$. The result then follows from part (iii) of Lemma 9.5 and the dominated convergence theorem.

Before showing that $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ is the Dirichlet form of a nice Markov process, we remark that $\mathcal{L}$, and thus also $\mathcal{D}(\mathcal{E})$, is quite a rich class of functions. We show in the proof of Theorem 9.8 below that $\mathcal{L}$ separates points of $\mathbf{T}^{\mathrm{wt}}$. Hence, if $\mathbf{K}$ is any compact subset of $\mathbf{T}^{\mathrm{wt}}$, then, by the Arzela-Ascoli theorem, the set of restrictions of functions in $\mathcal{L}$ to $\mathbf{K}$ is uniformly dense in the space of real-valued continuous functions on $\mathbf{K}$.

The following theorem states that there is a well-defined Markov process with the dynamics we would expect for a limit of the subtree prune and regraft chains.

Theorem 9.8. There exists a recurrent $\mathbf{P}$-symmetric Hunt process $X=$ $\left(X_{t}, \mathbb{P}^{T}\right)$ on $\mathbf{T}^{\mathrm{wt}}$ whose Dirichlet form is $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$.

Proof. We will check the conditions of Theorem A. 8 to establish the existence of $X$.

Because $\mathbf{T}^{\mathrm{wt}}$ is complete and separable (recall Theorem 4.44) there is a sequence $\mathbf{H}_{1} \subseteq \mathbf{H}_{2} \subseteq \ldots$ of compact subsets of $\mathbf{T}^{\text {wt }}$ such that $\mathbf{P}\left(\bigcup_{k \in \mathbb{N}} \mathbf{H}_{k}\right)=$ 1. Given $\alpha, \beta>0$, write $\mathcal{L}_{\alpha, \beta}$ for the subset of $\mathcal{L}$ consisting of functions $f$ such that

$$
\sup _{T \in \mathbf{T}^{\mathrm{wt}}}|f(T)| \leqslant \alpha
$$

and

$$
\sup _{S, T \in \mathbf{T}^{\mathrm{w}}, S \neq T} \frac{|f(S)-f(T)|}{\Delta_{\mathrm{GH}^{\mathrm{wt}}}(S, T)} \leqslant \beta
$$

By the separability of the continuous real-valued functions on each $\mathbf{H}_{k}$ with respect to the supremum norm, it follows that for each $k \in \mathbb{N}$ there is a countable set $L_{\alpha, \beta, k} \subseteq \mathcal{L}_{\alpha, \beta}$ such that for every $f \in \mathcal{L}_{\alpha, \beta}$

$$
\inf _{g \in L_{\alpha, \beta, k}} \sup _{T \in H_{k}}|f(T)-g(T)|=0
$$

Set $L_{\alpha, \beta}:=\bigcup_{k \in \mathbb{N}} L_{\alpha, \beta, k}$. Then for any $f \in \mathcal{L}_{\alpha, \beta}$ there exists a sequence $\left\{f_{n}\right\}_{n \in \mathbb{N}}$ in $L_{\alpha, \beta}$ such that $\lim _{n \rightarrow \infty} f_{n}=f$ pointwise on $\bigcup_{k \in \mathbb{N}} \mathbf{H}_{k}$, and, $a$ fortiori, $\mathbf{P}$-almost surely. By Lemma 9.7, the countable set $\bigcup_{m \in \mathbb{N}} L_{m, m}$ is dense in $\mathcal{L}$ and, a fortiori, in $\mathcal{D}(\mathcal{E})$, with respect to $\mathcal{E}(\cdot, \cdot)+(\cdot, \cdot)_{\mathbf{P}}$.

Now fix a countable dense subset $\mathbf{S} \subset \mathbf{T}^{\mathrm{wt}}$. Let $M$ denote the countable set of functions of the form

$$
T \mapsto p+q\left(\Delta_{\mathrm{GH}^{\mathrm{wt}}}(S, T) \wedge r\right)
$$

for some $S \in \mathbf{S}$ and $p, q, r \in \mathbb{Q}$. Note that $M \subseteq \mathcal{L}$, that $M$ separates the points of $\mathbf{T}^{\mathrm{wt}}$, and, for any $T \in \mathbf{T}^{\mathrm{wt}}$, that there is certainly a function $f \in M$ with $f(T) \neq 0$.

Consequently, if $\mathcal{C}$ is the algebra generated by the countable set $M \cup$ $\bigcup_{m \in \mathbb{N}} L_{m, m}$, then it is certainly the case that $\mathcal{C}$ is dense in $\mathcal{D}(\mathcal{E})$ with respect $\mathcal{E}(\cdot, \cdot)+(\cdot, \cdot)_{\mathbf{P}}$, that $\mathcal{C}$ separates the points of $\mathbf{T}^{\mathrm{wt}}$, and, for any $T \in \mathbf{T}^{\mathrm{wt}}$, that there is a function $f \in \mathcal{C}$ with $f(T) \neq 0$.

All that remains in verifying the conditions of Theorem A. 8 is to check the tightness condition that there exist compact subsets $\mathbf{K}_{1} \subseteq \mathbf{K}_{2} \subseteq \ldots$ of $\mathbf{T}^{\mathrm{wt}}$ such that $\lim _{n \rightarrow \infty} \operatorname{Cap}\left(\mathbf{T}^{\mathrm{wt}} \backslash \mathbf{K}_{n}\right)=0$ where Cap is the capacity associated with the Dirichlet form This convergence, however, is the content of Lemma 9.11 below.

Finally, because constants belongs to $\mathcal{D}(\mathcal{E})$, it follows from Theorem 1.6.3 in [72] that $X$ is recurrent.

The following results were needed in the proof of Theorem 9.8
Lemma 9.9. For $\varepsilon, a, \delta>0$, put $\mathbf{V}_{\varepsilon, a}:=\left\{T \in \mathbf{T}: \mu^{T}\left(R_{\varepsilon}(T)\right)>a\right\}$ and, as usual, $\mathbf{V}_{\varepsilon, a}^{\delta}:=\left\{T \in \mathbf{T}: d_{\mathrm{GH}}\left(T, \mathbf{V}_{\varepsilon, a}\right)<\delta\right\}$. Then, for fixed $\varepsilon>3 \delta$,

$$
\bigcap_{a>0} \mathbf{V}_{\varepsilon, a}^{\delta}=\varnothing
$$

Proof. Fix $S \in \mathbf{T}$. If $S \in \mathbf{V}_{\varepsilon, a}^{\delta}$, then there exists $T \in \mathbf{V}_{\varepsilon, a}$ such that $d_{\mathrm{GH}}(S, T)<\delta$. Observe that $R_{\varepsilon}(T)$ is not the trivial tree consisting of a single point because it has total length greater than $a$. Write $\left\{y_{1}, \ldots, y_{n}\right\}$ for the leaves of $R_{\varepsilon}(T)$. Note that $T \backslash R_{\varepsilon}(T)^{o}$ is the union of $n$ subtrees of diameter $\varepsilon$. The closure of each subtree contains a unique $y_{i}$. Choose $z_{i}$ in the subtree whose closure contains $y_{i}$ such that $d_{T}\left(y_{i}, z_{i}\right)=\varepsilon$.

Let $\Re$ be a correspondence between $S$ and $T$ with $\operatorname{dis}(\Re)<2 \delta$. Pick $x_{1}, \ldots, x_{n} \in S$ such that $\left(x_{i}, z_{i}\right) \in \Re$. Hence, $\left|d_{S}\left(x_{i}, x_{j}\right)-d_{T}\left(z_{i}, z_{j}\right)\right|<2 \delta$ for all $i, j$.

The distance in $R_{\varepsilon}(T)$ from the point $y_{k}$ to the segment $\left[y_{i}, y_{j}\right]$ is

$$
\frac{1}{2}\left(d_{S}\left(y_{k}, y_{i}\right)+d_{S}\left(y_{k}, y_{j}\right)-d_{S}\left(y_{i}, y_{j}\right)\right)
$$

Thus, the distance from $y_{k}, 3 \leqslant k \leqslant n$, to the subtree spanned by $y_{1}, \ldots, y_{k-1}$ is

$$
\bigwedge_{1 \leqslant i \leqslant j \leqslant k-1} \frac{1}{2}\left(d_{T}\left(y_{k}, y_{i}\right)+d_{T}\left(y_{k}, y_{j}\right)-d_{T}\left(y_{i}, y_{j}\right)\right)
$$

Hence,

$$
\begin{aligned}
\mu^{T}\left(R_{\varepsilon}(T)\right)= & d_{T}\left(y_{1}, y_{2}\right) \\
& +\sum_{k=3}^{n} \bigwedge_{1 \leqslant i \leqslant j \leqslant k-1} \frac{1}{2}\left(d_{T}\left(y_{k}, y_{i}\right)+d_{T}\left(y_{k}, y_{j}\right)-d_{T}\left(y_{i}, y_{j}\right)\right) .
\end{aligned}
$$

Now the distance in $S$ from the point $x_{k}$ to the segment $\left[x_{i}, x_{j}\right]$ is

$$
\begin{aligned}
& \frac{1}{2}\left(d_{S}\left(x_{k}, x_{i}\right)+d_{S}\left(x_{k}, x_{j}\right)-d_{S}\left(x_{i}, x_{j}\right)\right) \\
& \quad \geqslant \frac{1}{2}\left(d_{T}\left(z_{k}, z_{i}\right)+d_{T}\left(z_{k}, z_{j}\right)-d_{T}\left(z_{i}, z_{j}\right)-3 \times 2 \delta\right) \\
& \quad=\frac{1}{2}\left(d_{T}\left(y_{k}, y_{i}\right)+2 \varepsilon+d_{T}\left(y_{k}, y_{j}\right)+2 \varepsilon-d_{T}\left(y_{i}, y_{j}\right)-2 \varepsilon-6 \delta\right) \\
& \quad>0
\end{aligned}
$$

by the assumption that $\varepsilon>3 \delta$. In particular, $x_{1}, \ldots, x_{n}$ are leaves of the subtree spanned by $\left\{x_{1}, \ldots, x_{n}\right\}$, and $R_{\gamma}(S)$ has at least $n$ leaves when $0<$ $\gamma<2 \varepsilon-6 \delta$. Fix such a $\gamma$.

Now

$$
\begin{aligned}
\mu^{S} & \left(R_{\gamma}(S)\right) \\
\quad \geqslant & d_{S}\left(x_{1}, x_{2}\right)-2 \gamma \\
\quad & +\sum_{k=3}^{n} \bigwedge_{1 \leqslant i \leqslant j \leqslant k-1}\left[\frac{1}{2}\left(d_{S}\left(x_{k}, x_{i}\right)+d_{S}\left(x_{k}, x_{j}\right)-d_{S}\left(x_{i}, x_{j}\right)\right)-\gamma\right] \\
\quad \geqslant & \mu^{T}\left(R_{\varepsilon}(T)\right)+(2 \varepsilon-2 \delta-2 \gamma)+(n-2)(\varepsilon-3 \delta-\gamma) \\
\quad \geqslant & a+(2 \varepsilon-2 \delta-2 \gamma)+(n-2)(\varepsilon-3 \delta-\gamma)
\end{aligned}
$$

Because $\mu^{S}\left(R_{\gamma}(S)\right)$ is finite, it is apparent that $S$ cannot belong to $\mathbf{V}_{\varepsilon, a}^{\delta}$ when $a$ is sufficiently large.

Lemma 9.10. For $\varepsilon, a, \delta>0$, let $\mathbf{V}_{\varepsilon, a}$ be as in Lemma 9.9. Set $\mathbf{U}_{\varepsilon, a}:=$ $\left\{(T, \nu) \in \mathbf{T}^{\mathrm{wt}}: T \in \mathbf{V}_{\varepsilon, a}\right\}$. Then, for fixed $\varepsilon$,

$$
\lim _{a \rightarrow \infty} \operatorname{Cap}\left(\mathbf{U}_{\varepsilon, a}\right)=0
$$

Proof. Choose $\delta>0$ such that $\varepsilon>3 \delta$. Suppressing the dependence on $\varepsilon$ and $\delta$, define $u_{a}: \mathbf{T}^{\mathrm{wt}} \rightarrow[0,1]$ by

$$
u_{a}((T, \nu)):=\delta^{-1}\left(\delta-d_{\mathrm{GH}}\left(T, \mathbf{V}_{\varepsilon, a}\right)\right)_{+}
$$

Note that $u_{a}$ takes the value 1 on the open set $\mathbf{U}_{\varepsilon, a}$, and so $\operatorname{Cap}\left(\mathbf{U}_{\varepsilon, a}\right) \leqslant$ $\mathcal{E}\left(u_{a}, u_{a}\right)+\left(u_{a}, u_{a}\right)_{\mathbf{P}}$. Also, observe that

$$
\begin{aligned}
\left|u_{a}\left(\left(T^{\prime}, \nu^{\prime}\right)\right)-u_{a}\left(\left(T^{\prime \prime}, \nu^{\prime \prime}\right)\right)\right| & \leqslant \delta^{-1} d_{\mathrm{GH}}\left(T^{\prime}, T^{\prime \prime}\right) \\
& \leqslant \delta^{-1} \Delta_{\mathrm{GH}} \mathrm{wt}\left(\left(T^{\prime}, \nu^{\prime}\right),\left(T^{\prime \prime}, \nu^{\prime \prime}\right)\right)
\end{aligned}
$$

It suffices, therefore, by part (iii) of Lemma 9.5 and the dominated convergence theorem to show for each pair $\left(\left(T^{\prime}, \nu^{\prime}\right),\left(T^{\prime \prime}, \nu^{\prime \prime}\right)\right) \in \mathbf{T}^{\mathrm{wt}} \times \mathbf{T}^{\mathrm{wt}}$ that $u_{a}\left(\left(T^{\prime}, \nu^{\prime}\right)\right)-u_{a}\left(\left(T^{\prime \prime}, \nu^{\prime \prime}\right)\right)$ is 0 for $a$ sufficiently large and for each $T \in \mathbf{T}^{\mathrm{wt}}$ that $u_{a}((T, \nu))$ is 0 for $a$ sufficiently large. However, $u_{a}\left(\left(T^{\prime}, \nu^{\prime}\right)\right)-u_{a}\left(\left(T^{\prime \prime}, \nu^{\prime \prime}\right)\right) \neq 0$ implies that either $T^{\prime}$ or $T^{\prime \prime}$ belong to $\mathbf{V}_{\varepsilon, a}^{\delta}$, while $u_{a}((T, \nu)) \neq 0$ implies that $T$ belongs to $\mathbf{V}_{\varepsilon, a}^{\delta}$. The result then follows from Lemma 9.9.

Lemma 9.11. There is a sequence of compact sets $\mathbf{K}_{1} \subseteq \mathbf{K}_{2} \subseteq \ldots$ such that

$$
\lim _{n \rightarrow \infty} \operatorname{Cap}\left(\mathbf{T}^{\mathrm{wt}} \backslash \mathbf{K}_{n}\right)=0
$$

Proof. By Lemma 9.10, for $n=1,2, \ldots$ we can choose $a_{n}$ so that

$$
\operatorname{Cap}\left(\mathbf{U}_{2^{-n}, a_{n}}\right) \leqslant 2^{-n}
$$

Set

$$
\mathbf{F}_{n}:=\mathbf{T}^{\mathrm{wt}} \backslash \mathbf{U}_{2^{-n}, a_{n}}=\left\{(T, \nu) \in \mathbf{T}^{\mathrm{wt}}: \mu^{T}\left(R_{2^{-n}}(T)\right) \leqslant a_{n}\right\}
$$

and

$$
\mathbf{K}_{n}:=\bigcap_{m \geqslant n} \mathbf{F}_{n}
$$

By Proposition 4.43 and the analogue of Corollary 4.38 for unrooted trees, each set $\mathbf{K}_{n}$ is compact. By construction,

$$
\begin{aligned}
\operatorname{Cap}\left(\mathbf{T}^{\mathrm{wt}} \backslash \mathbf{K}_{n}\right) & =\operatorname{Cap}\left(\bigcup_{m \geqslant n} \mathbf{U}_{2^{-m}, a_{m}}\right) \\
& \leqslant \sum_{m \geqslant n} \operatorname{Cap}\left(\mathbf{U}_{2^{-m}, a_{m}}\right) \leqslant \sum_{m \geqslant n} 2^{-m}=2^{-(n-1)}
\end{aligned}
$$

## Summary of Dirichlet form theory

Our treatment in this appendix follows that of the standard reference [72] see also, $[106,3]$.

## A. 1 Non-negative definite symmetric bilinear forms

Let $H$ be a real Hilbert space with inner product $(\cdot, \cdot)$. We say $\mathcal{E}$ is a nonnegative definite symmetric bilinear form on $H$ with domain $\mathcal{D}(\mathcal{E})$ if

- $\mathcal{D}(\mathcal{E})$ is a dense linear subspace of $H$,
- $\mathcal{E}: \mathcal{D}(\mathcal{E}) \times \mathcal{D}(\mathcal{E}) \rightarrow \mathbb{R}$,
- $\mathcal{E}(u, v)=\mathcal{E}(v, u)$ for $u, v \in \mathcal{D}(\mathcal{E})$,
- $\mathcal{E}(a u+b v, w)=a \mathcal{E}(u, w)+b \mathcal{E}(v, w)$ for $u, v, w \in \mathcal{D}(\mathcal{E})$ and $a, b \in \mathbb{R}$,
- $\mathcal{E}(u, u) \geqslant 0$ for $u \in \mathcal{D}(\mathcal{E})$.

Given a non-negative definite symmetric bilinear form $\mathcal{E}$ on $H$ and $\alpha>0$, define another non-negative definite symmetric bilinear form $\mathcal{E}_{\alpha}$ on $H$ with domain $\mathcal{D}\left(\mathcal{E}_{\alpha}\right):=\mathcal{D}(\mathcal{E})$ by

$$
\mathcal{E}_{\alpha}(u, v):=\mathcal{E}(u, v)+\alpha(u, v), \quad u, v \in \mathcal{D}(\mathcal{E})
$$

Note that the space $\mathcal{D}(\mathcal{E})$ is a pre-Hilbert space with inner product $\mathcal{E}_{\alpha}$, and $\mathcal{E}_{\alpha}$ and $\mathcal{E}_{\beta}$ determine equivalent metrics on $\mathcal{D}(\mathcal{E})$ for different $\alpha, \beta>0$.

If $\mathcal{D}(\mathcal{E})$ is complete with respect to this metric, then $\mathcal{E}$ is said to be closed . In this case, $\mathcal{D}(\mathcal{E})$ is then a real Hilbert space with inner product $\mathcal{E}_{\alpha}$ for each $\alpha>0$.

## A. 2 Dirichlet forms

Now consider a $\sigma$-finite measure space $(X, \mathcal{B}, m)$ and take $H$ to be the Hilbert space $L^{2}(X, m)$ with the usual inner product

$$
(u, v):=\int_{X} u(x) v(x) m(d x), \quad u, v \in L^{2}(X, m)
$$

Call a non-negative definite symmetric bilinear form $\mathcal{E}$ on $L^{2}(X, m)$ Markovian if for each $\varepsilon>0$, there exists a real function $\phi_{\varepsilon}: \mathbb{R} \rightarrow \mathbb{R}$, such that

$$
\begin{aligned}
& \phi_{\varepsilon}(t)=t, \quad t \in[0,1] \\
& -\varepsilon \leqslant \phi_{\varepsilon}(t) \leqslant 1+\varepsilon, \quad t \in \mathbb{R} \\
& 0 \leqslant \phi_{\varepsilon}(t)-\phi_{\varepsilon}(s) \leqslant t-s, \quad-\infty<s<t<\infty
\end{aligned}
$$

and when $u$ belongs to $\mathcal{D}(\mathcal{E}), \phi_{\varepsilon} \circ u$ also belongs to $\mathcal{D}(\mathcal{E})$ with

$$
\mathcal{E}\left(\phi_{\varepsilon} \circ u, \phi_{\varepsilon} \circ u\right) \leqslant \mathcal{E}(u, u)
$$

A Dirichlet form is a non-negative definite symmetric bilinear form on $L^{2}(X, m)$ that is Markovian and closed.

A non-negative definite symmetric bilinear form $\mathcal{E}$ on $L^{2}(X, m)$ is certainly Markovian if whenever $u$ belongs to $\mathcal{D}(\mathcal{E})$, then $v=(0 \vee u) \wedge 1$ also belongs to $\mathcal{D}(\mathcal{E})$ and $\mathcal{E}(v, v) \leqslant \mathcal{E}(u, u)$. In this case say that the unit contraction acts on $\mathcal{E}$. It turns out the if the form is closed, then the form is Markovian if and only if the unit contraction acts on it.

Similarly, say that a function $v$ is called a normal contraction of a function $u$ if

$$
\begin{aligned}
|v(x)-v(y)| & \leqslant|u(x)-u(y)|, \quad x, y \in X, \\
|v(x)| & \leqslant|u(x)|, \quad x \in X,
\end{aligned}
$$

and say that $v \in L^{2}(X, m)$ a normal contraction of $u \in L^{2}(X, m)$ if some Borel version of $v$ is a normal contraction of some Borel version of $u$. Say that normal contractions act on $\mathcal{E}$ if whenever $v$ is a normal contraction of $u \in \mathcal{D}(\mathcal{E})$, then $v \in \mathcal{D}(\mathcal{E})$ and $\mathcal{E}(v, v) \leqslant \mathcal{E}(u, u)$. It also turns out that if the form is closed, then the form is Markovian if and only if the unit contraction acts on it.

Example A.1. Let $X \subseteq \mathbb{R}$ be an open subinterval and suppose that $m$ is a Radon measures on $X$ with support all of $X$. Define a non-negative definite symmetric bilinear form by

$$
\mathcal{E}(u, v):=\frac{1}{2} \int_{X} \frac{d u(x)}{d x} \frac{d v(x)}{d x} d x
$$

on the domain

$$
\mathcal{D}(\mathcal{E}):=\left\{u \in L^{2}(X, m): u \text { is absolutely continuous and } \mathcal{E}(u, u)<\infty\right\}
$$

We claim that $\mathcal{E}$ is a Dirichlet form on $L^{2}(X, m)$.

It is easy to check that the unit contraction acts on $\mathcal{E}$. To show the form is closed, take any $\mathcal{E}_{1}$-Cauchy sequence $\left\{u_{\ell}\right\}$. Then $\left\{d u_{\ell} / d x\right\}$ converges to some $f \in L^{2}(X, d x)$ in $L^{2}(X, d x)$. Also, $\left\{u_{\ell}\right\}$ converges to some $u \in L^{2}(X, m)$ in $L^{2}(X, m)$. From this and the inequality

$$
|u(a)-u(b)|^{2} \leqslant 2|a-b| \mathcal{E}(u, u), \quad a, b \in X
$$

we conclude that there is a subsequence $\left\{\ell_{k}\right\}$ such that $u_{\ell_{k}}$ converges to a continuous function $\tilde{u}$ uniformly on each bounded closed subinterval of $X$. Obviously $\tilde{u}=u$ m-a.e. For all infinitely differentiable compactly supported functions $\phi$ on $X$, an integration by parts shows that

$$
\begin{aligned}
& \int_{X} f(x) \phi(x) d x=\lim _{\ell_{k} \rightarrow \infty} \int_{X} \frac{d u_{\ell_{k}}(x)}{d x} \phi(x) d x \\
& \quad=-\lim _{\ell_{k} \rightarrow \infty} \int_{X} u_{\ell_{k}}(x) \phi^{\prime}(x) d x=-\int_{X} \tilde{u}(x) \phi^{\prime}(x) d x
\end{aligned}
$$

This implies that $\tilde{u}$ is absolutely continuous and $d \tilde{u} / d x=f$. Hence, $\tilde{u} \in \mathcal{D}(\mathcal{E})$ and $\left\{u_{\ell}\right\}$ is $\mathcal{E}_{1}$-convergent to $\tilde{u}$.

Example A.2. Consider a locally compact metric space ( $X, \rho$ ) equipped with a Radon measure $m$. Suppose that we are given a kernel $j$ on $X \times \mathcal{B}(X)$ satisfying the following conditions.

- For any $\varepsilon>0, j\left(x, X \backslash B_{\varepsilon}(x)\right)$ is, as a function of $x \in X$, locally integrable with respect to $m$. Here, as usual, $B_{\varepsilon}(x)$ is the ball around $x$ of radius $\varepsilon$.
- $\int_{X} u(x)(j v)(x) m(d x)=\int_{X}(j u)(x) v(x) m(d x)$ for all $u, v \in p \mathcal{B}(X)$.

Then, $j$ determines a symmetric Radon measure $J$ on $X \times X \backslash \Delta$, where $\Delta$ is the diagonal, by

$$
\int_{X \times X \backslash \Delta} f(x, y) J(d x, d y):=\int_{X}\left\{\int_{X} f(x, y) j(x, d y)\right\} m(d x)
$$

Put

$$
\mathcal{E}(u, v):=\int_{X \times X \backslash \Delta}(u(x)-u(y))(v(x)-v(y)) J(d x, d y)
$$

on the domain

$$
\mathcal{D}(\mathcal{E}):=\left\{u \in L^{2}(X, m): \mathcal{E}(u, u)<\infty\right\} .
$$

We claim that $\mathcal{E}$ is a Dirichlet form on $L^{2}(X, m)$ provided that $\mathcal{D}(\mathcal{E})$ is dense in $L^{2}(X, m)$.

It is clear that $\mathcal{E}$ is non-negative definite, symmetric, and bilinear. We next show that for a Borel function $u$ that $u=0 m$-a.e. implies that $\mathcal{E}(u, u)=0$. Suppose that $u=0 m$-a.e. Put $\Gamma_{K, \varepsilon}=\{(x, y) \in K \times K: \rho(x, y)>\varepsilon\}$ for $\varepsilon>0$ and $K$ compact. Then

$$
\begin{aligned}
\int_{\Gamma_{K, \varepsilon}} & (u(x)-u(y))^{2} J(d x, d y) \leqslant 2 \int_{\Gamma_{K, \varepsilon}}\left(u(x)^{2}+u(y)^{2}\right) J(d x, d y) \\
& =4 \int_{\Gamma_{K, \varepsilon}} u(x)^{2} J(d x, d y) \leqslant 4 \int_{K} u(x)^{2} j\left(x, X \backslash B_{\varepsilon}(x)\right) m(d x)=0 .
\end{aligned}
$$

Letting $\varepsilon \downarrow 0$ and $K \uparrow X$ gives $\mathcal{E}(u, u)=0$.
It is clear that every normal contraction operates on the form and so the form is Markovian. To prove that the form is closed, consider a sequence $\left\{u_{\ell}\right\}$ in $\mathcal{D}(\mathcal{E})$ such that $\lim _{\ell, m \rightarrow \infty} \mathcal{E}_{1}\left(u_{\ell}-u_{m}, u_{\ell}-u_{m}\right) \rightarrow 0$. Since $\left\{u_{\ell}\right\}$ converges in $L^{2}(X, m)$, there is a subsequence $\left\{\ell_{k}\right\}$ and a set $N \in \mathcal{B}(X)$ with $m(N)=0$ such that $\left\{u_{\ell_{k}}(x)\right\}$ converges on $X \backslash N$. Put $\tilde{u}_{\ell_{k}}(x)=u_{l_{k}}(x)$ on $X \backslash N$ and $\tilde{u}_{\ell_{k}}(x)=0$ on $N$. Then $\tilde{u}_{\ell_{k}}(x)$ has a limit $u(x)$ everywhere and $u_{\ell}$ converges to $u$ in $L^{2}(X, m)$. Moreover,

$$
\begin{aligned}
& \mathcal{E}\left(u-u_{m}, u-u_{m}\right) \\
& \quad=\int_{X \times X \backslash \Delta} \lim _{\ell_{k} \rightarrow \infty}\left\{\left(u_{\ell_{k}}(x)-u_{\ell_{k}}(y)\right)-\left(u_{m}(x)-u_{m}(y)\right)\right\}^{2} J(d x, d y) \\
& \quad \leqslant \liminf _{l_{k} \rightarrow \infty} \mathcal{E}\left(u_{l_{k}}-u_{m}, u_{l_{k}}-u_{m}\right)
\end{aligned}
$$

The last term can be made arbitrarily small for sufficiently large $m$. Thus, $u_{m}$ is $\mathcal{E}_{1}$-convergent to $u \in \mathcal{D}(\mathcal{E})$, as required.

## A. 3 Semigroups and resolvents

Suppose again that we have a real Hilbert space $H$ with inner product $(\cdot, \cdot)$. Consider a family $\left\{T_{t}\right\}_{t>0}$ of linear operators on $H$ satisfying the following conditions:

- each $T_{t}$ is a self-adjoint operator with domain $H$,
- $T_{s} T_{t}=T_{s+t}, s, t>0$ (that is, $\left\{T_{t}\right\}_{t>0}$ is a semigroup),
- $\left(T_{t} u, T_{t} u\right) \leqslant(u, u), t>0, u \in H$ (that is, each $T_{t}$ is a contraction).

We say that $\left\{T_{t}\right\}_{t>0}$ is strongly continuous if, in addition,

- $\lim _{t \downarrow 0}\left(T_{t} u-u, T_{t} u-u\right)=0$ for all $u \in H$.

A resolvent on $H$ is a family $\left\{G_{\alpha}\right\}_{\alpha>0}$ of linear operators on $H$ satisfying the following conditions:

- $G_{\alpha}$ is a self-adjoint operator with domain $H$,
- $G_{\alpha}-G_{\beta}+(\alpha-\beta) G_{\alpha} G_{\beta}=0$ (the resolvent equation),
- each operator $\alpha G_{\alpha}$ is a contraction.

The resolvent is said to be strongly continuous if, in addition,

- $\lim _{\alpha \rightarrow \infty}\left(\alpha G_{\alpha} u-u, \alpha G_{\alpha} u-u\right)=0$ for all $u \in H$.

Example A.3. Given a strongly continuous semigroup $\left\{T_{t}\right\}_{t>0}$ on $H$, the family of operators

$$
G_{\alpha} u:=\int_{0}^{\infty} e^{-\alpha t} T_{t} u d t
$$

is a strongly continuous resolvent on $H$ called the resolvent of the given semigroup. The semigroup may be recovered from the resolvent via the Yosida approximation

$$
T_{t} u=\lim _{\beta \rightarrow \infty} e^{-t \beta} \sum_{n=0}^{\infty} \frac{(t \beta)^{n}}{n!}\left(\beta G_{\beta}\right)^{n} u, \quad u \in H
$$

## A. 4 Generators

The generator $A$ of a strongly continuous semigroup $\left\{T_{t}\right\}_{t>0}$ on $H$ is defined by

$$
A u:=\lim _{t \downarrow 0} \frac{T_{t} u-u}{t}
$$

on the domain $\mathcal{D}(A)$ consisting of those $u \in H$ such that the limit exists.
Suppose that $\left\{G_{\alpha}\right\}_{\alpha>0}$ is a strongly continuous resolvent on $H$. Note that if $G_{\alpha} u=0$, then, by the resolvent equation, $G_{\beta} u=0$ for all $\beta>0$, and, by strong continuity, $u=\lim _{\beta \rightarrow \infty} \beta G_{\beta} u=0$. Thus, the operator $G_{\alpha}$ is invertible and we can set

$$
A u:=\alpha u-G_{\alpha}^{-1} u
$$

on the domain $\mathcal{D}(A):=G_{\alpha}(H)$. This operator $A$ is easily seen to be independent of $\alpha>0$ and is called the generator of the resolvent. $\left\{G_{\alpha}\right\}_{\alpha>0}$.

Lemma A.4. The generator of a strongly continuous semigroup on $H$ coincides with the generator of its resolvent, and the generator is a non-positive definite self-adjoint operator.

## A. 5 Spectral theory

A self-adjoint operator $S$ on $H$ with domain $H$ satisfying $S^{2}=S$ is called a projection. A family $\left\{E_{\lambda}\right\}_{\lambda \in \mathbb{R}}$ of projection operators on $H$ is called a spectral family if

$$
\begin{aligned}
& E_{\lambda} E_{\mu}=E_{\lambda}, \quad \lambda \leqslant \mu \\
& \lim _{\lambda^{\prime} \downarrow \lambda} E_{\lambda^{\prime}} u=E_{\lambda} u, \quad u \in H \\
& \lim _{\lambda \rightarrow-\infty} E_{\lambda} u=0, \quad u \in H \\
& \lim _{\lambda \rightarrow \infty} E_{\lambda} u=u, \quad u \in H
\end{aligned}
$$

Note that $0 \leqslant\left(E_{\lambda} u, u\right) \uparrow(u, u)$ as $\lambda \uparrow \infty$, for $u \in H$, and, by polarization, $\lambda \mapsto\left(E_{\lambda} u, v\right)$ is a function of bounded variation for $u, v \in H$.

Suppose we are given a spectral family $\left\{E_{\lambda}\right\}_{\lambda \in \mathbb{R}}$ on $H$ and a continuous function $\phi(\lambda)$ on $\mathbb{R}$. We can then define a self-adjoint operator $A$ on $H$, denoted by $\int_{-\infty}^{\infty} \phi(\lambda) d E_{\lambda}$, by requiring that

$$
(A u, v)=\int_{-\infty}^{\infty} \phi(\lambda) d\left(E_{\lambda} u, v\right), \quad \forall v \in H
$$

where the domain of $A$ is $\mathcal{D}(A):=\left\{u \in H: \int_{-\infty}^{\infty} \phi(\lambda) d\left(E_{\lambda} u, u\right)<\infty\right\}$.
Conversely, given a self-adjoint operator $A$ on $H$, there exists a unique spectral family $\left\{E_{\lambda}\right\}_{\lambda \in \mathbb{R}}$ such that $A=\int_{-\infty}^{\infty} \lambda d E_{\lambda}$. This is called the spectral representation of $A$. If $A$ is non-negative definite, then the corresponding spectral family satisfies $E_{\lambda}=0$ for $\lambda<0$.

Let $-A$ be a non-negative definite self-adjoint operator on $H$ and let $-A=\int_{0}^{\infty} \lambda d E_{\lambda}$ be its spectral representation. For any non-negative continuous function $\phi$ on $\mathbb{R}_{+}$, we define the self-adjoint operator $\phi(-A)$ by $\phi(-A):=\int_{0}^{\infty} \phi(\lambda) d E_{\lambda}$. Note that $\phi(-A)$ is again non-negative definite.

## A. 6 Dirichlet form, generator, semigroup, resolvent correspondence

Lemma A.5. Let $-A$ be a non-negative definite self-adjoint operator on $H$. The family $\left\{T_{t}\right\}_{t>0}:=\{\exp (t A)\}_{t>0}$ is a strongly continuous semigroup, and the family $\left\{G_{\alpha}\right\}_{\alpha>0}:=\left\{(\alpha-A)^{-1}\right\}_{\alpha>0}$ is a strongly continuous resolvent. The generator of $\left\{T_{t}\right\}_{t>0}$ is $A$ and $\left\{T_{t}\right\}_{t>0}$ is the unique strongly continuous semigroup with generator A. A similar statement holds for the resolvent.

Theorem A.6. There is a bijective correspondence between the family of closed non-negative definite symmetric bilinear forms $\mathcal{E}$ on $H$ and the family of non-positive definite self-adjoint operators $A$ on $H$. The correspondence is given by

$$
\mathcal{D}(\mathcal{E})=\mathcal{D}(\sqrt{-A})
$$

and

$$
\mathcal{E}(u, v)=(\sqrt{-A} u, \sqrt{-A} v)
$$

Consider a $\sigma$-finite measure space $(X, \mathcal{B}, m)$. A linear operator $S$ on $L^{2}(X, m)$ with domain $L^{2}(X, m)$ is Markovian if $0 \leqslant S u \leqslant 1 m$-a.e. whenever $u \in L^{2}(X, m)$ and $0 \leqslant u \leqslant 1 m$-a.e.

Theorem A.7. Let $\mathcal{E}$ be a closed non-negative definite symmetric bilinear form on $L^{2}(X, m)$. Write $\left\{T_{t}\right\}_{t>0}$ and $\left\{G_{\alpha}\right\}_{\alpha>0}$ for the corresponding strongly continuous semigroup and the strongly continuous resolvent on $L^{2}(X, m)$. The following five conditions are equivalent.
(a) $T_{t}$ is Markovian for each $t>0$.
(b) $\alpha G_{\alpha}$ is Markovian for each $\alpha>0$.
(c) $\mathcal{E}$ is Markovian.
(d) The unit contraction operates on $\mathcal{E}$.
(e) Normal contractions operate on $\mathcal{E}$.

## A. 7 Capacities

Suppose that $X$ is a Lusin space and $m$ is a Radon measure. There is a set function associated with a Dirichlet form $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ on $L^{2}(X, m)$ called the (1)-capacity and denoted by Cap. If $U \subseteq X$ is open, then

$$
\operatorname{Cap}(U):=\inf \left\{\mathcal{E}_{1}(f, f): f \in \mathcal{D}(\mathcal{E}), f(x) \geqslant 1, m-\text { a.e. } x \in U\right\}
$$

More generally, if $V \subseteq X$ is an arbitrary subset, then

$$
\operatorname{Cap}(V):=\inf \{\operatorname{Cap}(U): V \subseteq U, U \text { is open }\}
$$

The set function Cap is a Choquet capacity.
We say that some property holds quasi-everywhere or, equivalently, for quasi-every $x \in X$, if the set $x \in X$ where the property fails to hold has capacity 0 . We abbreviate this by saying that the property holds q.e. or for q.e. every $x \in X$.

## A. 8 Dirichlet forms and Hunt processes

A Hunt process is a strong Markov process

$$
\mathbf{X}=\left(\Omega, \mathcal{F},\left\{\mathcal{F}_{t}\right\}_{t \geqslant 0},\left\{\mathbb{P}^{x}\right\}_{x \in E},\left\{X_{t}\right\}_{t \geqslant 0}\right)
$$

on a Lusin state space $E$ that has right-continuous, left-limited sample paths and is also quasi-left-continuous. Write $\left\{P_{t}\right\}_{t \geqslant 0}$ for the transition semigroup of $\mathbf{X}$. That is, $P_{t} f(x)=\mathbb{P}^{x}\left[f\left(X_{t}\right)\right]$ for $f \in b \mathcal{B}(E)$. If $\mu$ is a Radon measure on $(E, \mathcal{B}(E))$, we say that $\mathbf{X}$ is $\mu$-symmetric if $\int_{E} f(x) P_{t} g(x) \mu(d x)=$ $\int_{E} P_{t} f(x) g(x) \mu(d x)$ for all $f, g \in b \mathcal{B}(E)$. Intuitively, if the process $\mathbf{X}$ is started according to the initial "distribution" $\mu$, then reversing the direction of time leaves finite-dimensional distributions unchanged.

Theorem A.8. Let $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ be a Dirichlet form on $L^{2}(E, \mu)$, where $E$ is Lusin and $\mu$ is Radon. Write $\left\{T_{t}\right\}_{t>0}$ for the associated strongly continuous contraction semigroup of Markovian operators. Suppose that there exists a collection $\mathcal{C} \subseteq L^{2}(E, \mu)$ and a sequence of compact sets $K_{1} \subseteq K_{2} \subseteq \ldots$ such that:
(a) $\mathcal{C}$ is a countably generated subalgebra of $\mathcal{D}(\mathcal{E}) \cap b C(E)$,
(b) $\mathcal{C}$ is $\mathcal{E}_{1}$-dense in $\mathcal{D}(\mathcal{E})$,
(c) $\mathcal{C}$ separates points of $E$ and, for any $x \in E$, there is an $f \in \mathcal{C}$ such that $f(x) \neq 0$,
(d) $\lim _{n \rightarrow \infty} \operatorname{Cap}\left(E \backslash K_{n}\right)=0$.

Then there is a $\mu$-symmetric Hunt process $\mathbf{X}$ on $E$ with transition semigroup $\left\{P_{t}\right\}_{t \geqslant 0}$ such that $P_{t} f(x)=T_{t} f(x)$ for $f \in b \mathcal{B}(E) \cap L^{2}(E, \mu)$.

Remark A.9. The theory in [72] for symmetric Hunt processes associated with Dirichlet forms is developed under the hypothesis that the state space is locally compact. However, the embedding results outlined in Section 7.3 of [72], shows that the results developed under the hypothesis of local compactness still holds if the state space is Lusin and the hypotheses of Theorem A. 8 hold.

Lemma A.10. Suppose that $\mathbf{X}$ is the $\mu$-symmetric Hunt process constructed from a Dirichlet form $(\mathcal{E}, \mathcal{D}(\mathcal{E}))$ satisfying the conditions of Theorem A. 8 and $B \in \mathcal{B}(E)$. Then $\mathbb{P}^{x}\left\{\exists t>0: X_{t} \in B\right\}=0$ for $\mu$-a.e. $x \in E$ if and only if $\operatorname{Cap}(B)=0$.

## Some fractal notions

This appendix is devoted to recalling briefly some definitions about various ways of assigning sizes and dimensions to metric spaces and then applying this theory to the ultrametric completions of $\mathbb{N}$ obtained in Example 3.41 from the $\mathbb{R}$-tree associated with a non-increasing family of partitions of $\mathbb{N}$.

## B. 1 Hausdorff and packing dimensions

Let $(\mathcal{X}, \rho)$ be a compact metric space. Given a set $A \subseteq \mathcal{X}$ and $\epsilon>0$, a countable collection of balls $\left\{B_{i}\right\}$ is said to be an $\epsilon$-covering of $A$ if $A \subseteq \bigcup_{i} B_{i}$ and each ball has diameter at most $\epsilon$. Note that if $\epsilon^{\prime}<\epsilon^{\prime \prime}$, then an $\epsilon^{\prime}$-covering of $A$ is also an $\epsilon^{\prime \prime}$-covering. For $\alpha>0$, the $\alpha$-dimensional Hausdorff measure on $\mathcal{X}$ is the Borel measure that assigns mass

$$
\mathcal{H}^{\alpha}(A):=\sup _{\epsilon>0} \inf \left\{\sum_{i} \operatorname{diam}\left(B_{i}\right)^{\alpha}:\left\{B_{i}\right\} \text { is an } \epsilon \text {-covering of } A\right\}
$$

to a Borel set $A$. The Hausdorff dimension of $A$ is the infimum of those $\alpha$ such that the corresponding $\alpha$-dimensional Hausdorff measure is zero.

A countable collection of balls $\left\{B_{i}\right\}$ is said to be an $\epsilon$-packing of a set $A \subseteq \mathcal{X}$ if the balls are disjoint, the center of each ball belongs to $A$, each ball has diameter at most $\epsilon$. Note that if $\epsilon^{\prime}<\epsilon^{\prime \prime}$, then an $\epsilon^{\prime \prime}$-packing of $A$ is also an $\epsilon^{\prime}$-packing. For $\alpha>0$, the $\alpha$-dimensional packing pre-measure on $\mathcal{X}$ assigns mass

$$
P^{\alpha}(A):=\inf _{\epsilon>0} \sup \left\{\sum_{i} \operatorname{diam}\left(B_{i}\right)^{\alpha}:\left\{B_{i}\right\} \text { is an } \epsilon \text {-packing of } A\right\}
$$

to a set $A$. The $\alpha$-dimensional packing measure on $\mathcal{X}$ is the Borel measure that assigns mass

$$
\mathcal{P}^{\alpha}(A):=\inf \left\{\sum_{i} P^{\alpha}\left(A_{i}\right): A \subseteq \bigcup_{i} A_{i}\right\}
$$

to a Borel set $A$ where the infimum is over all countable collections of Borel sets $\left\{A_{i}\right\}$ such that $A \subseteq \bigcup_{i} A_{i}$. The packing dimension of $A$ is the infimum of those $\alpha$ such that the corresponding $\alpha$-dimensional packing measure is zero.

Theorem B.1. The packing dimension of a set is always at least as great as its Hausdorff dimension.

We refer the reader to [107] for more about and properties of Hausdorff and packing dimension.

## B. 2 Energy and capacity

Let $(\mathcal{X}, \rho)$ be a compact metric space. Write $M_{1}(\mathcal{X})$ for the collection of Borel probability measures on $\mathcal{X}$. A gauge is a function $f:[0, \infty[\rightarrow[0, \infty]$, such that:

- $\quad f$ is continuous and non-increasing,
- $f(0)=\infty$,
- $f(r)<\infty$ for $r>0$,
- $\lim _{r \rightarrow \infty} f(r)=0$.

Given $\mu \in M_{1}(\mathcal{X})$ and a gauge $f$, the energy of $\mu$ in the gauge $f$ is the quantity

$$
\mathcal{E}_{f}(\mu):=\int \mu(d x) \int \mu(d y) f(\rho(x, y))
$$

The capacity of $\mathcal{X}$ in the gauge $f$ is the quantity

$$
\operatorname{Cap}_{f}(\mathcal{X}):=\left(\inf \left\{\mathcal{E}_{f}(\mu): \mu \in M_{1}(\mathcal{X})\right\}\right)^{-1}
$$

(note by our assumptions on $f$ that we need only consider diffuse $\mu \in M_{1}(\mathcal{X})$ in the infimum).

The capacity dimension of $\mathcal{X}$ is the supremum of those $\alpha>0$ such that $\mathcal{X}$ has strictly positive capacity in the gauge $f(x)=x^{-\alpha}$ (where we adopt the convention that the supremum of the empty set is 0 ).

Theorem B.2. The Hausdorff and capacity dimensions of a compact metric space always coincide.

We again refer to [107] for more about capacities and their connection to Hausdorff dimension.

## B. 3 Application to trees from coalescing partitions

Recall the construction in Example 3.41 of a $\mathbb{R}$-tree and an associated ultrametric completion $(\mathbb{S}, \delta)$ of $\mathbb{N}$ from a coalescing family $\{\Pi(t)\}_{t>0}$ of partitions of $\mathbb{N}$. We will assume that $\Pi(t)$ has finitely many blocks for $t>0$, so that $(\mathbb{S}, \delta)$ is compact.

Write $N(t)$ for the number of blocks of $\Pi(t)$ and for $k \in \mathbb{N}$ put $\sigma_{k}:=$ $\inf \{t \geqslant 0: N(t) \leqslant k\}$. The non-increasing function $\Pi$ is constant on each of the intervals $\left[\sigma_{k}, \sigma_{k-1}\left[, k>1\right.\right.$. Write $1=I_{1}(t)<\cdots<I_{N(t)}(t)$ for an ordered listing of the least elements of the various blocks of $\Pi(t)$.

We can associate each partition $\Pi(t)$ with an equivalence relation $\sim_{\Pi(t)}$ on $\mathbb{N}$ by declaring that $i \sim_{\Pi(t)} j$ if $i$ and $j$ are in the same block of $\Pi(t)$.

Given $B \subseteq \mathbb{S}$, write $\mathrm{cl} B$ for the closure of $B$. Each of the sets

$$
\begin{aligned}
U_{i}(t) & =\operatorname{cl}\left\{j \in \mathbb{N}: j \sim_{\Pi(t)} I_{i}(t)\right\} \\
& =\operatorname{cl}\left\{j \in \mathbb{N}: \delta\left(j, I_{i}(t)\right) \leqslant 2 t\right\} \\
& =\left\{y \in \mathbb{S}: \delta\left(y, I_{i}(t)\right) \leqslant 2 t\right\}
\end{aligned}
$$

is a closed ball with diameter at most $t$ (in an ultrametric space, the diameter and radius of a ball are equal). The closed balls of $\mathbb{S}$ are also the open balls and every ball is of the form $U_{i}(t)$ for some $t>0-$ see, for example, Proposition 18.4 of [123] - and, in fact, every ball is of the form $U_{i}\left(\sigma_{k}\right)$ for some $k \in \mathbb{N}$ and $1 \leqslant i \leqslant k$. In particular, the collection of balls is countable. Any ball of diameter at most $2 t$ is contained in a unique one of the $U_{i}(t)$, and any ball of diameter at least $2 t$ contains one or more of the $U_{i}(t)-s e e$, for example, Proposition 18.5 of [123].

We need to adapt to our setting the alternative expression for energy obtained by summation-by-parts in Section 2 of [112]. For $t>0$ write $\mathcal{U}(t)$ for the collection of balls $\left\{U_{1}(t), \ldots, U_{N(t)}(t)\right\}$. Let $\mathcal{U}$ denote the union of these collections over all $t>0$, so that $\mathcal{U}$ is just the countable collection of all balls of $\mathbb{S}$. Given $U \in \mathcal{U}$ with $U \neq \mathbb{S}$, let $U \rightarrow$ denote the unique element of $\mathcal{U}$ such that there exists no $V \in \mathcal{U}$ with $U \subsetneq V \subsetneq U \rightarrow$. More concretely, such a ball $U$ is in $\mathcal{U}\left(\sigma_{k}\right)$ but not in $\mathcal{U}\left(\sigma_{k-1}\right)$ for some unique $k>1$, and $U^{\rightarrow}$ is the unique element of $\mathcal{U}\left(\sigma_{k-1}\right)$ such that $U \subset U \rightarrow$. Define $\mathbb{S} \rightarrow:=\dagger$, where $\dagger$ is an adjoined symbol. Put $\operatorname{diam}(\dagger)=\infty$.

Given a gauge $f$, write $\varphi_{f}$ for the diffuse measure on $[0, \infty[$ such that $\varphi_{f}\left(\left[r, \infty[)=\varphi_{f}(] r, \infty[)=f(r), r \geqslant 0\right.\right.$. For a diffuse probability measure $\mu \in M_{1}(\mathbb{S})$ we have, with the convention $f(\infty)=0$,

$$
\begin{align*}
\mathcal{E}_{f}(\mu)= & \int \mu(d x) \int \mu(d y) f(\delta(x, y)) \\
= & \int \mu(d x) \int \mu(d y) \sum_{U \in \mathcal{U},\{x, y\} \subseteq U} f(\operatorname{diam}(U))-f\left(\operatorname{diam}\left(U^{\rightarrow}\right)\right) \\
= & \sum_{U \in \mathcal{U}}\left(f(\operatorname{diam}(U))-f\left(\operatorname{diam}\left(U^{\rightarrow}\right)\right)\right) \\
& \times \int \mu(d x) \int \mu(d y) \mathbf{1}\{\{x, y\} \subseteq U\}  \tag{B.1}\\
= & \sum_{U \in \mathcal{U}}\left(f(\operatorname{diam}(U))-f\left(\operatorname{diam}\left(U^{\rightarrow}\right)\right)\right) \mu(U)^{2} \\
= & \sum_{U \in \mathcal{U}} \int_{[0, \infty[ } \varphi_{f}(d t) \mathbf{1}\{U \in \mathcal{U}(t)\} \mu(U)^{2} \\
= & \int_{[0, \infty[ } \varphi_{f}(d t) \sum_{U \in \mathcal{U}(t)} \mu(U)^{2}
\end{align*}
$$

Proposition B.3. Suppose for all $t>0$ that the asymptotic block frequencies

$$
F_{i}(t):=\lim _{n \rightarrow \infty} n^{-1}\left|\left\{0 \leqslant j \leqslant n-1: j \sim_{\Pi(t)} I_{i}(t)\right\}\right|, 1 \leqslant i \leqslant N(t)
$$

exist and

$$
F_{1}(t)+\cdots+F_{N(t)}(t)=1
$$

Suppose also that for some $\alpha>0$ that

$$
0<\liminf _{t \downarrow 0} t^{\alpha} N(t) \leqslant \limsup _{t \downarrow 0} t^{\alpha} N(t)<\infty
$$

and

$$
0<\liminf _{t \downarrow 0} t^{-\alpha} \sum_{i=1}^{N(t)} F_{i}(t)^{2} \leqslant \limsup _{t \downarrow 0} t^{-\alpha} \sum_{i=1}^{N(t)} F_{i}(t)^{2}<\infty .
$$

Then the Hausdorff and packing dimensions of $\mathbb{S}$ are both $\alpha$ and there are constants $0<c^{\prime} \leqslant c^{\prime \prime}<\infty$ such that for any gauge $f$

$$
c^{\prime}\left(\int_{0}^{1} f(t) t^{\alpha-1} d t\right)^{-1} \leqslant \operatorname{Cap}_{f}(\mathbb{S}) \leqslant c^{\prime \prime}\left(\int_{0}^{1} f(t) t^{\alpha-1} d t\right)^{-1}
$$

Proof. In order to establish that both the Hausdorff and packing dimensions of $\mathbb{S}$ are at most $\alpha$ it suffices to consider the packing dimension, because packing dimension always dominates Hausdorff dimension.

By definition of packing dimension, in order to establish that the packing dimension is at most $\alpha$ it suffices to show for each $\eta>\alpha$ that there is a constant $c<\infty$ such that for any packing $B_{1}, B_{2}, \ldots$ of $\mathbb{S}$ with balls of diameter at most 1, we have $\sum_{k} \operatorname{diam}\left(B_{k}\right)^{\eta} \leqslant c$. If $2 \cdot 2^{-p} \leqslant \operatorname{diam}\left(B_{k}\right)<2 \cdot 2^{-(p-1)}$ for some $p \in\{0,1,2, \ldots\}$, then $B_{k}$ contains one or more of the balls $U_{i}\left(2^{-p}\right)$. Thus

$$
\left|\left\{k \in \mathbb{N}: 2 \cdot 2^{-p} \leqslant \operatorname{diam}\left(B_{k}\right)<2 \cdot 2^{-(p-1)}\right\}\right| \leqslant N\left(2^{-p}\right)
$$

and

$$
\sum_{k} \operatorname{diam}\left(B_{k}\right)^{\eta} \leqslant \sum_{p=0}^{\infty} N\left(2^{-p}\right) 2^{-(p-1) \eta}<\infty
$$

as required.
If we establish the claim regarding capacities, then this will establish that the capacity dimension of $\mathbb{S}$ is $\alpha$. This then gives the required lower bound on the packing and Hausdorff dimensions because the Hausdorff measure equals the capacity dimension and the packing dimension dominates Hausdorff dimension.

In order to establish the claimed lower bound on $\operatorname{Cap}_{f}(\mathbb{S})$ it appears, a priori, that for each gauge $f$ we might need to find a probability measure $\mu$ depending on $f$ such that $\left(\mathcal{E}_{f}(\mu)\right)^{-1}$ is at least the left-hand side of the inequality. It turns out, however, that we can find a measure that works simultaneously for all gauges $f$. We construct this measure as follows.

Let $\mathcal{A}$ denote the algebra of subsets of $\mathbb{S}$ generated by the collection of balls $\mathcal{U}$. Thus, $\mathcal{A}$ is just the countable collection of finite unions of balls. The $\sigma$-algebra generated by $\mathcal{A}$ is the Borel $\sigma$-algebra of $\mathbb{S}$. The sets in $\mathcal{A}$ are compact, and, moreover, for all $k \in \mathbb{N}$ and indices $1 \leqslant$ $i \leqslant k$ if $U_{i}\left(\sigma_{k}\right)=U_{i_{1}}\left(\sigma_{k+1}\right) \cup U_{i_{2}}\left(\sigma_{k+1}\right) \cup \cdots \cup U_{i_{m}}\left(\sigma_{k+1}\right)$ (that is, if $\left.\left\{I_{i_{1}}\left(\sigma_{k+1}\right), I_{i_{2}}\left(\sigma_{k+1}\right), \ldots, I_{i_{m}}\left(\sigma_{k+1}\right)\right\}=\left\{I_{\ell}\left(\sigma_{k+1}\right): I_{\ell}\left(\sigma_{k+1}\right) \sim_{\Pi\left(\sigma_{k}\right)} I_{i}\left(\sigma_{k}\right)\right\}\right)$, then $F_{i}\left(\sigma_{k}\right)=F_{i_{1}}\left(\sigma_{k+1}\right)+F_{i_{2}}\left(\sigma_{k+1}\right)+\cdots+F_{i_{m}}\left(\sigma_{k+1}\right)$. It is, therefore, possible to define a finitely additive set function $\nu$ on $\mathcal{A}$ such that

$$
\begin{equation*}
\nu\left(U_{i}(t)\right)=F_{i}(t), t>0,1 \leqslant i \leqslant N(t) \tag{B.2}
\end{equation*}
$$

and

$$
\begin{equation*}
\nu(\mathbb{S})=1 \tag{B.3}
\end{equation*}
$$

Furthermore, if $A_{1} \supseteq A_{2} \supseteq \ldots$ is a decreasing sequence of sets in the algebra $\mathcal{A}$ such that $\bigcap_{n} A_{n}=\varnothing$, then, by compactness, $A_{n}=\varnothing$ for all $n$ sufficiently large and it is certainly the case that $\lim _{n \rightarrow \infty} \nu\left(A_{n}\right)=0$. A standard extension theorem - see, for example, Theorems 3.1.1 and 3.1.4 of [48] - gives that the set function $\nu$ extends to a probability measure (also denoted by $\nu$ ) on the Borel $\sigma$-algebra of $\mathbb{S}$.

From (B.1) we see that for some constant $0<c^{\prime}<\infty$ (not depending on f) we have

$$
\begin{aligned}
\operatorname{Cap}_{f}(\mathbb{S}) & \geqslant\left(\mathcal{E}_{f}(\nu)\right)^{-1}=\left(\int \varphi_{f}(d t) \sum_{U \in \mathcal{U}(t)} \nu(U)^{2}\right)^{-1} \\
& =\left(\int \varphi_{f}(d t) \sum_{i=1}^{N(t)} F_{i}(t)^{2}\right)^{-1} \\
& \geqslant c^{\prime}\left(\int \varphi_{f}(d t)(t \wedge 1)^{\alpha}\right)^{-1}=c^{\prime}\left(\int_{0}^{1} f(t) t^{\alpha-1} d t\right)^{-1}
\end{aligned}
$$

Turning to the upper bound on $\operatorname{Cap}_{f}(\mathbb{S})$, note from the Cauchy-Schwarz inequality that for any $\mu \in M_{1}(\mathbb{S})$

$$
1=\left(\sum_{U \in \mathcal{U}(t)} \mu(U)\right)^{2} \leqslant N(t) \sum_{U \in \mathcal{U}(t)} \mu(U)^{2}
$$

and so, by (B.1),

$$
\begin{aligned}
\operatorname{Cap}_{f}(\mathbb{S}) & \leqslant\left(\int \varphi_{f}(d t) N(t)^{-1}\right)^{-1} \\
& \leqslant c^{\prime \prime}\left(\int \varphi_{f}(d t)(t \wedge 1)^{\alpha}\right)^{-1}=c^{\prime \prime}\left(\int_{0}^{1} f(t) t^{\alpha-1} d t\right)^{-1}
\end{aligned}
$$

for some constant $0<c^{\prime \prime}<\infty$.
Remark B.4. By the Cauchy-Schwarz inequality,

$$
1=\left(\sum_{i=1}^{N(t)} F_{i}(t)\right)^{2} \leqslant\left(\sum_{i=1}^{N(t)} F_{i}(t)^{2}\right) N(t)
$$

Thus,

$$
\limsup _{t \downarrow 0} t^{\alpha} N(t)<\infty \Longrightarrow \liminf _{t \downarrow 0} t^{-\alpha} \sum_{i=1}^{N(t)} F_{i}(t)^{2}>0
$$

and

$$
\limsup _{t \downarrow 0} t^{-\alpha} \sum_{i=1}^{N(t)} F_{i}(t)^{2}<\infty \Longrightarrow \liminf _{t \downarrow 0} t^{\alpha} N(t)>0
$$

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| Erwan SAINT LOUBERT BIÉ | Univ. Blaise Pascal, Clermont-Ferrand, F |
| Catherine SAVONA | Univ. Blaise Pascal, Clermont-Ferrand, F |
| François SIMENHAUS | Univ. Pierre et Marie Curie, Paris, F |
| Tommi SOTTINEN | Univ. Helsinki, Finland |

I. TORRECILLA-TARANTINO Univ. Barcelona, Spain<br>Gerónimo URIBE<br>Vincent VIGON<br>Matthias WINKEL<br>Marcus WUNSCH<br>Univ. Mexico<br>Univ. Strasbourg, F<br>Univ. Oxford, UK<br>Univ. Wien, Austria

## List of short lectures

| Larbi Alili | On some functional transformations and <br> an application to the boundary crossing <br> problem for a Brownian motion |
| :--- | :--- |
| Fabrice Baudoin | Stochastic differential equations and <br> differential operators |
| Hermine Biermé | Random fields: self-similarity, anisotropy <br> and directional analysis |
| François Bolley | Approximation of some diffusion PDE <br> by some interacting particle system |
| Francesco Caravenna | A renewal theory approach to <br> periodically inhomogeneous polymer <br> models |
| Loïc Chaumont | On positive self-similar Markov processes |
| Charles Cuthbertson | Multiple selective sweeps and multi-type <br> branching |
| Jérôme Demange | Porous media equation and Sobolev <br> inequalities |
| Anne Eyraud-Loisel | Backward and forward-backward <br> stochastic differential equations with <br> enlarged filtration |
| Neil Farricker | Spectrally negative Lévy processes |
| Uwe Franz | A probabilistic model for biological <br> clocks |


| Christina Goldschmidt | Random recursive trees and the Bolthausen-Sznitman coalescent |
| :---: | :---: |
| Cindy Greenwood | Some problem areas which invite probabilists |
| Bénédicte Haas | Equilibrium for fragmentation with immigration |
| Chris Howitt | Sticky particles and sticky flows |
| Aldéric Joulin | On maximal inequalities for $\alpha$-stable integrals: the case $\alpha$ close to two |
| Nathalie Krell | On the rates of decay of fragments in homogeneous fragmentations |
| Aline Kurtzmann | About reinforced diffusions |
| Krzysztof Latuszyński | Ergodicity of adaptive Monte Carlo |
| Christophe Leuridan | Constructive Markov chains indexed by $\mathbb{Z}$ |
| Stéphane Loisel | Differentiation of some functionals of risk processes and optimal reserve allocation |
| Yutao Ma | Convex concentration inequalities and forward-backward stochastic calculus |
| José Alfredo López-Mimbela | Finite time blowup of semilinear PDE's with symmetric $\alpha$-stable generators |
| Mike Ludkovski | Optimal switching with applications to finance |
| Philippe Marchal | Concentration inequalities for infinitely divisible laws |
| James Martin | Stationary distributions of multi-type exclusion processes |
| Marie-Amélie Morlais | An application of the theory of backward stochastic differential equations in finance |
| Jan Obłój | On local martingales which are functions of ... and their applications |
| Cyril Odasso | Exponential mixing for stochastic PDEs: the non-additive case |
| Juan Carlos Pardo-Millan | Asymptotic results for positive self-similar Markov processes |


| Robert Philipowski | Propagation du chaos pour l'équation <br> des milieux poreux |
| :--- | :--- |
| Tommi Sottinen | On the equivalence of multiparameter <br> Gaussian processes |
| Gerónimo Uribe | Markov bridges, backward times, and a <br> Brownian fragmentation |
| Vincent Vigon | Certains comportements des processus <br> de Lévy sont décryptables par la <br> factorisation de Wiener-Hopf |
| Matthias Winkel | Coupling construction of Lévy trees |
| Marcus Wunsch | A stability result for <br> drift-diffusion-Poisson systems |


[^0]:    ${ }^{1}$ See Stephen Jay Gould's essay "A special fondness for beetles" in his book [77] for a discussion of the occasions on which Haldane may or may not have made this remark.

