

Markov Processes*

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Contents

1	Introduction to Markov Processes	3
1.1	Probability Preliminaries	3
1.2	Markov Processes	3
1.3	Filtrations	4
1.4	Conditional Expectation	5
1.4.1	Properties	6
1.5	Solution to Exercises	9
2	The Markov Property	12
2.1	The Discrete Time Markov Property	12
2.2	Continuous Time Processes	13
2.3	Discrete Time Processes	15
2.4	Solution to Exercises	18
3	The Kolmogorov Extension Theorem	20
3.1	Stochastic Processes as Random Variables	20
3.2	Constructing Stochastic Processes	20
3.3	Stationarity	22
4	Transition Probabilities	23
4.1	The Chapman-Kolmogorov Equation	23
4.2	Constructing Markov Processes	24
4.3	Solution to Exercises	28
5	Discrete Time Markov Processes	31
5.1	Time-Homogeneous Processes	31
5.2	Stopping Times	32
5.3	The Strong Markov Property	35
5.4	Solution to Exercises	39
6	Discrete State Space Markov Processes	42
6.1	Markov Chains as Graphs	42
6.2	Recurrence and Transience	44
6.3	Recurrence Conditions	47
6.4	Constructing Invariant Probability Measures	52
6.5	Long Run Dynamics	56
6.6	Total Variation	60
6.7	Periodic Chains	63

*Chapter 8 of these notes contains material from Chapter 9 of Xue-Mei Lie's notes for the same course taught in the autumn of 2021. It is not examinable for the course taught in the autumn of 2023, however, it supplements the existing material.

6.8	Ergodic Theorem	64
6.9	Reversible Markov Chains	67
6.9.1	Markov Chain Monte Carlo	69
6.10	Finite State Space Markov Chains	70
6.10.1	Perron-Frobenius	72
6.11	Solution to Exercises	74
7	Continuous State Space Markov Processes	77
7.1	Weak Convergence	77
7.2	The Feller Property	77
7.3	Existence of Invariant Probability Measures	81
7.4	Random Dynamical Systems	86
7.4.1	Existence of Invariant Measures	86
7.4.2	Uniqueness of Invariant Probability Measures via Contraction	87
7.5	Uniqueness of Invariant Probability Measures via Minorisation	90
7.6	P -Invariant Sets	93
7.7	Solution to Exercises	98
8	Ergodic Theory	101
8.1	Birkhoff's Ergodic Theorem for Dynamical Systems	101
8.2	Birkhoff's Ergodic Theorem for Markov Chains	104
8.2.1	Constructing Two-Sided Markov Processes	104
8.2.2	Birkhoff's Ergodic Theorem for Stationary Markov Chains as Dynamical Systems	105
8.2.3	Birkhoff's Ergodic Theorem for Markov Chains	105
8.3	The Structure Theorem	107
8.4	Solution to Exercises	114
9	Appendix	115
9.1	Gaussian Measures	115
9.2	The Doeblin Coupling	116
9.3	Solution to Exercises	120

1 Introduction to Markov Processes

1.1 Probability Preliminaries

We will define randomness using Kolmogorov's framework. That is, we define an underlying probability space, which is a tuple $(\Omega, \mathcal{F}, \mathbb{P})$ with the following components.

- Ω is an abstract space.
- \mathcal{F} a σ -algebra.
- \mathbb{P} is a probability measure, namely a measure with unit mass.

Random quantities will take values in a state space \mathcal{X} which we assume to be a complete and separable metric space. We will use $\mathcal{B}(\mathcal{X})$ to denote its Borel σ -algebra, which is the σ -algebra generated by the open sets of \mathcal{X} . Recall that random quantities are represented as \mathcal{X} -valued random variables, which are measurable functions $X : (\Omega, \mathcal{F}) \rightarrow (\mathcal{X}, \mathcal{B}(\mathcal{X}))$. For X a \mathcal{X} -valued random variable $\text{Law}(X)$ or $X_*\mathbb{P}$ is the push forward of \mathbb{P} of X , where for each $A \subseteq \mathcal{B}(\mathcal{X})$ we have

$$X_*\mathbb{P}(A) = \mathbb{P}(X \in A) = \mathbb{P}(\{\omega \in \Omega : X(\omega) \in A\}).$$

A family of \mathcal{X} -valued random variables is denoted using an index set I as $(X_i)_{i \in I}$.

- $\mathbb{P}(X_i \in A, X_j \in B) := \mathbb{P}(\{\omega \in \Omega : X_i(\omega) \in A \text{ and } X_j(\omega) \in B\})$.
- For $I' \subseteq I$ the family $(X_i)_{i \in I'}$ is independent if for all $A_i \in \mathcal{B}(\mathcal{X})$ we have

$$\mathbb{P}(X_i \in A_i \text{ for all } i \in I') = \prod_{i \in I'} \mathbb{P}(X_i \in A_i).$$

- For dependent random variables, when $\mathbb{P}(X_j \in B) > 0$ we have

$$\mathbb{P}(X_i \in A | X_j \in B) := \frac{\mathbb{P}(X_i \in A, X_j \in B)}{\mathbb{P}(X_j \in B)}.$$

For I totally ordered, $X = (X_i)_{i \in I}$ defines a stochastic process. For I discrete we have a discrete-time process.

1.2 Markov Processes

Definition 1.2.1 (Intuitive). *A Markov process is a process that could be characterised by either of the following statements.*

1. *For any prediction of the future, knowledge of the present is just as good as knowledge of both the past and present.*
2. *Conditioned on the present, the past and present are independent.*

Definition 1.2.2 (Formal). *A process $(X_i)_{i=0}^{\infty}$ is Markov if it satisfies either one of the following characterisations.*

1. *For every $j \in \mathbb{N}$ and $A, B_1, \dots, B_j \in \mathcal{B}(\mathcal{X})$ such that $\mathbb{P}(X_1 \in B_1, \dots, X_j \in B_j) > 0$ we have*

$$\mathbb{P}(X_{j+1} \in A | X_1 \in B_1, \dots, X_j \in B_j) = \mathbb{P}(X_{j+1} \in A | X_j \in B_j).$$

2. For any $i < j < k$ and $A_0, \dots, A_i, B, C_{j+1}, \dots, C_k \in \mathcal{B}(\mathcal{X})$ with $\mathbb{P}(X_j \in B) > 0$ we have

$$\begin{aligned} \mathbb{P}(X_0 \in A_0, \dots, X_i \in A_i, X_{j+1} \in C_{j+1}, \dots, X_k \in C_k | X_j \in B) \\ = \mathbb{P}(X_0 \in A_0, \dots, X_i \in A_i | X_j \in B) \\ \cdot \mathbb{P}(X_{j+1} \in C_{j+1}, \dots, X_k \in C_k | X_j \in B). \end{aligned}$$

Exercise 1.2.3. Suppose that \mathcal{X} is a finite set. Show that the characterisations of Definition 1.2.2 are equivalent.

Exercise 1.2.4. Let $(X_j)_{j=0}^\infty$ be given by

$$X_j = X_0 + \sum_{i=1}^j Y_i,$$

where

- X_0 is a \mathbb{Z} -valued random variable, and
- $(Y_i)_{i=1}^\infty$ is an independent and identically distributed family of $\{\pm 1\}$ -valued random variables, independent of X_0 .

Show that $(X_j)_{j=0}^\infty$ is a Markov process.

Exercise 1.2.5. Let $(X_j)_{j=0}^\infty$ be as in Exercise 1.2.4 and consider

$$M_j := \max_{0 \leq i \leq j} X_i.$$

Taking $X_0 = 0$, show that $(M_j)_{j=0}^\infty$ is not a Markov process.

1.3 Filtrations

Definition 1.3.1. The σ -algebra generated by a family of random variables $(Y_j)_{j \in J}$, denoted $\sigma((Y_j)_{j \in J})$, is the smallest σ -algebra such that

$$\{Y_j^{-1}(A) : j \in J, A \in \mathcal{B}(\mathcal{X})\}$$

is contained within the σ -algebra. In other words, $\sigma((Y_j)_{j \in J})$ is the smallest σ -algebra such that the family $(Y_j)_{j \in J}$ is measurable.

Remark 1.3.2. A σ -algebra can be thought of as encoding information. If Z is a random variable with respect to $\sigma((Y_j)_{j \in J})$, then knowing the values of $(Y_j)_{j \in J}$ is sufficient for determining the value of Z . That is, $Z = g((Y_j)_{j \in J})$ for a measurable function g .

Exercise 1.3.3. Let Y_1 be a $\{-1, 0, 1\}$ -valued random variable and let Y_2 be a $\{-3, 3\}$ -valued random variable. What is the maximum value of $|\sigma(Y_1, Y_2)|$?

Definition 1.3.4. A filtration is a sequence of σ -algebras $(\mathcal{F}_n)_{n=0}^\infty$ with $\mathcal{F}_n \subseteq \mathcal{F}$ for all $n \in \mathbb{N}$ such that for $m \leq n$ we have $\mathcal{F}_m \subseteq \mathcal{F}_n$.

Definition 1.3.5. A stochastic process X with state space \mathcal{X} is a collection of \mathcal{X} -valued random variables $(X_n)_{n=0}^\infty$ on a probability space $(\Omega, \mathcal{F}, \mathbb{P})$.

Definition 1.3.6. A stochastic process $X = (X_n)_{n=0}^\infty$ is adapted to $(\mathcal{F}_n)_{n=0}^\infty$ if for every $n \in \mathbb{N}$ the random variable X_n is measurable with respect to \mathcal{F}_n .

Definition 1.3.7. For a stochastic process X , the filtration generated by X , denoted $\mathcal{F}^0 = (\mathcal{F}_n^0)_{n=0}^\infty$ is

$$\mathcal{F}_n^0 := \sigma((X_i)_{i=0}^n)$$

for every $n \in \mathbb{N}$.

Remark 1.3.8.

- The filtration \mathcal{F}^0 is the smallest filtration for which X is adapted.
- Intuitively, \mathcal{F}_n^0 is the information generated by the process up until time $n \in \mathbb{N}$.

1.4 Conditional Expectation

Throughout, $(\Omega, \mathcal{F}, \mathbb{P})$ will be the underlying probability space.

Definition 1.4.1. A random variable X is integrable if $\mathbb{E}(|X|) < \infty$.

Definition 1.4.2. Let X be an integrable random variable, and consider a sub- σ -algebra $\mathcal{F}' \subseteq \mathcal{F}$. The conditional expectation of X with respect to \mathcal{F}' is an \mathcal{F}' -measurable random variable X' such that for every $A \in \mathcal{F}'$ we have

$$\int_A X(\omega) \, d\mathbb{P}(\omega) = \int_A X'(\omega) \, d\mathbb{P}(\omega).$$

Proposition 1.4.3. In the setting of Definition 1.4.2, there exists, up to \mathbb{P} -null sets, a unique conditional expectation of X with respect to \mathcal{F}' .

Proof. Let μ be the restriction of \mathbb{P} to \mathcal{F}' . Let ν be the measure on \mathcal{F}' be given by

$$\nu(A) = \int_A X(\omega) \, d\mathbb{P}(\omega)$$

for all $A \in \mathcal{F}'$. Then if $\mu(A) = 0$ it follows that $\nu(A) = 0$ which implies that $\nu \ll \mu$. Therefore, by the Random-Nikodym theorem there exists a unique \mathcal{F}' -measurable function, X' , up to sets of zero measure such that

$$\nu(A) = \int_A X'(\omega) \, d\mathbb{P}(\omega)$$

for all $A \in \mathcal{F}'$. □

Remark 1.4.4.

- If we have a candidate for the conditional expectation, by uniqueness, it suffices to check the conditions of Definition 1.4.2 are satisfied to conclude that it is the conditional expectation.

- Note that $X(\omega)$ is not necessarily \mathcal{F}' measurable. So we cannot say that X' equals X pointwise.

We denote the conditional expectation of X with respect to \mathcal{F}' by $\mathbb{E}(X|\mathcal{F}')$.

- We also write $\mathbb{P}(A|\mathcal{F}') = \mathbb{E}(\mathbf{1}_A|\mathcal{F}')$.
- Given a random variable Y , we write $\mathbb{E}(\cdot|Y) = \mathbb{E}(\cdot|\sigma(Y))$.

Remark 1.4.5. Intuitively, we think of $\mathbb{E}(X|Y)$ as being an approximation of X when we only have the information of $\sigma(Y)$. As $\mathbb{E}(X|Y)$ is measurable with respect to $\sigma(Y)$ we can think of it as a measurable function $\phi(Y)$.

Example 1.4.6. Let $\mathcal{F}' = \{\emptyset, A, A^c, \Omega\}$ where $A \in \mathcal{F}$ and $\mathbb{P}(A) \in (0, 1)$. Let X be an integrable random variable. Then

$$\mathbb{E}(X|\mathcal{F}')(\omega) = \begin{cases} \frac{1}{\mathbb{P}(A)} \int_A X(\omega) d\mathbb{P}(\omega) & \omega \in A \\ \frac{1}{\mathbb{P}(A^c)} \int_{A^c} X(\omega) d\mathbb{P}(\omega) & \omega \in A^c. \end{cases}$$

To see this one just has to note that X' is \mathcal{F}' -measurable and

$$\int_B X'(\omega) d\mathbb{P} = \begin{cases} 0 & B = \emptyset \\ \int_A X(\omega) d\mathbb{P}(\omega) & B = A \\ \int_{A^c} X(\omega) d\mathbb{P}(\omega) & B = A^c \\ 1 & B = \Omega, \end{cases}$$

where all set equalities hold almost everywhere.

Example 1.4.7. Let Y be an \mathbb{N} -valued random variable. Let X be an integrable random variable. Then

$$\mathbb{E}(X|Y) = \sum_{i:\mathbb{P}(Y=i)>0} \mathbb{E}(X|Y=i) \mathbb{P}(Y=i).$$

Exercise 1.4.8. Let $(\Omega, \mathcal{F}, \mathbb{P}) = (\mathbb{R}, \mathcal{B}(\mathbb{R}), \mathcal{N}(0, 1))$. Let Y be an integrable random variable given by $Y(\omega) = \omega^2$. Let X be an integrable random variable. Give a formula for $\mathbb{E}(X|Y)$.

1.4.1 Properties

Proposition 1.4.9. Let X be an integrable random variable on the probability space $(\Omega, \mathcal{F}, \mathbb{P})$. Suppose $\mathcal{B} \subseteq \mathcal{F}$ is a σ -algebra such that $\sigma(X)$ and \mathcal{B} are independent. Then,

$$\mathbb{E}(X|\mathcal{B}) = \mathbb{E}(X).$$

Proof. Let $B \in \mathcal{B}$, then

$$\begin{aligned} \int_B X(\omega) d\mathbb{P}(\omega) &= \int_{\Omega} \mathbf{1}_B X(\omega) d\mathbb{P}(\omega) \\ &= \mathbb{E}(X \mathbf{1}_B) \\ &\stackrel{(1)}{=} \mathbb{E}(X) \mathbb{E}(\mathbf{1}_B) \\ &= \mathbb{E}(X) \mathbb{P}(B) \\ &= \int_B \mathbb{E}(X) d\mathbb{P}(\omega), \end{aligned}$$

where in (1) we have used the independence assumption. As $\mathbb{E}(X)$ is a constant, it is \mathcal{B} -measurable. Therefore, we conclude that $\mathbb{E}(X|\mathcal{B}) = \mathbb{E}(X)$. \square

Proposition 1.4.10. *Let $X, Y \in L^2(\Omega, \mathcal{F}, \mathbb{P})$. Let $\mathcal{B} \subseteq \mathcal{F}$ be a σ -algebra such that Y is \mathcal{B} -measurable. Then*

$$\mathbb{E}(XY|\mathcal{B}) = Y\mathbb{E}(X|\mathcal{B}).$$

Proof. Step 1: Let $Y = \mathbf{1}_A$ for $A \in \mathcal{B}$.

For any $B \in \mathcal{B}$ it follows that

$$\begin{aligned} \int_B Y(\omega)X(\omega) d\mathbb{P}(\omega) &= \int_{B \cap A} X(\omega) d\mathbb{P}(\omega) \\ &\stackrel{(1)}{=} \int_{B \cap A} \mathbb{E}(X|\mathcal{B})(\omega) d\mathbb{P}(\omega) \\ &= \int_B \mathbf{1}_A(\omega)\mathbb{E}(X|\mathcal{B})(\omega) d\mathbb{P}(\omega) \\ &= \int_B Y(\omega)\mathbb{E}(X|\mathcal{B})(\omega) d\mathbb{P}(\omega), \end{aligned}$$

where in (1) we use that $B \cap A \in \mathcal{B}$. As $A \in \mathcal{B}$ we note that $\mathbf{1}_A\mathbb{E}(X|\mathcal{B})$ is the product of \mathcal{B} -measurable functions and is therefore also \mathcal{B} -measurable. Hence, $\mathbb{E}(XY|\mathcal{B}) = Y\mathbb{E}(X|\mathcal{B})$.

Step 2: Let Y be a simple function.

Due to the linearity of the integral, we can use step 1 to extend the result to Y .

Step 3: Let Y be a non-negative \mathcal{B} -measurable function.

There exists a sequence of positive simple functions $(Y_n)_{n \in \mathbb{N}}$ such that $Y_n \nearrow Y$. Note that $Y_n X \rightarrow YX$ almost surely and $|Y_n X| \leq |YX|$. As X and Y are in L^2 we can use the Cauchy-Schwartz inequality to note that

$$\mathbb{E}(|XY|)^2 \leq \mathbb{E}(|X|)^2 \mathbb{E}(|Y|)^2 < \infty.$$

Therefore, we can use step 2 and apply the dominated convergence theorem to deduce that

$$\mathbb{E}(YX|\mathcal{B}) = \lim_{n \rightarrow \infty} \mathbb{E}(Y_n X|\mathcal{B}) \stackrel{(1)}{=} \lim_{n \rightarrow \infty} Y_n \mathbb{E}(X|\mathcal{B}) = Y\mathbb{E}(X|\mathcal{B}).$$

Step 4: Let Y be an arbitrary \mathcal{B} -measurable function.

We can use the decomposition $Y = Y^+ - Y^-$ to conclude the result from step 3. \square

Proposition 1.4.11 (The Tower Property). *Let X be an integrable random variable and let $\mathcal{A} \subseteq \mathcal{B}$ be sub- σ -algebras of \mathcal{F} . Then*

$$\mathbb{E}(\mathbb{E}(X|\mathcal{A})|\mathcal{B}) = \mathbb{E}(\mathbb{E}(X|\mathcal{B})|\mathcal{A}) = \mathbb{E}(X|\mathcal{A}).$$

Proof. As $\mathbb{E}(X|\mathcal{A})$ is \mathcal{A} -measurable, it is \mathcal{B} -measurable. Therefore, by Proposition 1.4.10 we get that

$$\mathbb{E}(\mathbb{E}(X|\mathcal{A})|\mathcal{B}) = \mathbb{E}(X|\mathcal{A}).$$

On the other hand, for $C \in \mathcal{A} \subseteq \mathcal{B}$ we have that

$$\int_C \mathbb{E}(X|\mathcal{A})(\omega) d\mathbb{P}(\omega) = \int_C X(\omega) d\mathbb{P}(\omega)$$

and

$$\int_C \mathbb{E}(X|\mathcal{B})(\omega) d\mathbb{P}(\omega) = \int_C X(\omega) d\mathbb{P}(\omega).$$

Therefore,

$$\int_C \mathbb{E}(X|\mathcal{A})(\omega) d\mathbb{P}(\omega) = \int_C \mathbb{E}(X|\mathcal{B})(\omega) d\mathbb{P}(\omega).$$

Thus, as $\mathbb{E}(X|\mathcal{A})$ is \mathcal{A} -measurable, we observe that $\mathbb{E}(X|\mathcal{A})$ satisfies the required conditions to be the conditional expectation of $\mathbb{E}(X|\mathcal{B})$ with respect to \mathcal{A} . Therefore, by Proposition 1.4.3 we conclude that

$$\mathbb{E}(\mathbb{E}(X|\mathcal{B})|\mathcal{A}) = \mathbb{E}(X|\mathcal{A}).$$

□

Remark 1.4.12. *Intuitively, Proposition 1.4.11 holds as \mathcal{A} has less information than \mathcal{B} , and so the conditioning on \mathcal{B} retains sufficient information for the conditioning on \mathcal{A} to take full effect.*

Proposition 1.4.13. *Suppose that X and Y are bounded \mathbb{R} -valued random variables, and let $f, g \in \mathcal{B}_b(\mathbb{R})$. Then*

$$\mathbb{E}(\mathbb{E}(Y|f(g(X))|g(X))) = \mathbb{E}(\mathbb{E}(Y|g(X))|f(g(X))) = \mathbb{E}(Y|f(g(X))).$$

Proof. Note that $f(g(X))$ and $g(X)$ are random variables as f and g are measurable functions. Specifically, we have that

$$\sigma(f(g(X))) := \mathcal{A}_{fg} = \{A : f(g(X(A))) \in \mathcal{B}(\mathbb{R})\}$$

and

$$\sigma(g(X)) := \mathcal{A}_g = \{A : g(X(A)) \in \mathcal{B}(\mathbb{R})\}.$$

Consider $A \in \mathcal{A}_{fg}$, then $f(g(X(A))) = B$ for some $B \in \mathcal{B}(\mathbb{R})$. As f is measurable it follows that $g(X(A)) = f^{-1}(B) \in \mathcal{B}(\mathbb{R})$ which implies that $A \in \mathcal{A}_g$. Therefore, $\mathcal{A}_{fg} \subseteq \mathcal{A}_g$. Hence, from Proposition 1.4.11 we deduce that

$$\mathbb{E}(\mathbb{E}(Y|g(X))|f(g(X))) = \mathbb{E}(Y|f(g(X))),$$

and

$$\mathbb{E}(\mathbb{E}(Y|f(g(X))|g(X))) = \mathbb{E}(Y|f(g(X))).$$

□

Proposition 1.4.14. *Suppose that X and Y are bounded \mathbb{R} -valued random variables. Then for any $f \in \mathcal{B}_b(\mathbb{R})$ we have*

$$\mathbb{E}\left((Y - \mathbb{E}(Y|X))^2\right) \leq \mathbb{E}\left((Y - f(X))^2\right).$$

Proof. On the one hand,

$$\begin{aligned} \mathbb{E}\left((Y - \mathbb{E}(Y|X))^2|X\right) &= \mathbb{E}(Y^2|X) - 2\mathbb{E}(Y\mathbb{E}(Y|X)|X) + \mathbb{E}(\mathbb{E}(Y|X)^2|X) \\ &= \mathbb{E}(Y^2|X) - 2\mathbb{E}(Y|X)^2 + \mathbb{E}(Y|X)^2 \\ &= \mathbb{E}(Y^2|X) - \mathbb{E}(Y|X)^2 \end{aligned}$$

On the other hand,

$$\begin{aligned} \mathbb{E}\left((Y - f(X))^2|X\right) &= \mathbb{E}(Y^2|X) - 2\mathbb{E}(Yf(X)|X) + \mathbb{E}(f(X)^2|X) \\ &= \mathbb{E}(Y^2|X) - 2f(X)\mathbb{E}(Y|X) + f(X)^2. \end{aligned}$$

Therefore,

$$\mathbb{E}\left((Y - f(X))^2|X\right) - \mathbb{E}\left((Y - \mathbb{E}(Y|X))^2|X\right) = (\mathbb{E}(Y|X) - f(X))^2 \geq 0$$

and hence

$$\mathbb{E}\left((Y - f(X))^2|X\right) \geq \mathbb{E}\left((Y - \mathbb{E}(Y|X))^2|X\right).$$

Taking the expectations of both sides we conclude that

$$\mathbb{E}\left((Y - f(X))^2\right) \geq \mathbb{E}\left((Y - \mathbb{E}(Y|X))^2\right).$$

□

Proposition 1.4.15. Suppose that $\mathcal{A} \subseteq \mathcal{B}$ are σ -algebras contained in \mathcal{F} . For a bounded random variable X we have that

$$\text{Var}(\mathbb{E}(X|\mathcal{B})) \geq \text{Var}(\mathbb{E}(X|\mathcal{A})).$$

Proof. Observe that,

$$\begin{aligned} \text{Var}(X) &= \mathbb{E}(\text{Var}(X|\mathcal{A})) + \text{Var}(\mathbb{E}(X|\mathcal{A})) \\ &= \mathbb{E}(\text{Var}(X|\mathcal{B})) + \text{Var}(\mathbb{E}(X|\mathcal{B})). \end{aligned}$$

Therefore,

$$\text{Var}(\mathbb{E}(X|\mathcal{B})) - \text{Var}(\mathbb{E}(X|\mathcal{A})) = \mathbb{E}(\text{Var}(X|\mathcal{A})) - \mathbb{E}(\text{Var}(X|\mathcal{B})).$$

Moreover,

$$\begin{aligned} \text{Var}(X|\mathcal{A}) &= \mathbb{E}((X - \mathbb{E}(X|\mathcal{A}))^2|\mathcal{A}) \\ &= \mathbb{E}((X - \mathbb{E}(\mathbb{E}(X|\mathcal{B})|\mathcal{A}))^2|\mathcal{A}) \\ &\stackrel{(1)}{\geq} \mathbb{E}((X - \mathbb{E}(X|\mathcal{B}))^2|\mathcal{A}) \\ &\stackrel{(2)}{=} \mathbb{E}(\mathbb{E}((X - \mathbb{E}(X|\mathcal{B}))^2|\mathcal{B})|\mathcal{A}) \\ &= \mathbb{E}(\text{Var}(X|\mathcal{B})|\mathcal{A}). \end{aligned} \tag{1.4.1}$$

Where in (1) we use Proposition 1.4.14 as $\mathbb{E}(\mathbb{E}(X|\mathcal{B})|\mathcal{A})$ is \mathcal{A} -measurable and so \mathcal{B} -measurable. In (2) the tower property for conditional expectations is used. As $\text{Var}(X|\mathcal{A})$ is an \mathcal{A} -measurable function, we have that $\mathbb{E}(\text{Var}(X|\mathcal{A})|\mathcal{A}) = \text{Var}(X|\mathcal{A})$. Using this and (1.4.1) we deduce that $\text{Var}(X|\mathcal{A}) \geq \text{Var}(X|\mathcal{B})$, which means that

$$\text{Var}(\mathbb{E}(X|\mathcal{B})) - \text{Var}(\mathbb{E}(X|\mathcal{A})) = \mathbb{E}(\text{Var}(X|\mathcal{A})) - \mathbb{E}(\text{Var}(X|\mathcal{B})) \geq 0$$

and hence $\text{Var}(\mathbb{E}(X|\mathcal{B})) \geq \text{Var}(\mathbb{E}(X|\mathcal{A}))$. □

1.5 Solution to Exercises

Exercise 1.2.3

Solution. (1) \Rightarrow (2). Let

- $A = \{X_0 \in A_0, \dots, X_i \in A_i\}$,
- $B = \{X_j \in B\}$, (we abuse notation slightly with the B 's) and
- $C = \{X_{j+1} \in C_{j+1}, \dots, X_k \in C_k\}$.

We assume that $\mathbb{P}(B), \mathbb{P}(A, B) > 0$. Then

$$\begin{aligned} \mathbb{P}(A, C|B) &= \frac{\mathbb{P}(A, C, B)}{\mathbb{P}(B)} \\ &= \frac{\mathbb{P}(C|A, B)\mathbb{P}(A, B)}{\mathbb{P}(B)} \\ &= \mathbb{P}(C|A, B)\mathbb{P}(A|B). \end{aligned}$$

Applying statement 1 we get that $\mathbb{P}(C|A, B) = \mathbb{P}(C|B)$ and so

$$\mathbb{P}(A, C|B) = \mathbb{P}(A|B)\mathbb{P}(C|B).$$

(2) \Rightarrow (1). Then

$$\begin{aligned}
\mathbb{P}(X_{j+1} \in A_{j+1} | X_1 \in B_1, \dots, X_j \in B_j) &= \frac{\mathbb{P}(X_{j+1} \in A_{j+1}, X_1 \in B_1, \dots, X_j \in B_j)}{\mathbb{P}(X_1 \in B_1, \dots, X_j \in B_j)} \\
&= \frac{\mathbb{P}(X_{j+1} \in A_{j+1}, X_1 \in B_1, \dots, X_{j-1} \in B_{j-1} | X_j \in B_j) \mathbb{P}(X_j \in B_j)}{\mathbb{P}(X_1 \in B_1, \dots, X_j \in B_j)} \\
&\stackrel{(2)}{=} \frac{\mathbb{P}(X_{j+1} \in A_{j+1} | X_j \in B_j) \mathbb{P}(X_1 \in B_1, \dots, X_{j-1} \in B_{j-1} | X_j \in B_j) \mathbb{P}(X_j \in B_j)}{\mathbb{P}(X_1 \in B_1, \dots, X_j \in B_j)} \\
&= \mathbb{P}(X_{j+1} \in A_{j+1} | X_j \in B_j).
\end{aligned}$$

□

Exercise 1.2.4

Solution. Note that $X_j = X_{j-1} + Y_j$. As $\mathcal{X} = \mathbb{Z}$ it suffices to consider the events $\{k\}$ for $k \in \mathbb{Z}$. Observe that,

$$\begin{aligned}
\mathbb{P}(X_{j+1} = x_{j+1} | X_0 = x_0, \dots, X_j = x_j) &= \mathbb{P}(X_{j+1} = x_j + Y_j | X_0 = x_0, \dots, X_j = x_j) \\
&= \mathbb{P}(X_{j+1} = x_j + Y_j | X_j = x_j),
\end{aligned}$$

where the last equality follows from the fact that the Y_j are independent and independent from X_0 . Therefore, $(X_j)_{j=1}^\infty$ satisfies statement 1 of Definition 1.2.2 and hence is a Markov process. □

Exercise 1.2.5

Solution. For the sequence (M_0, M_1, M_2, M_3) consider the following possibilities.

1. $(0, 0, 0, 1)$ which arises from the sequence $(X_0, X_1, X_2, X_3) = (0, -1, 0, 1)$.
2. $(0, 1, 1, 1)$ which may arise from the sequence $(X_0, X_1, X_2, X_3) = (0, 1, 0, -1)$ or $(X_0, X_1, X_2, X_3) = (0, 1, 0, 1)$.

Note that these are the only possibilities for which $M_3 = 1$. We can then compute the possibilities of M_4 in each case.

1. $\mathbb{P}(M_4 = 2 | M_0 = 0, M_1 = 0, M_2 = 0, M_3 = 1) = \frac{1}{2}$.
2. $\mathbb{P}(M_4 = 2 | M_0 = 0, M_1 = 1, M_2 = 1, M_3 = 1) = \frac{1}{2} \frac{1}{2} = \frac{1}{4}$,
 - The first $\frac{1}{2}$ comes from the fact that we must observe the sequence $(X_0, X_1, X_2, X_3) = (0, 1, 0, 1)$ in order to be able to reach $M_4 = 2$. The second $\frac{1}{2}$ is the probability of reaching $M_4 = 2$ from that sequence.

Therefore, we see that the future $j + 1 = 4$ is dependent on more than just the present $j = 3$. Hence, $(M_j)_{j=0}^\infty$ is not a Markov process. □

Exercise 1.3.3

Solution. There are $3 \times 2 = 6$ possibilities for the tuple (Y_1, Y_2) . Taking the power set of these gives a σ -algebra of size 2^6 , which is the largest σ -algebra as σ -algebra is simply a collection of subsets. □

Exercise 1.4.8

Solution. For a set A , let

$$f(A) := \{f(\omega) : \omega \in A\}.$$

Step 1: $\sigma(Y) = \{B \in \mathcal{B}(\mathbb{R}) : -B = B\}$.

Suppose $-B = B$, then for $A = B^2 \in \mathcal{B}(\mathbb{R})$ we have

$$Y^{-1}(A) = \pm B = B,$$

which implies that $B \in \sigma(Y)$. On the other hand, if $B \in \sigma(Y)$, then $B = Y^{-1}(A)$ for some $A \in \mathcal{B}(\mathbb{R})$ so that

$$\begin{aligned} -B &= \{-\omega : \omega \in Y^{-1}(A)\} \\ &= \{-\omega : -\omega \in Y^{-1}(A)\} \\ &= \{\omega : \omega \in Y^{-1}(A)\} \\ &= B, \end{aligned}$$

where we have used that $\omega \in Y^{-1}(A)$ if and only if $-\omega \in Y^{-1}(A)$.

Step 2: $\mathbb{E}(Y|X)(\omega) = \frac{1}{2}(X(\omega) + X(-\omega))$.

For $A \in \mathcal{B}(\mathbb{R})$ observe that

$$\left\{ \omega : \frac{1}{2}(X(\omega) + X(-\omega)) \in A \right\} = \left\{ -\omega : \frac{1}{2}(X(\omega) + X(-\omega)) \in A \right\},$$

which implies that

$$\left\{ \omega : \frac{1}{2}(X(\omega) + X(-\omega)) \in A \right\} \in \sigma(Y)$$

by step 1. Hence, $\frac{1}{2}(X(\omega) + X(-\omega))$ is $\sigma(Y)$ -measurable. Moreover, as $\mathcal{N}(0, 1)$ is a symmetric distribution, for $B \in \sigma(Y)$ we have

$$\begin{aligned} \int_B \frac{1}{2}(X(\omega) + X(-\omega)) \, d\mathbb{P}(\omega) &\stackrel{u=-\omega}{=} \frac{1}{2} \int_B X(\omega) \, d\mathbb{P}(\omega) + \frac{1}{2} \int_{-B} X(u) \, d\mathbb{P}(-u) \\ &= \frac{1}{2} \int_B X(\omega) \, d\mathbb{P}(\omega) + \frac{1}{2} \int_B X(u) \, d\mathbb{P}(u) \\ &= \int_B X(\omega) \, d\mathbb{P}(\omega). \end{aligned}$$

Therefore, $\mathbb{E}(Y|X) = \frac{1}{2}(X(\omega) + X(-\omega))$. □

2 The Markov Property

2.1 The Discrete Time Markov Property

We can now reformulate the Markov property with conditional expectations. The process $(X_n)_{n=0}^\infty$ has the Markov property if one of the following equivalent conditions holds.

1. For every $j \in \mathbb{N}$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$\mathbb{P}(X_{j+1} \in A | X_0, \dots, X_j) = \mathbb{P}(X_{j+1} \in A | X_j).$$

2. For any $j < k$, $A \in \mathcal{B}(\mathcal{X}^j)$ and $B \in \mathcal{B}(\mathcal{X}^{k-j})$, we have

$$\begin{aligned} \mathbb{P}((X_0, \dots, X_{j-1}) \in A, (X_{j+1}, \dots, X_k) \in B | X_j) &= \mathbb{P}((X_0, \dots, X_{j-1}) \in A | X_j) \\ &\quad \cdot \mathbb{P}((X_{j+1}, \dots, X_k) \in B | X_j). \end{aligned}$$

Theorem 2.1.1. Let \mathcal{X} and \mathcal{Y} be state spaces with \mathcal{X} being discrete. Let $F_n : \mathcal{X} \times \mathcal{Y} \rightarrow \mathcal{X}$ be a measurable map for each $n \in \mathbb{N}$. Furthermore, let X_0 be an \mathcal{X} -valued random variable and $(\zeta_n)_{n=1}^\infty$ be a family of independent \mathcal{Y} -valued random variables that are also independent of X_0 . Then the process $X = (X_n)_{n=0}^\infty$ given by

$$X_{n+1} = F_n(X_n, \zeta_{n+1})$$

for $n \in \mathbb{N}$ is a Markov process.

Proof. As \mathcal{X} discrete, it suffices to show the Markov property holds for singletons set. On the one hand,

$$\begin{aligned} \mathbb{P}(X_{n+1} = i_{n+1} | X_n = i_n, \dots, X_0 = i_0) &= \mathbb{P}(F_n(X_n, \zeta_{n+1}) | X_n = i_n, \dots, X_0 = i_0) \\ &= \mathbb{P}(F_n(i_n, \zeta_{n+1})), \end{aligned}$$

where in the second equality we use the fact that ζ_{n+1} is independent of the X_0, \dots, X_n . On the other hand,

$$\mathbb{P}(X_{n+1} = i_{n+1} | X_n = i_n) = \mathbb{P}(F_n(i_n, \zeta_{n+1})).$$

Therefore, X has the Markov property. □

We can generalise Theorem 2.1.1 to arbitrary state spaces. To do so, we require Proposition 2.1.2.

Proposition 2.1.2. Let \mathcal{X} and \mathcal{Y} be state spaces and let $\phi : \mathcal{X} \times \mathcal{Y} \rightarrow \mathbb{R}$ be a measurable function. Suppose X is an \mathcal{X} -valued random variable and Y is a \mathcal{Y} -valued random variable such that $\phi(X, Y)$ is integrable. Then for any σ -algebra $\mathcal{G} \subseteq \mathcal{F}$ such that X is \mathcal{G} -measurable and Y is independent of \mathcal{G} , it follows that

$$\mathbb{E}(\phi(X, Y) | \mathcal{G})(\omega) = \mathbb{E}(\phi(X(\omega), Y)).$$

Proof. Let $\phi(X, Y) = \mathbf{1}_{\{X \in A\}} \mathbf{1}_{\{Y \in B\}}$ where $A \in \mathcal{B}(\mathcal{X})$ and $B \in \mathcal{B}(\mathcal{Y})$. Then,

$$\begin{aligned} \mathbb{E}(\mathbf{1}_{\{X \in A\}} \mathbf{1}_{\{Y \in B\}} | \mathcal{G})(\omega) &\stackrel{(1)}{=} \mathbf{1}_{\{X \in A\}}(\omega) \mathbb{E}(\mathbf{1}_{\{Y \in B\}} | \mathcal{G})(\omega) \\ &\stackrel{(2)}{=} \mathbf{1}_{\{X \in A\}}(\omega) \mathbb{E}(\mathbf{1}_{\{Y \in B\}}) \\ &= \mathbb{E}(\mathbf{1}_{\{X \in A\}}(\omega) \mathbf{1}_{\{Y \in B\}}), \end{aligned}$$

where (1) follows as X is \mathcal{G} measurable and (2) follows as Y is independent of \mathcal{G} . Using the linearity of expectation we can extend this result to X and Y being simple random variables. Applying the monotone convergence theorem gives the result for X and Y non-negative random variables. To generalise to arbitrary random variables we utilize the decompositions $X = X^+ - X^-$ and $Y = Y^+ - Y^-$. □

Theorem 2.1.3. Let \mathcal{X} and \mathcal{Y} be state spaces. For each $n \in \mathbb{N}$, let $F_n : \mathcal{X} \times \mathcal{Y} \rightarrow \mathcal{X}$ be a measurable map. Let X_0 be an \mathcal{X} -valued random variable and let $(\zeta_n)_{n=0}^\infty$ be a family of independent \mathcal{Y} -valued random variables that are also independent of X_0 . Then the process $X = (X_n)_{n=0}^\infty$ given by

$$X_{n+1} = F_n(X_n, \zeta_{n+1})$$

for each $n \in \mathbb{N}$ is a Markov process.

Proof. Let $A \in \mathcal{F}$, then

$$\mathbb{P}(X_{n+1} \in A | X_n, \dots, X_0) = \mathbb{P}(F_n(X_n, \zeta_{n+1}) \in A | X_n, \dots, X_0),$$

where

- X_n is measurable with respect to $\sigma(X_0, \dots, X_n)$, and
- ζ_{n+1} is independent of $\sigma(X_0, \dots, X_n)$.

Therefore, by Proposition 2.1.2 it follows that

$$\begin{aligned} \mathbb{P}(X_{n+1} \in A | X_n, \dots, X_0) &= \mathbb{E}(\mathbf{1}_A(F_n(X_n, \zeta_{n+1})) | X_n, \dots, X_0) \\ &= \mathbb{E}(\mathbf{1}_A(F_n(X_n, \zeta_{n+1}))). \end{aligned}$$

A similar computation shows that $\mathbb{P}(X_{n+1} \in A | X_n) = \mathbb{E}(\mathbf{1}_A(F_n(X_n, \zeta_{n+1})))$ and so the process is Markov. \square

2.2 Continuous Time Processes

A continuous time process $(X_s)_{s \in I}$, is indexed by a continuous well-ordered set I , such as $I = [0, \infty)$.

Definition 2.2.1. A continuous time filtration is an uncountable family of σ -algebras $(\mathcal{F}_s : s \in I)$ with $\mathcal{F}_s \subseteq \mathcal{F}_t$ and $\mathcal{F}_s \subseteq \mathcal{F}_t$ for each $s \leq t$.

Definition 2.2.2. Given a continuous time process $(X_s)_{s \in I}$, the natural filtration $(\mathcal{F}_t^0 : t \in I)$ is

$$\mathcal{F}_t^0 = \sigma(\{X_s : s \leq t, s \in I\}).$$

Definition 2.2.3. A stochastic process $(X_s)_{s \in I}$ has the Markov property with respect to the filtration $(\mathcal{F}_s : s \in I)$ if $(X_s)_{s \in I}$ is adapted to $(\mathcal{F}_s : s \in I)$ and for all $s, t \in I$ with $s < t$ we have

$$\mathbb{P}(X_t \in A | \mathcal{F}_s) = \mathbb{P}(X_t \in A | X_s)$$

for every $A \in \mathcal{B}(\mathcal{X})$.

Remark 2.2.4.

1. Definition 2.2.3 says that our prediction of the future using information from the past and the present is as good as our prediction of the future using only information from the present.
2. We can reformulate the condition of Definition 2.2.3 as for all $s, t \in I$ with $s < t$, $A \in \mathcal{B}(\mathcal{X})$ and $C \in \mathcal{F}_s$ we have

$$\mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) = \mathbb{E}(\mathbb{P}(X_t \in A | X_s)\mathbf{1}_C).$$

Exercise 2.2.5. Verify that the reformulation of Definition 2.2.3 with statement 2 of Remark 2.2.4 is justified.

Definition 2.2.6. A non-empty collection of subsets \mathcal{D} , is a π -system if for any $A, B \in \mathcal{D}$ we have $A \cap B \in \mathcal{D}$.

Definition 2.2.7. A non-empty collection of subsets \mathcal{G} , is a λ -system if the following properties hold.

1. $\emptyset \in \mathcal{G}$.
2. For any $G \in \mathcal{G}$ we have $G^c \in \mathcal{G}$.
3. For $G_1 \subseteq G_2 \subseteq \dots \in \mathcal{G}$ we have $\bigcup_{i=1}^{\infty} G_i \in \mathcal{G}$.

Theorem 2.2.8 (Dynkin π - λ). If \mathcal{D} is a π -system and \mathcal{G} is a λ -system with $\mathcal{D} \subseteq \mathcal{G}$. Then $\sigma(\mathcal{D}) \subseteq \mathcal{G}$.

Proposition 2.2.9. Suppose a π -system \mathcal{D} generates $\mathcal{B}(\mathcal{X})$ and for each $s \in I$ we are given a π -system \mathcal{D}_s that generates \mathcal{F}_s . Moreover, suppose that for every $s < t$, $A \in \mathcal{D}$ and $C \in \mathcal{D}_s$ we have

$$\mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) = \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}_C).$$

Then $(X_s)_{s \in I}$ has the Markov property.

Proof. For fixed $s, t \in I$ such that $s < t$ consider the set

$$\mathcal{C} := \{C \in \mathcal{F}_s : \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) = \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}_C) \text{ for all } A \in \mathcal{D}\}.$$

Then $\emptyset \in \mathcal{C}$ and $\mathcal{X} \in \mathcal{C}$. If $C \in \mathcal{C}$ then $\mathbf{1}_{C^c} = \mathbf{1} - \mathbf{1}_C$, so,

$$\begin{aligned} \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_{C^c}) &= \mathbb{E}(\mathbf{1}_A(X_t)(\mathbf{1} - \mathbf{1}_C)) \\ &= \mathbb{E}(\mathbf{1}_A(X_t)) - \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) \\ &= \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}) - \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}_C) \\ &= \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}_{C^c}). \end{aligned}$$

Hence, $C^c \in \mathcal{C}$. Next, if $(C_n)_{n \in \mathbb{N}} \subseteq \mathcal{C}$ is such that $C_1 \subseteq C_2 \subseteq \dots$, then by the monotone convergence theorem it follows that $\bigcup_{i=1}^{\infty} C_i \in \mathcal{C}$. Therefore, \mathcal{C} defines a λ -system. As $\mathcal{D}_s \subseteq \mathcal{C}$ we can apply the Theorem 2.2.8 to infer that $\mathcal{F}_s = \sigma(\mathcal{D}_s) \subseteq \mathcal{C} \subseteq \mathcal{F}_s$ which implies that $\mathcal{F}_s = \mathcal{C}$. Similarly, we can consider

$$\mathcal{A} = \{A \in \mathcal{B}(\mathcal{X}) : \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) = \mathbb{E}(\mathbb{P}(X_t \in A|X_s)\mathbf{1}_C) \text{ for all } C \in \mathcal{F}_s\},$$

and show that \mathcal{A} defines a λ -system with $\mathcal{D} \subseteq \mathcal{A}$. Therefore, by Theorem 2.2.8 system we infer that $\mathcal{B}(\mathcal{X}) = \sigma(\mathcal{D}) \subseteq \mathcal{A} \subseteq \mathcal{B}(\mathcal{X})$. Therefore, $\mathcal{A} = \mathcal{B}(\mathcal{X})$ and so $(X_s)_{s \in I}$ has the Markov property. \square

Proposition 2.2.10. If $(X_s : s \in I)$ is a Markov process with respect to a filtration $(\mathcal{F}_s : s \in I)$, then it is a Markov process with respect to its natural filtration $(\mathcal{F}_s^0 : s \in I)$.

Proof. Note that for any $s < t$ we have $\sigma(X_s) \subseteq \mathcal{F}_s^0 \subseteq \mathcal{F}_s$. Therefore, for $t \in I$ such that $t > s$ it follows that

$$\begin{aligned} \mathbb{P}(X_t \in A | \mathcal{F}_s^0) &\stackrel{(1)}{=} \mathbb{E}(\mathbb{P}(X_t \in A | \mathcal{F}_s) | \mathcal{F}_s^0) \\ &\stackrel{(2)}{=} \mathbb{E}(\mathbb{P}(X_t \in A | X_s) | \mathcal{F}_s^0) \\ &\stackrel{(3)}{=} \mathbb{P}(X_t \in A | X_s), \end{aligned}$$

where (1) is an application of the tower property, (2) is applying the Markov property of $(X_s)_{s \in I}$ with respect to $(\mathcal{F}_s)_{s \in I}$ and (3) follows by the tower property. \square

Theorem 2.2.11. *The stochastic process $(X_s : s \in I)$ is a Markov process with filtration $(\mathcal{F}_s : s \in I)$ if and only if for any bounded measurable function $f : \mathcal{X} \rightarrow \mathbb{R}$ and $s < t$ we have that*

$$\mathbb{E}(f(X_t)|\mathcal{F}_s) = \mathbb{E}(f(X_t)|X_s).$$

Proof. (\Rightarrow). Step 1: Let $f = \mathbf{1}_A$ for $A \in \mathcal{B}(\mathcal{X})$.
Then

$$\begin{aligned} \mathbb{E}(f(X_t)|\mathcal{F}_s) &= \mathbb{P}(X_t \in A|\mathcal{F}_s) \\ &= \mathbb{P}(X_t \in A|X_s) \\ &= \mathbb{E}(f(X_t)|X_s). \end{aligned}$$

Step 2: Let f be a simple function.

By the linearity of expectation and step 1 the result extends to f .

Step 3: Let f be a non-negative bounded measurable function.

There is a sequence of simple functions $(f_n)_{n \in \mathbb{N}}$ such that $f_n \nearrow f$. Therefore, applying the monotone convergence theorem at the points (\star) we deduce that

$$\begin{aligned} \mathbb{E}(f(X_t)|\mathcal{F}_s) &= \mathbb{E}\left(\lim_{n \rightarrow \infty} f_n(X_t)|\mathcal{F}_s\right) \\ &\stackrel{(\star)}{=} \lim_{n \rightarrow \infty} (\mathbb{E}(f_n(X_t)|\mathcal{F}_s)) \\ &\stackrel{\text{Step 2}}{=} \lim_{n \rightarrow \infty} (\mathbb{E}(f_n(X_t)|X_s)) \\ &\stackrel{(\star)}{=} \mathbb{E}\left(\lim_{n \rightarrow \infty} f_n(X_t)|X_s\right) \\ &= \mathbb{E}(f(X_t)|X_s). \end{aligned}$$

For f an arbitrary measurable and bounded function, we can write $f = f^+ - f^-$ where f^+, f^- are bounded non-negative measurable functions. Hence, we can extend the equality to f using the linearity of expectation.

(\Leftarrow). For any $A \in \mathcal{B}(\mathcal{X})$ let $f = \mathbf{1}_A$, so that for $s < t$ we have

$$\begin{aligned} \mathbb{P}(X_t \in A|\mathcal{F}_s) &= \mathbb{E}(f(X_t)|\mathcal{F}_s) \\ &= \mathbb{E}(f(X_t)|X_s) \\ &= \mathbb{P}(X_t \in A|X_s). \end{aligned}$$

Therefore, $(X_s)_{s \in I}$ is a Markov process. □

2.3 Discrete Time Processes

Throughout, let I be a discrete well-ordered set. Furthermore, let $\mathcal{B}_b(\mathcal{X})$ be the set of bounded measurable functions on \mathcal{X} .

Theorem 2.3.1. *Suppose we have a process $(X_n)_{n=0}^\infty$ and $l < m < n$. Then the following are equivalent.*

1. For every $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_n)|X_l, X_m) = \mathbb{E}(f(X_n)|X_m).$$

2. For every $g \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(g(X_l)|X_m, X_n) = \mathbb{E}(g(X_l)|X_m).$$

3. For every $f, g \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_n)g(X_l)|X_m) = \mathbb{E}(f(X_n)|X_m)\mathbb{E}(g(X_l)|X_m).$$

Proof. Reversing the time of the process, we see that statement 3 remains the same whereas statements 1 and 2 switch. Therefore, it is sufficient to show the equivalence of statement 1 and statement 3. (1 \Rightarrow 3). Note that,

$$\begin{aligned}\mathbb{E}(f(X_n)g(X_l)|X_m) &= \mathbb{E}(\mathbb{E}(f(X_n)g(X_l)|X_m, X_l) | X_m) \\ &= \mathbb{E}(g(X_l)\mathbb{E}(f(X_n)|X_m, X_l) | X_m) \\ &\stackrel{1}{=} \mathbb{E}(g(X_l)\mathbb{E}(f(X_n)|X_m) | X_m) \\ &= \mathbb{E}(g(X_l)|X_m)\mathbb{E}(f(X_n)|X_m).\end{aligned}$$

(3 \Rightarrow 1). Consider $g, h \in \mathcal{B}_b(\mathcal{X})$. Then

$$\begin{aligned}\mathbb{E}(f(X_n)g(X_l)h(X_m)) &= \mathbb{E}(h(X_m)\mathbb{E}(f(X_n)g(X_l)|X_m)) \\ &\stackrel{3}{=} \mathbb{E}(h(X_m)\mathbb{E}(f(X_n)|X_m)\mathbb{E}(g(X_l)|X_m)) \\ &= \mathbb{E}(\mathbb{E}(g(X_l)h(X_m)\mathbb{E}(f(X_n)|X_m) | X_m)) \\ &= \mathbb{E}(g(X_l)h(X_m)\mathbb{E}(f(X_n)|X_m)).\end{aligned}$$

Therefore,

$$\int_A f(X_n) d\mathbb{P} = \int_A \mathbb{E}(f(X_n)|X_m) d\mathbb{P}$$

for every $A = A_1 \cap A_2$ where $A_1 \in \sigma(X_l)$ and $A_2 \in \sigma(X_m)$. Hence,

$$\mathbb{E}(f(X_n)|X_l, X_m) = \mathbb{E}(f(X_n)|X_m).$$

□

Remark 2.3.2.

- Statement 3 of Theorem 2.3.1 says that the future of the process is independent of its past, provided we know the present.
- Theorem 2.3.1 holds for generic stochastic processes. The statements are weaker than the Markov property as they only consider three points in time.

Lemma 2.3.3. Let $(X_n)_{n=0}^{\infty}$ be a Markov process, and suppose $t_1 < \dots < t_m = k$. Let $f, h \in \mathcal{B}_b(\mathcal{X})$. Then

$$\mathbb{E}(f(X_{k+2})h(X_{k+1})|X_{t_1}, \dots, X_{t_m}) = \mathbb{E}(f(X_{k+2})h(X_{k+1})|X_{t_m}).$$

Proof. Let $\mathcal{G} = \sigma(X_{t_1}, \dots, X_{t_m})$. Then

$$\begin{aligned}\mathbb{E}(f(X_{k+2})h(X_{k+1})|\mathcal{G}) &\stackrel{(1)}{=} \mathbb{E}(\mathbb{E}(\mathbb{E}(f(X_{k+2})h(X_{k+1})|\mathcal{F}_{k+1}^0) | \mathcal{F}_k^0) | \mathcal{G}) \\ &\stackrel{(2)}{=} \mathbb{E}(\mathbb{E}(\mathbb{E}(f(X_{k+2})h(X_{k+1})|\mathcal{F}_{k+1}^0) | X_k) | \mathcal{G}) \\ &\stackrel{(3)}{=} \mathbb{E}(\mathbb{E}(f(X_{k+2})h(X_{k+1})|X_k) | \mathcal{G}) \\ &\stackrel{(4)}{=} \mathbb{E}(f(X_{k+2})h(X_{k+1})|X_k),\end{aligned}$$

where the equalities are justified in the following ways.

- (1) comes from the tower property for conditional expectations applied to $\mathcal{F}_{k+1}^0 \supseteq \mathcal{F}_k^0 \supseteq \mathcal{G}$.

- (2) comes from the Markov property, as $\mathbb{E}(f(X_{k+2})h(X_{k+1})|\mathcal{F}_{k+1}^0)$ is a bounded measurable function of X_{k+1} , say $g(X_{k+1})$ such that $\mathbb{E}(g(X_{k+1})|\mathcal{F}_k^0) = \mathbb{E}(g(X_{k+1})|X_k)$.
- (3) comes from the tower property for conditional expectations applied to $\mathcal{F}_{k+1}^0 \supseteq \sigma(X_k)$.
- (4) comes from the tower property for conditional expectations applied to $\mathcal{G} \supseteq \sigma(X_k)$.

□

Corollary 2.3.4. Let $(X_n)_{n=0}^\infty$ be a Markov process, and suppose $t_1 < \dots < t_m = k$. Then for $A \in \mathcal{B}(\mathcal{X})$ we have that

$$\mathbb{P}(X_{k+2} \in A | X_{t_1}, \dots, X_{t_m}) = \mathbb{P}(X_{k+2} \in A | X_{t_m}).$$

Proof. This follows by taking $h = 1$ and $f = \mathbf{1}_A$ in Lemma 2.3.3. □

Through induction, Corollary 2.3.4 extends to larger difference time steps.

Exercise 2.3.5. Let $(X_n)_{n=0}^\infty$ be a Markov process, and suppose $s_1 < \dots < s_m < t_1 < \dots < t_n$ and $f_i \in \mathcal{B}_b(\mathcal{X})$ for $1 \leq i \leq n$. Then

$$\mathbb{E} \left(\prod_{i=1}^n f_i(X_{t_i}) | X_{s_1}, \dots, X_{s_m} \right) = \mathbb{E} \left(\prod_{i=1}^n f_i(X_{t_i}) | X_{s_m} \right).$$

Proposition 2.3.6. A process $(X_n)_{n=0}^\infty$ is a Markov process with respect to its natural filtration if and only if one of the following conditions holds.

1. For any $A_i \in \mathcal{B}(\mathcal{X})$ we have

$$\mathbb{P}(X_0 \in A_0, \dots, X_n \in A_n) = \int_{\Omega} \mathbb{P}(X_n \in A_n | X_{n-1})(\omega) \mathbf{1} \left(\left\{ (X_i)_{i=0}^{n-1} \in \prod_{i=0}^{n-1} A_i \right\} \right) (\omega) d\mathbb{P}(\omega).$$

2. For every $n \in \mathbb{N}$ and $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_n) | \mathcal{F}_{n-1}^0) = \mathbb{E}(f(X_n) | X_{n-1}).$$

3. For any $n \in \mathbb{N}$ and $f_i \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E} \left(\prod_{i=1}^n f_i(X_i) \right) = \mathbb{E} \left(\prod_{i=1}^{n-1} f_i(X_i) \mathbb{E}(f_n(X_n) | X_{n-1}) \right).$$

Proof.

1. (\Rightarrow). For $C = \{X_0 \in A_0, \dots, X_{n-1} \in A_{n-1}\} \in \mathcal{F}_{n-1}$ observe that

$$\begin{aligned} \mathbb{P}(X_0 \in A_0, \dots, X_n \in A_n) &= \mathbb{E}(\mathbf{1}_{A_n}(X_n) \mathbf{1}_C) \\ &= \mathbb{E}(\mathbb{P}(X_n \in A_n | X_{n-1}) \mathbf{1}_C) \\ &= \int_{\Omega} \mathbb{P}(X_n \in A_n | X_{n-1})(\omega) \mathbf{1}_C(\omega) d\mathbb{P}(\omega) \\ &= \int_{\Omega} \mathbb{P}(X_n \in A_n | X_{n-1})(\omega) \mathbf{1} \left(\left\{ (X_i)_{i=0}^{n-1} \in \prod_{i=0}^{n-1} A_i \right\} \right) (\omega) d\mathbb{P}(\omega). \end{aligned}$$

(\Leftarrow). Let $C = \{X_0 \in A_0, \dots, X_{n-1} \in A_{n-1}\}$. Then

$$\mathbb{E}(\mathbf{1}_{A_n}(X_n)\mathbf{1}_C) = \mathbb{E}(\mathbb{P}(X_n \in A_n | X_{n-1})\mathbf{1}_C).$$

As sets of the form C form a π -system that generates \mathcal{F}_{n-1} , it follows by Proposition 2.2.9 that Markov property holds for every $C \in \mathcal{F}_{n-1}$.

2. Follows the same arguments as those made in Theorem 2.2.11.

3. (\Rightarrow). Using Proposition 1.4.11 it follows that

$$\begin{aligned} \mathbb{E}\left(\prod_{i=1}^n f_i(x_i)\right) &= \mathbb{E}\left(\mathbb{E}\left(\prod_{i=1}^n f_i(X_i) \middle| \mathcal{F}_{n-1}\right)\right) \\ &= \mathbb{E}\left(\prod_{i=1}^{n-1} f_i(X_i) \mathbb{E}(f(X_n) | \mathcal{F}_{n-1})\right) \\ &= \mathbb{E}\left(\prod_{i=1}^{n-1} f_i(X_i) \mathbb{E}(f(X_n) | X_{n-1})\right). \end{aligned}$$

(\Leftarrow). For any $A_i \in \mathcal{B}(\mathcal{X})$ let $f_i = \mathbf{1}_{A_i}$. Then

$$\begin{aligned} \mathbb{P}(X_0 \in A_0, \dots, X_n \in A_n) &= \mathbb{E}\left(\prod_{i=1}^n f_i(X_i)\right) \\ &= \mathbb{E}\left(\prod_{i=1}^{n-1} f_i(X_i) \mathbb{E}(f_n(X_n) | X_{n-1})\right) \\ &= \int_{\Omega} \mathbb{P}(f(X_n) | X_{n-1})(\omega) \prod_{i=1}^{n-1} f_i(X_i)(\omega) d\mathbb{P}(\omega) \\ &= \int_{\Omega} \mathbb{P}(X_n \in A_n | X_{n-1})(\omega) \mathbf{1}\left(\left\{(X_i)_{i=0}^{n-1} \in \prod_{i=0}^{n-1} A_i\right\}\right)(\omega) d\mathbb{P}(\omega). \end{aligned}$$

Therefore, we deduce using statement 1 that $(X_n)_{n=0}^{\infty}$ has the Markov property. □

Exercise 2.3.7.

1. Suppose $(X_n)_{n=0}^{\infty}$ is a bounded \mathbb{R} -valued process which is Markov with respect to its natural filtration. Let $g \in \mathcal{B}_b(\mathbb{R})$ be a injective function. Show that $(Y_n)_{n=0}^{\infty}$ given by $Y_n = g(X_n)$ is Markov with respect to its natural filtration.
2. Show that the above statement is not true if we remove the assumption that g is injective.

2.4 Solution to Exercises

Exercise 2.2.5

Solution. Suppose the formulation of Definition 2.2.3 holds. Let $s, t \in I$, $A \in \mathcal{B}(\mathcal{X})$ and $C \in \mathcal{F}_s$. Then

$$\begin{aligned} \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C) &= \mathbb{E}(\mathbb{E}(\mathbf{1}_A(X_t) | \mathcal{F}_s)\mathbf{1}_C) \\ &= \mathbb{E}(\mathbb{P}(X_t \in A | \mathcal{F}_s)\mathbf{1}_C) \\ &= \mathbb{E}(\mathbb{P}(X_t \in A | X_s)\mathbf{1}_C). \end{aligned}$$

On the other hand, suppose the formulation of statement 2 of Remark 2.2.4 holds. Note that since $\mathbb{P}(X_t \in A|\mathcal{F}_s)$ and $\mathbb{P}(X_t \in A|X_s)$ are \mathcal{F}_s -measurable we have

$$C := \{\omega : \mathbb{P}(X_t \in A|\mathcal{F}_s)(\omega) \geq \mathbb{P}(X_t \in A|X_s)(\omega)\} \in \mathcal{F}_s$$

Therefore,

$$\begin{aligned} \mathbb{E}(|\mathbb{P}(X_t \in A|\mathcal{F}_s) - \mathbb{P}(X_t \in A|X_s)|) &= \mathbb{E}(|\mathbb{P}(X_t \in A|\mathcal{F}_s) - \mathbb{P}(X_t \in A|X_s)|(\mathbf{1}_C + \mathbf{1}_{C^c})) \\ &= \mathbb{E}((\mathbb{P}(X_t \in A|\mathcal{F}_s) - \mathbb{P}(X_t \in A|X_s))\mathbf{1}_C) \\ &\quad + \mathbb{E}((\mathbb{P}(X_t \in A|X_s) - \mathbb{P}(X_t \in A|\mathcal{F}_s))\mathbf{1}_{C^c}) \\ &= \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_C - \mathbb{P}(X_t \in A|X_s)\mathbf{1}_C) \\ &\quad + \mathbb{E}(\mathbf{1}_A(X_t)\mathbf{1}_{C^c} - \mathbb{P}(X_t \in A|X_s)\mathbf{1}_{C^c}) \\ &= 0. \end{aligned}$$

Therefore, $\mathbb{P}(X_t \in A|\mathcal{F}_s) = \mathbb{P}(X_t \in A|X_s)$. □

Exercise 2.3.5

Solution. Suppose that $t_1 = s_m + 1$. Let $\mathcal{G} = \sigma(X_{s_1}, \dots, X_{s_m})$. Then following the same arguments as those made in the proof of Lemma 2.3.3 we deduce that

$$\begin{aligned} \mathbb{E}\left(\prod_{i=1}^n f_i(X_{t_i})|\mathcal{G}\right) &= \mathbb{E}\left(\mathbb{E}\left(\mathbb{E}\left(\prod_{i=1}^n f_i(X_{t_i})|\mathcal{F}_{t_1}^0\right)|\mathcal{F}_{s_m}^0\right)|\mathcal{G}\right) \\ &= \mathbb{E}\left(\mathbb{E}\left(\mathbb{E}\left(\prod_{i=1}^n f_i(X_{t_i})|\mathcal{F}_{t_1}^0\right)|X_{s_m}\right)|\mathcal{G}\right) \\ &= \mathbb{E}\left(\prod_{i=1}^n f_i(X_{t_i})|X_{s_m}\right). \end{aligned}$$

When $t_1 \neq s_m + 1$, one just takes $f_i = 1$ for $i \in \{s_m + 1, \dots, t_1 - 1\}$. □

Exercise 2.3.7

Solution.

- Let $A \in \mathcal{B}(\mathbb{R})$. As g is injective we know that

$$\{Y_n \in A\} = \{X_n \in g^{-1}(A)\}.$$

Therefore,

$$\begin{aligned} \mathbb{P}(Y_{n+1} \in A|X_0, \dots, X_n) &= \mathbb{P}(X_{n+1} \in g^{-1}(A)|X_0, \dots, X_n) \\ &\stackrel{(1)}{=} \mathbb{P}(X_{n+1} \in g^{-1}(A)|X_n) \\ &= \mathbb{P}(Y_{n+1} \in A|X_n), \end{aligned}$$

where (1) is an application of the Markov property of $(X_n)_{n=0}^\infty$ with respect to its natural filtration. Consequently, $(Y_n)_{n=0}^\infty$ is Markov with respect to the natural filtration of $(X_n)_{n=0}^\infty$ so by a discrete-time analogue of Proposition 2.2.10 $(Y_n)_{n=0}^\infty$ is Markov with respect to its own natural filtration.

- Consider the symmetric random walk, $(X_n)_{n=0}^\infty$, on \mathbb{Z} , which is a Markov process. Let

$$Y_n = g(X_n) = \left\lfloor \frac{X_n}{3} \right\rfloor.$$

Note $\mathbb{P}(Y_n = 0|Y_{n-1} = 1) > 0$ as we could have $X_{n-1} = 3$ and then $X_n = 2$. However, $\mathbb{P}(Y_n = 0|Y_{n-1} = 1, Y_{n-2} = 2) = 0$ as $Y_{n-2} = 2$ implies that $X_{n-2} \in \{6, 7, 8\}$ but $Y_n = 0$ implies that $X_n \in \{0, 1, 2\}$. We cannot transition between these sets with two steps and hence the probability is zero. Note that the conditioning is well-defined as $\{Y_{n-1} = 1, Y_{n-2} = 2\}$ is an event with non-zero probability. Therefore, the process $(Y_n)_{n=0}^\infty$ is not Markovian. □

3 The Kolmogorov Extension Theorem

3.1 Stochastic Processes as Random Variables

We can view a stochastic process $(X_i)_{i=0}^{\infty}$ as a $\mathcal{X}^{\mathbb{N}}$ -valued random variable, instead of as a sequence of \mathcal{X} -valued random variables. Using this interpretation we have a map $\Omega \rightarrow \mathcal{X}^{\mathbb{N}}$ given by

$$\omega \mapsto (X_i(\omega))_{i=0}^{\infty}.$$

For an index set Λ , each $m \in \Lambda$ has a projection map $\pi_m : \prod_{i \in \Lambda} \mathcal{X}_i \rightarrow \mathcal{X}_m$ given by

$$\prod_{i \in \Lambda} \mathcal{X}_i \ni a = (a_i : a \in \Lambda) \mapsto \pi_m(a) = a_m \in \mathcal{X}_m.$$

Definition 3.1.1. Given an index set Λ and measurable spaces $(\mathcal{X}_i, \mathcal{F}^{(i)})$ for each $i \in \Lambda$, the product σ -algebra, denoted $\bigotimes_{i \in \Lambda} \mathcal{F}^{(i)}$, is the smallest σ -algebra on $\prod_{i \in \Lambda} \mathcal{X}_i$ such that the projection maps π_m for $m \in \Lambda$ are measurable. That is,

$$\bigotimes_{i \in \Lambda} \mathcal{F}^{(i)} = \sigma \left(\left\{ \pi_m^{-1}(A_m) : A_m \in \mathcal{F}^{(m)}, m \in \Lambda \right\} \right).$$

Remark 3.1.2.

1. Sets that are finite intersections of sets of the form $\pi_m^{-1}(A_m)$ with $m \in \Lambda$ and $A_m \in \mathcal{F}^{(m)}$ are called cylinder sets.
 - (a) Cylinder sets generate $\bigotimes_{i \in \Lambda} \mathcal{F}^{(i)}$.
 - (b) Cylinder sets $A \subseteq \prod_{i \in \Lambda} \mathcal{X}_i$ are of the form $A = \prod_{i \in \Lambda} A_i$ where each $A_i \in \mathcal{F}^{(i)}$ and all but finitely many $A_i \neq \mathcal{X}_i$.
2. When Λ is countable, we know that $\prod_{i \in \Lambda} E_i$ for measurable sets $E_i \in \mathcal{F}^{(i)}$ is measurable.
3. When Λ is uncountable, we cannot assume that $\prod_{i \in \Lambda} E_i$ for $E_i \in \mathcal{F}^{(i)}$ is measurable.

3.2 Constructing Stochastic Processes

Throughout, we will be working with a discrete-time stochastic process, and usually have $\prod_{i \in \Lambda} \mathcal{X}_i = \mathcal{X}^{\mathbb{N}}$ with \mathcal{X} a complete metric space.

Proposition 3.2.1. Suppose that we have a countable product of measurable spaces $\prod_{i=0}^{\infty} \mathcal{X}_i$ where each \mathcal{X}_i is equipped with the σ -algebra $\mathcal{F}^{(i)}$. Also suppose for each $i \in \mathbb{N}$ we have $\mathcal{F}^{(i)} = \sigma(\mathcal{D}_i)$, then

$$\bigotimes_{i=0}^{\infty} \mathcal{F}^{(i)} = \sigma \left(\prod_{i=0}^{\infty} E_i : E_i \in \mathcal{D}_i \right).$$

If the \mathcal{X}_i are separable metric spaces, then the Borel σ -algebra of the product topological space $\prod_{i=0}^{\infty} \mathcal{X}_i$ is the product of Borel σ -algebras of the \mathcal{X}_i , that is

$$\mathcal{B} \left(\prod_{i=0}^{\infty} \mathcal{X}_i \right) = \bigotimes_{i=0}^{\infty} \mathcal{B}(\mathcal{X}_i).$$

Corollary 3.2.2. *If $(X_n)_{n=0}^\infty$ is a \mathcal{X} -valued stochastic process, the sequence $(X_n)_{n=0}^\infty$ is a $(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathcal{X}^\mathbb{N}))$ random variable. Moreover, the process $(X_n)_{n=0}^\infty$ induces a probability measure $\text{Law}((X_n)_{n=0}^\infty)$ on $\mathcal{B}(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathcal{X}^\mathbb{N}))$.*

Supposing we have a stochastic process $(X_n)_{n=0}^\infty$, we construct a canonical probability space for the process as $(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathcal{X}^\mathbb{N}), \text{Law}((X_n)_{n=0}^\infty))$. Therefore, if we are only interested in the process $(X_n)_{n=0}^\infty$ we can forget about $(\Omega, \mathcal{F}, \mathbb{P})$ and focus on this canonical space instead. Note that $(X_n)_{n=0}^\infty$ corresponds to the identity map on this space. Supposing we do not have a stochastic process, we can only guess what its law should look like. With the equivalence observation we made above, we can construct this process by constructing its canonical probability space. More specifically, if we know the values that $(X_n)_{n=0}^\infty$ takes in $\mathcal{X}^\mathbb{N}$, then we just need to build $\text{Law}((X_n)_{n=0}^\infty)$ on $\mathcal{B}(\mathcal{X}^\mathbb{N})$.

Definition 3.2.3. *A finite dimensional distribution on the first $n + 1$ time steps, $\{0, \dots, n\}$, is a measure μ_n on $(\mathcal{X}^{n+1}, \mathcal{B}(\mathcal{X}^{n+1}))$.*

Finite-dimensional distributions constitute our guess on what the law of our process should look like.

Definition 3.2.4. *A family of probability measures $(\mu_n)_{n=0}^\infty$, where each μ_n a measure on $(\mathcal{X}^{n+1}, \mathcal{B}(\mathcal{X}^{n+1}))$, is consistent if for any $n \in \mathbb{N}$ and $A_0, \dots, A_n \in \mathcal{B}(\mathcal{X})$ we have*

$$\mu_n(A_0 \times \dots \times A_n) = \mu_{n+1}(A_0 \times \dots \times A_n \times \mathcal{X}).$$

That is to say the measures are marginals of each other.

Example 3.2.5. *Let $(X_n)_{n=0}^\infty$ be a \mathcal{X} -valued stochastic process on a probability space $(\Omega, \mathcal{F}, \mathbb{P})$. If we set $\mu_n = \text{Law}((X_j)_{j=0}^n)$, then $(\mu_n)_{n=0}^\infty$ is a consistent family of finite-dimensional distributions.*

Remark 3.2.6. *Example 3.2.5 shows that for our guess of finite-dimensional distributions to yield a valid stochastic process, we ought to ensure our guess is consistent.*

Theorem 3.2.7 (Kolmogorov's Extension). *Let $(\mu_n)_{n=0}^\infty$ be a consistent family of finite-dimensional distributions. Then there exists a unique probability measure μ on $(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathcal{X}^\mathbb{N}))$ such that for any $n \in \mathbb{N}$ and $A \in \mathcal{B}(\mathcal{X}^{n+1})$ we have*

$$\mu \left(A \times \left(\prod_{i=n+1}^\infty \mathcal{X} \right) \right) = \mu_n(A).$$

Corollary 3.2.8. *Given a stochastic process $(X_n)_{n=0}^\infty$, the family of laws*

$$(\text{Law}(X_0, \dots, X_n))_{n=0}^\infty$$

uniquely characterises $\text{Law}((X_n)_{n=0}^\infty)$.

From Theorem 3.2.7, we can construct a stochastic process from a consistent family of finite-dimensional distributions.

Corollary 3.2.9. *Given any consistent family of finite dimensional distributions $(\mu_n)_{n=0}^\infty$ there exists a process $(X_n)_{n=0}^\infty$ with*

$$\text{Law}(X_0, \dots, X_n) = \mu_n$$

for every $n \in \mathbb{N}$.

3.3 Stationarity

Working with sequences of random variables gives us a way to define stationarity.

Definition 3.3.1. For each $n \in \mathbb{N}$, the shift map is the $\mathcal{B}(\mathcal{X}^{\mathbb{N}})$ -measurable function $\theta_n : \mathcal{X}^{\mathbb{N}} \rightarrow \mathcal{X}^{\mathbb{N}}$ given by

$$(a_0, a_1, \dots) \mapsto (a_n, a_{n+1}, \dots).$$

Definition 3.3.2. A stochastic process $X = (X_n)_{n=0}^{\infty}$ is stationary if for all $n \in \mathbb{N}$ the processes $\theta_n X$ and X have the same law.

4 Transition Probabilities

Our method for constructing stochastic processes in Section 3 does not guarantee the Markov property is satisfied by the resulting sequence. Indeed, the self-avoiding random walk is a process with consistent finite-dimensional distributions that is not a Markov process.

4.1 The Chapman-Kolmogorov Equation

Definition 4.1.1. A Markov process $(X_n)_{n=0}^\infty$ is time-homogeneous if for every $A \in \mathcal{B}(\mathcal{X})$ there is a choice of $P(\bullet, A) \in \mathcal{B}_b(\mathcal{X})$ such that

$$\mathbb{P}(X_{n+1} \in A | X_n = \bullet) \sim P(\bullet, A),$$

for every $n \in \mathbb{N}$, modulo $\text{Law}(X_n)$ -null sets.

Remark 4.1.2.

1. The main property of time-homogeneity is the independence of n . Many results we will discuss tend to hold without this assumption, however, the assumption simplifies notation.
2. The function $P(\bullet, A)$ describes the conditional probabilities and hence captures some information about the process. Eventually, we want to work back from this data and understand what properties $P(\bullet, A)$ needs to have to construct a Markov process.

Definition 4.1.3. A family $P = (P(x, A) : x \in \mathcal{X}, A \in \mathcal{B}(\mathcal{X}))$ is a family of transition probabilities if the following hold.

1. For each $x \in \mathcal{X}$, the function $P(x, \bullet)$ is a probability measure on $(\mathcal{X}, \mathcal{B}(\mathcal{X}))$.
2. For each $A \in \mathcal{B}(\mathcal{X})$, the function $x \mapsto P(x, A)$ is Borel measurable.

Remark 4.1.4. Equivalently, we can say that there exists a measurable map $P : \mathcal{X} \rightarrow \mathcal{P}(\mathcal{X})$ such that for all $A \in \mathcal{B}(\mathcal{X})$ and $x \in \mathcal{X}$ we have

$$(P(x))(A) = P(x, A).$$

Exercise 4.1.5. Consider the random dynamical system as given in Theorem 2.1.3 but with $F_n = F$ for all $n \in \mathbb{N}$. That is, $X_{n+1} = F(X_n, \zeta_{n+1})$ where the $(\zeta_n)_{n=0}^\infty$ are independent and identically distributed random variables also independent of X_0 . Let $\text{Law}(X_0) = \nu$ and $\text{Law}(\zeta_n) = \mu$. Show that $X = (X_n)_{n=0}^\infty$ is a time-homogeneous Markov process and compute its transition probabilities.

Exercise 4.1.6. Let P be the set of transition probability built from a time-homogeneous Markov process $X = (X_n)_{n=0}^\infty$. Show that $(X_{2n})_{n=0}^\infty$ and $(X_{3n})_{n=0}^\infty$ are time-homogeneous Markov processes and compute their transition probabilities P^2 and P^3 . Show that for all $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$P^2(x, A) = \int_{\mathcal{X}} P(y, A)P(x, dy) \tag{4.1.1}$$

and

$$P^3(x, A) = \int_{\mathcal{X}} P(y, A)P^2(x, dy) = \int_{\mathcal{X}} P^2(y, A)P(x, dy). \tag{4.1.2}$$

Equations (4.1.1) and (4.1.2) are instances of the Chapman-Kolmogorov equation.

Recall, that we think about $P(x, A)$ as encoding conditional probabilities. That is,

$$P(x, A) = \mathbb{P}(X_{n+1} \in A | X_n = x).$$

As we are operating in the time-homogeneous setting, we can consider $P(x, A)$ as a one-step conditional probability.

Definition 4.1.7. A sequence of transition probabilities $(P^n)_{n=0}^\infty$ is a transition function if the following statements hold.

1. $P^0(x, \bullet) = \delta_x$. Here δ_x is the Dirac measure at x . That is,

$$\delta_x(A) = \begin{cases} 1 & x \in A \\ 0 & x \notin A. \end{cases}$$

2. The family satisfies the Chapman-Kolmogorov equations, that is, for every $n, m \in \mathbb{N}$, $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$P^{m+n}(x, A) = \int_{\mathcal{X}} P^n(y, A) P^m(x, dy).$$

Remark 4.1.8.

1. We can intuitively think of P^n as encoding an n -step conditional probability.
2. The Chapman-Kolmogorov equation says that a step into a set is consistent with taking smaller intermediary steps to get to the set.

A sequence of transition functions can be built from a one-step conditional probability, P . More specifically, given any transition probability P , we can construct a transition function $(P^n)_{n=0}^\infty$ by

1. setting $P^0(x, \bullet) = \delta_x$,
2. $P^1(x, \bullet) = P(x, \bullet)$, and
3. for $n > 1$ setting

$$P^n(x, A) = \int_{\mathcal{X}} P(y, A) P^{n-1}(x, dy).$$

4.2 Constructing Markov Processes

Theorem 4.2.1. Let $(X_n)_{n=0}^\infty$ be a time-homogeneous Markov process with transition probability P . Let $(P^n)_{n=0}^\infty$ be the transition function built from P . Then the following statements hold.

1. For any $n, m \geq 0$ and $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_{n+m}) | X_m) = \int_{\mathcal{X}} f(y) P^n(X_m, dy).$$

2. If $X_0 \sim \mu$, then for any $n \geq 0$ and $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_n)) = \int_{\mathcal{X}} \int_{\mathcal{X}} f(y) P^n(x, dy) \mu(dx).$$

Proof.

1. Proceed by induction on n .

- For $n = 0$ we have

$$\begin{aligned}\mathbb{E}(f(X_m)|X_m) &= f(X_m) \\ &= \int_{\mathcal{X}} f(y)P^0(X_m, dy).\end{aligned}$$

- Suppose the result holds for $n \leq k$. Then,

$$\begin{aligned}\mathbb{E}(f(X_{k+1+m})|X_m) &\stackrel{\text{Tot Prop.}}{=} \mathbb{E}(\mathbb{E}(f(X_{k+1+m})|\mathcal{F}_{k+m}^0)|X_m) \\ &\stackrel{\text{Markov.}}{=} \mathbb{E}(\mathbb{E}(f(X_{k+1+m})|X_{k+m})|X_m) \\ &\stackrel{\text{Time Hom.}}{=} \mathbb{E}\left(\int_{\mathcal{X}} f(y)P(X_{m+k}, dy)|X_m\right) \\ &\stackrel{\text{Ind Hyp.}}{=} \int_{\mathcal{X}} \int_{\mathcal{X}} f(y)P(z, dy)P^k(X_m, dz) \\ &\stackrel{\text{Fubini.}}{=} \int_{\mathcal{X}} f(y)P^{k+1}(X_m, dy).\end{aligned}$$

2. Using statement 1 it follows that

$$\begin{aligned}\mathbb{E}(f(X_n)) &= \mathbb{E}(\mathbb{E}(f(X_n)|X_0)) \\ &= \mathbb{E}\left(\int_{\mathcal{X}} f(y)P^n(X_0, dy)\right) \\ &= \int_{\mathcal{X}} \int_{\mathcal{X}} f(y)P^n(x, dy)\mu(dx),\end{aligned}$$

where the last equality follows as $\text{Law}(X_0) = \mu$. □

Remark 4.2.2.

1. For clarity we can take $f = \mathbf{1}_A$ with $A \in \mathcal{B}(\mathcal{X})$ so that the statements of Theorem 4.2.1 reduce to

- $\mathbb{P}(X_{n+m} \in A|X_m) = P^n(X_m, A)$, and
- $\mathbb{P}(X_n \in A) = \int_{\mathcal{X}} P^n(x, A)\mu(dx)$.

2. X being a Markov process is a sufficient condition for the conclusions Theorem 4.2.1, but it is not a necessary condition.

Proposition 4.2.3. Let $X = (X_n)_{n=0}^{\infty}$ be a process with $\mu = \text{Law}(X_0)$ and transition probability P . Then X is Markov if and only if for all $n \in \mathbb{N}$ and $f_i \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}\left(\prod_{i=0}^n f_i(X_i)\right) = \underbrace{\int_{\mathcal{X}} \cdots \int_{\mathcal{X}}}_{n+1} \prod_{i=0}^n f_i(y_i) \prod_{i=0}^{n-1} P(y_i, dy_{i+1})\mu(dy_0).$$

Proof. (\Rightarrow). We proceed by induction on n .

- For $n = 0$, as $\text{Law}(X_0) = \mu$ it follows that

$$\mathbb{E}(f_0(X_0)) = \int_{\mathcal{X}} f_0(y_0)\mu(dy_0).$$

- Now suppose the result holds for $k \leq n - 1$. Then

$$\begin{aligned}
\mathbb{E} \left(\prod_{i=0}^n f_i(X_i) \right) &\stackrel{\text{Tower.}}{=} \mathbb{E} \left(\mathbb{E} \left(\prod_{i=0}^n f_i(X_i) \mid \mathcal{F}_{n-1} \right) \right) \\
&= \mathbb{E} \left(\prod_{i=0}^{n-1} f_i(X_i) \mathbb{E}(f_n(X_n) \mid \mathcal{F}_{n-1}) \right) \\
&\stackrel{\text{Markov.}}{=} \mathbb{E} \left(\prod_{i=0}^{n-1} f_i(X_i) \mathbb{E}(f_n(X_n) \mid X_{n-1}) \right) \\
&\stackrel{(1)}{=} \mathbb{E} \left(\prod_{i=0}^{n-1} f_i(X_i) \int_{\mathcal{X}} f_n(y_n) P(X_{n-1}, dy_n) \right) \\
&\stackrel{\text{Ind Hyp.}}{=} \int_{\mathcal{X}} \cdots \int_{\mathcal{X}} \left(\prod_{i=0}^{n-1} f_i(y_i) \int_{\mathcal{X}} f_n(y_n) P(y_{n-1}, dy_n) \right) \prod_{i=0}^{n-2} P(y_i, dy_{i+1}) \mu(dy_0) \\
&= \int_{\mathcal{X}} \cdots \int_{\mathcal{X}} \prod_{i=0}^n f_i(y_i) \prod_{i=0}^{n-1} P(y_i, dy_{i+1}) \mu(dy_0),
\end{aligned}$$

where (1) follows as the conditional one-step probabilities is given by P .

(\Leftarrow). Refer to Proposition 2.3.6. □

By taking each f_i to be an indicator function we arrive at Corollary 4.2.4.

Corollary 4.2.4. *Let $X = (X_n)_{n=0}^{\infty}$ be a process with $\mu = \text{Law}(X_0)$ and transition probability P . Then X is Markov if and only if for all $n \in \mathbb{N}$ and $A_i \in \mathcal{B}_b(\mathcal{X})$ we have*

$$\mathbb{P}(X_0 \in A_0, \dots, X_n \in A_n) = \int_{A_0} \cdots \int_{A_n} \prod_{i=0}^{n-1} P(y_i, dy_{i+1}) \mu(dy_0).$$

Remark 4.2.5. *As the transition probability determines one-step conditional probabilities, the Markov process of Corollary 4.2.4 is time-homogeneous.*

Given P and μ for a Markov process X we can compute the finite-dimensional distributions of the process.

Proposition 4.2.6. *Given a transition probability P and measure μ on \mathcal{X} , there exists a unique (up to law) Markov process $X = (X_n)_{n=0}^{\infty}$ with transition probabilities P and $\text{Law}(X_0) = \mu$.*

Proof. Consider the sequence of measures $(\mu_n)_{n \in \mathbb{N}}$ on \mathcal{X}^n given by

$$\mu_n(A_0 \times \cdots \times A_{n-1}) = \int_{A_0} \cdots \int_{A_{n-1}} \prod_{i=1}^{n-1} P(y_{i-1}, dy_i) \mu(dy_0).$$

Observe that,

$$\begin{aligned}
\mu_{n+1}(A_0 \times \cdots \times A_{n-1} \times \mathcal{X}) &= \int_{A_0} \cdots \int_{A_{n-1}} \int_{\mathcal{X}} \prod_{i=1}^n P(y_{i-1}, dy_i) \mu(dy_0) \\
&= \int_{A_0} \cdots \int_{A_{n-1}} \int_{\mathcal{X}} P(y_{n-1}, dy_n) \prod_{i=1}^{n-1} P(y_{i-1}, dy_i) \mu(dy_0) \\
&= \int_{A_0} \cdots \int_{A_{n-1}} \prod_{i=1}^{n-1} P(y_{i-1}, dy_i) \mu(dy_0),
\end{aligned}$$

where the last equality follows from the fact the $P(y_{n-1}, dy_n)$ is a probability measure. Therefore, $(\mu_n)_{n \in \mathbb{N}}$ form a consistent sequence of measures, and so we can apply Theorem 3.2.7 to deduce that there exists a unique measure \mathbb{P}_μ on \mathcal{X}^∞ such that its restriction to \mathcal{X}^n is μ_n . Choosing $\Omega = \mathcal{X}^\infty$ and our probability measure to be \mathbb{P}_μ , the canonical process $(\pi_n)_{n \in \mathbb{N}}$ with $\pi_n((\omega_0, \omega_1, \dots)) = \omega_n$ is such that

$$\begin{aligned} \mathbb{P}(\pi_0 \in A_0, \dots, \pi_n \in A_n) &= \mathbb{P}_\mu(A_0 \times \dots \times A_n \times \mathcal{X}^\infty) \\ &= \mu_n(A_0 \times \dots \times A_n). \end{aligned}$$

Therefore, $(\pi_n)_{n \in \mathbb{N}}$ has finite-dimensional distributions $(\mu_n)_{n \in \mathbb{N}}$. Therefore, by Corollary 4.2.4 the process $(\pi_n)_{n \in \mathbb{N}}$ is a Markov process with transition probability P and $\text{Law}(X_0) = \mu$. \square

Remark 4.2.7.

1. When $\mu = \delta_x$ with $x \in \mathcal{X}$ we write \mathbb{P}_x instead of \mathbb{P}_μ . Similarly, we write \mathbb{E}_x or \mathbb{E}_μ to denote the expectation over $(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathcal{X}^\mathbb{N}))$ with respect to \mathbb{P}_x and \mathbb{P}_μ respectively.
2. Note that by construction, the Markov process of Proposition 4.2.6 is time-homogeneous.

We can see P as propagating the law μ through time. Hence, we can view the transition probability as an operator on $\mathcal{P}(\mathcal{X})$.

Definition 4.2.8. Given a transition probability P , we let $T^* : \mathcal{P}(\mathcal{X}) \rightarrow \mathcal{P}(\mathcal{X})$ be the operator where

$$\mu \mapsto (T^* \mu) = \int_{\mathcal{X}} P(x, \bullet) \mu(dx)$$

where

$$(T^* \mu)(A) = \int_{\mathcal{X}} P(x, A) \mu(dx).$$

Exercise 4.2.9. Formalise the intuition that P propagates the law of X_0 through time. That is, in the context of the Proposition 4.2.6 show that

$$\text{Law}(X_n) = T(\dots(T(\text{Law}(X_0))\dots)) = T^n \text{Law}(X_0).$$

Definition 4.2.10. A measure μ is invariant for the transition probability P if $T\mu = \mu$.

Example 4.2.11. Let X_0 have a Gaussian measure with mean zero and variance a^2 . Let $(\xi_n)_{n \in \mathbb{N}}$ be a sequence of independent and identically distributed Gaussian random variables, independent of X_0 , with mean zero and variance b^2 , where $a > b$. Then the Gaussian measure with mean zero and variance a^2 is invariant for the Markov process $(X_n)_{n=0}^\infty$ where

$$X_n = \frac{a^2 - b^2}{a^2} X_{n-1} + \xi_n$$

for $n \geq 1$.

- By construction $\text{Law}(X_0)$ is Gaussian with mean zero and variance a^2 .
- Suppose $\text{Law}(X_k)$ is Gaussian with mean zero and variance a^2 for $k \leq n - 1$. Then $\text{Law}(X_n)$ has mean zero and variance

$$\left(\frac{a^2 - b^2}{a^2}\right) a^2 + b^2 = a^2.$$

Remark 4.2.12. Let π be an invariant measure of P , then \mathbb{P}_π constructed by applying Proposition 4.2.6 to π is the law of a stationary process. Indeed, by the invariant nature of π , for $n \in \mathbb{N}$, the process $\theta_n X$ has transition probability P and initial distribution π . Therefore, by Proposition 4.2.6, the shifted process has law \mathbb{P}_π . Hence, the process is a stationary process.

Above we have considered P as an operator on measures. We can also view P as an operator on $\mathcal{B}_b(\mathcal{X})$.

Definition 4.2.13. Let P be a transition probability, we let $T_* : \mathcal{B}_b(\mathcal{X}) \rightarrow \mathcal{B}_b(\mathcal{X})$ be the operator where

$$f \mapsto (T_* f) = \int_{\mathcal{X}} f(y) P(\bullet, dy)$$

where

$$(T_* f)(x) = \int_{\mathcal{X}} f(y) P(x, dy).$$

Exercise 4.2.14. For a fixed transition probability P , show that the operators T^* and T_* are dual. That is, for any $f \in \mathcal{B}_b(\mathcal{X})$ and $\mu \in \mathcal{P}(\mathcal{X})$ we have

$$\int_{\mathcal{X}} (T_* f)(x) \mu(dx) = \int_{\mathcal{X}} f(x) (T^* \mu)(dx).$$

Remark 4.2.15. Henceforth, we will write $T^* = T$.

4.3 Solution to Exercises

Exercise 4.1.5

Solution. From Theorem 2.1.3 we know that the process is Markov. Moreover, we observe that

$$\begin{aligned} \mathbb{P}(X_{n+1} \in A | X_n) &= \mathbb{E}(\mathbf{1}_A(F(X_n, \zeta_{n+1}))) \\ &= \int_{\mathcal{Y}} \mathbf{1}_A(F(X_n, y)) \mu(dy) \\ &= \mu(\{y \in \mathcal{Y} : F(X_n, y) \in A\}), \end{aligned}$$

meaning the process is a time-homogeneous Markov process. In particular, $P(x, A) = \mu(\{y \in \mathcal{Y} : F(x, y) \in A\})$ because by the measurability assumptions on F , one can show that $P(x, \bullet)$ satisfies the required measurability properties. \square

Exercise 4.1.6

Solution. Observe that

$$\mathbb{P}(X_{2n+2} \in A | X_{2n}, X_{2n-2}, \dots, X_0) \stackrel{\text{Cor. 2.3.4}}{=} \mathbb{P}(X_{2n+2} \in A | X_{2n}).$$

Similarly, for $(X_{3n})_{n=0}^\infty$. Therefore, $(X_{2n})_{n=0}^\infty$ and $(X_{3n})_{n=0}^\infty$ are Markov processes. Moreover,

$$\begin{aligned} \mathbb{P}(X_{2n+2} \in A | X_{2n}) &\stackrel{(1)}{=} \mathbb{E}(\mathbb{P}(X_{2n+2} \in A | X_{2n+1}, X_{2n}) | X_{2n}) \\ &\stackrel{(2)}{=} \mathbb{E}(\mathbb{P}(X_{2n+2} \in A | X_{2n+1}) | X_{2n}) \\ &= \mathbb{E}(P(X_{2n+1}, A) | X_{2n}) \\ &\stackrel{(3)}{=} \int_{\mathcal{X}} P(y, A) P(X_{2n}, dy) \end{aligned}$$

where (1) is an application of the tower rule and (2) is applying the Markov property of $(X_n)_{n=0}^\infty$. Therefore,

$$\mathbb{P}(X_{2n+2} \in A | X_{2n} = x) = \int_{\mathcal{X}} P(y, A) P(x, dy).$$

In (3) we have used the fact that for f a measurable function we have

$$\mathbb{E}(f(X_{n+1}) | X_n = x) = \int_{\mathcal{X}} f(y) P(x, dy).$$

Similarly,

$$\begin{aligned} \mathbb{P}(X_{3n+3} \in A | X_{3n}) &= \mathbb{E}(\mathbb{P}(X_{3n+3} \in A | X_{3n+1}, X_{3n}) | X_{3n}) \\ &\stackrel{\text{Cor. 2.3.4}}{=} \mathbb{E}(\mathbb{P}(X_{3n+3} \in A | X_{3n+1}) | X_{3n}) \\ &= \mathbb{E}(P^2(X_{3n+1}, A) | X_{3n}) \\ &= \int_{\mathcal{X}} P^2(y, A) P(X_{3n}, dy), \end{aligned}$$

therefore,

$$P^3(x, A) = \int_{\mathcal{X}} P^2(y, A) P(x, dy).$$

□

Exercise 4.2.9

Solution. For ease of notation let $\text{Law}(X_n) = \mu_n$ for $n \in \mathbb{N}$. Proceed by induction on n .

- For $n = 1$ we have

$$\begin{aligned} \mu_1(A) &= \int_{\mathcal{X}} P(x, A) \mu_0(dx) \\ &= T\mu_0(A), \end{aligned}$$

hence, $\mu_1 = T\mu_0$.

- Suppose $\mu_k = T^k \mu_0$ for $k \leq n$. Then

$$\begin{aligned} \mu_{n+1}(A) &= \int_{\mathcal{X}} P^{n+1}(x, A) \mu_0(dx) \\ &= \int_{\mathcal{X}} P(y, A) P^n(x, dy) \mu_0(dx) \\ &= \int_{\mathcal{X}} P(y, A) \mu_n(dy) \\ &= \int_{\mathcal{X}} P(y, A) T^n \mu_0(dy) \\ &= T(T^n \mu_0)(A) \\ &= T^{n+1} \mu_0(A). \end{aligned}$$

Therefore, we conclude by induction that $\text{Law}(X_n) = T^n \text{Law}(X_0)$ for $n \in \mathbb{N}$.

□

Exercise 4.2.14

Proof. We proceed directly, noting that the equivalence arises from changing the order of integration,

$$\begin{aligned}\int_{x \in \mathcal{X}} f(x)(T^* \mu)(dx) &= \int_{x \in \mathcal{X}} f(x) \int_{y \in \mathcal{X}} P(y, dx) \mu(dy) \\ &= \int_{x \in \mathcal{X}} \int_{y \in \mathcal{X}} f(x) P(y, dx) \mu(dy) \\ &= \int_{y \in \mathcal{X}} (T_* f)(y) \mu(dy).\end{aligned}$$

□

5 Discrete Time Markov Processes

5.1 Time-Homogeneous Processes

On a discrete space, a probability measure is just a, potentially infinite, vector

$$\mu(dx) \longrightarrow \mu = (\mu_j : j \in \mathcal{X})$$

with $\mu_j \geq 0$ and $\sum_{j \in \mathcal{X}} \mu_j = 1$. In such a case, the transition probability P can be represented as a, potentially infinite, matrix

$$P(i, \{j\}) \longrightarrow P = (P_{ij} : i, j \in \mathcal{X})$$

where $P_{ij} \geq 0$ and $\sum_{j \in \mathcal{X}} P_{ij} = 1$. Note that if $X = (X_n)_{n=0}^{\infty}$ is a time-homogeneous Markov process with transition probability P then

$$\mathbb{P}(X_{n+1} = j | X_n = i) = P_{ij}.$$

Definition 5.1.1. A matrix $P = (P_{ij} : i, j \in \mathcal{X})$ is called a stochastic matrix if

1. $P_{ij} \geq 0$, and
2. $\sum_{j \in \mathcal{X}} P_{ij} = 1$.

For discrete state spaces \mathcal{X} , we have a direct correspondence between transition probabilities and stochastic matrices.

Exercise 5.1.2. Previously it was for shown that from a transition probability P , by setting $P^0(x, \bullet) = \delta_x$, we can construct a transition function by letting

$$P^n(x, A) = \int_{\mathcal{X}} P(y, A) P^{n-1}(x, dy).$$

In discrete spaces, show that $P^0 = I$, where I is the matrix with entries $I_{ij} = \delta_{ij}$, and

$$P^n = \underbrace{P \times \cdots \times P}_n$$

where \times is matrix multiplication. Moreover, verify that P^n is a stochastic matrix.

With Exercise 5.1.2, we can formulate our previous results in the specific context of a discrete state space.

Theorem 5.1.3. Let $(X_n)_{n=0}^{\infty}$ be a time-homogeneous discrete Markov process with stochastic matrix P . Then the following statements hold.

- For any $n, m \in \mathbb{N}$ and $f \in \mathcal{B}_b(\mathcal{X})$, we have

$$\mathbb{E}(f(X_{n+m}) | X_m = i) = \sum_{j \in \mathcal{X}} P_{ij}^n f(j).$$

- If $X_0 \sim \mu$, then for any $n \in \mathbb{N}$ and $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_n)) = \sum_{i, j \in \mathcal{X}} \mu_i P_{ij}^n f(j).$$

Again, we can contextualise the statements of Theorem 5.1.3 by letting $f = \delta_j$ for $j \in \mathcal{X}$. In this case

- $\mathbb{P}(X_{n+m} = j | X_m = i) = P_{ij}^n$, and
- $\mathbb{P}(X_n = j) = \sum_{i \in \mathcal{X}} \mu_i P_{ij}^n$.

Proposition 5.1.4. Let $X = (X_n)_{n=0}^\infty$ be a discrete process with $\mu = \text{Law}(X_0)$ and stochastic matrix P . Then X is Markov if and only if for all $n \in \mathbb{N}$ we have

$$\mathbb{E} \left(\prod_{m=0}^n f_m(X_m) \right) = \sum_{i_0, \dots, i_n \in \mathcal{X}} f_0(i_0) \dots f_n(i_n) \mu_{i_0} P_{i_0 i_1} \dots P_{i_{n-1} i_n},$$

where each $f_m \in \mathcal{B}_b(\mathcal{X})$.

Corollary 5.1.5. Let $X = (X_n)_{n=0}^\infty$ be a discrete process with $\mu = \text{Law}(X_0)$ and transition probability P . Then X is Markov if and only if for all $n \in \mathbb{N}$ and $i_0, \dots, i_n \in \mathcal{X}$ we have

$$\mathbb{P}(X_0 = i_0, \dots, X_n \in i_n) = \mu_{i_0} P_{i_0 i_1} \dots P_{i_{n-1} i_n}.$$

In the discrete settings the operator $T : \mathcal{P}(\mathcal{X}) \rightarrow \mathcal{P}(\mathcal{X})$ manifests as

$$(T\mu)(\{j\}) = \sum_{i \in \mathcal{X}} \mu_i P_{ij} = (\mu P)_j.$$

Similarly, the operator $T_* : \mathcal{B}_b(\mathcal{X}) \rightarrow \mathcal{B}_b(\mathcal{X})$ is given by

$$(T_* f)(i) = \sum_{j \in \mathcal{X}} P_{ij} f_j = (Pf)_i.$$

Therefore, in the discrete setting, the behaviour of our process as $n \rightarrow \infty$ can be reduced to understanding the behaviour of P^n as $n \rightarrow \infty$.

5.2 Stopping Times

Definition 5.2.1. Given a filtration $(\mathcal{F}_n)_{n=0}^\infty$, a $\mathbb{N} \cup \{\infty\}$ -valued random variable T is an $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time if for every $n \in \mathbb{N}$ we have $\{T \leq n\} \in \mathcal{F}_n$.

Recalling that a filtration tells us what information we have at time n , Definition 5.2.1 says that by time n there is enough information to determine whether T has occurred. One can think of T as a random alarm clock that cannot look into the future. By induction, the above is equivalent to $\{T = n\} \in \mathcal{F}_n$ or all $n \in \mathbb{N}$. One can interchange between using $\{T = n\}$ and $\{T \leq n\}$ when working with stopping times, with each being useful in different contexts. We include infinity to allow the possibility that the "alarm clock" T never rings.

Exercise 5.2.2. Let $(X_n)_{n=0}^\infty$ be a stochastic process, let $A \in \mathcal{B}(\mathcal{X})$. Show that $\tau_A = \inf\{n \in \mathbb{N} : X_n \in A\}$ is a $(\mathcal{F}_n^0)_{n=0}^\infty$ -stopping time. The random variable τ_A is the hitting time of A .

Example 5.2.3.

1. A deterministic time T is a stopping time, including $T = \infty$.
2. In general, for $A \in \mathcal{B}(\mathcal{X})$, the random variable $\ell_A = \sup\{n \geq 0 : X_n \in A\}$ is not a stopping time. Intuitively, the value of ℓ_A is dependent on the future.

Definition 5.2.4. Given a stochastic process $(X_n)_{n=0}^\infty$ and a $\mathbb{N} \cup \{\infty\}$ -valued random variable T , the stopped

process $(X_n^T)_{n=0}^\infty$ is given by

$$X_n^T(\omega) = X_{n \wedge T}(\omega) = \begin{cases} X_n(\omega) & n \leq T(\omega) \\ X_{T(\omega)}(\omega) & \text{otherwise.} \end{cases}$$

Exercise 5.2.5. Let $(\mathcal{F}_n)_{n=0}^\infty$ be a filtration, let T be a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time, and let $(X_n)_{n=0}^\infty$ be a $(\mathcal{F}_n)_{n=0}^\infty$ -adapted process. Show that $(X_n^T)_{n=0}^\infty$ is adapted to $(\mathcal{F}_n)_{n=0}^\infty$.

When filtrations are not mentioned in a statement, just assume there is some fixed filtration operating in the background.

Proposition 5.2.6. Let S and T be stopping times, and let $(T_n)_{n=0}^\infty$ be a sequence of stopping times. Then the following hold.

1. $S \vee T$ and $S \wedge T$ are stopping times.
2. $\sup_{n \in \mathbb{N}}(T_n)$, $\inf_{n \in \mathbb{N}}(T_n)$, $\liminf_{n \rightarrow \infty}(T_n)$ and $\limsup_{n \rightarrow \infty}(T_n)$ are stopping times.

Proof.

1. Note that

$$\{S \vee T \leq n\} = \{S \leq n\} \cap \{T \leq n\} \in \mathcal{F}_n$$

and

$$\{S \wedge T \leq n\} = \{S \leq n\} \cup \{T \leq n\} \in \mathcal{F}_n.$$

2. It suffices to show that $\sup_{n \in \mathbb{N}}(T_n)$ and $\inf_{n \in \mathbb{N}}(T_n)$ are stopping times as $\liminf_{n \rightarrow \infty}(T_n) = \sup_{n \in \mathbb{N}} \inf_{m \geq n}(T_m)$ and $\limsup_{n \rightarrow \infty}(T_n) = \inf_{n \in \mathbb{N}} \sup_{m \geq n}(T_m)$. Observe that

$$\left\{ \sup_{j \in \mathbb{N}}(T_j) \leq n \right\} = \bigcap_{j=0}^{\infty} \{T_j \leq n\} \in \mathcal{F}_n$$

and

$$\left\{ \inf_{j \in \mathbb{N}}(T_j) \leq n \right\} = \bigcup_{j=0}^{\infty} \{T_j \leq n\} \in \mathcal{F}_n.$$

□

Example 5.2.7. If S and T are stopping times, then it is not necessarily the case that $T - S$ is a stopping time. Consider the Markov chain of Figure 1 with the initial distribution δ_1 . Let $S = \tau_{\{X_n=3\}}$ and $T = \tau_{\{X_n=5\}}$. Then $\{T - S = 1\} \notin \mathcal{F}_1$, since $T \geq 2$.

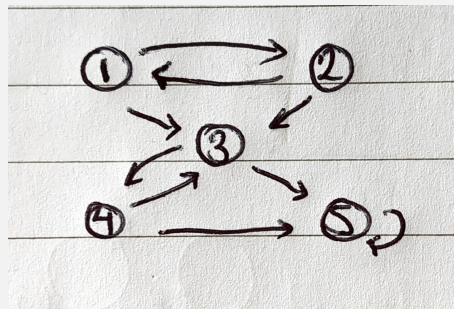


Figure 1

Given a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time T we can consider the σ -algebra that represents the information we have at time T . Let $\mathcal{F}_\infty = \bigvee_{n=0}^\infty \mathcal{F}_n$, which is the σ -algebra generated by $\bigcup_{n=0}^\infty \mathcal{F}_n$.

Exercise 5.2.8. Let T be a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time, show that $\{T = \infty\} \in \mathcal{F}_\infty$.

Definition 5.2.9. For a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time T , the stopped σ -algebra is

$$\mathcal{F}_T = \{A \in \mathcal{F}_\infty : \text{for all } n \in \mathbb{N}, A \cap \{T = n\} \in \mathcal{F}_n\}.$$

One can think of the stopped σ -algebra as containing events A , that when conditioned on the event $\{T = n\}$ can be determined by the information available up to time n . The event $\{T = n\}$ can be replaced with $\{T \leq n\}$, as we operate in the discrete setting.

Lemma 5.2.10. If T is a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time, then T is \mathcal{F}_T -measurable.

Proof. For $m, n \in \mathbb{N}$ we have

$$\{T = m\} \cap \{T = n\} = \begin{cases} \emptyset & m \neq n \\ \{T = n\} & m = n. \end{cases}$$

In either case $\{T = m\} \cap \{T = n\} \in \mathcal{F}_n$. Therefore, $\{T = m\} \in \mathcal{F}_T$ for every $m \in \mathbb{N}$, meaning that T is \mathcal{F}_T -measurable. \square

Exercise 5.2.11. Let S and T be $(\mathcal{F}_n)_{n=0}^\infty$ -stopping times.

1. Show that if $S \leq T$, then $\mathcal{F}_S \subseteq \mathcal{F}_T$.
2. Suppose $S \leq T$ and $A \in \mathcal{F}_S$. Show that $S\mathbf{1}_A + T\mathbf{1}_{A^c}$ is a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time.
3. Show that $\{A \cap \{S \leq T\} : A \in \mathcal{F}_S\} \subseteq \mathcal{F}_{S \wedge T}$.
4. For X a bounded random variable, show that

$$\mathbb{E}(\mathbb{E}(X|\mathcal{F}_T) | \mathcal{F}_S) = \mathbb{E}(\mathbb{E}(X|\mathcal{F}_S) | \mathcal{F}_T) = \mathbb{E}(X|\mathcal{F}_{S \wedge T}).$$

Definition 5.2.12. A stopping time T is finite if $\mathbb{P}(T < \infty) = 1$. When we have the stronger condition that $\{T = \infty\} = \emptyset$ we write this as $T < \infty$.

If $T < \infty$, then for a stochastic process $X = (X_n)_{n=0}^\infty$ the random variable X_T is well-defined.

Lemma 5.2.13. Let $(X_n)_{n=0}^\infty$ be adapted $(\mathcal{F}_n)_{n=0}^\infty$ and let T be a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time. Then for any $m \in \mathbb{N}$, the random variable $X_{T \wedge m}$ is \mathcal{F}_T -measurable. Moreover, suppose that $T < \infty$, then X_T is \mathcal{F}_T -measurable.

Proof. Let $A \in \mathcal{B}(\mathcal{X})$, then for $m, n \in \mathbb{N}$ we have

$$\{X_{T \wedge m} \in A\} \cap \{T = n\} = \{X_{n \wedge m} \in A\} \cap \{T = n\}.$$

Note $\{X_{n \wedge m} \in A\} \in \mathcal{F}_{n \wedge m} \subseteq \mathcal{F}_n$. Moreover, $\{T = n\} \in \mathcal{F}_n$. Therefore, the right-hand side is in \mathcal{F}_n which implies that $X_{T \wedge m}$ is \mathcal{F}_T -measurable. One shows in a similar way that X_T is \mathcal{F}_T -measurable. \square

Recall, that $(\mathcal{F}_n^0)_{n=0}^\infty$ denotes the natural filtration of a stochastic process $X = (X_n)_{n=0}^\infty$.

Lemma 5.2.14. Let $(X_n)_{n=0}^\infty$ be a stochastic process and let $T < \infty$ be a $(\mathcal{F}_n^0)_{n=0}^\infty$ -stopping time. Then for any $k \in \mathbb{N}$ we have

$$\{T = k\} \in \sigma(X_{T \wedge 0}, \dots, X_{T \wedge k}).$$

Proof. We proceed by induction on k .

- For $k = 0$ we have

$$\{T = 0\} \in \mathcal{F}_0^0 \in \sigma(X_0) = \sigma(X_{T \wedge 0}).$$

- Suppose the result holds for $n \leq k - 1$. Then

$$\begin{aligned} \mathbf{1}_{\{T=k\}} &= \mathbf{1}_{\{T=k\}} \mathbf{1}_{\{T>k-1\}} \\ &\stackrel{(1)}{=} \tilde{\varphi}(X_0, \dots, X_k) \mathbf{1}_{\{T>k-1\}} \\ &\stackrel{(2)}{=} \tilde{\varphi}(X_{0 \wedge T}, \dots, X_{k \wedge T}) \mathbf{1}_{\{T>k-1\}} \\ &= \tilde{\varphi}(X_{0 \wedge T}, \dots, X_{k \wedge T}) (1 - \mathbf{1}_{\{T \leq k-1\}}) \\ &\stackrel{\text{Ind Hyp.}}{=} \tilde{\varphi}(X_{0 \wedge T}, \dots, X_{k \wedge T}) \varphi(X_{0 \wedge T}, \dots, X_{(k-1) \wedge T}) \\ &\in \sigma(X_{0 \wedge T}, \dots, X_{k \wedge T}). \end{aligned}$$

In (1) we have used the fact that T is a $(\mathcal{F}_n^0)_{n=0}^\infty$ -stopping time, and so $\{T = k\} \in \mathcal{F}_k^0$. The equality of (2) follows from the fact that $i = i \wedge T$ for $i \in \{0, \dots, k\}$ on the domain of $\mathbf{1}_{\{T>k-1\}}$. □

Proposition 5.2.15. Let $(X_n)_{n=0}^\infty$ be a stochastic process and $T < \infty$ be a $(\mathcal{F}_n^0)_{n=0}^\infty$ -stopping time. Then

$$\mathcal{F}_T = \sigma(X_{T \wedge n} : n \in \mathbb{N}).$$

Proof. As $X_{T \wedge n}$ is \mathcal{F}_T -measurable for any $n \in \mathbb{N}$ by Lemma 5.2.13, it is clear that $\sigma(X_{T \wedge n} : n \in \mathbb{N}) \subseteq \mathcal{F}_T$. On the other hand, let $A \in \mathcal{F}_T$. Then for any $n \in \mathbb{N}$ we know that $A \cap \{T = n\} \in \mathcal{F}_n$ which implies that

$$\mathbf{1}_{A \cap \{T=n\}} = \varphi(X_0, \dots, X_n)$$

for some $\varphi \in \mathcal{B}_b(\mathcal{X}^{n+1})$. Note that

$$\varphi(X_0, \dots, X_n) = \varphi(X_0, \dots, X_n) \mathbf{1}_{\{T=n\}} = \varphi(X_{0 \wedge T}, \dots, X_{n \wedge T}) \mathbf{1}_{\{T=n\}}.$$

Due to Lemma 5.2.14 we know that $\mathbf{1}_{\{T=n\}} \in \sigma(X_{0 \wedge T}, \dots, X_{n \wedge T})$, therefore, $\mathbf{1}_{A \cap \{T=n\}}$ is measurable with respect to $\sigma(X_{T \wedge n} : n \in \mathbb{N})$. Which implies that $\mathcal{F}_T \subseteq \sigma(X_{T \wedge n} : n \in \mathbb{N})$ which completes the proof. □

5.3 The Strong Markov Property

Recall, that we can interpret a stochastic process $X = (X_n)_{n=0}^\infty$ as a random element of the canonical probability space $(\mathcal{X}^\mathbb{N}, \mathcal{B}(\mathbb{X}^\mathbb{N}), \text{Law}(X))$. Furthermore, for each $j \in \mathbb{N}$ we have a measurable map $\theta_j : \mathcal{X}^\mathbb{N} \rightarrow \mathcal{X}^\mathbb{N}$ given by

$$(a_0, \dots, a_j, a_{j+1}, \dots) \mapsto (a_j, a_{j+1}, \dots).$$

Suppose we want a function $F \in \mathcal{B}_b(\mathcal{X}^\mathbb{N})$ that only depends of times $n \geq j$. Then we can formulate it as a function $\Phi_F \in \mathcal{B}_b(\mathcal{X}^\mathbb{N})$ where

$$F(\cdot) = \Phi_F(\theta_j \cdot).$$

Definition 5.3.1. A process $X = (X_n)_{n=0}^\infty$ has the strong Markov property if for every finite stopping time T

and every bounded measurable function $\Phi \in \mathcal{B}_b(\mathcal{X}^{\mathbb{N}})$ we have

$$\mathbb{E}(\Phi(\theta_T X) | \mathcal{F}_T) = \mathbb{E}(\Phi(\theta_T X) | X_T).$$

Remark 5.3.2. In the setting of Definition 5.3.1, we can consider another process $Y = (Y_n)_{n=0}^{\infty}$ given by $Y_n = X_{T+n}$, that is, $Y = \theta_T X$ with the filtration $(\mathcal{G}_n)_{n=0}^{\infty}$ where $\mathcal{G}_n = \mathcal{F}_{T+n}$.

Exercise 5.3.3. For Y and $(\mathcal{G}_n)_{n=0}^{\infty}$ as in Remark 5.3.2, show that Y is a Markov process on $(\mathcal{G}_n)_{n=0}^{\infty}$.

Lemma 5.3.4. Let X be a time-homogeneous Markov process with transition probability P . Then, for any finite stopping time T , fixed $n \in \mathbb{N}$, and $A \in \mathcal{B}(\mathcal{X})$ we have

$$\mathbb{P}(X_{T+n} \in A | \mathcal{F}_T) = P^n(X_T, A).$$

Proof. It suffices to show the equivalent statement that for all $f \in \mathcal{B}_b(\mathcal{X})$ we have

$$\mathbb{E}(f(X_{T+n}) | \mathcal{F}_T) = \int_{\mathcal{X}} f(y) P^n(X_T, dy).$$

In one direction we set $f = \mathbf{1}_A$ and in the other we use an approximation argument to show the equivalence between the statements. Note that $\int_{\mathcal{X}} f(y) P^n(X_T, dy)$ is \mathcal{F}_T -measurable. Moreover, for $B \in \mathcal{F}_T$ we can write

$$B = \left(\bigcup_{m=0}^{\infty} B \cap \{T = m\} \right) \cup C$$

where $\mathbb{P}(C) = 0$ as T is a finite stopping time. Let $B_m = B \cap \{T = m\}$, then

$$\begin{aligned} \int_B f(X_{T+n}) d\mathbb{P} &= \sum_{m=0}^{\infty} \int_{B_m} f(X_{T+n}) d\mathbb{P} \\ &= \sum_{m=0}^{\infty} \int_{B_m} f(X_{m+n}) d\mathbb{P} \\ &= \sum_{m=0}^{\infty} \mathbb{E}(\mathbf{1}_{B_m} f(X_{m+n})) \\ &= \sum_{m=0}^{\infty} \mathbb{E}(\mathbb{E}(\mathbf{1}_{B_m} f(X_{m+n}) | \mathcal{F}_m)) \\ &= \sum_{m=0}^{\infty} \mathbb{E}(\mathbf{1}_{B_m} \mathbb{E}(f(X_{m+n}) | X_m)) \\ &= \sum_{m=0}^{\infty} \mathbb{E} \left(\mathbf{1}_{B_m} \int_{\mathcal{X}} f(y) P^n(X_m, dy) \right) \\ &= \mathbb{E} \left(\int_{\mathcal{X}} f(y) P^n(X_T, dy) \sum_{m=0}^{\infty} \mathbf{1}_{B_m} \right) \\ &= \mathbb{E} \left(\int_{\mathcal{X}} f(y) P^n(X_T, dy) (\mathbf{1}_B - \mathbf{1}_C) \right) \\ &= \int_B \int_{\mathcal{X}} f(y) P^n(X_T, dy) d\mathbb{P}. \end{aligned}$$

Therefore, $\int_{\mathcal{X}} f(y) P^n(X_T, dy)$ satisfies both conditions to be the conditional expectation of $f(X_{T+n})$ with respect to \mathcal{F}_T . \square

Theorem 5.3.5. Let $X = (X_n)_{n=0}^\infty$ be a time-homogeneous Markov process with transition probability P . If T is a finite stopping time, then the process $(\theta_T X_n)_{n \in \mathbb{N}}$ is also a time-homogeneous Markov process with transition probability P . In particular, for all $\Phi \in \mathcal{B}_b(\mathcal{X}^\mathbb{N})$ we have

$$\mathbb{E}(\Phi(\theta_T X) | \mathcal{F}_T) = \mathbb{E}(\Phi(\theta_T X) | X_T). \quad (5.3.1)$$

Moreover, for any $n > 0$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$\mathbb{P}(X_{n+T} \in A | \mathcal{F}_T) = P^n(X_T, A) \quad (5.3.2)$$

almost surely. It follows that $X = (X_n)_{n=0}^\infty$ has the strong Markov property.

Proof. For any $k \in \mathbb{N}$ and $A \in \mathcal{B}(\mathcal{X})$ one can consider the stopping time $\tilde{T} = T + k$ and apply Lemma 5.3.4 with $n = 1$ to deduce that

$$\mathbb{P}(X_{\tilde{T}+1} \in A | \mathcal{F}_{\tilde{T}}) = P(X_{\tilde{T}}, A)$$

which is exactly

$$\mathbb{P}((\theta_T X)_{k+1} \in A | \mathcal{F}_k) = P((\theta_T X)_k, A).$$

This is saying that $\theta_T X$ is a time homogeneous Markov process with transition probability P . Moreover, Lemma 5.3.4 shows that

$$\mathbb{P}(X_{n+T} \in A | \mathcal{F}_T) = P^n(X_T, A)$$

holds almost surely for any $n > 0$ and $A \in \mathcal{B}(\mathcal{X})$. To show (5.3.1) it suffices to show that the equation holds for all functions of the form $\Phi(a) = \prod_{i=0}^k f_i(a_i)$, where $a = (a_1, a_2, \dots)$, $f_i \in \mathcal{B}_b(\mathcal{X})$ and $k \in \mathbb{N}$. The reason why this is sufficient is because these functions approximate all functions in $\mathcal{B}_b(\mathcal{X})$. Let $\Phi(\theta_T X) = \prod_{i=0}^k f_i(X_{T+i})$. We proceed by induction on k .

- For $k = 0$ the result holds as $f_0(X_T)$ is measurable with respect to \mathcal{F}_T and X_T .
- Suppose the result holds true for all $n \leq k - 1$. Then

$$\begin{aligned} \mathbb{E}\left(\prod_{i=0}^k f_i(X_{T+i}) \middle| \mathcal{F}_T\right) &= \mathbb{E}\left(\mathbb{E}\left(\prod_{i=0}^k f_i(X_{T+i}) \middle| \mathcal{F}_{T+k-1}\right) \middle| \mathcal{F}_T\right) \\ &= \mathbb{E}\left(\prod_{i=0}^{k-1} f_i(X_{T+i}) \mathbb{E}(f_k(X_{T+k}) | \mathcal{F}_{T+k-1}) \middle| \mathcal{F}_T\right) \\ &= \mathbb{E}\left(\prod_{i=0}^{k-1} f_i(X_{T+i}) \int_{\mathcal{X}} f_k(y_k) P(X_{T+k-1}, dy_k) \middle| \mathcal{F}_T\right). \end{aligned}$$

Let

$$\tilde{f}_{k-1}(X_{T+k-1}) = f_{k-1}(X_{T+k-1}) \int_{\mathcal{X}} f_k(y_k) P(X_{T+k-1}, dy_k).$$

Then $\tilde{f}_{k-1} \in \mathcal{B}_b(\mathcal{X})$ and so we can apply our induction hypothesis to deduce that

$$\mathbb{E}\left(\tilde{f}_{k-1}(X_{T+k-1}) \prod_{i=0}^{k-2} f_i(X_{T+i}) \middle| \mathcal{F}_T\right) = \mathbb{E}\left(\tilde{f}_{k-1}(X_{T+k-1}) \prod_{i=0}^{k-2} f_i(X_{T+i}) \middle| X_T\right).$$

Therefore,

$$\begin{aligned}
\mathbb{E} \left(\prod_{i=0}^k f_i(X_{T+i}) \middle| \mathcal{F}_T \right) &= \mathbb{E} \left(\tilde{f}_{k-1}(X_{T+k-1}) \prod_{i=0}^{k-2} f_i(X_{T+i}) \middle| X_T \right) \\
&= \mathbb{E} \left(\prod_{i=0}^{k-1} f_i(X_{T+i}) \int_{\mathcal{X}} f_k(y_k) P(X_{T+k-1}, dy_k) \middle| X_T \right) \\
&= \mathbb{E} \left(\prod_{i=0}^{k-1} f_i(X_{T+i}) \mathbb{E}(f_k(X_{T+k}) | \mathcal{F}_{T+k-1}) \middle| X_T \right) \\
&= \mathbb{E} \left(\mathbb{E} \left(\prod_{i=0}^k f_i(X_{T+i}) \middle| \mathcal{F}_{T+k-1} \right) \middle| X_T \right) \\
&= \mathbb{E} \left(\prod_{i=0}^k f_i(X_{T+i}) \middle| X_T \right).
\end{aligned}$$

□

Remark 5.3.6.

- *There are continuous time Markov processes which are not strong Markov processes.*
- *The time-homogeneous condition for the discrete-time case is not necessary. We only state it here to simplify the proofs.*

If we do not have T finite, then we can condition on the event $\{T < \infty\}$.

Theorem 5.3.7. *Let $X = (X_n)_{n=0}^\infty$ be a time-homogeneous Markov process with transition probability P . Then for all $\Phi \in \mathcal{B}_b(\mathcal{X}^\mathbb{N})$ we have*

$$\mathbb{E}(\Phi(\theta_T X) \mathbf{1}_{\{T < \infty\}} | \mathcal{F}_T) = \mathbb{E}_{X_T}(\Phi(X)) \mathbf{1}_{\{T < \infty\}}.$$

Proof. As before it suffices to show that the equation holds for all functions of the form $\Phi(a) = \prod_{i=0}^k f_i(a_i)$, where $a = (a_1, a_2, \dots)$, $f_i \in \mathcal{B}_b(\mathcal{X})$ and $k \in \mathbb{N}$. We consider a single fixed coordinate. Let $f \in \mathcal{B}_b(\mathcal{X})$ and $B \in \mathcal{F}_T$, then

$$\begin{aligned}
\int_{B \cap \{T < \infty\}} f(X_{T+n}) d\mathbb{P} &= \sum_{m=0}^{\infty} \int_{B \cap \{T=m\}} f(X_{T+n}) d\mathbb{P} \\
&= \sum_{m=0}^{\infty} \int_{B \cap \{T=m\}} f(X_{m+n}) d\mathbb{P} \\
&= \sum_{m=0}^{\infty} \int_{B \cap \{T=m\}} \mathbb{E}(f(X_{m+n}) | \mathcal{F}_m) d\mathbb{P} \\
&= \sum_{m=0}^{\infty} \int_{B \cap \{T=m\}} \mathbb{E}(f(X_{m+n}) | X_m) d\mathbb{P} \\
&= \int_{B \cap \{T < \infty\}} \mathbb{E}(X_{T+n} \mathbf{1}_{\{T < \infty\}} | X_T) d\mathbb{P}.
\end{aligned}$$

Therefore,

$$\begin{aligned}
\mathbb{E}(f(X_{T+n}) \mathbf{1}_{\{T < \infty\}} | \mathcal{F}_T) &= \mathbb{E}(f(X_{T+n}) \mathbf{1}_{\{T < \infty\}} | X_T) \\
&= \mathbb{E}_{X_T}(f(X_{T+n})) \mathbf{1}_{\{T < \infty\}}.
\end{aligned}$$

One can extend this argument to a finite number of coordinates to complete the proof. □

Remark 5.3.8. Recall, that $\mathbb{E}_{X_T}(\cdot)$ denotes the expectation when X_T is used to start the Markov process.

5.4 Solution to Exercises

Exercise 5.1.2

Solution. Proceed by induction on n .

- By construction

$$P_{ij} = \mathbb{P}(X_{n+1} = j | X_n = i) \geq 0.$$

Moreover, by the law of total probability we have

$$\sum_{j \in \mathcal{X}} P_{ij} = \sum_{j \in \mathcal{X}} \mathbb{P}(X_{n+1} = j | X_n = i) = 1.$$

Therefore, P is a stochastic matrix.

- Suppose the result holds for $k \leq n - 1$. Then

$$\begin{aligned} P^n(x, \{j\}) &= \sum_{i \in \mathcal{X}} P(i, \{j\}) P^{n-1}(x, \{i\}) \\ &= \sum_{i \in \mathcal{X}} P_{ji} \underbrace{(P \times \cdots \times P)_{ix}}_{n-1} \\ &= \underbrace{(P \times \cdots \times P)_{jx}}_n. \end{aligned}$$

Consequently,

$$\begin{aligned} \sum_{k \in \mathcal{X}} P_{ik}^n &= \sum_{k \in \mathcal{X}} \sum_{j \in \mathcal{X}} P_{ij}^{n-1} P_{jk} \\ &= \sum_{j \in \mathcal{X}} \sum_{k \in \mathcal{X}} P_{ij}^{n-1} P_{jk} \\ &= \sum_{j \in \mathcal{X}} P_{ij}^{n-1} \sum_{k \in \mathcal{X}} P_{jk} \\ &= \sum_{j \in \mathcal{X}} P_{ij}^{n-1} (1) \\ &= 1. \end{aligned}$$

Therefore, by induction, we conclude that $P^n = \underbrace{P \times \cdots \times P}_n$ and P^n is a stochastic matrix. □

Exercise 5.2.2

Solution. This follows directly from observing that

$$\{\tau_A = n\} = \left(\bigcap_{k=0}^{n-1} \{X_k \notin A\} \right) \cap \{X_n \in A\} \in \mathcal{F}_n^0.$$

□

Exercise 5.2.5

Note that for any $n \in \mathbb{N}$ we can write

$$X_n^T(\omega) = X_n(\omega) \mathbf{1}_{\{T(\omega) > n\}} + X_T \mathbf{1}_{\{T(\omega) \leq n\}}$$

where

- $X_n(\omega)$ is \mathcal{F}_n -measurable,
- $\mathbf{1}_{\{T(\omega) > n\}} = 1 - \mathbf{1}_{\{T(\omega) \leq n\}}$ is \mathcal{F}_n -measurable as T is a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time,
- $X_{T(\omega)}$ is \mathcal{F}_n -measurable as it is $\mathcal{F}_{T(\omega)}$ -measurable, and
- $\mathbf{1}_{\{T(\omega) \leq n\}}$ is \mathcal{F}_n -measurable as T is a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time.

Therefore, $X_n^T(\omega) \in \mathcal{F}_n$ which implies the stopped process is adapted to $(\mathcal{F}_n)_{n=0}^\infty$.

Exercise 5.2.8

Solution. Note

$$\{T = \infty\} = \bigcap_{n=1}^{\infty} \{T > n\} = \bigcap_{n=1}^{\infty} \{T \leq n\}^c.$$

As $\{T \leq n\} \in \mathcal{F}_n$, it follows that $\{T = \infty\} \in \mathcal{F}_\infty$. □

Exercise 5.2.11

Solution.

1. Let $A \in \mathcal{F}_S$. Then

$$A \cap \{T \leq n\} = (A \cap \{S \leq n\}) \cap \{T \leq n\}.$$

In particular, $A \in \mathcal{F}_S$ implies that $A \cap \{S \leq n\} \in \mathcal{F}_n$ and T being a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time implies $\{T \leq n\} \in \mathcal{F}_n$. Therefore, $A \cap \{T \leq n\} \in \mathcal{F}_n$ which implies that $A \in \mathcal{F}_T$.

2. Note that

$$\{S\mathbf{1}_A + T\mathbf{1}_{A^c} \leq n\} = \underbrace{(\{S \leq n\} \cap A)}_{(1)} \cup \underbrace{(\{T \leq n\} \cap A^c)}_{(2)}.$$

As $A \in \mathcal{F}_S$, we have that (1) $\in \mathcal{F}_n$. By statement 1 we know $A \in \mathcal{F}_T$ and so (2) $\in \mathcal{F}_n$ as $A \in \mathcal{F}_S$. Hence,

$$\{S\mathbf{1}_A + T\mathbf{1}_{A^c} \leq n\} \in \mathcal{F}_n$$

which implies that $S\mathbf{1}_A + T\mathbf{1}_{A^c}$ is a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time.

3. For $A \in \mathcal{F}_S$ it follows that $A \cap \{S \leq n\} \in \mathcal{F}_n$. Recall, that $S \wedge T$ is itself a $(\mathcal{F}_n)_{n=0}^\infty$ -stopping time, so that $\{S \wedge T \leq n\} \in \mathcal{F}_n$. Moreover, $\{S \leq T\} \cap \{S \wedge T \leq n\} \subseteq \{S \leq n\}$. Therefore,

$$(A \cap \{S \leq T\}) \cap \{S \wedge T \leq n\} = A \cap \{S \leq n\} \cap \{S \wedge T \leq n\} \in \mathcal{F}_n.$$

4. Step 1: Show that $\mathcal{F}_{S \wedge T} = \mathcal{F}_S \cap \mathcal{F}_T$.

If $A \in \mathcal{F}_S \cap \mathcal{F}_T$, then

$$A \cap \{S \wedge T \leq n\} = (A \cap \{S \leq n\}) \cup (A \cap \{T \leq n\}) \in \mathcal{F}_n,$$

which implies that $A \in \mathcal{F}_{S \wedge T}$. Suppose instead that $A \in \mathcal{F}_{S \wedge T}$. Then

$$\begin{aligned} A \cap \{S = n\} &= A \cap \{S = n\} \cap (\{S \leq T\} \cup \{S > T\}) \\ &= A \cap \{S \wedge T = n\} \cup A \cap \{S \wedge T \leq n-1\} \in \mathcal{F}_n. \end{aligned}$$

Therefore, $\mathcal{F}_S \cap \mathcal{F}_T = \mathcal{F}_{S \wedge T}$.

Step 2: Show that $\mathbb{E}(\mathbb{E}(X|\mathcal{F}_S)|\mathcal{F}_T) = \mathbb{E}(X|\mathcal{F}_{S \wedge T})$.

Consider $\mathbb{E}(X|\mathcal{F}_{S \wedge T})$. It is $\mathcal{F}_{S \wedge T}$ -measurable and thus \mathcal{F}_T -measurable by step 1. Let $G \in \mathcal{F}_T$. Then $G \cap \{T \leq S\} \in \mathcal{F}_{S \wedge T} \subseteq \mathcal{F}_T$, again using step 1. Hence,

$$\begin{aligned} \int_G \mathbf{1}_{\{T \leq S\}} \mathbb{E}(X|\mathcal{F}_{S \wedge T}) \, d\mathbb{P} &= \int_{G \cap \{T \leq S\}} \mathbb{E}(X|\mathcal{F}_{S \wedge T}) \, d\mathbb{P} \\ &= \int_{G \cap \{T \leq S\}} X \, d\mathbb{P} \\ &= \int_G \mathbf{1}_{\{T \leq S\}} X \, d\mathbb{P} \end{aligned}$$

and similarly

$$\int_G \mathbf{1}_{\{T \leq S\}} \mathbb{E}(X | \mathcal{F}_S) d\mathbb{P} = \int_G \mathbf{1}_{\{T \leq S\}} X d\mathbb{P}.$$

Therefore,

$$\int_G \mathbf{1}_{\{T \leq S\}} \mathbb{E}(X | \mathcal{F}_{S \wedge T}) d\mathbb{P} = \int_G \mathbf{1}_{\{T \leq S\}} \mathbb{E}(X | \mathcal{F}_S) d\mathbb{P}. \quad (5.4.1)$$

As $G \cap \{S < T\} \in \mathcal{F}_S \cap \mathcal{F}_T$, by the same reasoning we deduce that

$$\int_G \mathbf{1}_{\{T < S\}} \mathbb{E}(X | \mathcal{F}_{S \wedge T}) d\mathbb{P} = \int_G \mathbf{1}_{\{S < T\}} \mathbb{E}(X | \mathcal{F}_S) d\mathbb{P}. \quad (5.4.2)$$

Adding (5.4.1) and (5.4.2) together we deduce that

$$\int_G \mathbb{E}(X | \mathcal{F}_{S \wedge T}) d\mathbb{P} = \int_G \mathbb{E}(X | \mathcal{F}_S) d\mathbb{P}$$

for all $G \in \mathcal{F}_T$ which implies that $\mathbb{E}(\mathbb{E}(X | \mathcal{F}_S) | \mathcal{F}_T) = \mathbb{E}(X | \mathcal{F}_{S \wedge T})$.

Step 3: Show that $\mathbb{E}(\mathbb{E}(X | \mathcal{F}_T) | \mathcal{F}_S) = \mathbb{E}(X | \mathcal{F}_{S \wedge T})$.

This follows by the similar arguments as those made in step 2.

□

Exercise 5.3.3

Solution. For every $f \in \mathcal{B}_b(\mathcal{X})$ and $n, m \in \mathbb{N}$ it follows that

$$\begin{aligned} \mathbb{E}(f(Y_{n+m}) | \mathcal{G}_m) &= \mathbb{E}(f(X_{n+m+T}) | \mathcal{F}_{T+m}) \\ &\stackrel{(1)}{=} \mathbb{E}(f(X_{n+m+T}) | X_{T+m}) \\ &= \mathbb{E}(f(Y_{n+m}) | Y_m), \end{aligned}$$

where (1) is an application of the strong Markov property of X .

□

6 Discrete State Space Markov Processes

We will identify the setting of discrete state spaces by saying that \mathcal{X} is countable. The notions developed here, do not necessarily have a direct analogue for continuous state space Markov processes. We will continue to consider discrete-time Markov processes.

6.1 Markov Chains as Graphs

Given a stochastic matrix P on \mathcal{X} , that is $P = (P_{ij} : i, j \in \mathcal{X})$, we can build an oriented graph on \mathcal{X} where an edge is drawn from i to j if and only if $P_{ij} > 0$. As $P_{ij} = \mathbb{P}(X_{n+1} = j | X_n = i)$, we can think of the graph where the nodes are states and the edges represent paths between the states that are admissible in the process defined by P .

Example 6.1.1. *The stochastic matrix*

$$P = \frac{1}{10} \begin{pmatrix} 0 & 5 & 5 & 0 \\ 3 & 7 & 0 & 0 \\ 0 & 10 & 0 & 0 \\ 2 & 8 & 0 & 0 \end{pmatrix}$$

has the corresponding oriented graph depicted in Figure 2. We can use the graph to help compute probabilities such as $\mathbb{P}(X_2 = 2 | X_0 = 1)$. We see that the only paths which contribute to this probability are $1 \rightarrow 3 \rightarrow 2$ and $1 \rightarrow 2 \rightarrow 2$. Therefore,

$$\mathbb{P}(X_2 = 2 | X_0 = 1) = \frac{1}{2}(1) + \frac{1}{2} \left(\frac{7}{10} \right).$$

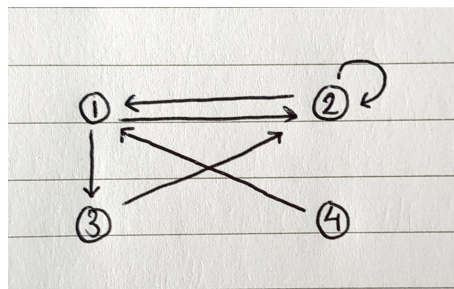


Figure 2: The oriented graph of Example 6.1.1.

Definition 6.1.2. *Let \mathcal{X} be countable and P be a stochastic matrix on \mathcal{X} .*

1. *The state $j \in \mathcal{X}$ is accessible from $i \in \mathcal{X}$ if $P_{ij}^n > 0$ for some $n \in \mathbb{N}$. We denote this $i \rightarrow j$.*
2. *States $i, j \in \mathcal{X}$ communicate if $i \rightarrow j$ and $j \rightarrow i$. We denote this $i \leftrightarrow j$.*
3. *Given a state $i \in \mathcal{X}$, we let $[i] = \{j \in \mathcal{X} : i \leftrightarrow j\}$ denote the communication class of i .*
4. *The stochastic matrix P is irreducible if $[i] = \mathcal{X}$ for some $i \in \mathcal{X}$. Otherwise, P is reducible.*

Exercise 6.1.3. *Show that \leftrightarrow is an equivalence relation on \mathcal{X} .*

Example 6.1.4. *Consider Example 6.1.1. The set $\{1, 2, 3\}$ is a communicating class, and $\{4\}$ is another. Therefore, P is reducible.*

When \mathcal{X} is infinite, only finite paths in the incidence graph guarantee accessibility, infinite paths may not.

Lemma 6.1.5. *If $i \rightarrow j$, then for any $i' \in [i]$ and $j' \in [j]$ we have $i' \rightarrow j'$.*

Proof. By assumption there exists $n_1, n_2, n_3 \in \mathbb{N}$ such that

- $P_{i'i}^{n_1} > 0$,
- $P_{ij}^{n_2} > 0$, and
- $P_{j'j'}^{n_3} > 0$.

Therefore,

$$P_{i'j'}^{n_1+n_2+n_3} \stackrel{(1)}{\geq} P_{i'i}^{n_1} P_{ij}^{n_2} P_{j'j'}^{n_3} > 0$$

where (1) follows from the fact that P^{n_1}, P^{n_2} and P^{n_3} have non-negative entries. Hence, $i' \rightarrow j'$. □

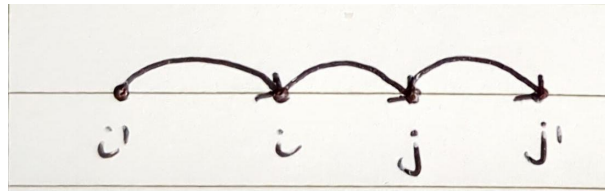


Figure 3: A graphical representation for the proof of Lemma 6.1.5.

Exercise 6.1.6. *Show that the relation $[i] \leq [j]$ if and only if $j \rightarrow i$ is well-defined and a partial order. That is, \leq is reflexive, transitive, and anti-symmetric.*

Definition 6.1.7. *An equivalence class $[i]$ is minimal, or closed, if there is no $j \in \mathcal{X}$ such that $[j] \leq [i]$ and $[j] \neq [i]$.*

Example 6.1.8. *Consider the stochastic matrix identified by the graph in Figure 4.*

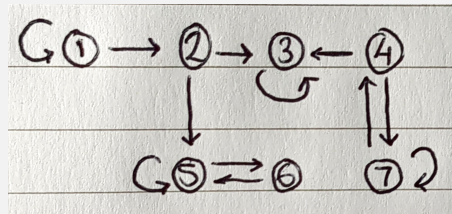


Figure 4: The oriented graph of the stochastic matrix referred to in Example 6.1.8.

In this case, the communication classes are the following.

- $[1] = \{1\}$.
- $[2] = \{2\}$.
- $[3] = \{3\}$.
- $[4] = \{4, 7\}$.

- $[5] = \{5, 6\}$.

Note that we have the relations

- $[5] \leq [2] \leq [1]$,
- $[3] \leq [4]$, and
- $[3] \leq [2] \leq [1]$.

However, there is no relation between $[4]$ and $[2]$. Therefore, \leq cannot be considered as a total order.

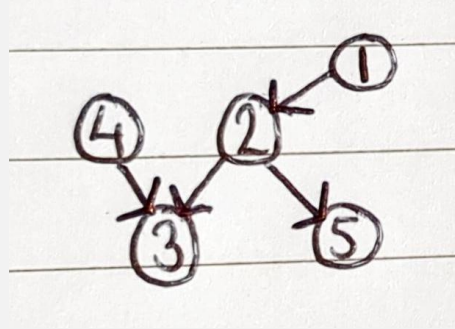


Figure 5: One can think of the equivalence classes as macro states, for which the process can only transition to a state that is less than or equal to it.

6.2 Recurrence and Transience

Definition 6.2.1. Given a state $i \in \mathcal{X}$, let

$$T_i = \inf\{n \geq 1 : X_n = i\}.$$

If $X_0 = i$ then T_i is called the first return time to state i .

For a state $i \in \mathcal{X}$ we adopt the notation

- $\mathbb{E}_i(\cdot) = \mathbb{E}(\cdot | X_0 = i)$, and
- $\mathbb{P}_i(\cdot) = \mathbb{P}(\cdot | X_0 = i)$.

Example 6.2.2. Suppose $\mathcal{X} = \{1, 2\}$, $P_{12} = \alpha \in [0, 1]$ and $P_{21} = \beta \in [0, 1]$ so that P is given by

$$P = \begin{pmatrix} 1 - \alpha & \alpha \\ \beta & 1 - \beta \end{pmatrix}.$$

Then,

$$\mathbb{P}_1(T_1 = n) = \begin{cases} 1 - \alpha & n = 1 \\ \alpha(1 - \beta)^{n-2}\beta & n \geq 2. \end{cases}$$

Hence,

$$\begin{aligned}\mathbb{P}_1(T_1 < \infty) &= \sum_{n=1}^{\infty} \mathbb{P}(T_1 = n) \\ &= (1 - \alpha) + \alpha\beta \sum_{n=0}^{\infty} (1 - \beta)^n \\ &= \begin{cases} 1 - \alpha + \alpha\beta \frac{1}{1-\beta} & \beta \neq 0 \\ 1 - \alpha & \beta = 0. \end{cases}\end{aligned}$$

Definition 6.2.3. A state $i \in \mathcal{X}$ is recurrent if $\mathbb{P}_i(T_i < \infty) = 1$. If a state $i \in \mathcal{X}$ is not recurrent it is transient.

Definition 6.2.4. A Markov chain, that is a stochastic matrix P , on \mathcal{X} is recurrent if every state $i \in \mathcal{X}$ is recurrent and it is transient if every state is transient.

Remark 6.2.5. We will see that if $i \in \mathcal{X}$ is recurrent, then every $i' \in [i]$ is recurrent. This means recurrence and transience are properties of communication classes.

Lemma 6.2.6. For states $i, j \in \mathcal{X}$ we have that $i \rightarrow j$ if and only if $\mathbb{P}_i(T_j < \infty) > 0$. Moreover,

$$\mathbb{P}_i(T_j < \infty) \leq \sum_{n=1}^{\infty} P_{ij}^n.$$

Proof. Observe that if $\mathbb{P}_i(T_j = n) > 0$ then

$$P_{ij}^n \stackrel{(\star)}{\geq} \mathbb{P}_i(T_j = n) > 0,$$

where (\star) is justified by the fact that P_{ij}^n is the probability of any path of length n between i and j , with $\mathbb{P}_i(T_j = n)$ only considering the subset of such paths that do not previously encounter j . Moreover, if $P_{ij}^n = 0$ then no path of length n exists from i to j with a non-zero probability and so $\mathbb{P}_i(T_j = n) = 0$. Hence, $\mathbb{P}_i(T_j = n) > 0$ if and only if $P_{ij}^n > 0$. Next, observe that

$$\mathbb{P}_i(T_j < \infty) = \mathbb{P}_i\left(\bigcup_{n=1}^{\infty} \{T_j = n\}\right). \quad (6.2.1)$$

So that $\mathbb{P}_i(T_j < \infty) > 0$ if and only if there exists an $n \geq 1$ such that $\mathbb{P}_i(T_j = n) > 0$ which happens if and only if $P_{ij}^n > 0$, that is $i \rightarrow j$. Applying a union bound to equation (6.2.1) we conclude that

$$\mathbb{P}_i(T_j < \infty) \leq \sum_{n=1}^{\infty} \mathbb{P}_i(T_j = n) \leq \sum_{n=1}^{\infty} P_{ij}^n.$$

□

Lemma 6.2.7. Let $j \in \mathcal{X}$ be recurrent. For $i \in \mathcal{X}$ if $\mathbb{P}_j(T_i < \infty) > 0$ then $\mathbb{P}_i(T_j < \infty) = 1$.

Proof. Assume that $\mathbb{P}_i(T_j = \infty) > 0$. This means that there is a set of infinite paths with non-zero probability that start at i and do not reach j . The condition that $\mathbb{P}_j(T_i < \infty) > 0$ says that i is accessible from j , Lemma 6.2.6. In particular, $m := \min(\{n : P_{ji}^n > 0\})$ is finite as the set is not empty. Note by construction that paths

of length m from j to i , reach i from j without returning to j beforehand. Otherwise, the Markov assumptions would imply that there exists a $k < m$ such that $P_{ji}^k > 0$. With this, it follows that

$$\begin{aligned}
\mathbb{P}_j(T_j = \infty) &\geq \mathbb{P}_j(T_j = \infty, T_i = m) \\
&= \mathbb{P}\left(\bigcap_{k=1}^{\infty} \{X_k \neq j\}, T_i = m \mid X_0 = j\right) \\
&= \mathbb{P}\left(\bigcap_{k=1}^{\infty} \{X_{m+k} \neq j\}, T_i = m \mid X_0 = j\right) \\
&= \mathbb{P}\left(\bigcap_{k=1}^{\infty} \{X_{m+k} \neq j\} \mid T_i = m, X_0 = j\right) \mathbb{P}(T_i = m \mid X_0 = j) \\
&= \mathbb{P}\left(\bigcap_{k=1}^{\infty} \{X_k \neq j\} \mid X_0 = i\right) \mathbb{P}_j(T_i = m) \\
&= \mathbb{P}_i(T_j = \infty) \mathbb{P}_j(T_i = m) \\
&> 0,
\end{aligned}$$

which contradicts j being a recurrent state. □

Exercise 6.2.8. Show in the context of Lemma 6.2.7 that $\mathbb{P}_\mu(T_j < \infty) = 1$ for any μ supported on $[j]$.

Definition 6.2.9. Passage times are constructed inductively from hitting times.

- $T_j^0 = 0$.
- $T_j^1 = T_j$.
- $T_j^{n+1} = \inf \{k > T_j^n : X_k = j\}$ for $n \geq 1$.

Exercise 6.2.10. Show that T_j^n is a $(\mathcal{F}_k)_{k=0}^\infty$ -stopping time.

Lemma 6.2.11. Let X have an initial distribution μ , and suppose $\mathbb{P}_\mu(T_j < \infty) = 1$ for a recurrent state $j \in \mathcal{X}$. Then the random variables $\{T_j^n - T_j^{n-1}\}_{n=1}^\infty$ are independent with

$$\mathbb{P}(T_j^k - T_j^{k-1} = m) = \mathbb{P}_j(T_j = m)$$

for any $m, k \in \mathbb{N}$.

Proof. By the strong Markov property for $k \geq 0$ we have that

$$\begin{aligned}
\mathbb{P}(T_j^{k+1} - T_j^k = m \mid \mathcal{F}_{T_j^k}) (\omega) &= \mathbb{P}_{T_j^k(\omega)}(T_j = m) \\
&= \mathbb{P}_j(T_j = m).
\end{aligned}$$

Taking the expectation we deduce that

$$\mathbb{P}_j(T_j = m) = \mathbb{P}(T_j^{k+1} - T_j^k = m)$$

by the tower property and the fact that the left-hand side is a constant. Consequently, for any $A \in \mathcal{F}_{T_j^k}$ we have

$$\begin{aligned} \mathbb{E} \left(\mathbf{1}_A \mathbf{1}_{\{T_j^{k+1} - T_j^k = m\}} \right) &= \mathbb{E} \left(\mathbb{E} \left(\mathbf{1}_A \mathbf{1}_{\{T_j^{k+1} - T_j^k = m\}} \middle| \mathcal{F}_{T_j^k} \right) \right) \\ &= \mathbb{E} \left(\mathbf{1}_A \mathbb{E} \left(\mathbf{1}_{\{T_j^{k+1} - T_j^k = m\}} \middle| \mathcal{F}_{T_j^k} \right) \right) \\ &= \mathbb{P}(A) \mathbb{P}_j(T_j = m) \\ &= \mathbb{P}(A) \mathbb{P}(T_j^{k+1} - T_j^k = m), \end{aligned}$$

which shows that the random variables $\{T_j^n - T_j^{n-1}\}_{n=1}^\infty$ are independent. \square

Remark 6.2.12. *The assumption on μ in Lemma 6.2.11 can be removed by conditioning. One can show that*

$$\mathbb{P} \left(T_j^k - T_j^{k-1} = m, T_j^k < \infty \middle| \mathcal{F}_{T_j^k} \right) = \mathbb{P}_j(T_j = m) \mathbf{1}_{T_j^k < \infty}$$

and an analogous statement of independence.

Note the following,

- $T_j^n = \sum_{k=1}^n T_j^k - T_j^{k-1}$, and
- $\{T_j^n < \infty\} = \bigcap_{k=1}^n \{T_j^k - T_j^{k-1} < \infty\}$.

Lemma 6.2.13. *For any $i, j \in \mathcal{X}$ and $k \in \mathbb{N}$ it follows that*

$$\mathbb{P}_i(T_j^{k+1} < \infty) = \mathbb{P}_i(T_j < \infty) \mathbb{P}_j(T_j^k < \infty).$$

Consequently,

$$\mathbb{P}_j(T_j^{k+1} < \infty) = \mathbb{P}_j(T_j < \infty)^{k+1}.$$

Proof. Let $\Phi \in \mathcal{B}_b(\mathcal{X}^\mathbb{N})$ be given by $\Phi(X) = \mathbf{1}_A(X)$, where

$$A := \{X \in \mathcal{X}^\mathbb{N} : |\{n \geq 1 : X_n = j\}| \geq k\}.$$

Note that

$$\mathbf{1}_{\{T_j^{k+1} < \infty\}}(\omega) = \Phi(\theta_{T_j(\omega)} X(\omega)) \mathbf{1}_{\{T_j < \infty\}}(\omega).$$

Taking the conditional expectation of both sides with respect to \mathcal{F}_{T_j} we deduce using the strong Markov property that

$$\begin{aligned} \mathbb{E} \left(\mathbf{1}_{\{T_j^{k+1} < \infty\}} \middle| \mathcal{F}_{T_j} \right) &= \mathbf{1}_{\{T_j < \infty\}} \mathbb{E}_{X_{T_j}}(\Phi(X)) \\ &= \mathbf{1}_{\{T_j < \infty\}} \mathbb{P}_j(T_j^k < \infty). \end{aligned}$$

Applying $\mathbb{E}_i(\cdot)$ completes the proof. \square

6.3 Recurrence Conditions

Definition 6.3.1. *The occupation time of a state $j \in \mathcal{X}$ is the random variable*

$$\eta_j = \sum_{n=1}^{\infty} \mathbf{1}_{\{X_n = j\}}.$$

Remark 6.3.2. *Occupation times are not stopping times.*

Theorem 6.3.3. *A state $j \in \mathcal{X}$ is transient if and only if $\sum_{n=1}^{\infty} P_{jj}^n < \infty$. Equivalently, $j \in \mathcal{X}$ is recurrent if and only if $\sum_{n=1}^{\infty} P_{jj}^n = \infty$.*

Proof. Note that $\mathbb{E}_j(\mathbf{1}_{\{X_n=j\}}) = P_{jj}^n$ so that

$$\begin{aligned} \sum_{n=1}^{\infty} P_{jj}^n &= \mathbb{E}_j(\eta_j) \\ &= \sum_{n=1}^{\infty} n \mathbb{P}_j(\eta_j = n) \\ &= \sum_{n=1}^{\infty} \mathbb{P}_j(\eta_j \geq n) \\ &= \sum_{n=1}^{\infty} \mathbb{P}_j(T_j^n < \infty) \\ &= \sum_{n=1}^{\infty} \mathbb{P}_j(T_j < \infty)^n. \end{aligned}$$

The geometric series is summable if and only if $\mathbb{P}_j(T_j < \infty) < 1$, which is to say that j is transient. \square

Corollary 6.3.4. *Suppose $j \in [i]$, then i is recurrent (transient) if and only if j is recurrent (transient).*

Proof. Since $j \leftrightarrow j$ there exists $m_1, m_2 \in \mathbb{N}$ such that $P_{ji}^{m_1}, P_{ij}^{m_2} > 0$. Therefore,

$$\begin{aligned} \sum_{k=m_1+m_2}^{\infty} P_{jj}^k &\geq \sum_{n=1}^{\infty} P_{ji}^{m_1} P_{ii}^n P_{ij}^{m_2} \\ &= P_{ji}^{m_1} P_{ij}^{m_2} \sum_{n=1}^{\infty} P_{ii}^n, \end{aligned}$$

so if i is recurrent then j is recurrent. By symmetry, the same holds for transience. \square

Example 6.3.5. *Consider the process $(X_n)_{n=0}^{\infty}$ where $X_0 = 0$ and*

$$X_n = X_{n-1} + \xi_n$$

for $n \geq 1$, where ξ_n is a $\{\pm 1\}$ -valued random variable with mean zero. For the chain to return to the state zero after n steps, the variables $(\xi_k)_{k=1, \dots, n}$ must take the values ± 1 equally often. Consequently, the chain can return to zero only after an even number of steps. More specifically,

$$\sum_{n=1}^{\infty} P_{00}^n = \sum_{n=1}^{\infty} P_{00}^{2n} = \sum_{n=1}^{\infty} \binom{2n}{n} \frac{1}{2^{2n}}.$$

As $\binom{2n}{n} \sim \frac{2^{2n}}{\sqrt{\pi n}}$, it follows that $\sum_{n=1}^{\infty} P_{00}^n = \infty$. Therefore, by Theorem 6.3.3, the state zero is recurrent. In particular, by Corollary 6.3.4 it follows that every state is recurrent as the chain is irreducible.

Lemma 6.3.6. Let $k \in \mathcal{X}$, then one of the following holds.

1. $\sum_{n=1}^{\infty} P_{ij}^n = \infty$, for every $i, j \in [k]$.
2. $\sum_{n=1}^{\infty} P_{ij}^n < \infty$ for every $i, j \in [k]$.

In particular, if $[k]$ has a finite number of elements and is a minimal class then statement 1 must hold so that every element of $[k]$ is recurrent.

Proof.

1. Suppose that $\sum_{n=1}^{\infty} P_{ij}^n = \infty$ for $i, j \in [k]$. Then for $i', j' \in [k]$ it follows that there exists $m_1, m_2 \in \mathbb{N}$ such that $P_{i'i}^{m_1}, P_{j'j}^{m_2} > 0$. Therefore,

$$\begin{aligned} \sum_{n=m_1+m_2}^{\infty} P_{i'j'}^n &\geq \sum_{n=1}^{\infty} P_{ij}^n P_{i'i}^{m_1} P_{j'j}^{m_2} \\ &= P_{i'i}^{m_1} P_{j'j}^{m_2} \sum_{n=1}^{\infty} P_{ij}^n \\ &= \infty. \end{aligned}$$

2. Suppose that $\sum_{n=1}^{\infty} P_{ij}^n < \infty$ for $i, j \in [k]$. Then for $i', j' \in [k]$ it follows that there exists $m_1, m_2 \in \mathbb{N}$ such that $P_{i'i}^{m_1}, P_{j'j}^{m_2} > 0$. Hence, we deduce that

$$\sum_{n=m_1+m_2}^{\infty} P_{ij}^n \geq P_{i'i}^{m_1} P_{j'j}^{m_2} \sum_{n=1}^{\infty} P_{i'j'}^n,$$

which implies that $\sum_{n=1}^{\infty} P_{i'j'}^n < \infty$.

If $[k]$ is a finite minimal class with $i \in [k]$ observe that

$$\begin{aligned} \sum_{j \in [k]} \sum_{n=1}^{\infty} P_{ij}^n &\stackrel{(1)}{=} \sum_{n=1}^{\infty} \sum_{j \in [k]} P_{ij}^n \\ &\stackrel{(2)}{=} \sum_{n=1}^{\infty} 1 \\ &= \infty, \end{aligned}$$

where (1) is just changing the ordering of a finite and infinite sum and (2) follows from the fact that $[k]$ is minimal and so $P_{ij}^n > 0$ if and only if $j \in [k]$. Therefore, we know that for at least one $j \in [k]$ the sum $\sum_{n=1}^{\infty} P_{ij}^n$ is infinite which implies that $\sum_{n=1}^{\infty} P_{ij}^n = \infty$ for all $j \in [k]$ by statement 1. Hence, every element of $[k]$ is recurrent. \square

Theorem 6.3.7.

- A state $j \in \mathcal{X}$ is recurrent if and only if $\mathbb{P}_j(X_n = j \text{ i.o.}) = 1$.
- A state $j \in \mathcal{X}$ is transient if and only if $\mathbb{P}_j(X_n = j \text{ i.o.}) = 0$.

Proof. Note that $\{X_n = j \text{ i.o.}\} = \{\eta_j = \infty\}$. Moreover,

$$\{\eta_j = \infty\} = \bigcap_{n=1}^{\infty} \{\eta_j \geq n\}.$$

The sets $\{\eta_j \geq n\}$ are decreasing in $n \in \mathbb{N}$, so

$$\begin{aligned} \mathbb{P}_j(\eta_j = \infty) &= \lim_{n \rightarrow \infty} \mathbb{P}_j(\eta_j \geq n) \\ &= \lim_{n \rightarrow \infty} \mathbb{P}_j(T_j^n < \infty) \\ &\stackrel{\text{Lem 6.2.13}}{=} \lim_{n \rightarrow \infty} \mathbb{P}_j(T_j < \infty)^n \\ &= \begin{cases} 1 & j \text{ recurrent} \\ 0 & j \text{ transient.} \end{cases} \end{aligned}$$

□

Lemma 6.3.8. *Suppose \mathcal{X} is finite, then a state is recurrent if and only if it is in a minimal class. In particular, there always exists a recurrent state.*

Proof. (\Rightarrow). As \mathcal{X} is finite, the partial order \leq is defined on a finite number of communication classes, and so there must exist a minimal class. In particular, this minimal class must be finite and so using Lemma 6.3.6 the class must be recurrent. Thus a recurrent state exists.

(\Leftarrow). Let $i \in \mathcal{X}$ be a recurrent state and suppose for contradiction that $i \in [k]$, where $[k]$ is not a minimal class. Then there exists a $j \in [k']$ such that $i \rightarrow j$ but $j \not\rightarrow i$. In particular, there exists an n_j such that $P_{ij}^{n_j} > 0$ and $P_{ji}^n = 0$ for all $n \in \mathbb{N}$. Therefore,

$$\mathbb{P}_i(T_i = \infty) \geq P_{ij}^{n_j} > 0.$$

This implies that $\mathbb{P}_i(T_i < \infty) < 1$ and so the state i is not recurrent, which is a contradiction. □

Proposition 6.3.9. *Suppose that $i, j \in \mathcal{X}$ are states such that $i \rightarrow j$ but $j \not\rightarrow i$, then i must be transient. In particular, if $[i]$ is not minimal, then it consists of transient states.*

Proof. Let $m = \min(\{n : P_{ij}^n > 0\})$. It follows that paths of length m from i to j never return to i before time m . Moreover, as $j \not\rightarrow i$ such a path never returns to i after time m either. Consequently,

$$\mathbb{P}_i(T_i = \infty) \geq P_{ij}^m > 0.$$

Therefore, i is transient. In particular, if $[i]$ is not minimal then a $j \in \mathcal{X}$ such that $i \rightarrow j$ but $j \not\rightarrow i$ exists and so $[i]$ contains transient states. □

Lemma 6.3.10. *For states $i, j \in \mathcal{X}$ we have that*

$$\sum_{n=1}^{\infty} P_{ij}^n = \frac{\mathbb{P}_i(T_j < \infty)}{1 - \mathbb{P}_j(T_j < \infty)}$$

with the understanding that the right-hand side is infinite if $\mathbb{P}_i(T_j < \infty) \neq 0$ and $\mathbb{P}_j(T_j < \infty) = 1$.

Proof. Recall that for any $j \in \mathcal{X}$, the occupation time is given by $\eta_j = \sum_{n=1}^{\infty} \mathbf{1}_{\{X_n=j\}}$. Hence, we can write

$$\begin{aligned}
\sum_{n=1}^{\infty} P_{ij}^n &= \mathbb{E}_i(\eta_j) \\
&= \sum_{k=1}^{\infty} \mathbb{P}_i(\eta_j \geq k) \\
&= \sum_{k=1}^{\infty} \mathbb{P}_i(T_j^k < \infty) \\
&\stackrel{\text{Lem 6.2.13}}{=} \sum_{k=1}^{\infty} \mathbb{P}_i(T_j < \infty) \mathbb{P}_j(T_j^{k-1} < \infty) \\
&\stackrel{\text{Lem 6.2.13}}{=} \sum_{k=1}^{\infty} \mathbb{P}_i(T_j < \infty) \mathbb{P}_j(T_j < \infty)^{k-1} \\
&= \frac{\mathbb{P}_i(T_j < \infty)}{1 - \mathbb{P}_j(T_j < \infty)}.
\end{aligned}$$

□

Theorem 6.3.11. *If a state $j \in \mathcal{X}$ is transient, then*

$$\sum_{n=1}^{\infty} P_{ij}^n < \infty$$

for all $i \in \mathcal{X}$. In particular, $\lim_{n \rightarrow \infty} P_{ij}^n = 0$.

Proof. If $j \in \mathcal{X}$ is transient then $\mathbb{P}_j(T_j < \infty) < 1$, so by Lemma 6.3.10 we have that

$$\sum_{n=1}^{\infty} P_{ij}^n = \frac{\mathbb{P}_i(T_j < \infty)}{1 - \mathbb{P}_j(T_j < \infty)} < \infty$$

which implies that $\lim_{n \rightarrow \infty} (P_{ij}^n) = 0$ for all $i \in \mathcal{X}$.

□

Remark 6.3.12. *Intuitively, Theorem 6.3.11 says that transient states are difficult to reach.*

Theorem 6.3.13. *If P has an invariant probability measure π , then for any transient state $j \in \mathcal{X}$ we must have $\pi(j) = 0$.*

Proof. Without loss of generality, we can assume that $\mathcal{X} = \mathbb{N}$. Suppose $\pi(j) > 0$ for a transient state $j \in \mathcal{X}$. Then as $\sum_{k \in \mathbb{N}} \pi(k) = 1 < \infty$, there exists an $N \in \mathbb{N}$ such that

$$\sum_{k=N+1}^{\infty} \pi(k) < \frac{\pi(j)}{2}.$$

Moreover, by Theorem 6.3.11, we can find an $n \in \mathbb{N}$ such that for all $0 \leq k \leq N$ we have $P_{kj}^n < \frac{\pi(j)}{2}$. As π is invariant we can write

$$\pi(j) = \sum_{k=0}^{\infty} \pi(k) P_{kj}^n.$$

Therefore,

$$\begin{aligned}
\pi(j) &= \sum_{k=0}^N \pi(k)P_{kj}^n + \sum_{k=N+1}^{\infty} \pi(k)P_{kj}^n \\
&\stackrel{(1)}{<} \frac{\pi(j)}{2} + \sum_{k=N+1}^{\infty} \pi(k)P_{kj}^n \\
&\stackrel{(2)}{<} \frac{\pi(j)}{2} + \sum_{k=N+1}^{\infty} \pi(k) \\
&< \frac{\pi(j)}{2} + \frac{\pi(j)}{2} \\
&= \pi(j),
\end{aligned}$$

where in (1) we have used the fact that π is a probability measure, and so the sum $\sum_{k=0}^N \pi(k)P_{kj}^n$ can be thought of averaging the P_{kj}^n , however, it is not a full average as $\sum_{k=0}^N \pi(k) \leq 1$. In (2) we have used the fact that P is a stochastic matrix and so $P_{kj} \leq 1$. Thus, we get a contradiction. \square

Corollary 6.3.14. *A transient Markov chain has no invariant probability measures.*

6.4 Constructing Invariant Probability Measures

Given any recurrent state $i \in \mathcal{X}$ we can consider the measure μ^i on \mathcal{X} , where for $j \in \mathcal{X}$ we let

$$\begin{aligned}
\mu^i(j) &= \mathbb{E}_i \left(\sum_{n=0}^{T_i-1} \mathbf{1}_{\{X_n=j\}} \right) \\
&= \mathbb{E}_i \left(\sum_{n=0}^{\infty} \mathbf{1}_{\{n < T_i\}} \mathbf{1}_{\{X_n=j\}} \right) \\
&= \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = j, T_i > n).
\end{aligned}$$

Remark 6.4.1. *Note that μ^i need not be a finite measure. Even if $\mathbb{P}_i(T_i < \infty) = 1$ we have*

$$\begin{aligned}
\sum_{j \in \mathcal{X}} \mu^i(j) &= \sum_{j \in \mathcal{X}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = j, T_i > n) \\
&= \sum_{n=0}^{\infty} \mathbb{P}_i(T_i > n) \\
&= \mathbb{E}_i(T_i).
\end{aligned}$$

Which is not necessarily finite as a random variable can be finite almost everywhere without having a finite expectation, take the Cauchy distribution.

Theorem 6.4.2. *If $i \in \mathcal{X}$ is recurrent for P , then μ^i is invariant for P .*

Proof. Fix a recurrent state $i \in \mathcal{X}$ and let $\mu = \mu^i$. We want to show that $(\mu P)(j) = \mu(j)$ for all $j \in \mathcal{X}$.

- For $j \neq i$ we have

$$\begin{aligned}
\mu(j) &= \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = j, T_i > n) \\
&= \sum_{k \in \mathcal{X}} \sum_{n=1}^{\infty} \mathbb{P}_i(X_n = j, T_i > n, X_{n-1} = k, T_i > n-1) \\
&= \sum_{k \in \mathcal{X}} \sum_{n=1}^{\infty} \mathbb{P}_i(X_n = j, T_i > n | X_{n-1} = k, T_i > n-1) \mathbb{P}_i(X_{n-1} = k, T_i > n-1) \\
&= \sum_{k \in \mathcal{X}} \sum_{n=1}^{\infty} \mathbb{P}_i(X_n = j | X_{n-1} = k, T_i > n-1) \mathbb{P}_i(X_{n-1} = k, T_i > n-1) \\
&\stackrel{(1)}{=} \sum_{k \in \mathcal{X}} P_{kj} \mu(k) \\
&= (\mu P)(j),
\end{aligned}$$

where (1) is an application of the Markov property.

- For $j = i$, on the one hand,

$$\begin{aligned}
\mu(i) &= \mathbb{E}_i \left(\sum_{n=0}^{T_i-1} \mathbf{1}_{\{X_n=i\}} \right) \\
&\stackrel{(1)}{=} \mathbb{E}_i (\mathbf{1}_{\{X_0=i\}}) \\
&= 1,
\end{aligned}$$

where (1) follows as by construction of T_i it must be that $\{X_k \neq i\}$ for $1 \leq k \leq T_i - 1$. On the other hand,

$$\begin{aligned}
(\mu P)(i) &= \sum_{k \in \mathcal{X}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = k, T_i > n) P_{ki} \\
&= \sum_{k \in \mathcal{X} \setminus \{i\}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = k, T_i > n) P_{ki} \\
&= \sum_{k \in \mathcal{X} \setminus \{i\}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = k, T_i > n) \mathbb{P}(X_{n+1} = i | X_n = k) \\
&= \sum_{k \in \mathcal{X} \setminus \{i\}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = k, T_i > n) \mathbb{P}(X_{n+1} = i | X_n = k, T_i > n) \\
&= \sum_{k \in \mathcal{X} \setminus \{i\}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_{n+1} = i, X_n = k, T_i > n) \\
&= \sum_{n=0}^{\infty} \mathbb{P}_i(X_{n+1} = i, T_i > n) \\
&= \sum_{n=0}^{\infty} \mathbb{P}_i(T_i = n+1) \\
&= \mathbb{P}_i(T_i < \infty) \\
&= 1,
\end{aligned}$$

where the last equality follows from the fact that i is recurrent. Therefore, $\mu(i) = (\mu P)(i)$.

□

Definition 6.4.3.

- A recurrent state $i \in \mathcal{X}$ is positive recurrent if $\mathbb{E}_i(T_i) < \infty$.
- A recurrent state $i \in \mathcal{X}$ is null recurrent if $\mathbb{E}_i(T_i) = \infty$.

Corollary 6.4.4. If $i \in \mathcal{X}$ is positive recurrent then μ^i is an invariant finite measure.

Proof. The invariance of μ^i follows directly from Theorem 6.4.2. The finiteness of μ^i follows from the positive recurrence on i . More specifically,

$$\begin{aligned} \sum_{j \in \mathcal{X}} \mu^i(j) &= \sum_{j \in \mathcal{X}} \sum_{n=0}^{\infty} \mathbb{P}_i(X_n = j, T_i > n) \\ &= \sum_{n=0}^{\infty} \mathbb{P}_i(T_i > n) \\ &= \mathbb{E}_i(T_i) \\ &< \infty. \end{aligned}$$

□

Lemma 6.4.5. Let $i \in \mathcal{X}$ be recurrent, then for any invariant measure ν and $k \in \mathcal{X}$ we have

$$\nu(k) \geq \nu(i)\mu^i(k).$$

Proof.

- If $k = i$, then $\mu^i(i) = 1$ and so the result holds clearly.
- When $k \neq i$, note that

$$\begin{aligned} P_{jk}^n &= \mathbb{P}_j(X_n = k) \\ &\stackrel{(1)}{\geq} \sum_{m=0}^{n-1} \mathbb{P}_j \left(X_n = k, \{X_m = i\} \cup \bigcap_{l=m+1}^{n-1} \{X_l \neq i\} \right) \\ &= \sum_{m=0}^{n-1} \mathbb{P} \left(X_n = k, \bigcap_{l=m+1}^{n-1} \{X_l \neq i\} \mid X_m = i \right) P_{ji}^m \\ &= \sum_{m=0}^{n-1} \mathbb{P}_i(X_{n-m} = k, T_i > n - m) P_{ji}^m, \end{aligned}$$

where in (1) we are using the fact that

$$\{X_m = i\} \cup \bigcap_{l=m+1}^{n-1} \{X_l \neq i\}$$

are disjoint events for $m = 0, \dots, n-1$, whose union is not necessarily the whole sample space. Intuitively,

these sets are the events that m is last visit to i before n . It follows using the invariance of ν that

$$\begin{aligned}
\nu(k) &= (\nu P^n)(k) \\
&= \sum_{j \in \mathcal{X}} \nu(j) P_{jk}^n \\
&\geq \sum_{j \in \mathcal{X}} \nu(j) \left(\sum_{m=0}^{n-1} \mathbb{P}_i(X_{n-m} = k, T_i > n-m) P_{ji}^m \right) \\
&\geq \sum_{m=0}^{n-1} \mathbb{P}_i(X_{n-m} = k, T_i > n-m) \sum_{j \in \mathcal{X}} \nu(j) P_{ji}^m \\
&\geq \nu(i) \sum_{m=0}^{n-1} \mathbb{P}_i(X_{n-m} = k, T_i > n-m).
\end{aligned}$$

By re-indexing the sum we get

$$\sum_{m=0}^{n-1} \mathbb{P}_i(X_{n-m} = k, T_i > n-m) = \sum_{l=1}^n \mathbb{P}_i(X_l = k, T_i > l).$$

Moreover, for $k \neq i$ by the construction of μ^i we know that

$$\begin{aligned}
\mu^i(k) &= \sum_{l=1}^{\infty} \mathbb{P}_i(X_l = k, T_i > l) \\
&= \lim_{n \rightarrow \infty} \sum_{l=1}^n \mathbb{P}_i(X_l = k, T_i > l) \\
&=: \lim_{n \rightarrow \infty} \mu_n^i(k).
\end{aligned}$$

Consequently, we have shown that

$$\nu(k) \geq \nu(i) \mu_n^i(k)$$

and so taking the limit gives

$$\nu(k) \geq \nu(i) \mu^i(k).$$

□

Theorem 6.4.6. *If a Markov chain is irreducible and recurrent, then its invariant measure is unique up to a multiplicative constant.*

Proof. Let ν be invariant and set $\mu = \mu^i$ for some $i \in \mathcal{X}$. By Theorem 6.4.2 the measure μ is invariant as $i \in \mathcal{X}$ is recurrent. Moreover, as $\mu(i) = 1$ we can write

$$\begin{aligned}
0 &= \nu(i) - \nu(i) \mu(i) \\
&= (\nu P^n)(i) - \nu(i) (\mu P^n)(i) \\
&= \sum_{k \in \mathcal{X}} (\nu(k) - \nu(i) \mu(k)) P_{ki}^n,
\end{aligned}$$

for any $n \in \mathbb{N}$. Note that $\nu(k) - \nu(i) \mu(k) \geq 0$ which implies that all the individual terms of the sum are zero. As the chain is irreducible, for any $k \in \mathcal{X}$ there is an $n \in \mathbb{N}$ such that $P_{ki}^n > 0$. Therefore, for each $k \in \mathcal{X}$ we can deduce that $\nu(k) = \nu(i) \mu(k)$. Thus, invariant measures are the same up to some multiplicative constant, $\nu(i)$. □

Theorem 6.4.7. *Suppose we have an irreducible Markov chain.*

1. *If the chain has an invariant probability measure π , then all the states are positive recurrent and*

$$\pi(i) = \frac{1}{\mathbb{E}_i(T_i)}$$

for $i \in \mathcal{X}$.

2. *If there exists a positive recurrent state, then the chain has a unique invariant probability measure and every state is positive recurrent.*

Proof.

1. As π is an invariant probability measure, there exists a state $i \in \mathcal{X}$ such that $\pi(i) > 0$. Therefore, the contrapositive of Theorem 6.3.13 tells us that $i \in \mathcal{X}$ is a recurrent state. Therefore, as recurrence is a communication class property, and the chain is irreducible, we deduce that every state is recurrent. For a fixed state $i \in \mathcal{X}$ we know by Theorem 6.4.2 that μ^i is an invariant measure. By Theorem 6.4.6 we know that μ^i is equal to π up to a multiplicative constant. That is, $\pi(j) = k\mu^i(j)$ for all $j \in \mathcal{X}$ and some $k \in \mathbb{R}$. Hence, μ^i is also a finite measure. Consequently,

$$\infty > \sum_{j \in \mathcal{X}} \mu^i(j) = \mathbb{E}_i(T_i),$$

and so i is a positive recurrent state. Moreover,

$$1 = \sum_{j \in \mathcal{X}} \pi(j) = \sum_{j \in \mathcal{X}} k\mu^i(j) = k\mathbb{E}_i(T_i)$$

implies that $k = \frac{1}{\mathbb{E}_i(T_i)}$. As $\mu^i(i) = 1$ we deduce that

$$\pi(i) = \frac{1}{\mathbb{E}_i(T_i)}.$$

Repeating this for each $i \in \mathcal{X}$ we arrive at the same conclusion for each state of the chain.

2. Suppose $i \in \mathcal{X}$ is a positive recurrent state of the chain, then we know μ^i is an invariant finite measure by Corollary 6.4.4. Therefore, we can normalise μ^i and apply the previous step to deduce that all states are positive recurrent and that the invariant probability measure is unique. □

6.5 Long Run Dynamics

In the setting of irreducible, positive recurrent chains we know there exists a unique invariant probability measure π . Now we want to answer the question as to whether measures will converge under the dynamics of the chain to this unique invariant probability measure. Formally, given a measure ν on \mathcal{X} we would like to understand under what constraints $(\nu P^n)_{n \in \mathbb{N}}$ converges to π . Of course, this requires an understanding of what we mean by convergence. For a positive result, we require added constraints on the structure of the chain. To see why we require additional constraints refer to Example 6.5.1.

Example 6.5.1. *Let $\mathcal{X} = \{1, 2\}$ and $P = \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix}$. Then P is irreducible and positive recurrent. Note that $P^{2n} = I$ and $P^{2n+1} = P$. Hence, $\lim_{n \rightarrow \infty} P_{ij}^n$ does not exist for any $i, j \in \mathcal{X}$. Therefore, we cannot make any conclusion about the convergence of $(\nu P^n)_{n \in \mathbb{N}}$. However,*

$$\frac{1}{n} \sum_{k=1}^n P_{ij}^k \xrightarrow{n \rightarrow \infty} \frac{1}{2}$$

for each $i, j \in \mathcal{X}$.

For a state $i \in \mathcal{X}$, the return times to i is the set

$$R(i) := \{n > 0 : P_{ii}^n > 0\}.$$

Definition 6.5.2. The period of a state $i \in \mathcal{X}$ denoted $d(i)$, is

$$d(i) := \begin{cases} \gcd(R(i)) & R(i) \neq \emptyset \\ \infty & R(i) = \emptyset. \end{cases}$$

Definition 6.5.3. For $i \in \mathcal{X}$, if $d(i) = 1$ then i is called aperiodic, while if $d(i) > 1$ then i is called periodic.

Remark 6.5.4.

- If $R(i) = \emptyset$, and so $d(i) = \infty$, it must be the case that i is in its own communication class.
- Note that if $d(i) < \infty$ then one can have $d(i) \notin R(i)$.

Definition 6.5.5. A chain is aperiodic if every state is aperiodic, and it is periodic with period d if every state has period d .

Example 6.5.6.

1. Consider the chain depicted in Figure 6. The chain is irreducible, and positive recurrent. However, it is not aperiodic as the state 1, for example, has a period of 4.

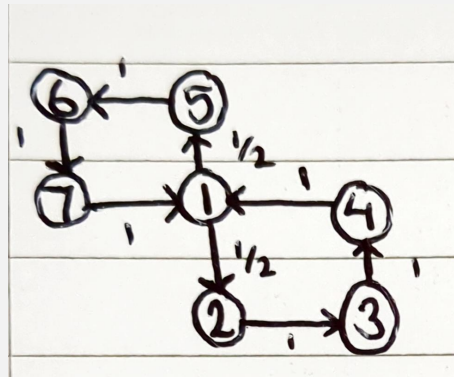


Figure 6: A chain that is irreducible, positive recurrent but not aperiodic.

2. Consider the chain depicted in Figure 7. Note that

$$R(1) = \{3n + 4m : n, m \in \mathbb{N}\}.$$

In particular, $3, 4 \in R(1)$ which implies that $d(1) = 1$. In this case, $d(i) \notin R(i)$.

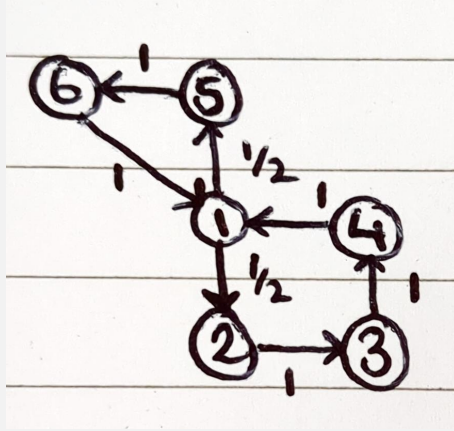


Figure 7: A chain that is irreducible, positive recurrent and aperiodic.

Proposition 6.5.7. For $i, j \in \mathcal{X}$ in the same communication class, if $i \neq j$ then $d(i) = d(j) < \infty$.

Proof. The condition that i and j are distinct elements in the same communication class implies that $d(i), d(j) < \infty$. As i and j are in the same communication class, we know there exists an $n \in \mathbb{N}$ such that $P_{ij}^n > 0$ and an $m \in \mathbb{N}$ such that $P_{ji}^m > 0$. Consequently, by the Chapman-Kolmogorov equation we know that $P_{ii}^{m+n}, P_{jj}^{m+n} > 0$ which implies that $m+n \in R(i) \cap R(j)$. For $k \in R(i)$, using the Chapman-Kolmogorov equation we note that $k+m+n \in R(j)$. Therefore, as $d(j)|n+m$ and $d(j)|k+n+m$ it follows that $d(j)|k$. As $k \in R(i)$ was arbitrary we conclude that $d(j)$ is a common divisor of $R(i)$. Hence, $d(j) \leq d(i)$. By symmetry, we also deduce that $d(i) \leq d(j)$ and so $d(i) = d(j)$. \square

Corollary 6.5.8. An irreducible chain is either periodic or aperiodic.

Theorem 6.5.9. Suppose P is irreducible, aperiodic, and positive recurrent. Let π denote its unique invariant probability measure. Then

$$\lim_{n \rightarrow \infty} \left(\sum_{j \in \mathcal{X}} |P_{ij}^n - \pi(j)| \right) = 0$$

for all $i \in \mathcal{X}$.

Proof. Let $(X_n)_{n \in \mathbb{N}}$ and $(X'_n)_{n \in \mathbb{N}}$ be independent Markov process with transition probabilities P and initial distributions μ and ν respectively. Then by Lemma 9.2.8 we know that $Z_n = (X_n, X'_n)$ is a time homogeneous Markov process on $\mathcal{X} \times \mathcal{X}$ with initial distribution $\mu \otimes \nu$ and transition probabilities

$$Q_{(i,i'),(j,j')} = P_{ij}P_{i'j'}$$

for all $i, j, i', j' \in \mathcal{X}$. Let $T = \inf \{n \geq 1 : X_n = X'_n\}$. Using Lemma 9.2.6 we have

$$\sum_{j \in \mathcal{X}} |\mathbb{P}(X_n = j) - \mathbb{P}(X'_n = j)| \leq 2\mathbb{P}(T > n). \quad (6.5.1)$$

By Lemma 9.2.12 we know that $\mathbb{P}(T < \infty) = 1$, which implies that $\mathbb{P}(T > n) \rightarrow 0$ as $n \rightarrow \infty$. Hence, from (6.5.1) we deduce that

$$\sum_{j \in \mathcal{X}} |\mathbb{P}(X_n = i) - \mathbb{P}(X'_n = i)| \xrightarrow{n \rightarrow \infty} 0.$$

In particular, we can take any $i \in \mathcal{X}$ and let $\mu = \delta_i$ and $\nu = \pi$ to see that

$$\lim_{n \rightarrow \infty} \left(\sum_{j \in \mathcal{X}} |P_{ij}^n - \pi(j)| \right) = 0.$$

□

Remark 6.5.10. With the notation used in the introduction to this section, Theorem 6.5.9 establishes the convergence of the measures $\nu = \delta_i$ for $i \in \mathcal{X}$ to π , in the sense outlined in the theorem, under the dynamics of the chain. This will be useful to use when we try to generalise to arbitrary probability measures.

Corollary 6.5.11. Suppose that X is irreducible, positive recurrent and aperiodic on a discrete state space \mathcal{X} . Let π denote its unique invariant probability measure. Then for any $f : \mathcal{X} \rightarrow \mathbb{R}$ which is π -integrable, and probability measure μ on \mathcal{X} we have that

$$\mathbb{E}_\mu(f(X_k)) \rightarrow \int_{\mathcal{X}} f \, d\pi$$

as $k \rightarrow \infty$. Moreover,

$$\frac{1}{n} \sum_{k=1}^n \mathbb{E}_\mu(f(X_k)) \rightarrow \int_{\mathcal{X}} f \, d\pi \quad (6.5.2)$$

as $n \rightarrow \infty$.

Proof. Suppose $f = \mathbf{1}_j$ for some $j \in \mathcal{X}$, then

$$\mathbb{E}_\mu(f(X_k)) = \sum_{i \in \mathcal{X}} \mu(i) P_{ij}^k.$$

We know by Theorem 6.5.9 that

$$\sum_{i \in \mathcal{X}} |\mu(i) P_{ij}^k - \pi(i)| \rightarrow 0$$

as $k \rightarrow \infty$. Therefore,

$$\begin{aligned} \left| \mathbb{E}_\mu(f(X_k)) - \int_{\mathcal{X}} f \, d\pi \right| &= \left| \sum_{i \in \mathcal{X}} \mu(i) P_{ij}^k - \pi(j) \right| \\ &= \left| \sum_{i \in \mathcal{X}} \mu(i) P_{ij}^k - \sum_{i \in \mathcal{X}} P_{ij} \pi(i) \right| \\ &\leq \sum_{i \in \mathcal{X}} |\mu(i) P_{ij}^k - P_{ij} \pi(i)| \\ &\leq \sum_{i \in \mathcal{X}} |\mu(i) P_{ij}^{k-1} - \pi(i)| \\ &\xrightarrow{k \rightarrow \infty} 0. \end{aligned}$$

Now for $f : \mathcal{X} \rightarrow \mathbb{R}$ non-negative and π -integrable, we can write $f = \sum_{i \in \mathcal{X}} f(i) \mathbf{1}_i$. Let $f_n = \sum_{i=1}^n f(i) \mathbf{1}_i$. Then by the algebra of limits we know that

$$\mathbb{E}_\mu(f_n(X_k)) \rightarrow \int_{\mathcal{X}} f_n \, d\pi$$

as $k \rightarrow \infty$. As f is non-negative we note that

$$\mathbb{E}_\mu(f_n(X_k)) \leq \mathbb{E}_\mu(f_{n+1}(X_k))$$

and so by monotone convergence we have that $\mathbb{E}_\mu(f_n(X_k)) \rightarrow \mathbb{E}_\mu(f(X_k))$ as $n \rightarrow \infty$. Similarly, we have that $\int_{\mathcal{X}} f_n d\pi \rightarrow \int_{\mathcal{X}} f d\pi$. Therefore,

$$\mathbb{E}_\mu(f(X_k)) \rightarrow \int_{\mathcal{X}} f d\pi$$

as $k \rightarrow \infty$. We can then extend this to general π -integrable functions $f : \mathcal{X} \rightarrow \mathbb{R}$ by considering $f = f^+ - f^-$. Note that (6.5.2) follows as for a converging sequence, the limit of the partial averages converges to the limit of the sequence. \square

Remark 6.5.12. For $i, j \in \mathcal{X}$, by letting $\mu = \delta_i$ and $f = \mathbf{1}_j$ in Corollary 6.5.11 we deduce that

$$\frac{1}{n} \sum_{k=1}^n P_{ij}^k \rightarrow \pi(j).$$

6.6 Total Variation

Definition 6.6.1. The total variation distance between probability measures μ and ν , on some measurable space \mathcal{X} , is

$$\|\mu - \nu\|_{\text{TV}} = 2 \sup_{A \subseteq \mathcal{X}} |\mu(A) - \nu(A)|,$$

where the supremum is over measurable subsets $A \subseteq \mathcal{X}$.

Remark 6.6.2.

1. $\|\cdot\|_{\text{TV}}$ has the dual formula,

$$\|\mu - \nu\|_{\text{TV}} = \sup_{f \in \mathcal{B}_b(\mathcal{X}), \|f\|_\infty \leq 1} \left| \int_{\mathcal{X}} f d\mu - \int_{\mathcal{X}} f d\nu \right|.$$

2. Note that $\|\mu - \nu\|_{\text{TV}} \in [0, 2]$ with $\|\mu - \nu\|_{\text{TV}} = 0$ if and only if $\mu = \nu$ and $\|\mu - \nu\|_{\text{TV}} = 2$ if and only if μ and ν are mutually singular. Where by mutually singular we mean that there exists a $A \subseteq \mathcal{X}$ such that $\mu(A) = 1$ and $\nu(A) = 0$.

Lemma 6.6.3. For \mathcal{X} discrete

$$\|\mu - \nu\|_{\text{TV}} = \sum_{i \in \mathcal{X}} |\mu(i) - \nu(i)| = \|\mu - \nu\|_1.$$

Proof. Let $B = \{i \in \mathcal{X} : \mu(i) \geq \nu(i)\}$. Then since μ and ν are probability measures we note that

$$\begin{aligned} 0 &= 1 - 1 \\ &= \left(\sum_{i \in B} + \sum_{i \in B^c} \right) (\mu(i) - \nu(i)) \\ &= \sum_{i \in B} \mu(i) - \nu(i) + \sum_{i \in B^c} \mu(i) - \nu(i) \\ &= \sum_{i \in B} (\mu(i) - \nu(i)) - \sum_{i \in B^c} (\nu(i) - \mu(i)), \end{aligned}$$

which implies that

$$\sum_{i \in B} \mu(i) - \nu(i) = \sum_{i \in B^c} \nu(i) - \mu(i).$$

Note that the terms of these sums are non-negative by the construction of B so that

$$\|\mu - \nu\|_1 = \sum_{i \in B} \mu(i) - \nu(i) + \sum_{i \in B^c} \nu(i) - \mu(i).$$

In particular,

$$\begin{aligned} \|\mu - \nu\|_1 &= \sum_{i \in B} \mu(i) - \nu(i) + \sum_{i \in B^c} \nu(i) - \mu(i) \\ &= \mu(B) - \nu(B) + (\mu(B^c) - \nu(B^c)) \\ &= 2(\mu(B) - \nu(B)) \\ &\leq \|\mu - \nu\|_{\text{TV}}. \end{aligned} \tag{6.6.1}$$

For any $A \subseteq \mathcal{X}$ observe that

$$\begin{aligned} |\mu(A) - \nu(A)| &= |\mu(A \cap B) - \nu(A \cap B) - (\mu(A \cap B^c) - \nu(A \cap B^c))| \\ &\leq 2 \max(|\mu(A \cap B) - \nu(A \cap B)|, |\mu(A \cap B^c) - \nu(A \cap B^c)|) \\ &\leq 2 \max(|\mu(B) - \nu(B)|, |\mu(B^c) - \nu(B^c)|) \\ &= 2|\mu(B) - \nu(B)| \\ &= \sum_{i \in \mathcal{X}} |\mu(i) - \nu(i)|. \end{aligned}$$

Hence, taking the supremum of both sides we deduce that

$$\|\mu - \nu\|_{\text{TV}} \leq \|\mu - \nu\|_1.$$

Combined with (6.6.1) we conclude. \square

Definition 6.6.4. Let $(\nu_n)_{n \in \mathbb{N}}$ be a sequence of measures and let ν be a measure.

1. The sequence $(\nu_n)_{n \in \mathbb{N}}$ converges in total variation to ν if $\|\nu_n - \nu\|_{\text{TV}} \rightarrow 0$.
2. The sequence $(\nu_n)_{n \in \mathbb{N}}$ converges strongly to ν if $\nu_n(A) \rightarrow \nu(A)$ for every measurable A .
3. The sequence $(\nu_n)_{n \in \mathbb{N}}$ converges weakly if

$$\int_{\mathcal{X}} f \, d\nu_n \rightarrow \int_{\mathcal{X}} f \, d\nu$$

for every $f \in \mathcal{C}_b(\mathcal{X})$.

Remark 6.6.5. Note that convergence in total variation implies strong convergence which in turn implies weak convergence. However, the reverse implication in each of these cases does not hold.

Example 6.6.6.

1. For $A \in \mathcal{B}([-1, 1])$, let

$$\nu_n(A) = \int_A 1 + \sin(nx) \, dx$$

and ν be the Lebesgue measure on $[-1, 1]$. Then by the Riemann-Lebesgue measure it follows that

$\nu_n(A) \rightarrow \nu(A)$ for every $A \in \mathcal{B}(\mathcal{X})$. That is, the sequence $(\nu_n)_{n \in \mathbb{N}}$ converges to ν strongly. However,

$$\begin{aligned} \|\nu_n - \nu\|_{\text{TV}} &= \sup_{f \in \mathcal{B}_b(\mathcal{X}), \|f\|_\infty \leq 1} \left| \int_{[-1,1]} f(x)(1 + \sin(nx)) \, dx - \int_{[-1,1]} f(x) \, dx \right| \\ &= \sup_{f \in \mathcal{B}_b(\mathcal{X}), \|f\|_\infty \leq 1} \left| \int_{[-1,1]} f(x) \sin(nx) \, dx \right| \\ &\stackrel{n \rightarrow \infty}{\not\rightarrow} 0, \end{aligned}$$

meaning the sequence $(\nu_n)_{n \in \mathbb{N}}$ does not converge in total variation to ν .

2. Consider the sequence of measures $(\nu_n)_{n \in \mathbb{N}}$ on \mathbb{R} where $\nu_n = \delta_{\frac{1}{n}}$. Then for any $f \in \mathcal{C}_b(\mathcal{X})$ we have that

$$\int_{\mathcal{X}} f \delta_{\frac{1}{n}} = f\left(\frac{1}{n}\right) \xrightarrow{n \rightarrow \infty} f(0) = \int_{\mathcal{X}} f \delta_0.$$

Therefore, the sequence $(\nu_n)_{n \in \mathbb{N}}$ converges weakly to δ_0 . However, $\nu_n(\{0\}) = 0$ for every $n \in \mathbb{N}$ and $\delta_0(\{0\}) = 1$, meaning the sequence $(\nu_n)_{n \in \mathbb{N}}$ does not converge strongly to δ_0 . Moreover, $\|\nu_n - \delta_0\|_{\text{TV}} = 2$ for all $n \in \mathbb{N}$, and so does not converge in total variation.

Theorem 6.6.7. Suppose P is irreducible, aperiodic, and positive recurrent. Let π denote the unique invariant probability measure. Then for any probability measure ν on \mathcal{X} we have that $\nu P^n \rightarrow \pi$ in total variation.

Proof. Note that

$$\begin{aligned} \|\mu P^n - \pi\|_{\text{TV}} &= \sum_{j=1}^{\infty} \left| \sum_{i=1}^{\infty} \mu(i) P_{ij}^n - \pi(j) \right| \\ &= \sum_{j=1}^{\infty} \left| \sum_{i=1}^{\infty} \mu(i) P_{ij}^n - \sum_{i=1}^{\infty} \mu(i) \pi(j) \right| \\ &\leq \sum_{j=1}^{\infty} \sum_{i=1}^{\infty} \mu(i) |P_{ij}^n - \pi(j)|. \end{aligned}$$

Given an $\epsilon > 0$, we can choose $N \in \mathbb{N}$ such that

$$\sum_{i=N+1}^{\infty} \mu(i) < \frac{\epsilon}{4}$$

as μ is a probability measure meaning $\sum_{i \in \mathbb{N}} \mu(i) < \infty$. Consequently,

$$\sum_{i=N+1}^{\infty} \mu(i) \sum_{j=1}^{\infty} |P_{ij}^n - \pi(j)| < \frac{\epsilon}{2},$$

where we have just applied the triangle inequality to the inner sum, and the fact that P_{ij}^n and $\pi(j)$ are bounded by one. On the other hand, by Theorem 6.5.9 we can choose $M \in \mathbb{N}$ such that for all $i \leq N$ we have

$$\sum_{j=1}^{\infty} |P_{ij}^n - \pi(j)| \leq \frac{\epsilon}{2}$$

for $n \geq M$. Therefore,

$$\sum_{j=1}^{\infty} \sum_{i=1}^N \mu(i) |P_{ij}^n - \pi(j)| \leq \sum_{i=1}^N \mu(i) \frac{\epsilon}{2} \leq \frac{\epsilon}{2}.$$

Hence,

$$\begin{aligned}
\|\mu P^n - \pi\|_{\text{TV}} &\leq \sum_{j=1}^{\infty} \sum_{i=1}^{\infty} \mu(i) |P_{ij}^n - \pi(j)| \\
&= \sum_{j=1}^{\infty} \sum_{i=1}^N \mu(i) |P_{ij}^n - \pi(j)| + \sum_{j=1}^{\infty} \sum_{i=N+1}^{\infty} \mu(i) |P_{ij}^n - \pi(j)| \\
&\leq \frac{\epsilon}{2} + \frac{\epsilon}{2} \\
&= \epsilon.
\end{aligned}$$

□

Remark 6.6.8. The proof of Theorem 6.6.7 utilises the result of Theorem 6.5.9, however, we could instead capitalise on the proof of Theorem 6.5.9. Note that Theorem 6.6.7 is just a stronger version of Theorem 6.5.9, which we have already proved. In fact, at the last step of the proof of Theorem 6.5.9 we can choose the initial distributions of our chains to be ν and π to arrive at Theorem 6.6.7.

6.7 Periodic Chains

We now provide an alternative, but equivalent, definition of the period of an irreducible Markov chain.

Lemma 6.7.1. The period of an irreducible stochastic matrix P is equal to the largest $d \in \mathbb{N}_{>0}$ such that one can partition the state space as

$$\mathcal{X} = A_0 \sqcup \cdots \sqcup A_{d-1}$$

where if $i \in A_n$ then for $j \in \mathcal{X}$ such that $P_{ij} > 0$ we have $j \in A_{n+1 \bmod d}$.

Proof. Suppose P has period d . Fix a state $i \in \mathcal{X}$ and let

$$A_n = \{j \in \mathcal{X} : P_{ij}^{kd+n} > 0 \text{ for some } k \in \mathbb{N}\}$$

for $n = 0, \dots, d-1$. By irreducibility we know that $(A_n)_{n=0}^{d-1}$ forms a cover of \mathcal{X} . Suppose $j \in A_{n_1} \cap A_{n_2}$. This implies that there exists $k_1, k_2 \in \mathbb{N}$ such that $P_{ij}^{k_1 d + n_1}, P_{ij}^{k_2 d + n_2} > 0$. However, by irreducibility, there exists a $q \in \mathbb{N}$ such that $P_{ji}^q > 0$. So $k_1 d + n_1 + q$ and $k_2 d + n_2 + q$ are in $R(i)$ and so d divides $n_1 - n_2$ which implies $n_1 = n_2$. Therefore, $A_{n_1} \cap A_{n_2} = \emptyset$ for $n_1 \neq n_2$. Now assume that p is the largest number for which such a decomposition of p disjoint sets exists. For $j \in \mathcal{X}$, if j returns to itself in q steps then we must have that $p|q$. Therefore, p is a divisor of the set of return times for j . Hence, by the definition of d we know that $p \leq d$ and thus $p = d$ if it is the largest number where such a decomposition exists. □

Example 6.7.2. For a d -periodic, irreducible stochastic matrix P , we note that by Lemma 6.7.1 the chain with the stochastic matrix P^d is restricted to one of the A^n for $n = 0, \dots, d-1$. Consider the chain depicted in Figure 8. In this case, the chain has a period of 3 and a corresponding decomposition is given by

- $A_0 = \{2\}$,
- $A_1 = \{4\}$, and
- $A_2 = \{1, 3\}$.

The stochastic matrix for the chain has the form

$$P = \begin{pmatrix} 0 & 1 & 0 & 0 \\ 0 & 0 & 0 & 1 \\ 0 & 1 & 0 & 0 \\ q & 0 & 1-q & 0 \end{pmatrix}$$

for some $q \in (0, 1)$. We note that

$$P^3 = \begin{pmatrix} q & 0 & 1-q & 0 \\ 0 & 1 & 0 & 0 \\ q & 0 & 1-q & 0 \\ 0 & 0 & 0 & 1 \end{pmatrix}.$$

which is no longer irreducible. The dynamics are constrained to the sets $\{2\}$, $\{4\}$ and $\{1, 3\}$ as expected.

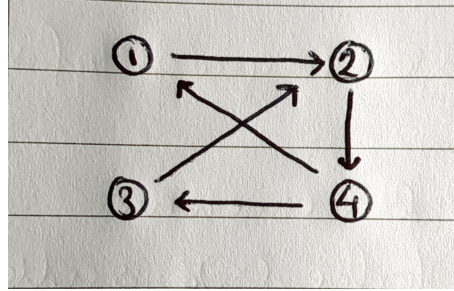


Figure 8: An example of a periodic, irreducible chain whose three-step dynamics is reducible.

Proposition 6.7.3. Suppose $T^n \mu = \mu$ for some fixed $n \in \mathbb{N}$, and let $\hat{\mu} = \frac{1}{n} \sum_{k=1}^n T^k \mu$. Then $T \hat{\mu} = \hat{\mu}$.

Proof. Let $A \in \mathcal{B}(\mathcal{X})$. Then

$$\begin{aligned} T \hat{\mu}(A) &= \frac{1}{n} \sum_{k=1}^n T^{k+1} \mu(A) \\ &= \frac{1}{n} \sum_{k=1}^{n-1} T^{k+1} \mu(A) + \frac{1}{n} T^{n+1} \mu(A) \\ &= \frac{1}{n} \sum_{k=2}^n T^k \mu(A) + \frac{1}{n} T \mu(A) \\ &= \hat{\mu}(A). \end{aligned}$$

□

Remark 6.7.4. Suppose we have a period d chain on \mathcal{X} with a decomposition

$$\mathcal{X} = A_0 \sqcup \cdots \sqcup A_{d-1}$$

as in Lemma 6.7.1, and an invariant measure μ for P^d on A^n . Then

$$\hat{\mu} = \frac{1}{d} \sum_{k=1}^d \mu P^k$$

is an invariant measure for P on \mathcal{X} .

6.8 Ergodic Theorem

Theorem 6.8.1 (Strong Law of Large Numbers). Let $(\xi_n)_{n=1}^\infty$ be a sequence of independent and identically

distributed real-valued random variables with $\mathbb{E}(|\xi_1|) < \infty$. Then

$$\lim_{n \rightarrow \infty} \left(\frac{1}{n} \sum_{k=1}^n \xi_k \right) = \mathbb{E}(\xi_k)$$

almost surely.

Exercise 6.8.2. Suppose X is a time-homogeneous Markov process with initial distribution δ_i for $i \in \mathcal{X}$. Let T^k be the k^{th} return time to state i . Show that the random variables

$$\left\{ \sum_{l=T^k+1}^{T^{k+1}} f(X_l) : k \in \mathbb{N} \right\}$$

are independent and identically distributed.

Theorem 6.8.3 (Ergodic Theorem). Let X be an irreducible, positive recurrent Markov chain on a discrete state space \mathcal{X} . Let π denote its unique invariant probability measure. Then for any π -integrable function $f : \mathcal{X} \rightarrow \mathbb{R}$ it follows that

$$\lim_{n \rightarrow \infty} \left(\frac{1}{n} \sum_{k=1}^n f(X_k) \right) = \sum_{j \in \mathcal{X}} f(j) \pi(j) = \int_{\mathcal{X}} f \, d\pi \quad (6.8.1)$$

almost surely.

Proof. Consider $f \geq 0$. Fix $i \in \mathcal{X}$. Let $T = T_i$ and $T^k = T_i^k$ be the passage times to i . It is sufficient to prove the statement for X initially distributed according to \mathbb{P}_i , as for an arbitrary initial distribution μ we have

$$\mathbb{P}_\mu(\cdot) = \sum_{i \in \mathcal{X}} \mathbb{P}_i(\cdot) \mu(i).$$

Let $\mu = \mu^i$ be the finite measure on \mathcal{X} given by

$$\mu(j) = \mathbb{E}_i \left(\sum_{k=1}^T \mathbf{1}_{\{X_k=j\}} \right) = \mathbb{E}_i(T) \pi(j).$$

Observe that,

$$\begin{aligned} \mathbb{E}_i \left(\sum_{l=1}^T f(X_l) \right) &= \mathbb{E}_i \left(\sum_{l=1}^T \sum_{j \in \mathcal{X}} \mathbf{1}_{\{X_l=j\}} f(j) \right) \\ &\stackrel{(1)}{=} \sum_{j \in \mathcal{X}} f(j) \mathbb{E}_i \left(\sum_{l=1}^T \mathbf{1}_{\{X_l=j\}} \right) \\ &= \mathbb{E}_i(T) \sum_{j \in \mathcal{X}} f(j) \pi(j) \\ &= \mathbb{E}_i(T) \int_{\mathcal{X}} f \, d\pi \\ &< \infty. \end{aligned} \quad (6.8.2)$$

We can exchange the order of summation at (1) as the sum is absolutely convergent. Using Exercise 6.8.2 we can apply the strong law of large numbers to deduce that,

$$\lim_{n \rightarrow \infty} \left(\frac{1}{n} \sum_{k=1}^n \sum_{T^{k-1}+1}^{T^k} f(X_l) \right) = \lim_{n \rightarrow \infty} \left(\frac{1}{n} \sum_{l=1}^{T^n} f(X_l) \right) \stackrel{(6.8.2)}{=} \mathbb{E}_i(T) \int_{\mathcal{X}} f \, d\pi. \quad (6.8.3)$$

almost surely. Since the differences of consecutive passage times are independent and identically distributed, Lemma 6.2.11, we can apply the strong law of large numbers to deduce that

$$\lim_{n \rightarrow \infty} \left(\sum_{k=1}^n (T^k - T^{k-1}) \right) = \lim_{n \rightarrow \infty} \left(\frac{1}{n} T^n \right) = \mathbb{E}_i(T)$$

almost surely. Now let

$$\eta(n) = \sum_{k=1}^n \mathbf{1}_{\{X_k=i\}}$$

for $n \in \mathbb{N}$. Observe that

$$T^{\eta(n)} \leq n < T^{\eta(n)+1}.$$

It follows that

$$\frac{1}{\eta(n)} \sum_{l=1}^{T^{\eta(n)}} f(X_l) \leq \frac{1}{\eta(n)} \sum_{l=1}^n f(X_l) \leq \frac{1}{\eta(n)} \sum_{l=1}^{T^{\eta(n)+1}} f(X_l). \quad (6.8.4)$$

As i is recurrent, by Theorem 6.3.7, we know the event $\{X_k = i\}$ occurs infinitely often with probability one and so $\eta(n) \rightarrow \infty$ almost surely. Therefore,

$$\begin{aligned} \lim_{n \rightarrow \infty} \frac{1}{\eta(n)} \sum_{l=1}^{T^{\eta(n)}} f(X_l) &= \lim_{n \rightarrow \infty} \frac{1}{\eta(n)} \sum_{l=1}^{T^{\eta(n)+1}} f(X_l) \\ &= \mathbb{E}_i(T) \int_{\mathcal{X}} f \, d\pi \end{aligned}$$

almost surely. Hence, by (6.8.4) we have that

$$\lim_{n \rightarrow \infty} \frac{1}{\eta(n)} \sum_{l=1}^n f(X_l) = \mathbb{E}_i(T) \int_{\mathcal{X}} f \, d\pi \quad (6.8.5)$$

almost surely. Taking $f = 1$ in (6.8.5) we note that

$$\lim_{n \rightarrow \infty} \frac{n}{\eta(n)} = \mathbb{E}_i(T) > 0$$

almost surely. Therefore,

$$\begin{aligned} \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{l=1}^n f(X_l) &= \lim_{n \rightarrow \infty} \frac{\eta(n)}{n} \frac{1}{\eta(n)} \sum_{l=1}^n f(X_l) \\ &= \int_{\mathcal{X}} f \, d\pi. \end{aligned}$$

To extend the result to general π -integrable function f we consider the decomposition $f = f^+ - f^-$. □

Remark 6.8.4.

1. Theorem 6.8.3 can be thought of as translating time averages into space averages.
2. Note that the statement of Theorem 6.8.3 holds for every initial distribution of X .

3. Note that Theorem 6.8.3 does not require the chain to be aperiodic, whereas Theorem 6.6.7 and Corollary 6.5.11 do.

4. By taking $f = \mathbf{1}_j$ in (6.8.3) we get that

$$\lim_{n \rightarrow \infty} \left(\frac{1}{n} \sum_{l=1}^{T^n} \mathbf{1}_j(X_l) \right) = \frac{\pi(j)}{\pi(i)}$$

almost surely. In other words, $\frac{\pi(j)}{\pi(i)}$ is the average time spent at state j during one excursion starting and ending at i .

5. By taking $f = \mathbf{1}_j$ in Theorem 6.8.3 we deduce that

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n \mathbf{1}_{\{X_k=j\}} = \pi(j).$$

That is, the average number of times the chain arrives at state $j \in \mathcal{X}$ converges to $\pi(j)$. Note the distinction between this statement and Remark 6.5.12

6.9 Reversible Markov Chains

Exercise 6.9.1. For a, not necessarily discrete, state space \mathcal{X} , suppose we have a transition probability P and a probability measure μ on \mathcal{X} such that $\mu P = \mu$. Show that one can construct a two-sided Markov process $(X_n : n \in \mathbb{Z})$ with

- $\text{Law}(X_n) = \mu$, and
- $\mathbb{P}(X_{n+1} \in A | X_n = x) = P(x, a)$

for all $n \in \mathbb{Z}$, $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$.

Definition 6.9.2. For a two-sided stochastic process $X = (X_n)_{n \in \mathbb{Z}}$, the stochastic process $\hat{X} = (\hat{X}_m)_{m \in \mathbb{Z}}$ with $\hat{X}_m = X_{-m}$ is the reversed stochastic process.

Theorem 6.9.3. Let P be an irreducible and positive recurrent stochastic matrix with an invariant probability measure π . Let $X = (X_n)_{n \in \mathbb{Z}}$ be the two-sided Markov process constructed in Exercise 6.9.1 with transition probability P and $\text{Law}(X_n) = \pi$. Let $\hat{X} = (\hat{X}_m)_{m \in \mathbb{Z}}$ be the reversed stochastic process of X . Then \hat{X} is a time-homogeneous Markov chain with stochastic matrix

$$\hat{P}_{ji} = P_{ij} \frac{\pi(i)}{\pi(j)}.$$

Proof. As P is irreducible and positive recurrent there exists an invariant probability measure π . In particular, as

every state $j \in \mathcal{X}$ is positive recurrent we have $\pi(j) > 0$. Given consecutive times $n_0 < n_1 < \dots < n_l$, we have

$$\begin{aligned}
 \mathbb{P}(\hat{X}_{n_0} = i_0, \dots, \hat{X}_{n_l} = i_l) &= \mathbb{P}(X_{-n_0} = i_l, \dots, X_{-n_l} = i_0) \\
 &= \pi(i_l)P_{i_l i_{l-1}} \dots P_{i_1 i_0} \\
 &= \left(\frac{\pi(i_l)}{\pi(i_{l-1})} P_{i_l i_{l-1}} \right) \dots \left(\frac{\pi(i_1)}{\pi(i_0)} P_{i_1 i_0} \right) \pi(i_0) \\
 &= \hat{P}_{i_{l-1} i_l} \dots \hat{P}_{i_0 i_1} \pi(i_0) \\
 &= \pi(i_0) \hat{P}_{i_0 i_1} \dots \hat{P}_{i_{l-1} i_l}.
 \end{aligned}$$

Therefore, using Corollary 5.1.5 we deduce that \hat{X} is Markov with transition probabilities \hat{P} . □

Definition 6.9.4. A stochastic matrix P and a measure π satisfy detailed balance if

$$\pi(i)P_{ij} = \pi(j)P_{ji}$$

for all $i, j \in \mathcal{X}$.

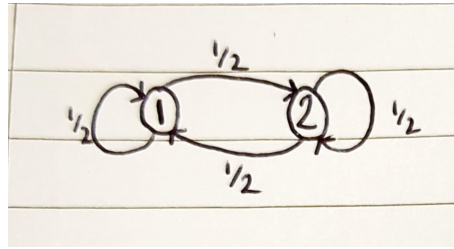


Figure 9: An example of a chain satisfying detailed balance.

Proposition 6.9.5. Let π be a probability measure that satisfies detailed balance with respect to P . Then π is P -invariant.

Proof. Observe that

$$\begin{aligned}
 (\pi P)(i) &= \sum_{j \in \mathcal{X}} \pi(j)P_{ji} \\
 &= \sum_{j \in \mathcal{X}} \pi(i)P_{ij} \\
 &= \pi(i) \sum_{j \in \mathcal{X}} P_{ij} \\
 &= \pi(i).
 \end{aligned}$$

□

Example 6.9.6. Consider the chain of Example 6.2.2. Then a measure $\pi = (\pi(1), \pi(2))$ satisfies detailed balance if

$$\pi(1)\alpha = \pi(2)\beta.$$

If π is a probability measure then $\pi(1) + \pi(2) = 1$ so that

$$\begin{cases} \pi(1) = \frac{\beta}{\alpha+\beta} \\ \pi(2) = \frac{\alpha}{\alpha+\beta}. \end{cases}$$

One can then verify Proposition 6.9.5 with

$$\begin{aligned} \pi P &= \begin{pmatrix} \frac{\beta}{\alpha+\beta} & \frac{\alpha}{\alpha+\beta} \end{pmatrix} \begin{pmatrix} 1-\alpha & \alpha \\ \beta & 1-\beta \end{pmatrix} \\ &= \begin{pmatrix} \frac{\beta(1-\alpha)+\alpha\beta}{\alpha+\beta} & \frac{\alpha\beta+\alpha(1-\beta)}{\alpha+\beta} \end{pmatrix} \\ &= \pi. \end{aligned}$$

Theorem 6.9.7. In the setting of Theorem 6.9.3, if π and P satisfy detailed balance, then $\hat{P} = P$ with $\text{Law}(\hat{X}) = \text{Law}(X)$.

A process satisfying the statement of Theorem 6.9.7 is called reversible. Intuitively, in an irreducible chain that satisfies detailed balance, one cannot tell whether the chain is being propagated forward or backwards in time.

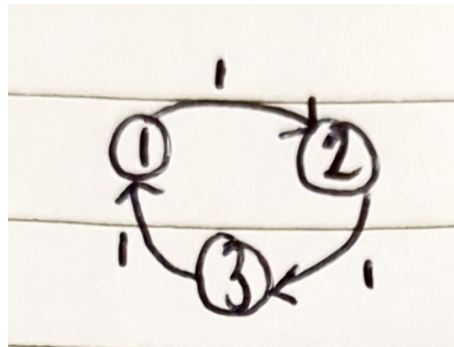


Figure 10: We know this chain is not a detailed balanced chain, as we can distinguish whether the chain is moving forward or backwards in time.

6.9.1 Markov Chain Monte Carlo

Suppose we want to simulate a probability measure π on a large but finite state space \mathcal{X} . That is, we want to calculate $\sum_{i \in \mathcal{X}} f(i)\pi(i)$ for observables $f : \mathcal{X} \rightarrow \mathbb{R}$. One approach is to construct a Markov chain for which π is its invariant distribution and use the ergodic law of large numbers to get that

$$\sum_{i \in \mathcal{X}} f(i)\pi(i) = \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n f(X_k).$$

Despite not knowing the specific values of $\pi(i)$ for $i \in \mathcal{X}$, one can often model them up to constants of proportionality. We note that the ratios

$$\delta(j, i) = \frac{\pi(j)}{\pi(i)}$$

eliminate those constants. Ideally, one would sum these ratios over $j \in \mathcal{X}$ to determine $\frac{1}{\pi(i)}$, however, in practice this sum is expensive. Moreover, $\delta(i, j)$ may be intractable to compute for certain (i, j) .

Example 6.9.8. Let $\mathcal{X} = \{-1, 1\}^\Lambda$ with $\Lambda = [-N, N]^d \cap \mathbb{Z}^d$. For a configuration $\sigma = (\sigma_x : x \in \Lambda)$, let

$$\pi(\sigma) \propto \exp\left(\frac{\beta}{2} \sum_{x, y \in \Lambda, |x-y|=1} \sigma_x \sigma_y\right)$$

for some $\beta > 0$. This model is called the Ising model. An application of the Ising model is to approximate the dynamics of molecular spin within a material. In this setting, calculating the normalisation constant to determine $\pi(\sigma)$ is difficult as materials contain a large number of molecules. However, if σ and σ' differ at exactly one $\bar{x} \in \Lambda$ then

$$\delta(\sigma, \sigma') = \exp\left(\beta \sum_{y \in \Lambda, |x-y|=1} (\sigma_x - \sigma'_x) \sigma_y\right).$$

Using these ideas we can now consider constructing a suitable Markov chain. Start with some irreducible Markov chain with transition probability Q and then set

1. $P_{ij} = Q_{ij} \wedge \delta(j, i)Q_{ji}$ for $i \neq j$, and
2. $P_{ii} = 1 - \sum_{j \neq i} P_{ij}$.

The chain with transition matrix P on \mathcal{X} is not necessarily irreducible. However, showing P and π satisfy the detailed balance tells us that π is an invariant measure of the chain,

$$\begin{aligned} \pi(i)P_{ij} &= \pi(i)Q_{ij} \wedge \pi(j)Q_{ji} \\ &= \pi(j)Q_{ji} \wedge \pi(i)Q_{ij} \\ &= \pi(j)(Q_{ji} \wedge \delta(i, j)Q_{ij}) \\ &= \pi(j)P_{ji}. \end{aligned}$$

As multiple invariant measures may exist, due to P potentially not being irreducible, the chain may converge to a different invariant measure. However, supposing that the chain is irreducible, or that it will converge to π , we know by Theorem 6.8.3 that

$$\frac{1}{n} \sum_{k=1}^n f(X_k) \approx \sum_{i \in \mathcal{X}} f(i)\pi(i)$$

for large n .

6.10 Finite State Space Markov Chains

Throughout we will suppose that $\mathcal{X} = \{1, \dots, N\}$. Let

$$\delta_n = \min_{1 \leq i, j \leq N} P_{ij}^n$$

for $n \in \mathbb{N}$.

Proposition 6.10.1. Let P be a stochastic matrix on \mathcal{X} . Then δ_n is increasing in n . Moreover, the following are equivalent.

1. P is irreducible and aperiodic.
2. P^n is irreducible for every $n \geq 1$.
3. There exists an $n_0 \in \mathbb{N}$ such that $\delta_{n_0} > 0$.

Proof. The Chapman-Kolmogorov equation tells us that

$$\begin{aligned} P_{ij}^{m+n} &\geq \sum_{k \in \mathcal{X}} P_{ik}^m P_{kj}^n \\ &\geq \delta_n \sum_{k \in \mathcal{X}} P_{ik}^m \\ &= \delta_n \end{aligned}$$

for all $i, j \in \mathcal{X}$. Hence, $\delta_{m+n} = \min_{1 \leq i, j \leq n} P_{ij}^{m+n} \geq \delta_n$ and so δ_n is increasing in n

(2) \Rightarrow (1). If P were not aperiodic, then as the chain is irreducible and finite it follows that it must be periodic. Let P be d -periodic, then using the reasoning of Example 6.7.2 we see that P^d is reducible, which is a contradiction.

(3) \Rightarrow (1). By assumption, P^{n_0} has strictly positive entries, which implies that P is irreducible. As δ_n is increasing we know that $\delta_n > \delta_{n_0} > 0$ for all $n \geq n_0$. Therefore, P^n also has strictly positive entries for $n \geq n_0$. Consequently, for any $i \in \mathcal{X}$ we have that $\{n_0, n_0 + 1, \dots\} \subseteq R(i)$ which implies that $d(i) = 1$ and so P is aperiodic.

(3) \Rightarrow (2). For $n < n_0$ there exists a $k \in \mathbb{N}$ such that $nk \geq n_0$. Hence, $\delta_{nk} \geq \delta_{n_0} > 0$, which implies that P^n is irreducible.

(1) \Rightarrow (3). By Lemma 9.2.9 we know that for all $1 \leq i \leq N$ there exists a k_i such that $kd(i) \in R(i)$ for $k \geq k_i$. As P is aperiodic $d(i) = 1$ for each $i \in \mathcal{X}$ and so for all $n \geq N_0$ it follows that $n \in R(i)$ for each $i \in \mathcal{X}$. That is, $P_{ii}^n > 0$ for each $i \in \mathcal{X}$ and $n \geq N_0$. As P is irreducible it follows that for $i, j \in \mathcal{X}$ there exists a $m(i, j) \in \mathbb{N}$ such that $P_{ij}^{m(i, j)} > 0$. Using the Chapman-Kolmogorov equation we $P_{ij}^{n+m(i, j)} \geq P_{ii}^n P_{ij}^{m(i, j)}$. Hence, $n_0 = N_0 + \max_{i, j \in \mathcal{X}} (m(i, j))$ is such that $\delta_{n_0} > 0$. \square

Exercise 6.10.2. Let P be a stochastic matrix. Then for a recurrent state $i \in \mathcal{X}$ and $j \in [i]$, show that $\mathbb{P}_j(T_i < \infty) = 1$.

Lemma 6.10.3. Let P be an irreducible, aperiodic stochastic matrix on a finite state space. Then for any $i, j \in \mathcal{X}$ and $\alpha > 0$ we have that

$$\mathbb{E}_j(T_i^\alpha) < \infty.$$

Proof. Using Exercise 6.10.2 we know that $\mathbb{P}_j(T_i^\alpha = \infty) = 0$ and so we can write

$$\mathbb{E}_j(T_i^\alpha) = \sum_{n=0}^{\infty} n^\alpha \mathbb{P}_j(T_i = n).$$

Hence,

$$\mathbb{E}_j(T_i^\alpha) \leq \sum_{n=0}^{\infty} n^\alpha \mathbb{P}_j(T_i > n - 1).$$

By Proposition 6.10.1 we know that there exists a $n_0 \in \mathbb{N}$ such that $\delta_{n_0} > 0$, consequently,

$$\begin{aligned} \mathbb{P}(X_{n_0(k+1)} \neq i | X_{n_0k} \neq i) &= \sum_{l \neq i} \mathbb{P}(X_{n_0(k+1)} \neq i | X_{n_0k} = l) \frac{\mathbb{P}(X_{n_0k} = l)}{\mathbb{P}(X_{n_0k} \neq i)} \\ &\leq \sum_{l \neq i} (1 - \delta_{n_0}) \frac{\mathbb{P}(X_{n_0k} = l)}{\mathbb{P}(X_{n_0k} \neq i)} \\ &\leq 1 - \delta_{n_0} \end{aligned}$$

It follows that

$$\begin{aligned}
\mathbb{P}_j(T_i > n_0(k+1)) &\leq \mathbb{P}_j(X_{n_0(k+1)} \neq i, T_i > n_0k) \\
&= \mathbb{P}(X_{n_0(k+1)} \neq i | T_i > n_0k) \mathbb{P}_j(T_i > n_0k) \\
&\stackrel{(1)}{=} \mathbb{P}(X_{n_0(k+1)} \neq i | X_{n_0k} \neq i) \mathbb{P}_j(T_i > n_0k) \\
&\leq (1 - \delta_{n_0}) \mathbb{P}_j(T_i > n_0k) \\
&\stackrel{(2)}{\leq} (1 - \delta_{n_0})^{k+1}
\end{aligned}$$

where in (1) we used the fact that $\{T_i > n_0k\} \in \mathcal{F}_{n_0k}$ and so we can apply the Markov property to condition with respect to $\{X_{n_0k} \neq i\} \in \sigma(X_{n_0k})$ instead. In (2), we are just iterating the previous computations. Therefore,

$$\begin{aligned}
\sum_{n=2n_0}^{\infty} n^\alpha \mathbb{P}_j(T_i > n-1) &= \sum_{k=2}^{\infty} \sum_{j=n_0k}^{n_0(k+1)-1} j^\alpha \mathbb{P}_j(T_i > j-1) \\
&\leq \sum_{k=2}^{\infty} \sum_{j=n_0k}^{n_0(k+1)-1} (n_0(k+1))^\alpha \mathbb{P}_j(T_i > n_0k-1) \\
&\leq \sum_{k=2}^{\infty} \sum_{j=n_0k}^{n_0(k+1)-1} (n_0(k+1))^\alpha (1 - \delta_{n_0})^{k+1} \\
&\leq \sum_{k=2}^{\infty} n_0^{\alpha+1} (k+1)^\alpha (1 - \delta_{n_0})^{k-1} \\
&< \infty.
\end{aligned}$$

□

6.10.1 Perron-Frobenius

Let $\mathbb{R}_+^N = \{\eta \in \mathbb{R}^N : \eta(i) \geq 0 \text{ for all } 1 \leq i \leq N\}$.

Lemma 6.10.4. *Let P be irreducible and aperiodic on $\mathcal{X} = \{1, \dots, N\}$. Then there exists some $n \in \mathbb{N}$ and $\delta > 0$ such that for every $\eta \in \mathbb{R}_+^N$ we have*

$$(\eta P^n)(i) \geq \delta \|\eta\|_1$$

for every $i \in \mathcal{X}$, where $\|\eta\|_1 = \sum_{i \in \mathcal{X}} \eta(i)$.

Proof. Take $n = n_0$ and $\delta = \delta_{n_0}$ as in Proposition 6.10.1, then

$$\begin{aligned}
(\eta P^n)(i) &= \sum_{j=1}^N \eta(j) P_{ji}^n \\
&\geq \delta_n \sum_{j=1}^N \eta(j) \\
&= \delta_n \|\eta\|_1.
\end{aligned}$$

□

Lemma 6.10.5. *Suppose that P is an irreducible stochastic matrix on $\mathcal{X} = \{1, \dots, N\}$. Then there exists*

an $n \in \mathbb{N}$ and $\delta > 0$ such that $T^n = \frac{1}{n} \sum_{j=1}^n P^j$ satisfies

$$\min_{1 \leq i, j \leq N} T_{ij}^n \geq \delta.$$

Proof. If P is aperiodic then this result follows from Lemma 6.10.4. So let P not be aperiodic, then as we are operating on a finite state space, and P is irreducible, it follows that P is periodic. Suppose that P has period d , then we can write

$$\mathcal{X} = A_0 \sqcup \cdots \sqcup A_{d-1}$$

where P^d is irreducible and aperiodic on A_j for $0 \leq j \leq d-1$. In particular, for $0 \leq l \leq d-1$, there exists a $m_l \in \mathbb{N}$ such that

$$\min_{i, j \in A_l} (P^d)_{ij}^{m_l} > 0.$$

Now let $m = \max_{0 \leq l \leq d-1} m_l > 0$ and $n = 2dm + d$. Suppose $i \in A_l$ and $j \in A_{l'}$. If $l = l'$ then $P_{ij}^d > 0$ and $d \leq 2dm + d$. Suppose instead that $|l - l'| = r > 0$. By irreducibility we know that there exists an $i' \in A_l$ and $j' \in A_{l'}$ such that $P_{i'j'}^r > 0$. Therefore,

$$P_{ij}^{dm_l + dm_{l'} + r} \geq P_{ii'}^{dm_l} P_{i'j'}^r P_{j'j}^{dm_{l'}} > 0$$

where $dm_l + dm_{l'} + r \leq 2dm + d$. Thus, $T^n = \frac{1}{n} \sum_{j=1}^n P^j$ has the property that

$$\min_{1 \leq i, j \leq N} T_{ij}^n > 0,$$

where we maintain the strict inequality as our state space is finite. \square

Theorem 6.10.6 (Perron-Frobenius). *Let P be an $N \times N$ irreducible stochastic matrix on a finite state space \mathcal{X} . Then all the eigenvalues of P satisfy $|\lambda| \leq 1$. Moreover, one is a left-eigenvalue with a unique real left-eigenvector π , up to multiplication by a constant, that is $\pi P = \pi$. In particular, π can be chosen so that $\pi(i) > 0$ for every $i \in \mathcal{X}$ and $\sum_{i=1}^N \pi(i) = 1$.*

Proof. As P is a stochastic matrix it follows that

$$\|\eta P\|_1 \leq \|\eta\|_1$$

for every $\eta \in \mathbb{C}^N$ which shows that $|\lambda| \leq 1$ for any eigenvalue λ of P . Observe that $\frac{1}{N}(1, \dots, 1)$ is a right-eigenvector of P with eigenvalue one, and so there must be a left-eigenvector π of P with eigenvalue one. Since P is real we can take π to be real, moreover, we can normalize π such that $\|\pi\|_1 = 1$. Now suppose that π_+ and π_- are both non-zero, where $\pi_{\pm}(i) = \max(\pm\pi(i), 0)$ for $1 \leq i \leq N$. That is, π contains both positive and negative entries. Let $\alpha = \min(\|\pi_+\|_1, \|\pi_-\|_1) > 0$. By the irreducibility of P we can consider a T^n and $\delta > 0$ satisfying the statement of Lemma 6.10.5. Note that π is an eigenvector of T^n with eigenvalue one. For $\eta \in \mathbb{R}^N$, we can write $\eta = \eta_+ - \eta_-$. Thus,

$$\begin{aligned} \|\pi\|_1 &= \|\pi T^n\|_1 \\ &= \|\pi_+ T^n - \pi_- T^n\|_1 \\ &\leq \|\pi_+ T^n - \delta\alpha \mathbf{1}\|_1 + \|\pi_- T^n - \delta\alpha \mathbf{1}\|_1 \\ &\stackrel{(1)}{\leq} \|\pi_+ T^n\|_1 + \|\pi_- T^n\|_1 - 2\|\alpha\delta \mathbf{1}\|_1 \\ &= \|\pi_+ T^n\|_1 + \|\pi_- T^n\|_1 - 2\alpha\delta N \\ &\stackrel{(2)}{\leq} \|\pi_+\|_1 + \|\pi_-\|_1 - 2\alpha\delta N \\ &\stackrel{(3)}{=} \|\pi\|_1 - 2\alpha\delta N, \end{aligned}$$

which is a contradiction for $\alpha > 0$. In (1) we have used Lemma 6.10.5 to deduce that

$$\|\pi_{\pm} T^n - \delta\alpha \mathbf{1}\|_1 = \|\pi_{\pm} T^n\|_1 - \|\delta\alpha \mathbf{1}\|_1.$$

In (2) we have used that $\|\eta T^n\| \leq \|\eta\|_1$. In (3), we have used $\|\pi\|_1 = \|\pi_+\|_1 + \|\pi_-\|_1$. Consequently, we establish that $\pi \in \mathbb{R}_+^N$. Therefore, by construction of the T^n it follows that

$$\pi(i) = (\pi T^n)(i) \geq \|\pi\|_1,$$

that is, all its entries are positive. For uniqueness, we suppose that $\tilde{\pi} \in \mathbb{R}_+^N$ is another real left-eigenvector with eigenvalue one. Moreover, we can assume that $\tilde{\pi}$ has positive entries and $\|\tilde{\pi}\|_1 = 1$. Then $\theta = \pi - \tilde{\pi}$ is another real left-eigenvector with eigenvalues one, and so its entries must all have the same sign by our previous arguments. However, note that

$$\sum_{i \in \mathcal{X}} \theta(i) = \sum_{i \in \mathcal{X}} \pi(i) - \sum_{i \in \mathcal{X}} \tilde{\pi}(i) = \|\pi\|_1 - \|\tilde{\pi}\|_1 = 0.$$

Hence, $\theta(i) = 0$ for all $i \in \mathcal{X}$, which implies that $\pi = \tilde{\pi}$. □

Remark 6.10.7.

- One of the consequences of Theorem 6.10.6, is that any irreducible Markov chain on a finite state space has an invariant probability measure.
- The specific π outlined in Theorem 6.10.6 is known as the Perron-Frobenius vector of P .

6.11 Solution to Exercises

Exercise 6.1.3

Solution.

- As $P_{ii}^0 = 1$ for all $i \in \mathcal{X}$ it follows that $i \leftrightarrow i$.
- If $i \leftrightarrow j$ then $i \rightarrow j$ and $j \rightarrow i$, hence $j \leftrightarrow i$.
- If $i \leftrightarrow j$ and $j \leftrightarrow k$, then there exists an $n_1 \in \mathbb{N}$ such that $P_{ij}^{n_1} > 0$ and an $n_2 \in \mathbb{N}$ such that $P_{jk}^{n_2} > 0$. Therefore,

$$P_{ik}^{n_1+n_2} \geq P_{ij}^{n_1} P_{jk}^{n_2} > 0$$

and so $i \rightarrow k$. Similarly, $k \rightarrow i$ so that $i \leftrightarrow k$. □

Exercise 6.1.6

Solution. Let $i, j \in \mathcal{X}$ be such that $[i] \leq [j]$. Consider $i' \in [i]$ and $j' \in [j]$. Then by Lemma 6.1.5 we have $j' \rightarrow i'$ and thus $[i'] \leq [j']$. Therefore, \leq is well-defined.

- As $i \rightarrow i$ it follows that $[i] \leq [i]$.
- If $[i] \leq [j]$ and $[j] \leq [k]$ then $j \rightarrow i$ and $k \rightarrow j$. Therefore, $k \rightarrow i$ which implies that $[i] \leq [k]$.
- If $[i] \leq [j]$ and $[j] \leq [i]$ then $i \leftrightarrow j$ so that $[i] = [j]$. □

Exercise 6.2.8

Solution. Note from Lemma 6.2.6 that if $i \in [j]$ then $j \rightarrow i$ so that $\mathbb{P}_j(T_i < \infty) > 0$. Therefore,

$$\begin{aligned}\mathbb{P}_\mu(T_j < \infty) &= \sum_{i \in [j]} \mathbb{P}_i(T_j < \infty) \mu(\{i\}) \\ &\stackrel{\text{Lem. 6.2.7}}{=} \sum_{i \in [j]} \mu(\{i\}) \\ &= 1.\end{aligned}$$

□

Exercise 6.2.10

Solution. Proceed by induction on n .

- As $T_j^0 = 0$, it follows that

$$\{T_j^0 \leq k\} = \mathcal{X} \in \mathcal{F}_k$$

for every $k \in \mathbb{N}$. Therefore, T_j^0 is a $(\mathcal{F}_k)_{k=0}^\infty$ -stopping time.

- Assume that T_j^m is a $(\mathcal{F}_k)_{k=0}^\infty$ -stopping time for $m \leq n-1$. Observe that $\{T_j^n \leq k\} = \emptyset \in \mathcal{F}_k$ for $k \leq n-1$. Therefore, suppose that $k \geq n$ then

$$\{T_j^n \leq k\} = \bigcup_{l=n-1}^k \left(\{T_j^{n-1} = l\} \cap \bigcup_{p=l+1}^k \{X_p = j\} \right).$$

We know that $\{X_p = j\} \in \mathcal{F}_p \subseteq \mathcal{F}_k$ and by the inductive assumption we know that $\{T_j^{n-1} = l\} \in \mathcal{F}_l \subseteq \mathcal{F}_k$. Therefore, $\{T_j^n \leq k\} \in \mathcal{F}_k$, which means that T_j^n is a $(\mathcal{F}_k)_{k=0}^\infty$ -stopping time.

□

Exercise 6.8.2

Solution. Let $Y_k = \sum_{j=T^{k+1}}^{T^{k+1}} f(X_j)$, and consider $g, h \in \mathcal{B}(\mathbb{R})$. Without loss of generality suppose that $k' = k+n$ for $n \geq 1$. Then,

$$\begin{aligned}\mathbb{E}(g(Y_k)h(Y_{k'})) &= \mathbb{E}(\mathbb{E}(g(Y_k)h(Y_{k'})|\mathcal{F}_{T^k})) \\ &\stackrel{\text{SMP}}{=} \mathbb{E}_i(g(Y_0)h(Y_n)) \\ &= \mathbb{E}_i(\mathbb{E}_i(g(Y_0)h(Y_n)|\mathcal{F}_{T^n})) \\ &\stackrel{(1)}{=} \mathbb{E}_i(g(Y_0)\mathbb{E}_i(h(Y_n)|\mathcal{F}_{T^n})) \\ &\stackrel{\text{SMP}}{=} \mathbb{E}_i(g(Y_0)\mathbb{E}_i(h(Y_0))) \\ &= \mathbb{E}_i(g(Y_0))\mathbb{E}_i(h(Y_0)) \\ &\stackrel{\text{SMP}}{=} \mathbb{E}(g(Y_k))\mathbb{E}(h(Y_k)),\end{aligned}$$

where SMP denotes an application of the strong Markov property, and (1) follows as $g(Y_0)$ is \mathcal{F}_{T^n} -measurable. As $g, h \in \mathcal{B}(\mathbb{R})$ were arbitrary this shows the mutual independence of Y_k and $Y_{k'}$. One easily extends the above argument to show that

$$\mathbb{E} \left(\prod_{i=1}^m g_i(Y_i) \right) = \prod_{i=1}^m \mathbb{E}(g_i(Y_i))$$

where $g_i \in \mathcal{B}(\mathbb{R})$ and $m \in \mathbb{N}$. This then shows the independence of the random variables $(Y_k)_{k \in \mathbb{N}}$. Moreover, for any $k \in \mathbb{N}$ we have that

$$\begin{aligned} \mathbb{P}(Y_k \in A) &= \mathbb{E}(\mathbf{1}_A(Y_k)) \\ &= \mathbb{E}(\mathbb{E}(\mathbf{1}_A(Y_k) | \mathcal{F}_{T^k})) \\ &\stackrel{\text{SMP}}{=} \mathbb{E}_i(\mathbf{1}_A(Y_0)) \\ &= \mathbb{P}_i(Y_0 \in A) \end{aligned}$$

where SMP denotes an application of the strong Markov property. Therefore, each Y_k is identically distributed. Hence, we have shown that the random variables

$$\left\{ \sum_{l=T^{k+1}}^{T^{k+1}} f(X_l) : k \in \mathbb{N} \right\}$$

are independent and identically distributed. □

Exercise 6.9.1

Solution. For any $m \in \mathbb{Z}$, as μ is stationary with respect to P we note that

$$\mathbb{P}(X_m \in A_0, \dots, X_{m+n} \in A_n) = \mathbb{P}(X_0 \in A_0, \dots, X_n \in A_n) \quad (6.11.1)$$

for any $n \in \mathbb{N}$. Note that the family of measures $(\mu_n)_{n \in \mathbb{N}}$ given by $\mu_n = \text{Law}(X_0, \dots, X_n)$ is consistent. Consequently, we can apply Theorem 3.2.7 to construct a stochastic process such that

- $\text{Law}(X_n) = \mu$, and
- $\mathbb{P}(X_{n+1} | X_n = x) = P(x, a)$

for $n \in \mathbb{N}$, $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$. Using (6.11.1) it is clear that we can extend this stochastic process to \mathbb{Z} , with the properties detailed above now holding for $n \in \mathbb{Z}$, $x \in \mathcal{X}$ and $A \in \mathcal{B}$ as required. □

Exercise 6.10.2

Solution. Let $\eta_i = \sum_{n=1}^{\infty} \mathbf{1}_{\{X_n=i\}}$. Then as $i \in \mathcal{X}$ is a recurrent state it follows that $\mathbb{P}_i(\eta_i = \infty) = 1$. Note that by the Markov property we have

$$\{\eta_i = \infty\} = \{X_n = i, \text{ for some } n \geq m\}$$

for any $m \in \mathbb{N}$. Observe that

$$\begin{aligned} 1 &= \mathbb{P}_i(X_n = i, \text{ for some } n \geq m) \\ &= \sum_{k \in \mathcal{X}} P_{ik}^m \mathbb{P}_i(X_n = i, \text{ for some } n \geq m | X_m = k) \\ &= \sum_{k \in \mathcal{X}} P_{ik}^m \mathbb{P}_k(T_i < \infty) \\ &\leq \sum_{k \in \mathcal{X}} P_{ik}^m \\ &= 1. \end{aligned}$$

Therefore, for the $k \in \mathcal{X}$ for which $P_{ik}^m \neq 0$ it follows that $\mathbb{P}_k(T_i < \infty) = 1$. Hence, for $j \in [i]$ we can let $m \in \mathbb{N}$ be such that $P_{ij}^m > 0$ and deduce that $\mathbb{P}_j(T_i < \infty) = 1$. □

7 Continuous State Space Markov Processes

We now generalise to complete separable metric spaces \mathcal{X} , such as $\mathcal{X} = \mathbb{R}^n$ or the infinite dimensional Banach space $\mathcal{X} = \mathcal{C}([0, 1])$.

7.1 Weak Convergence

In the time-homogeneous setting, transition probabilities $P(\cdot, \bullet)$ are such that for $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$P(x, A) = \mathbb{P}(X_{n+1} \in A | X_n = x).$$

We can use P to act on functions and measures.

- On functions $T_* : \mathcal{B}_b(\mathcal{X}) \rightarrow \mathcal{B}_b(\mathcal{X})$ is given by $(T_*f)(\cdot) = \int_{\mathcal{X}} f(y)P(\cdot, dy)$. Equivalently,

$$(T_*f)(x) = \mathbb{E}(f(X_{n+1}) | X_n = x).$$

- On measures $T^* : \mathcal{P}(\mathcal{X}) \rightarrow \mathcal{P}(\mathcal{X})$ is given by $(T^*\mu)(\bullet) = \int_{\mathcal{X}} P(y, \bullet)\mu(dy)$. Equivalently,

$$(T^*\mu)(A) = \int_{\mathcal{X}} \mathbb{P}(X_{n+1} \in A | X_n = y)\mu(dy).$$

Remark 7.1.1. In most cases $P(\bullet, A) = 0$ for $A = \{y\}$ a singleton. Therefore, notions developed in Section 6 regarding irreducibility, recurrence and transience do not generalise to this setting.

Henceforth, we will denote the set of bounded and continuous functions on \mathcal{X} as $\mathcal{C}_b(\mathcal{X})$.

Lemma 7.1.2. Let $\mu, \mu' \in \mathcal{P}(\mathcal{X})$ be such that

$$\int_{\mathcal{X}} f d\mu = \int_{\mathcal{X}} f d\mu'$$

for every $f \in \mathcal{C}_b(\mathcal{X})$. Then $\mu = \mu'$.

As $\mathcal{C}_b(\mathcal{X})$ can distinguish probability measures, it is natural to use them to define a notion of convergence.

Definition 7.1.3. A sequence of probability measures $(\mu_n)_{n \in \mathbb{N}} \subseteq \mathcal{P}(\mathcal{X})$ converges weakly to $\mu \in \mathcal{P}(\mathcal{X})$ if

$$\int_{\mathcal{X}} f(x) \mu_n(dx) \xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}} f(x) \mu(dx)$$

for every $f \in \mathcal{C}_b(\mathcal{X})$.

Remark 7.1.4.

- On $\mathcal{P}(\mathcal{X})$ there exists a metric $d(\cdot, \cdot)$, known as the Levy-Prokhorov metric, that topologizes weak convergence. That is, μ_n converges weakly to μ if and only if $d(\mu_n, \mu) \rightarrow 0$.
- Let $(Z_n)_{n \in \mathbb{N}}$ be a sequence of random variables, then $Z_n \rightarrow Z$ almost everywhere implies $Z_n \rightarrow Z$ in probability which implies that $\text{Law}(Z_n) \rightarrow \text{Law}(Z)$ weakly.

7.2 The Feller Property

Definition 7.2.1. A transition probability P , or T , is Feller if T^* maps $C_b(\mathcal{X}) \subseteq \mathcal{B}_b(\mathcal{X})$ into $C_b(\mathcal{X})$. That is, $T^*(C_b(\mathcal{X})) \subseteq C_b(\mathcal{X})$. Moreover, P , or T , is strong Feller if $T^*(\mathcal{B}_b(\mathcal{X})) \subseteq C_b(\mathcal{X})$.

Remark 7.2.2. A transition probability P being Feller is equivalent to the map $\mathcal{X} \rightarrow \mathcal{P}(\mathcal{X})$ given by $x \mapsto P(x, \bullet)$ being continuous, where $\mathcal{P}(\mathcal{X})$ is equipped with the topology of weak convergence. Indeed, let $f \in C_b(\mathcal{X})$ and let $(x_n)_{n \in \mathbb{N}}$ be a sequence converging to $x \in \mathcal{X}$. Then

$$\begin{aligned} \lim_{n \rightarrow \infty} (Tf)(x_n) &= \lim_{n \rightarrow \infty} \int_{\mathcal{X}} f(y) P(x_n, dy) \\ &\stackrel{(1)}{=} \int_{\mathcal{X}} f(y) P(x, dy) \\ &= (Tf)(x), \end{aligned}$$

where (1) is justified either by the continuity of P with respect to the weak topology or by the Feller property of T .

Lemma 7.2.3. Let $f \in \mathcal{B}_b(\mathbb{R}^n)$ and $g \in L^1(\mathbb{R}^n)$, then $f \star g \in C_b(\mathbb{R}^n)$, where

$$(f \star g)(x) = \int_{\mathbb{R}^n} f(y) g(x - y) dy.$$

Proof. Step 1: Show that $\|f \star g\|_{L^\infty} \leq \|f\|_{L^\infty} \|g\|_{L^1}$.
Observe that

$$\begin{aligned} |(f \star g)(x)| &\stackrel{T.1}{\leq} \int_{\mathbb{R}^n} |f(y)| |g(x - y)| dy \\ &\leq \|f\|_{L^\infty} \int_{\mathbb{R}^n} |g(x - y)| dy \\ &\stackrel{(1)}{=} \|f\|_{L^\infty} \|g\|_{L^1}, \end{aligned}$$

where the translational invariance of the Lebesgue measure is used in (1). Therefore, $f \star g$ is bounded.

Step 2: For $g \in C_c^\infty(\mathbb{R}^n)$ show that $f \star g \in C_b(\mathbb{R}^n)$.

Let $g \in C_c^\infty$, then for any $x, x' \in \mathbb{R}^n$ we have

$$(f \star g)(x) - (f \star g)(x') = \int_{\mathbb{R}^n} f(y) (g(x - y) - g(x' - y)) dy.$$

By the continuity of g we know that $g(x - y) - g(x' - y) \rightarrow 0$ as $x \rightarrow x'$. Moreover, we know that

$$|f(y) (g(x - y) - g(x' - y))| \leq 2\|f\|_{L^\infty} \|g\|_{L^\infty} < \infty,$$

where the finiteness follows from the fact that f is bounded and g is continuous with compact support and so is also bounded. Therefore, using the dominated convergence theorem,

$$|(f \star g)(x) - (f \star g)(x')| \rightarrow 0$$

as $x \rightarrow x'$.

Step 3: Given $g \in L^1(\mathbb{R}^n)$, find a sequence $(g_n)_{n \in \mathbb{N}} \subseteq C_c^\infty(\mathbb{R}^n)$ such that $g_n \rightarrow g$ in $L^1(\mathbb{R}^n)$.

Recall that $C_c^\infty(\mathbb{R}^n)$ is dense in $L^1(\mathbb{R}^n)$. So for $g \in L^1(\mathbb{R}^n)$ there exists a sequence $(g_n)_{n \in \mathbb{N}} \subseteq C_c^\infty(\mathbb{R}^n)$ such that $g_n \rightarrow g$ in $L^1(\mathbb{R}^n)$.

Step 4: Argue that $f \star g_n \rightarrow f \star g$ in $L^\infty(\mathbb{R}^n)$.

Let $x \in \mathbb{R}^n$ with $(x_n)_{n \in \mathbb{N}} \subseteq \mathbb{R}^n$ such that $x_n \rightarrow x$. Note that

$$\begin{aligned} |(f \star g)(x) - (f \star g)(x_n)| &\leq |(f \star g)(x) - (f \star g_k)(x)| + |(f \star g_k)(x) - (f \star g_k)(x_n)| \\ &\quad + |(f \star g_k)(x_n) - (f \star g)(x_n)|. \end{aligned} \tag{7.2.1}$$

Note that,

$$\begin{aligned} |(f \star g)(x) - (f \star g_k)(x)| &\leq \|f\|_{L^\infty} \int_{\mathbb{R}^n} |g(x) - g_k(x)| \, dx \\ &= \|f\|_{L^\infty} \|g - g_k\|_{L^1}. \end{aligned}$$

and,

$$|(f \star g_k)(x_n) - (f \star g)(x_n)| \leq \|f\|_{L^\infty} \|g - g_k\|_{L^1}.$$

Therefore, given an $\epsilon > 0$ there exists an $N_1 \in \mathbb{N}$ such that

$$|(f \star g)(x) - (f \star g_k)(x)| + |(f \star g_k)(x_n) - (f \star g)(x_n)| < \frac{2\epsilon}{3}$$

for all $n \geq N_1$. Using step 3, there exists a $\delta > 0$ such that for $|x - x'| < \delta$ we have

$$|(f \star g_k)(x) - (f \star g_k)(x_n)| < \frac{\epsilon}{3}.$$

As $x_n \rightarrow x$ we can choose an $N_2 \in \mathbb{N}$ such that $|x - x_n| < \delta$ for all $n \geq N_2$. Therefore, for all $n \geq \max(N_1, N_2)$ we have

$$|(f \star g)(x) - (f \star g)(x_n)| \leq \frac{2\epsilon}{3} + \frac{\epsilon}{3} = \epsilon,$$

which means that $f \star g$ is continuous. Recall that $f \star g$ is bounded from step 1 and so $f \star g \in \mathcal{C}_b(\mathbb{R}^n)$. \square

Consequently, we obtain a criterion for the strong Feller property.

Corollary 7.2.4. *Suppose $\mathcal{X} = \mathbb{R}^n$. If there exists $g \in L^1(\mathbb{R}^n)$ such that for every $f \in \mathcal{B}_b(\mathcal{X})$ we have $Tf = f \star g$, then T is strong Feller.*

Proof. Follows directly from Lemma 7.2.3. \square

Remark 7.2.5. *From Corollary 7.2.4 it follows that if a process has a law which is absolutely continuous with an integrable density, then the operator T is strong Feller.*

Example 7.2.6.

1. For $\mathcal{X} = \mathbb{R}$ consider the transition probability

$$P(x, \cdot) = \begin{cases} \delta_1 & x > 0 \\ \delta_0 & x \leq 0. \end{cases}$$

Then

$$(T_*f)(x) = \begin{cases} f(1) & x > 0 \\ f(0) & x \leq 0. \end{cases}$$

Hence, P is not Feller.

2. Let X be a homogeneous Markov process given by $X_n = X_{n-1} + Y_n$ where $(Y_n)_{n=1}^\infty$ are independent and identically distributed with law μ .

(a) If μ is the law taking values ± 1 with mean zero, then P is Feller but not strong Feller. To see this note that

$$(T_*f)(x) = \frac{f(x+1) + f(x-1)}{2}.$$

(b) If $\mu = \mathcal{N}(0, 1)$, then P is strong Feller. To see this note that

$$(T_*f)(x) = \frac{1}{\sqrt{2\pi}} \int f(y) e^{-\frac{(x-y)^2}{2}} dy,$$

hence one can conclude by using Corollary 7.2.4.

Definition 7.2.7. Let \mathcal{X} be a separable metric space. Given $\mu \in \mathcal{P}(\mathcal{X})$ the support of μ denoted $\text{supp}(\mu)$, is the intersection of all closed sets $C \subseteq \mathcal{X}$ with $\mu(C) = 1$.

Lemma 7.2.8. Given a separable metric space \mathcal{X} and $\mu \in \mathcal{P}(\mathcal{X})$ one has

$$\text{supp}(\mu) = \{x \in \mathcal{X} : \mu(B(x, \epsilon)) > 0, \text{ for every } \epsilon > 0\}.$$

Proof. Let $x \in E := \{x \in \mathcal{X} : \mu(B(x, \epsilon)) > 0, \text{ for every } \epsilon > 0\}$. Let C be a closed set with $\mu(C) = 1$. Suppose that $x \notin C$, so that there exists an $\epsilon > 0$ such that $B(x, \epsilon) \not\subseteq C$. It follows that

$$\mu(C \cup B(x, \epsilon)) = \mu(C) + \mu(B(x, \epsilon)) > 1,$$

which contradicts μ being a probability measure. Hence, $x \in C$ which implies that $E \subseteq \text{supp}(\mu)$. Now consider $x \in \text{supp}(\mu)$ and suppose that there exists an $\epsilon > 0$ for which $\mu(B(x, \epsilon)) = 0$. For any closed set C with $\mu(C) = 1$ we know that $x \in C$. In particular, $C \cap B(x, \epsilon)^c$ is also closed and such that

$$\mu(C \cap B(x, \epsilon)^c) = \mu(C) + \mu(B(x, \epsilon)^c) - \mu(C \cup B(x, \epsilon)^c) \geq 1 + 1 - 1 = 1.$$

Hence, $C \cap B(x, \epsilon)^c$ is a closed set with $\mu(C \cap B(x, \epsilon)^c) = 1$, but $x \notin C \cap B(x, \epsilon)^c$ which contradicts $x \in \text{supp}(\mu)$. Therefore, $\mu(B(x, \epsilon)) > 0$ for all $\epsilon > 0$ and so $\text{supp}(\mu) \subseteq E$. \square

Proposition 7.2.9. For a separable metric space \mathcal{X} and $\mu \in \mathcal{P}(\mathcal{X})$ it follows that $\mu(\text{supp}(\mu)) = 1$. In other words, $\text{supp}(\mu)$ is the smallest closed set of \mathcal{X} with full μ -measure.

Proof. The set $V := \mathcal{X} \setminus \text{supp}(\mu)$ is separable as \mathcal{X} is separable. Let $Q \subseteq V$ be countably dense in V . By Lemma 7.2.8 we know that for all $q \in Q$ there exists an $\tilde{\epsilon}_q > 0$ such that $\mu(B(q, \tilde{\epsilon}_q)) = 0$. In particular, $\mu(B(q, \epsilon)) = 0$ for all $\epsilon < \tilde{\epsilon}_q$. Hence, as V is open we can choose $\epsilon_q > 0$ such that $\mu(B(q, \epsilon_q)) = 0$ and $B(q, \epsilon_q) \subseteq V$. As Q is dense we have $V \subseteq \bigcup_{q \in Q} B(q, \epsilon_q)$, hence, as Q is countable it follows by countable additivity that

$$\mu(V) \leq \sum_{q \in Q} \mu(B(q, \epsilon_q)) = 0.$$

Hence, $\mu(\text{supp}(\mu)) = 1$. Therefore, by construction, $\text{supp}(\mu)$ is the smallest closed set with full support. \square

Exercise 7.2.10. Recall that $\mu, \nu \in \mathcal{P}(\mathcal{X})$ are mutually singular, denoted $\mu \perp \nu$, if there exists $A \in \mathcal{B}(\mathcal{X})$ such that $\mu(A) = 1$ and $\nu(A) = 0$.

1. Show that $\text{supp}(\mu) \cap \text{supp}(\nu) = \emptyset$ implies that $\mu \perp \nu$.
2. Show that μ and ν being mutually singular does not guarantee that $\text{supp}(\mu) \cap \text{supp}(\nu) = \emptyset$.

Theorem 7.2.11. Let $\mu, \nu \in \mathcal{P}(\mathcal{X})$ be invariant for a transition operator T . Suppose that T has the strong Feller property, then $\mu \perp \nu$ implies that $\text{supp}(\mu) \cap \text{supp}(\nu) = \emptyset$.

Proof. Due to mutual singularity there exists a measurable set $F \in \mathcal{B}(\mathcal{X})$ such that $\mu(F) = 1$ and $\nu(F) = 0$. Let $\psi = T\mathbf{1}_F$, so that $0 \leq \psi \leq 1$ and $\psi \in \mathcal{C}_b(\mathcal{X})$ by the strong Feller property of T . By the invariance of ν it follows that

$$\int_{\mathcal{X}} \psi(y) \nu(dy) = \int_{\mathcal{X}} \mathbf{1}_F(y) \nu(dy) = \nu(F) = 0. \quad (7.2.2)$$

Similarly,

$$\int_{\mathcal{X}} \psi(y) \mu(dy) = \int_{\mathcal{X}} \mathbf{1}_F(y) \mu(dy) = \mu(F) = 1. \quad (7.2.3)$$

Consider the disjoint closed sets $A = \psi^{-1}(\{0\})$ and $B = \psi^{-1}(\{1\})$. As $0 \leq \psi \leq 1$, by (7.2.2) we must have $\nu(A) = 1$ and similarly by (7.2.3) $\mu(B) = 1$. Consequently, $\text{supp}(\nu) \subseteq A$ and $\text{supp}(\mu) \subseteq B$. As A and B are disjoint it follows that $\text{supp}(\nu) \cap \text{supp}(\mu) = \emptyset$. \square

7.3 Existence of Invariant Probability Measures

For continuous state spaces, the Krylov-Bogoliubov theorem, Theorem 7.3.7, is an argument for the existence of invariant measures.

Definition 7.3.1. Let A be a topological space.

1. A subset $K \subseteq A$ is (sequentially) compact if every sequence $(a_n)_{n \in \mathbb{N}} \subseteq K$ has a convergent subsequence in K .
2. A subset $J \subseteq A$ is relatively compact if its closure is compact.

As we are only interested in the existence of invariant measures, it will be sufficient to consider relative compactness.

Definition 7.3.2. A subset $\mathcal{M} \subseteq \mathcal{P}(\mathcal{X})$ is tight if for every $\epsilon > 0$ there exists a compact set $K_\epsilon \subseteq \mathcal{X}$ such that

$$\mu(K_\epsilon) > 1 - \epsilon$$

for all $\mu \in \mathcal{M}$.

Example 7.3.3. Let $\mathcal{M} = (\delta_n)_{n \in \mathbb{N}}$ for δ_n the delta measure at $n \in \mathbb{N}$. Then for every compact set $K \subseteq \mathbb{R}$ there exists an $n \in \mathbb{N}$ such that $n \notin K$. Therefore, $\sup_{n \in \mathbb{N}} (\delta_n(\mathbb{R} \setminus K)) = 1$, meaning \mathcal{M} is not tight.

Lemma 7.3.4. If $\mathcal{M} \subseteq \mathcal{P}(\mathcal{X})$ is finite then it is tight.

Proof. As the finite union of compact sets is compact, it suffices to consider $\mathcal{M} = \{\mu\}$. Since \mathcal{X} is separable it has a countably dense subset, which we enumerate as $(r_k)_{k=1}^\infty \subseteq \mathcal{X}$. Let $B(x, \delta) = \{y \in \mathcal{X} : d(x, y) < \delta\}$, then

$$\bigcup_{k=1}^{\infty} B\left(r_k, \frac{1}{n}\right) = \mathcal{X}$$

for every $n \geq 1$. Which means that

$$\lim_{N \nearrow \infty} \mu\left(\bigcup_{k=1}^N B\left(r_k, \frac{1}{n}\right)\right) = 1.$$

Let $\epsilon > 0$. For each $m \geq 1$ we can find an $N_m \in \mathbb{N}$ such that

$$\mu\left(\bigcup_{k=1}^{N_m} B\left(r_k, \frac{1}{m}\right)\right) > 1 - 2^{-m}\epsilon.$$

If \mathcal{X} is locally compact, then one can take $K = \bar{J}$ where

$$J = \bigcup_{k=1}^{N_1} B(r_k, 1).$$

If \mathcal{X} is not locally compact, then take

$$J = \bigcap_{m=1}^{\infty} \bigcup_{k=1}^{N_m} B\left(r_k, \frac{1}{m}\right).$$

In either case, J is totally bounded and so \bar{J} is compact. In particular,

$$\mu(J^c) < \sum_{m=1}^{\infty} 2^{-m} \epsilon = \epsilon.$$

Therefore, we set $K = \bar{J}$ to deduce that

$$\mu(K) \geq \mu(J) > 1 - \epsilon$$

for all $\mu \in \mathcal{M}$. □

It turns out that we can use tightness to show relative compactness.

Theorem 7.3.5 (Prokhorov). *Let \mathcal{X} be a complete separable metric space. Then $\mathcal{M} \subseteq \mathcal{P}(\mathcal{X})$ is relatively compact if and only if \mathcal{M} is tight.*

In practice, to obtain subsequential limits one often tries to show tightness. Moreover, as $\mathcal{P}(\mathcal{X})$ with weak convergence is metrizable, we can go from convergent subsequences to the convergence of the full sequence if the limits of subsequential limits are unique.

Exercise 7.3.6. *Suppose $(\mu_n)_{n \in \mathbb{N}} \subseteq \mathcal{P}(\mathcal{X})$ is tight, and every subsequence has the same weak limit μ . Show that $\mu_n \rightarrow \mu$ weakly.*

Theorem 7.3.7 (Krylov-Bogoliubov). *Let P be a Feller transition probability, and suppose that there exists an $x_0 \in \mathcal{X}$ such that sequence of measures $(P^n(x_0, \bullet))_{n=1}^{\infty}$ is tight. Then there exists an invariant probability measure for P .*

Proof. Let $\mu_N \in \mathcal{P}(\mathcal{X})$ be given by

$$\mu_N(\bullet) = \frac{1}{N} \sum_{n=1}^N P^n(x_0, \bullet).$$

By the tightness of $(P^n(x_0, \bullet))_{n=1}^{\infty}$, given an $\epsilon > 0$ there exists a compact set $K_\epsilon \subseteq \mathbb{R}$ such that

$$\sup_{n \geq 1} (P^n(x_0, \mathbb{R} \setminus K_\epsilon)) < \epsilon.$$

Consequently,

$$\begin{aligned} \sup_{N \geq 1} (\mu_N(\mathbb{R} \setminus K_\epsilon)) &= \sup_{N \geq 1} \left(\frac{1}{N} \sum_{n=1}^N P^n(x_0, \mathbb{R} \setminus K_\epsilon) \right) \\ &\leq \sup_{N \geq 1} \left(\frac{1}{N} \sum_{n=1}^N \epsilon \right) \\ &= \sup_{N \geq 1} (\epsilon) \\ &= \epsilon. \end{aligned}$$

Hence, the set of measures $(\mu_N)_{N \in \mathbb{N}}$ is tight. Therefore, there exists a subsequence $(\mu_{N_k})_{k \in \mathbb{N}}$ and $\mu \in \mathcal{P}(\mathcal{X})$ such that $\mu_{N_k} \rightarrow \mu$ weakly. Since P is Feller, $T_*\phi \in \mathcal{C}_b(\mathcal{X})$ for every $\phi \in \mathcal{C}_b(\mathcal{X})$, hence,

$$\begin{aligned}
\int_{\mathcal{X}} T_*\phi \, d\mu &= \lim_{k \rightarrow \infty} \int_{\mathcal{X}} T_*\phi \, d\mu_{N_k} \\
&= \lim_{k \rightarrow \infty} \frac{1}{N_k} \sum_{n=1}^{N_k} \int_{\mathcal{X}} \phi(y) P^{n+1}(x_0, dy) \\
&= \lim_{k \rightarrow \infty} \left(\frac{1}{N_k} \sum_{n=1}^{N_k} \int_{\mathcal{X}} \phi(y) P^n(x_0, dy) + \frac{1}{N_k} \int_{\mathcal{X}} \phi(y) P^{N_k+1}(x_0, dy) - \frac{1}{N_k} \int_{\mathcal{X}} \phi(y) P(x_0, dy) \right) \\
&= \lim_{k \rightarrow \infty} \left(\int_{\mathcal{X}} \phi \, d\mu_{N_k} + \frac{1}{N_k} \int_{\mathcal{X}} \phi(y) (P^{N_k+1}(x_0, dy) - P(x_0, dy)) \right) \\
&= \lim_{k \rightarrow \infty} \int_{\mathcal{X}} \phi \, d\mu_{N_k} \\
&\stackrel{(1)}{=} \int_{\mathcal{X}} \phi \, d\mu
\end{aligned}$$

where (1) follows by the weak convergence of the $(\mu_{N_k})_{k \in \mathbb{N}}$. Therefore,

$$\begin{aligned}
\int_{\mathcal{X}} \phi \, d(T^*\mu) &\stackrel{\text{Ex 4.2.14}}{=} \int_{\mathcal{X}} T_*\phi \, d\mu \\
&= \int_{\mathcal{X}} \phi \, d\mu
\end{aligned}$$

for all $\phi \in \mathcal{C}_b(\mathcal{X})$. Hence, μ is an invariant measure for P . \square

Corollary 7.3.8. *Let \mathcal{X} be a compact state space. Then a Feller transition function P has an invariant probability measure.*

Proof. Note that $\mathcal{P}(\mathcal{X})$ is compact as \mathcal{X} is compact. Hence, the tightness of measures follows. Therefore, we can apply Theorem 7.3.7 to conclude. \square

Example 7.3.9.

1. When \mathcal{X} is discrete we note that $\mathcal{B}_b(\mathcal{X}) = \mathcal{C}_b(\mathcal{X})$, which means that every transition probability is strong Feller.
2. When \mathcal{X} is finite it follows from statement 1 and Corollary 7.3.8 that every transition function P has an invariant probability measure.

Exercise 7.3.10. *Suppose X is a Markov process on \mathbb{R}^n with Feller transition function P . Moreover, suppose that there exists $G : \mathbb{R}_+ \rightarrow (0, \infty)$ an increasing function with $\lim_{r \rightarrow \infty} G(r) = \infty$ with*

$$\sup_{n \in \mathbb{N}} (\mathbb{E}_x (G(|X_n|))) < \infty$$

for some $x \in \mathcal{X}$. Then there exists an invariant probability measure for P .

Example 7.3.11. *Consider a process $(X_n)_{n=0}^{\infty}$, where $\mathbb{E}(|X_0|) < \infty$, $X_n = aX_{n-1} + \xi_n$ for $n \geq 1$ where $a \in (0, 1)$ and $(\xi_n)_{n \in \mathbb{N}}$ is a sequence of independent and identically distributed random variables with $\mathbb{E}(|\xi_n|) =$*

$b < \infty$. Then for $n \geq 1$ we have

$$\begin{aligned}
 \mathbb{E}(|X_n|) &= a\mathbb{E}(|X_{n-1}|) + \mathbb{E}(|\xi_n|) \\
 &\leq a\mathbb{E}(|X_{n-1}|) + b \\
 &\leq a(a\mathbb{E}(|X_{n-2}|) + b) + b \\
 &\quad \vdots \\
 &\leq b \sum_{k=0}^{n-1} a^k + \mathbb{E}(|X_0|) \\
 &\leq \frac{b}{1-a} + \mathbb{E}(|X_0|).
 \end{aligned}$$

Therefore, by Exercise 7.3.10 with $G(x) = x$, it follows that the process $(X_n)_{n=0}^{\infty}$ has an invariant probability measure.

We can systematize the ideas of Exercise 7.3.10 using Lyapunov functions.

Lemma 7.3.12. Let P be a transition function on \mathcal{X} and let $V : \mathcal{X} \rightarrow [0, \infty]$ be a measurable function. Suppose that there exists a $\gamma \in (0, 1)$ and $C > 0$ such that

$$(TV)(x) \leq \gamma V(x) + C. \quad (7.3.1)$$

Then

$$T^n V(x) \leq \gamma^n V(x) + \frac{C}{1-\gamma}.$$

Proof. Iterating (7.3.1) gives

$$\begin{aligned}
 T^n V(x) &= (T \circ T^{n-1} V)(x) \\
 &\leq \gamma T^{n-1} V(x) + C \\
 &\leq \gamma^2 T^{n-2} V(x) + \gamma C + C \\
 &\quad \vdots \\
 &\leq \gamma^n V(x) + C \sum_{j=0}^{n-1} \gamma^j \\
 &\leq \gamma^n V(x) + C \sum_{j=0}^{\infty} \gamma^j \\
 &= \gamma^n V(x) + \frac{C}{1-\gamma}.
 \end{aligned}$$

□

Definition 7.3.13. Let \mathcal{X} be a complete separable metric space and let P be a transition probability on \mathcal{X} . A Borel measurable function $V : \mathcal{X} \rightarrow [0, \infty]$ is called a Lyapunov function for P if it satisfies the following.

1. $V^{-1}([0, \infty)) \neq \emptyset$.
2. For every $a \in [0, \infty)$ the set $K_a = \{y : V(y) \leq a\}$ is compact.
3. There exists a $\gamma \in (0, 1)$ and $C > 0$ such that

$$(TV)(x) \leq \gamma V(x) + C.$$

Theorem 7.3.14. *If a transition function P is Feller and has a Lyapunov function, then it has an invariant measure.*

Proof. Fix $x_0 \in \mathcal{X}$ such that $V(x_0) \neq \infty$. Then for $a > 0$ we have

$$\begin{aligned} \sup_{n \in \mathbb{N}} (P^n(x_0, \mathbb{R} \setminus K_a)) &\leq \sup_{n \in \mathbb{N}} \left(\int_{\mathcal{X}} \frac{V(y)}{a} P^n(x_0, dy) \right) \\ &= \sup_{n \in \mathbb{N}} (T^n V(x_0)) \\ &\stackrel{\text{Lem 7.3.12.}}{\leq} \frac{1}{a} \left(V(x_0) + \frac{C}{1-\gamma} \right) \\ &\xrightarrow{a \rightarrow \infty} 0. \end{aligned}$$

Hence, the family of measures $(P^n(x_0, \bullet))_{n \in \mathbb{N}}$ is tight and so we can conclude by applying Theorem 7.3.7. \square

Proposition 7.3.15. *Let P be a transition probability on \mathcal{X} , and let $V : \mathcal{X} \rightarrow [0, \infty)$ be a measurable function which satisfies*

$$(TV)(x) \leq \gamma V(x) + C$$

for $\gamma \in (0, 1)$ and $C > 0$. Then for every P -invariant probability measure π , we have

$$\int_{\mathcal{X}} V(x) \pi(dx) \leq \frac{C}{1-\gamma}.$$

Proof. Let $M \geq 0$ and let $V_M(x) = V(x) \wedge M$. Note that $r \mapsto r \wedge M$ is concave, so we can apply Jensen's inequality to get

$$\begin{aligned} \int_{\mathcal{X}} V_M(y) P(x, dy) &\leq \left(\int_{\mathcal{X}} V(y) P(x, dy) \right) \wedge M \\ &= (\gamma V(x) + C) \wedge M. \end{aligned}$$

Iterating this bound it follows that

$$\begin{aligned} \int_{\mathcal{X}} V_M(y) \pi(dx) &\stackrel{(1)}{=} \int_{\mathcal{X}} V_M(y) T^n \pi(dx) \\ &= \int_{\mathcal{X}} (T^n V_M)(y) \pi(dx) \\ &\leq \int_{\mathcal{X}} \left(\gamma^n V(x) + \frac{C}{1-\gamma} \right) \wedge M \pi(dx) \end{aligned}$$

for $n \geq 1$. In (1) we use the fact that π is P -invariant. By taking $n \rightarrow \infty$ we get

$$\int_{\mathcal{X}} V_M(y) \pi(dx) \leq \frac{C}{1-\gamma} \wedge M \leq \frac{C}{1-\gamma}.$$

Then taking $M \rightarrow \infty$ we can apply the monotone convergence theorem to deduce that

$$\int_{\mathcal{X}} V(y) \pi(dx) \leq \frac{C}{1-\gamma}.$$

\square

Remark 7.3.16. *Theorem 7.3.14 gives us sufficient conditions a test function with respect to a transition probability P ought to have to ensure the existence of a P -invariant probability measure. Proposition 7.3.15 considers weaker test functions, and hence, cannot guarantee the existence of P -invariant probability measures. We note that the test functions in Proposition 7.3.15 are not necessarily Lyapunov functions as their image is not the extended non-negative real line, and there are no conditions on the compactness of sub-level sets. However,*

if we assume that such P -invariant probability measures exist then the weaker test functions considered in Proposition 7.3.15 can still provide information on the behaviour of this P -invariant probability measure.

7.4 Random Dynamical Systems

Recall that a random dynamical system is a Markov process $X = (X_n)_{n=0}^\infty$ where $X_0 = \xi_0$ and $X_{n+1} = F(X_n, \xi_n)$ for $(\xi_j)_{j=0}^\infty$ a sequence of independent and identically distributed random variables, with $\nu = \text{Law}(\xi_i)$, on a measurable space \mathcal{Y} , and $F : \mathcal{X} \times \mathcal{Y} \rightarrow \mathcal{X}$ a measurable function.

Remark 7.4.1. *It turns out that any discrete-time Markov process can be formulated as a random dynamical system. However, in practice, the functions and random variables involved may be complicated.*

7.4.1 Existence of Invariant Measures

To determine the existence of an invariant measure of a random dynamical system we need to investigate the Feller property and the tightness of the n -step transition probabilities. For the tightness, we will utilise Lyapunov functions.

Exercise 7.4.2. *Suppose that $X = (X_n)_{n=0}^\infty$ is a random dynamical system. Show that*

$$(Tf)(x) = \int_{\mathcal{Y}} f(F(x, y)) \nu(dy)$$

for any $f \in \mathcal{B}_b(\mathcal{X})$.

Theorem 7.4.3. *Suppose that $X = (X_n)_{n=0}^\infty$ is a random dynamical system and that there is a measurable set $A \subseteq \mathcal{Y}$ with $\nu(A) = 1$ and $x \mapsto F(x, y)$ continuous for every $y \in A$. Then T is Feller.*

Proof. Let $\phi \in \mathcal{C}_b(\mathcal{X})$ and suppose that $(x_n)_{n \in \mathbb{N}} \subseteq \mathcal{X}$ is convergent to $x \in \mathcal{X}$. Then the continuity of ϕ , and the almost everywhere continuity of $F(\cdot, y)$, we have that

$$\phi(F(x_n, y)) \rightarrow \phi(F(x, y))$$

for $y \in \mathcal{Y}$, ν almost everywhere. So by Exercise 7.4.2 it follows that

$$\begin{aligned} \lim_{n \rightarrow \infty} (T\phi)(x_n) &= \lim_{n \rightarrow \infty} \int \phi(F(x_n, y)) \nu(dy) \\ &\stackrel{(1)}{=} \int \lim_{n \rightarrow \infty} \phi(F(x_n, y)) \nu(dy) \\ &= \int \phi(F(x, y)) \nu(dy) \\ &= (T\phi)(x), \end{aligned}$$

where (1) follows as ϕ is bounded. This implies that $T\phi \in \mathcal{C}_b(\mathcal{X})$ and T is Feller. \square

Theorem 7.4.4. *Suppose that $X = (X_n)_{n=0}^\infty$ is a random dynamical system. Suppose that there is a measurable set $A \subseteq \mathcal{Y}$ with $\nu(A) = 1$ such that $x \mapsto F(x, y)$ continuous for every $y \in A$. Furthermore, suppose that there is some Borel measurable function $V : \mathcal{X} \rightarrow [0, \infty]$, with compact sub-level sets, V is finite at some point, and*

$$\int_{\mathcal{Y}} V(F(x, y)) \nu(dy) \leq \gamma V(x) + C$$

for all $x \in \mathcal{X}$, for some $\gamma \in (0, 1)$ and $C > 0$. Then X has at least one invariant probability measure.

Proof. From Theorem 7.4.3 we have that T is Feller, and we note that V is a Lyapunov function. Therefore, we can conclude by applying Theorem 7.3.14. \square

We have here established conditions for the existence of invariant measures for random dynamical systems. We will now investigate the uniqueness of these invariant measures. For this, we will use the contraction properties of the function F .

7.4.2 Uniqueness of Invariant Probability Measures via Contraction

Definition 7.4.5. For $i = 1, 2$ let $p_i : \mathcal{X}^2 \rightarrow \mathcal{X}$ be the projection maps given by $(x_1, x_2) \mapsto x_i$. For $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$, a measure $\mu \in \mathcal{P}(\mathcal{X}^2)$ is a coupling of π_1 and π_2 if

$$(p_i)_* \mu = \pi_i$$

for $i = 1, 2$. That is, if $Z = (X, Y) \sim \mu$, then $X \sim \pi_1$ and $Y \sim \pi_2$.

Lemma 7.4.6. Let $\Delta = \{(x, x) \in \mathcal{X} \times \mathcal{X} : x \in \mathcal{X}\}$. If there exists a coupling $\mu \in \mathcal{P}(\mathcal{X}^2)$ of $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$ with $\mu(\Delta) = 1$, then $\pi_1 = \pi_2$. In particular, if

$$\int_{\mathcal{X} \times \mathcal{X}} 1 \wedge d(x, y) \mu(dx, dy) = 0 \quad (7.4.1)$$

where d is the metric on \mathcal{X} , then $\pi_1 = \pi_2$.

Proof. Let $A \in \mathcal{B}(\mathcal{X})$, then

$$\begin{aligned} \pi_1(A) &= \mu(A \times \mathcal{X}) \\ &\stackrel{(\star)}{=} \mu((A \times \mathcal{X}) \cap \Delta) \\ &\stackrel{(\star\star)}{=} \mu((\mathcal{X} \times A) \cap \Delta) \\ &\stackrel{(\star)}{=} \mu(\mathcal{X} \times A) \\ &= \pi_2(A), \end{aligned}$$

where (\star) follows from the fact that $\mu(\Delta) = 1$, and $(\star\star)$ follows from how Δ is constructed. In particular, note that $1 \wedge d(x, y) \geq 0$ and

$$\{(x, y) \in \mathcal{X} \times \mathcal{X} : 1 \wedge d(x, y) = 0\} = \Delta.$$

Hence, by (7.4.1) it follows that $\mu(\Delta) = 1$. Therefore, by the above arguments, $\pi_1 = \pi_2$. \square

Lemma 7.4.7. Let $(\mu_n)_{n \in \mathbb{N}}$ be a family of couplings of $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$, then $(\mu_n)_{n \in \mathbb{N}}$ is tight.

Proof. Given an $\epsilon > 0$, there exists compact sets $K_1, K_2 \subseteq \mathcal{X}$ such that

$$\pi_i(K_i) > 1 - \frac{\epsilon}{2}$$

for $i = 1, 2$. Hence,

$$\begin{aligned} \mu_n(\mathcal{X}^2 \setminus K_1 \times K_2) &\leq \mu_n((\mathcal{X} \setminus K_1) \times \mathcal{X}) + \mu_n(\mathcal{X} \times (\mathcal{X} \setminus K_2)) \\ &= \pi_1(K_1^c) + \pi_2(K_2^c) \\ &< \epsilon, \end{aligned}$$

which shows that $(\mu_n)_{n \in \mathbb{N}}$ is tight. \square

Lemma 7.4.8. Let $(\nu_n)_{n \in \mathbb{N}}$ be a sequence of measures on \mathcal{X} which converges weakly to a measure ν . Then for any continuous map $\phi : \mathcal{X} \rightarrow \mathcal{Y}$, it follows that $\phi_*\nu_n$ converges weakly to $\phi_*\nu$.

Proof. Let $f \in \mathcal{C}_b(\mathcal{Y})$. Then

$$\begin{aligned} \int_{\mathcal{Y}} f \, d(\phi_*\nu_n) &= \int_{\mathcal{X}} (f \circ \phi) \, d\nu_n \\ &\xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}} (f \circ \phi) \, d\nu \\ &= \int_{\mathcal{Y}} f \, d(\phi_*\nu) \end{aligned}$$

where the convergence follows as $f \circ \phi \in \mathcal{C}_b(\mathcal{X})$ and ν_n converges weakly to ν . \square

Lemma 7.4.9. Let $(\mu_n)_{n \in \mathbb{N}}$ be a sequence of couplings of $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$ converging weakly to μ . Then μ is a coupling of π_1 and π_2 .

Proof. The projection maps $p_i : \mathcal{X}^2 \rightarrow \mathcal{X}$ for $i = 1, 2$ are continuous functions. So by Lemma 7.4.8 it follows that $(p_i)_*\mu_n$ converges weakly to $(p_i)_*\mu$. By construction $(p_i)_*\mu_n = \pi_i$, so

$$\int_{\mathcal{X}} f \, d\pi_i = \int_{\mathcal{X}} f \, d((p_i)_*\mu)$$

for all $f \in \mathcal{C}_b(\mathcal{X})$. Hence, by Lemma 7.1.2 it follows that $(p_i)_*\mu = \pi_i$, which implies that μ is a coupling. \square

Remark 7.4.10. Lemma 7.4.7 along with Theorem 7.3.5 shows that the set of all couplings is relatively compact. Lemma 7.4.9 shows that the set of couplings is closed, and hence the set of all couplings for $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$ is compact.

Definition 7.4.11. Let X_0 and X'_0 be random dynamical systems driven by the independent and identically distributed random variables $(\xi_n)_{n=0}^\infty$. Then the synchronized coupling $Z = (X, X')$ is the Markov process on $\mathcal{X} \times \mathcal{X}$ where

$$Z_{n+1} = (X_{n+1}, X'_{n+1}) = (F(X_n, \xi_n), F(X'_n, \xi_n))$$

and $\mu_n = \text{Law}(Z_n)$.

Lemma 7.4.12. In the setting of Definition 7.4.11 suppose that for some $\gamma \in (0, 1)$ we have

$$\int_{\mathcal{X}} d(F(x, y), F(x', y)) \, \nu(dy) \leq \gamma d(x, x')$$

where d is the metric on \mathcal{X} . Then for the synchronized coupling we have

$$\lim_{n \rightarrow \infty} \mathbb{E}(1 \wedge d(X_n, X'_n)) = \lim_{n \rightarrow \infty} \int_{\mathcal{X}} 1 \wedge d(x, x') \, \mu_n(dx, dx') = 0.$$

Proof. Note that we can write

$$\mathbb{E}(1 \wedge d(X_n, X'_n)) = \mathbb{E}(\mathbb{E}(1 \wedge d(X_n, X'_n) | X_{n-1}, X'_{n-1})).$$

Observe that

$$\begin{aligned}
\mathbb{E} (1 \wedge d(X_n, X'_n) | X_{n-1}, X'_{n-1}) &\leq 1 \wedge \mathbb{E} (d(X_n, X'_n) | X_{n-1}, X'_{n-1}) \\
&= 1 \wedge \mathbb{E} (d(F(X_{n-1}, \xi_{n-1}), F(X'_{n-1}, \xi_{n-1})) | X_{n-1}, X'_{n-1}) \\
&= 1 \wedge \int_{\mathcal{Y}} d(F(X_{n-1}, \xi_{n-1}), F(X'_{n-1}, \xi_{n-1})) \nu(dy) \\
&\leq 1 \wedge \gamma d(X_{n-1}, X'_{n-1}).
\end{aligned}$$

Hence,

$$\mathbb{E} (1 \wedge d(X_n, X'_n)) \leq \mathbb{E} (1 \wedge \gamma d(X_n, X'_n)).$$

Iterating this argument we get

$$\begin{aligned}
\mathbb{E} (1 \wedge \gamma d(X_{n-1}, X'_{n-1})) &= \mathbb{E} (\mathbb{E} (1 \wedge \gamma d(X_{n-1}, X'_{n-1}) | X_{n-2}, X'_{n-2})) \\
&\leq \mathbb{E} (1 \wedge \gamma^2 d(X_{n-2}, X'_{n-2})) \\
&\vdots \\
&\leq \mathbb{E} (1 \wedge \gamma^n d(X_0, X'_0)).
\end{aligned}$$

Therefore,

$$\mathbb{E} (1 \wedge \gamma d(X_{n-1}, X'_{n-1})) \leq \mathbb{E} (1 \wedge \gamma^n d(X_0, X'_0)) \xrightarrow{n \rightarrow \infty} 0.$$

□

Theorem 7.4.13. *In the setting of Definition 7.4.11 suppose that for some $\gamma \in (0, 1)$ we have*

$$\int_{\mathcal{X}} d(F(x, y), F(x', y)) \nu(dy) \leq \gamma d(x, x')$$

where d is the metric on \mathcal{X} . Then the random dynamical system has at most one invariant probability measure.

Proof. Let π_1 and π_2 be invariant distributions for a random dynamical system. Let $\text{Law}(X_0, X'_0) = \pi_1 \otimes \pi_2$, such that $\text{Law}(X_n) = \pi_1$ and $\text{Law}(X'_n) = \pi_2$ by their invariant property. Then $\mu_n = \text{Law}(X_n, X'_n)$ is a coupling of π_1 and π_2 . Moreover, it is a synchronized coupling as X_0 and X'_0 are independent. As we know the set of couplings is compact there exists a subsequence $(\mu_{n_k})_{k \in \mathbb{N}}$ weakly converging to a coupling $\mu \in \mathcal{P}(\mathcal{X}^2)$. As each μ_n is a synchronized coupling, using Lemma 7.4.12 we know that

$$\lim_{n \rightarrow \infty} \int_{\mathcal{X}} 1 \wedge d(x, x') \mu_n(dx, dx') = 0.$$

In particular, by the boundedness and continuity of $1 \wedge d(\cdot, \cdot)$ it follows that

$$\int_{\mathcal{X}} 1 \wedge d(x, x') \mu(dx, dx') = 0.$$

Hence, we deduce that $\pi_1 = \pi_2$ using Lemma 7.4.6.

□

Example 7.4.14. *Recall the random dynamical system from Example 7.3.11. Observe that*

$$\mathbb{E} (|F(x, y) - F(x', y)|) = \mathbb{E} (a|x - x'|) = a|x - x'|,$$

so that by Theorem 7.4.13 this process has at most one invariant probability measure. In Example 7.3.11 it was shown that an invariant measure does exist, hence this random dynamical system has exactly one invariant probability measure.

7.5 Uniqueness of Invariant Probability Measures via Minorisation

Uniqueness via minorization is a probabilistic argument for the uniqueness of invariant probability measures. It is useful to also understand uniqueness arguments from this perspective as the random dynamical system formulation of a process may involve complicated functions and random variables.

Definition 7.5.1. Let μ, ν be positive measures on a measurable space Ω . Suppose that $\mu, \nu \ll \eta$, then

$$\|\mu - \nu\|_{\text{TV}} := \int_{\Omega} \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta.$$

Remark 7.5.2.

1. The existence of η in Definition 7.5.1 is not restrictive as $\eta = \mu + \nu$ is such that $\mu, \nu \ll \eta$. Furthermore, Definition 7.5.1 is independent of the choice η . One can see this, as for η such that $\mu, \nu \ll \eta$ then $(\mu + \nu) \ll \eta$, so

$$\begin{aligned} \int_{\Omega} \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta &= \int_{\Omega} \left| \frac{d\mu}{d(\mu + \nu)} - \frac{d\nu}{d(\mu + \nu)} \right| \frac{d(\mu + \nu)}{d\eta} d\eta \\ &= \int_{\Omega} \left| \frac{d\mu}{d(\mu + \nu)} - \frac{d\nu}{d(\mu + \nu)} \right| d(\mu + \nu). \end{aligned}$$

2. Note how Definition 7.5.1 contains the definition of the total variation for the discrete case, given in Lemma 6.6.3, as one can just take η to be the measure that is 1 at each singleton point of the discrete space Ω .

Exercise 7.5.3. Show that if μ and ν are probability measures then Definition 7.5.1 gives

$$\|\mu - \nu\|_{\text{TV}} = 2 \sup (\{|\mu(A) - \nu(A)| : A \subseteq \Omega, \text{ measurable}\}).$$

Lemma 7.5.4. Let $a, b \geq 0$ then $|a - b| = a + b - 2a \wedge b$.

With μ, ν and η as in Definition 7.5.1 and using Lemma 7.5.4 it follows that

$$\left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| = \frac{d\mu}{d\eta} + \frac{d\nu}{d\eta} - 2 \left(\frac{d\mu}{d\eta} \wedge \frac{d\nu}{d\eta} \right). \quad (7.5.1)$$

Definition 7.5.5. For positive measures μ and ν let $\mu \wedge \nu$ be the positive measure given by

$$(\mu \wedge \nu)(A) = \int_A \left(\frac{d\mu}{d(\mu + \nu)} \wedge \frac{d\nu}{d(\mu + \nu)} \right) d(\mu + \nu).$$

The measure $\mu \wedge \nu$ constructed in Definition 7.5.5 is the minimum of μ and ν and so can be thought of as their intersection.

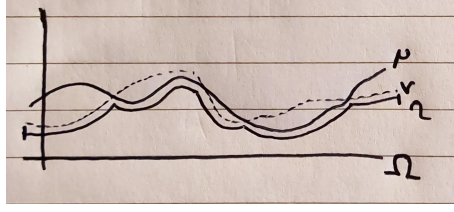


Figure 11: An illustration of the measure $\eta = \mu \wedge \nu$ as constructed in Definition 7.5.5 from measures μ and ν

Lemma 7.5.6. For $\mu, \nu \in \mathcal{P}(\Omega)$ we have

$$\|\mu - \nu\|_{\text{TV}} = 2(1 - (\mu \wedge \nu)(\Omega)).$$

Proof. Using (7.5.1) it follows that

$$\begin{aligned} \|\mu - \nu\|_{\text{TV}} &= \int_{\Omega} \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta \\ &= \int_{\Omega} \frac{d\mu}{d\nu} + \frac{d\nu}{d\eta} - 2 \left(\frac{d\mu}{d\eta} \wedge \frac{d\nu}{d\eta} \right) d\eta \\ &= 1 + 1 - 2(\mu \wedge \nu)(\Omega) \\ &= 2(1 - (\mu \wedge \nu)(\Omega)). \end{aligned}$$

□

Lemma 7.5.7. Let $\mu, \nu \in \mathcal{P}(\Omega)$ be distinct and let

$$\bar{\mu} = \frac{\mu - (\mu \wedge \nu)}{\frac{1}{2}\|\mu - \nu\|_{\text{TV}}}$$

and

$$\bar{\nu} = \frac{\nu - (\mu \wedge \nu)}{\frac{1}{2}\|\mu - \nu\|_{\text{TV}}}.$$

Then $\bar{\mu}, \bar{\nu} \in \mathcal{P}(\Omega)$ with

$$\mu - \nu = \frac{1}{2}\|\mu - \nu\|_{\text{TV}}(\bar{\mu} - \bar{\nu}). \quad (7.5.2)$$

Proof. The equality (7.5.2) is immediate. Observe that

$$(\mu - (\mu \wedge \nu))(A) = \int_A \frac{d\mu}{d(\mu + \nu)} - \left(\frac{d\mu}{d(\mu + \nu)} \wedge \frac{d\nu}{d(\mu + \nu)} \right) d(\mu + \nu) \geq 0$$

and so $\bar{\mu}$ and $\bar{\nu}$ are positive measures. Moreover, by Lemma 7.5.6 we have

$$\begin{aligned} \frac{1}{2}\|\mu - \nu\|_{\text{TV}} &= 1 - (\mu \wedge \nu)(\Omega) \\ &= \mu(\Omega) - (\mu \wedge \nu)(\Omega) \\ &= \nu(\Omega) - (\mu \wedge \nu)(\Omega) \end{aligned}$$

and so $\bar{\mu}(\Omega) = \bar{\nu}(\Omega) = 1$ which implies that $\bar{\mu}, \bar{\nu} \in \mathcal{P}(\Omega)$. □

Lemma 7.5.8. Let T be a linear operator on measures defined on Ω such that $T(\mathcal{P}(\Omega)) \subseteq \mathcal{P}(\Omega)$. Then for

any $\mu, \nu \in \mathcal{P}(\Omega)$ we have

$$\|T\mu - T\nu\|_{\text{TV}} = \frac{1}{2} \|\mu - \nu\|_{\text{TV}} \|T\bar{\mu} - T\bar{\nu}\|_{\text{TV}} \leq \|\mu - \nu\|_{\text{TV}}. \quad (7.5.3)$$

Proof. Note by the linearity of T that $\overline{T\mu} = T\bar{\mu}$. Thus, applying T to both sides of (7.5.2) it follows that

$$\|T\mu - T\nu\|_{\text{TV}} = \frac{1}{2} \|\mu - \nu\|_{\text{TV}} \|T\bar{\mu} - T\bar{\nu}\|_{\text{TV}}.$$

As $T(\mathcal{P}(\Omega)) \subseteq \mathcal{P}(\Omega)$, we can apply the triangle inequality to get that $\|T\bar{\mu} - T\bar{\nu}\|_{\text{TV}} \leq 2$. Therefore,

$$\|T\mu - T\nu\|_{\text{TV}} = \frac{1}{2} \|\mu - \nu\|_{\text{TV}} \|T\bar{\mu} - T\bar{\nu}\|_{\text{TV}} \leq \|\mu - \nu\|_{\text{TV}}.$$

□

Remark 7.5.9. The transition operator, T , of a time homogeneous Markov chain, namely

$$(T\mu)(A) = \int_{\mathcal{X}} P(x, A) \mu(dx)$$

satisfies the requirements of Lemma 7.5.8.

Definition 7.5.10. A transition probability P is minorized by $\eta \in \mathcal{P}(\mathcal{X})$ with constant $\alpha > 0$ if

$$P(x, A) \geq \alpha \eta(A)$$

for all $x \in \mathcal{X}$ and $A \in \mathcal{B}(\mathcal{X})$.

Remark 7.5.11. A transition probability being minorized can be thought of as the continuous analogue of a stochastic matrix being irreducible. Indeed the property of minorization plays an equivalent role in showing the uniqueness of invariant probability measures for continuous state space Markov chains.

Exercise 7.5.12. Let $(\mu_n)_{n \in \mathbb{N}} \subseteq \mathcal{P}(\Omega)$. Show that $\eta = \sum_{n=1}^{\infty} \frac{1}{2^n} \mu_n \in \mathcal{P}(\mathcal{X})$.

Lemma 7.5.13. The metric space $(\mathcal{P}(\Omega), \|\cdot\|_{\text{TV}})$ is a complete.

Proof. Let $(\mu_n)_{n \in \mathbb{N}} \subseteq \mathcal{P}(\Omega)$ be a Cauchy sequence in $\|\cdot\|_{\text{TV}}$. Using Exercise 7.5.12 we know that $\eta = \sum_{n=1}^{\infty} 2^{-n} \mu_n \in \mathcal{P}(\mathcal{X})$. As $\mu_n \ll \eta$ for all $n \in \mathbb{N}$, it follows that

$$\|\mu_m - \mu_n\|_{\text{TV}} = \left\| \frac{d\mu_m}{d\eta} - \frac{d\mu_n}{d\eta} \right\|_{L^1(\eta)}.$$

Therefore, $\left(\frac{d\mu_n}{d\eta} \right)_{n \in \mathbb{N}}$ is $L^1(\eta)$ -Cauchy. Since $L^1(\eta)$ is complete, we know that $\left(\frac{d\mu_n}{d\eta} \right)_{n \in \mathbb{N}}$ has an $L^1(\eta)$ -limit f with $f \geq 0$ and $\|f\|_{L^1(\eta)} = 1$. Setting $\mu \in \mathcal{P}(\mathcal{X})$ to be $\frac{d\mu}{d\eta} = f$ it follows that $\mu_n \rightarrow \mu$ in $\|\cdot\|_{\text{TV}}$. □

Theorem 7.5.14. Let P be a transition probability on \mathcal{X} and suppose P is minorized by $\eta \in \mathcal{P}(\mathcal{X})$ with constant $\alpha \in (0, 1)$, then the following hold.

1. P has a unique invariant probability measure π .

2. For any $\mu, \nu \in \mathcal{P}(\mathcal{X})$ it follows that

$$\|T^n \mu - T^n \nu\|_{\text{TV}} \leq (1 - \alpha)^n \|\mu - \nu\|_{\text{TV}}.$$

Proof. Recall that $(\mathcal{P}(\Omega), \|\cdot\|_{\text{TV}})$ is a complete metric space, from Lemma 7.5.13. Hence, as statement 2 shows that T is a strict contraction we can show statement 1 by using Banach's fixed point theorem. To show statement 2 we note that for any measure $\lambda \in \mathcal{P}(\mathcal{X})$ and $A \in \mathcal{B}(\mathcal{X})$ we have

$$\begin{aligned} (T\lambda)(A) &= \int_{\mathcal{X}} P(x, A) \lambda(dx) \\ &\geq \int_{\mathcal{X}} \alpha \eta(A) \lambda(dx) \\ &= \alpha \eta(A). \end{aligned}$$

Hence, for any $\lambda \in \mathcal{P}(\mathcal{X})$ it follows that

$$\frac{1}{1 - \alpha} (T\lambda - \alpha \eta) \in \mathcal{P}(\mathcal{X}).$$

Thus, for any $\mu, \nu \in \mathcal{P}(\mathcal{X})$ we have

$$\begin{aligned} \|T\mu - T\nu\|_{\text{TV}} &= (1 - \alpha) \left\| \frac{T\mu - \alpha \eta}{1 - \alpha} - \frac{T\nu - \alpha \eta}{1 - \alpha} \right\|_{\text{TV}} \\ &\leq (1 - \alpha)(2). \end{aligned}$$

Using (7.5.3) it follows that

$$\begin{aligned} \|T\mu - T\nu\|_{\text{TV}} &= \frac{1}{2} \|\mu - \nu\|_{\text{TV}} \|T\bar{\mu} - T\bar{\nu}\|_{\text{TV}} \\ &\leq \frac{1}{2} \|\mu - \nu\|_{\text{TV}} (2(1 - \alpha)) \\ &= (1 - \alpha) \|\mu - \nu\|_{\text{TV}}. \end{aligned}$$

Iterating this we arrive at statement 2. □

Corollary 7.5.15. Under the assumptions of Theorem 7.5.14, let $\pi, \mu \in \mathcal{P}(\mathcal{X})$ with $T\pi = \pi$. Then

$$\|T^n \mu - \pi\|_{\text{TV}} \leq (1 - \alpha)^n \|\mu - \pi\|_{\text{TV}}.$$

7.6 P-Invariant Sets

Now we introduce a continuous analogue of closed communication classes.

Definition 7.6.1. Let P be a transition function on \mathcal{X} , then $A \in \mathcal{B}(\mathcal{X})$ is P -invariant if $P(x, A) = 1$ for every $x \in A$.

Exercise 7.6.2. If $A \in \mathcal{B}(\mathcal{X})$ is P -invariant, and X is a Markov process with transition function P and initial distribution $\pi \in \mathcal{P}(\mathcal{X})$, then

$$\mathbb{P}(X_0 \in A, \dots, X_n \in A) = \pi(A).$$

If $A \subseteq \mathcal{X}$ is closed, and \mathcal{X} is a complete separable metric space, then A is a complete separable metric space. Thus, we can think of A as being a possible state space. Note that we can extend any $\tilde{\pi} \in \mathcal{P}(A)$ to an element $\pi \in \mathcal{P}(\mathcal{X})$ by letting

$$\pi(F) = \tilde{\pi}(F \cap A)$$

for all $F \in \mathcal{B}(\mathcal{X})$. Similarly, given $\pi \in \mathcal{P}(\mathcal{X})$ we can restrict π to a measure, $\tilde{\pi}$, on A . Namely, $\tilde{\pi} := \pi|_{\mathcal{B}(A)}$. However, it may not be the case that $\tilde{\pi}$ is a probability measure. If A is P -invariant then we can restrict a transition function P on \mathcal{X} to a transition function \tilde{P} on A . For any $F \in \mathcal{B}(\mathcal{X})$ one can let

$$\tilde{P}(x, B) = P(x, B \cap A) = P(x, B)$$

for any $x \in A$. For \tilde{P} we have the corresponding operator \tilde{T} .

Lemma 7.6.3. *Let P be a transition function on \mathcal{X} . Let $A \subseteq \mathcal{X}$ be closed and P -invariant.*

1. *If $\tilde{\pi} \in \mathcal{P}(A)$ is \tilde{P} -invariant, then the extension π of $\tilde{\pi}$ to \mathcal{X} is P -invariant.*
2. *If $\pi \in \mathcal{P}(\mathcal{X})$ is P -invariant, then its restriction $\tilde{\pi}$ to A is \tilde{P} -invariant.*

Proof.

1. As $\tilde{\pi}$ is \tilde{P} -invariant we have

$$\tilde{\pi}(C) = \int_A \tilde{P}(x, C) \tilde{\pi}(dx)$$

for any $C \in \mathcal{B}(A) = A \cap \mathcal{B}(\mathcal{X})$. For $B \in \mathcal{B}(\mathcal{X})$ we have

$$\begin{aligned} (T\pi)(B) &= \int_{\mathcal{X}} P(x, B) \pi(dx) \\ &\stackrel{(1)}{=} \int_A P(x, B) \pi(dx) \\ &= \int_A P(x, B \cap A) \pi(dx) \\ &\stackrel{(2)}{=} \int_A \tilde{P}(x, B \cap A) \tilde{\pi}(dx) \\ &= \tilde{\pi}(B \cap A) \\ &= \pi(B), \end{aligned}$$

where in (1) we use that $\text{supp}(\pi) \subseteq A$, and in (2) we use the fact that \tilde{P} and $\tilde{\pi}$ agree on $A \cap \mathcal{B}(\mathcal{X})$.

2. Let $B \in \mathcal{B}(A) = A \cap \mathcal{B}(\mathcal{X})$, so that in particular we have $B \subseteq A$. Then

$$\begin{aligned} \pi(B) &= \int_{\mathcal{X}} P(x, B) \pi(dx) \\ &= \int_A P(x, B) \pi(dx) + \int_{\mathcal{X} \setminus A} P(x, B) \pi(dx). \end{aligned}$$

By the invariance of A we have

$$\begin{aligned} \pi(A) &= \int_A P(x, A) \pi(dx) + \int_{\mathcal{X} \setminus A} P(x, A) \pi(dx) \\ &= \pi(A) + \int_{\mathcal{X} \setminus A} P(x, A) \pi(dx) \end{aligned}$$

Therefore,

$$\int_{\mathcal{X} \setminus A} P(x, A) \pi(dx) = 0$$

which implies that

$$\int_{\mathcal{X} \setminus A} P(x, B) \pi(dx) = 0.$$

Hence,

$$\pi(B) = \int_A P(x, B) \pi(dx).$$

As $\tilde{\pi}$ and \tilde{P} coincide with π and P on A respectively, it follows that $\tilde{\pi}$ is \tilde{P} -invariant. \square

Theorem 7.6.4. *Let P be Feller on \mathcal{X} and A be a compact P -invariant set, then the restriction of P to A is Feller. In particular, there exists a P -invariant $\pi \in \mathcal{P}(\mathcal{X})$.*

Proof. Let $\tilde{f} \in \mathcal{C}_b(A)$. Then by the Tietze Extension theorem, there exists $f \in \mathcal{C}_b(\mathcal{X})$ such that $f|_A = \tilde{f}$. Then for $x \in A$ we have

$$\begin{aligned} (\tilde{T}\tilde{f})(x) &= \int_A \tilde{f}(y) \tilde{P}(x, dy) \\ &= \int_A f(y) P(x, dy) \\ &\stackrel{(1)}{=} \int_{\mathcal{X}} f(y) P(x, dy) \\ &= (Tf)(x), \end{aligned}$$

where in (1) we have used the fact that A is P -invariant and $x \in A$. As P is Feller it follows that $Tf \in \mathcal{C}_b(\mathcal{X})$ as $f \in \mathcal{C}_b(\mathcal{X})$. Therefore, as $(Tf)|_A = \tilde{T}\tilde{f}$ we deduce that $\tilde{T}\tilde{f} \in \mathcal{C}_b(A)$ which shows that \tilde{P} is Feller on A . Hence, by Corollary 7.3.8 there exists a \tilde{P} -invariant probability measure $\tilde{\pi} \in \mathcal{P}(A)$. Using Lemma statement 1 of 7.6.3 we can conclude that π , the extension of $\tilde{\pi}$ to \mathcal{X} , is P -invariant. \square

We have seen with Theorem 7.6.4 that P -invariance can be used to deduce the existence of invariant measures. On the other hand, we would like to see how P -invariance can be used to make statements regarding all invariant measures, thus working toward characterising the uniqueness of invariance measures. Consequently, for a P -invariant set A we introduce the sequence of sets $(A_n)_{n \in \mathbb{N}}$ where $A_0 = A$ and

$$A_n = \{x \in \mathcal{X} : P(x, A_{n-1}) > 0\} \quad (7.6.1)$$

for $n \geq 1$. It can be argued inductively that the sequence of sets $(A_n)_{n \in \mathbb{N}}$ is nested. More specifically, if we assume $A_{n-1} \subseteq A_n$ then for $x \in A_n$ we have that,

$$P(x, A_n) \stackrel{A_{n-1} \subseteq A_n}{\geq} P(x, A_{n-1}) \stackrel{x \in A_n}{>} 0,$$

which implies that $x \in A_{n+1}$.

Lemma 7.6.5. *Let $A \subseteq \mathcal{X}$ be P -invariant and let $(A_n)_{n \in \mathbb{N}}$ be as in (7.6.1). Then for every $n \geq 1$ and $x \in A_n$ we have $P^n(x, A) > 0$.*

Proof. We proceed by induction on n .

- The case $n = 1$ follows by construction of A_1 .
- Suppose the statement holds for $n - 1$, then

$$\begin{aligned} P^n(x, A) &\stackrel{(1)}{=} \int_{\mathcal{X}} P^{n-1}(y, A) P(x, dy) \\ &\geq \int_{A_{n-1}} P^{n-1}(y, A) P(x, dy) \\ &\stackrel{(2)}{>} 0, \end{aligned}$$

where (1) is an application of the Chapman-Kolmogorov equation. With (2) following by the inductive assumption as $y \in A_{n-1}$ implies that $P^{n-1}(y, A) > 0$, and we know that $P(x, A_{n-1}) > 0$ as $x \in A_n$.

□

Remark 7.6.6. Where the original construction of A_n was the set of points which with positive probability reach A_{n-1} in one step, Lemma 7.6.5 provides an alternative characterisation of the set A_n . Namely, A_n is the set of points which with positive probability reach the set A in n steps.

Proposition 7.6.7. Let $A \subseteq \mathcal{X}$ be P -invariant and consider the sequence $(A_n)_{n \in \mathbb{N}}$ as given by (7.6.1). Suppose $\mathcal{X} = \bigcup_{n \in \mathbb{N}} A_n$, then for every P -invariant probability measure $\pi \in \mathcal{P}(\mathcal{X})$ we have $\pi(A) = 1$.

Proof. Suppose that π is P -invariant with $\pi(A) < 1$. Then since $\pi(A_n) \nearrow 1$, there exists an $N \in \mathbb{N}$ such that $\pi(A_N \setminus A) > 0$. Therefore,

$$\begin{aligned} \pi(A) &= \int_{\mathcal{X}} P^N(x, A) \pi(dx) \\ &\geq \int_{A_N} P^N(x, A) \pi(dx) \\ &\geq \underbrace{\int_A P^N(x, A) \pi(dx)}_{\pi(A)} + \underbrace{\int_{A_N \setminus A} P^N(x, A) \pi(dx)}_{\stackrel{(1)}{>0}} \\ &> \pi(A), \end{aligned}$$

where (1) follows by construction of A_N and because $x \in A_N$ implies that $P^N(x, A) > 0$ by Lemma 7.6.5. Therefore, we arrive at a contradiction and so $\pi(A) = 1$. □

Corollary 7.6.8. Consider a random dynamical system

$$X_{n+1} = F(X_n, \xi_n).$$

Suppose A is a P -invariant set, where P is a transition probability on \mathcal{X} with the Feller property. Moreover, suppose that A is compact with $\bigcup_{n \in \mathbb{N}} A_n = \mathcal{X}$, and there exists a $\gamma \in (0, 1)$ such that

$$\mathbb{E}(d(F(x, \xi), F(y, \xi))) \leq \gamma d(x, y) \tag{7.6.2}$$

for all $x, y \in A$. Then there exists a unique P -invariant measure on \mathcal{X} .

Proof. The existence of an invariant measure follows from Theorem 7.6.4. By Proposition 7.6.7 any P -invariant probability measure $\pi \in \mathcal{P}(\mathcal{X})$ is an extension of $\tilde{\pi} \in \mathcal{P}(A)$ which is \tilde{P} -invariant by statement 2 of Lemma 7.6.3. Therefore, if two P -invariant measures on \mathcal{X} existed, we would have two \tilde{P} -invariant measures on A . However, using (7.6.2) we know by Theorem 7.4.13 that \tilde{P} -invariant probability measure on A is unique. Therefore, P -invariant probability measures on \mathcal{X} are unique. □

Example 7.6.9. Let $(\xi_n)_{n \in \mathbb{N}}$ be a sequence of independent and identically distribution $\mathcal{C}([0, 1])$ -valued random variables such that $\sup_{t \in [0, 1]} |\xi_n(t)| \leq 1$ almost surely. Let $\varphi_t(x, f)$ be the solution to the differential equation

$$\begin{cases} \frac{dx}{dt} = \frac{1}{x(t)} - 2 + f(t) & t \in (0, \infty) \\ x(0) = x. \end{cases}$$

With $F(x, f) = \varphi_1(x, f)$ consider the Markov process $(X_n)_{n \in \mathbb{N}}$ where $X_0 \in (0, \infty)$ and

$$X_n = F(X_{n-1}, \xi_{n-1})$$

for $n \geq 1$. One can show that $\varphi_t(x, f)$ is indeed well-defined, that is, it exists and is unique. Recall that $\xi_n \in [-1, 1]$ almost surely, hence, the deterministic maps $F_-(x) = F(x, -1)$ and $F_+(x) = F(x, 1)$ are such that

$$X_{n+1} \in [F_-(X_n), F_+(X_n)] \quad (7.6.3)$$

almost surely. Observe the following.

1. Note that $F_+(x)$ is the solution to $\frac{dz}{dt} = \frac{1}{z(t)} - 1$ with $z(0) = x$. For which $z(t) \equiv 1$ is an equilibrium solution, and

$$\lim_{t \rightarrow \infty} \varphi_t(x, 1) = \lim_{t \rightarrow \infty} z(t) = 1.$$

In particular, we note that $\varphi_n(x, 1) = \underbrace{(F_+ \circ \dots \circ F_+)}_n(x) =: F_+^n(x)$ and so $\lim_{n \rightarrow \infty} F_+^n(x) = 1$.

2. Note that $F_-(x)$ is the solution to $\frac{dz}{dt} = \frac{1}{z(t)} - 3$ with $z(0) = x$. For which $z(t) \equiv \frac{1}{3}$ is an equilibrium solution, and

$$\lim_{t \rightarrow \infty} \varphi_t(x, -1) = \lim_{t \rightarrow \infty} z(t) = \frac{1}{3}.$$

In particular, we note that $\varphi_n(x, -1) = \underbrace{(F_- \circ \dots \circ F_-)}_n(x) =: F_-^n(x)$ and so $\lim_{n \rightarrow \infty} F_-^n(x) = \frac{1}{3}$.

Now for a fixed $\epsilon > 0$ let $A = [\frac{1}{3} - \epsilon, 1 + \epsilon]$.

1. On $(1, \infty)$ we note that $\frac{1}{z(t)} - 1 < 0$ and so it must be the case that $(F_+)^{-1}(1 + \epsilon) > 1 + \epsilon$.
2. On $(0, \frac{1}{3})$ we have that $\frac{1}{z(t)} - 3 > 0$ and so it must be the case that $(F_-)^{-1}(\frac{1}{3} - \epsilon) < \frac{1}{3} - \epsilon$.

Consequently,

$$A \subseteq \left[(F_-)^{-1} \left(\frac{1}{3} - \epsilon \right), (F_+)^{-1} (1 + \epsilon) \right].$$

In particular, using (7.6.3) this means that if $X_0 \in \left[(F_-)^{-1} \left(\frac{1}{3} - \epsilon \right), (F_+)^{-1} (1 + \epsilon) \right]$ then $X_1 \in [\frac{1}{3} - \epsilon, 1 + \epsilon] = A$. Therefore, A is a P -invariant set. Recalling that

$$A_{n+1} := \{x \in \mathcal{X} : P(x, A_n) > 0\}$$

we can use similar arguments to show that

$$\left[(F_-^n)^{-1} \left(\frac{1}{3} - \epsilon \right), (F_+^n)^{-1} (1 + \epsilon) \right] \subseteq A_n.$$

Moreover, using the observation regarding convergence to equilibrium solutions made previously, we note that

$$\bigcup_{n \in \mathbb{N}} A_n = (0, \infty).$$

Therefore, using Proposition 7.6.7 we deduce that for any invariant probability measure π we have that $\pi([\frac{1}{3} - \epsilon, 1 + \epsilon]) = 1$. As $\epsilon > 0$ was arbitrary we conclude that $\pi([\frac{1}{3}, 1]) = 1$. Note that since $F(x, \xi_n)$ is differentiable in x , it is continuous and so $(X_n)_{n \in \mathbb{N}}$ is Feller by Theorem 7.4.3. In particular, let $v(t) = \frac{d\varphi_t(x, f)}{dx}$ such that

$$\frac{dv(t)}{dt} = -\frac{1}{x^2(t)}v(t)$$

with $v(0) = 1$. Consequently, we have that

$$v(t) = \exp\left(-\int_0^t \frac{1}{x^2(s)} ds\right).$$

Therefore,

$$F'(X_n, \xi_n) = v(t) = \exp\left(-\int_0^t \frac{1}{x^2(s)} ds\right) \leq \exp\left(\frac{1}{1+\epsilon}\right) < 1.$$

So,

$$\mathbb{E}(|F(x, \xi) - F(y, \xi)|) < |x - y|.$$

Thus we have P -invariance on a compact set A such that $\bigcup_{n \in \mathbb{N}} A_n = \mathcal{X}$, allowing us to applying Corollary 7.6.8 to conclude that the Markov chain $(X_n)_{n \in \mathbb{N}}$ on $(0, \infty)$ has a unique invariant probability measure.

7.7 Solution to Exercises

Exercise 7.2.10

Solution.

1. Let $A = \text{supp}(\mu)$. Then by Proposition 7.2.9 we know that $\mu(A) = 1$. As $\text{supp}(\nu) \subseteq \mathcal{X} \setminus A$ and $\nu(\text{supp}(\nu)) = 1$ it follows that $\nu(\mathcal{X} \setminus A) = 1$ which implies $\nu(A) = 0$.
2. Let $\mathcal{X} = [0, 1]$, $\mu = \delta_0$ and ν the Lebesgue measure. Then $\mu \perp \nu$ as one can take $A = \{0\}$, however, $\text{supp}(\mu) \cap \text{supp}(\nu) = \{0\}$. □

Exercise 7.3.6

Solution. Suppose that $\mu_n \not\rightarrow \mu$ weakly. Then as $\mathcal{P}(\mathcal{X})$ with the weak topology is metrizable, say with the metric $d(\cdot, \cdot)$, there exists an $\epsilon > 0$ and a subsequence $(\mu_{n_k})_{k \in \mathbb{N}}$ such that $d(\mu, \mu_{n_k}) \geq \epsilon$ for all $k \in \mathbb{N}$. However, as $(\mu_n)_{n \in \mathbb{N}}$ is tight it is relatively compact. Therefore, the sequence $(\mu_{n_k})_{k \in \mathbb{N}} \subseteq (\mu_n)_{n \in \mathbb{N}}$ contains a convergent subsequence, which by assumption has limit μ . This contradicts the fact that $d(\mu, \mu_{n_k}) \geq \epsilon$ for all $k \in \mathbb{N}$. □

Exercise 7.3.10

Solution. For any $M > 0$ we have that

$$\begin{aligned} \sup_{n \in \mathbb{N}} (P^n(x, B(0, M)^c)) &= \sup_{n \in \mathbb{N}} (\mathbb{P}_x(|X_n| \geq M)) \\ &\stackrel{\text{Chebyshev.}}{\leq} \frac{1}{G(M)} \sup_{n \in \mathbb{N}} (\mathbb{E}_x(G(|X_n|))) \\ &\xrightarrow{M \rightarrow \infty} 0. \end{aligned}$$

Hence, we deduce that $(P^n(x, \bullet))_{n \in \mathbb{N}}$ is tight and so we can apply Theorem 7.3.7 to conclude. □

Exercise 7.4.2

Solution. Let $f \in \mathcal{B}_b(\mathcal{X})$, then

$$\begin{aligned} (Tf)(x) &= \mathbb{E}(f(X_1)|X_0 = x) \\ &= \mathbb{E}(f(F(X_0, \xi_0))|X_0 = x) \\ &\stackrel{(1)}{=} \int_{\mathcal{Y}} f(F(x, y)) \nu(dy) \end{aligned}$$

where in (1) we use the independence of the random variables $(\xi_n)_{n \in \mathbb{N}}$. □

Exercise 7.5.3

Solution. On the one hand, for any measurable set $A \subseteq \Omega$ we have

$$\begin{aligned} \|\mu - \nu\|_{\text{TV}} &= \int_{\Omega} \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta \\ &= \int_A \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta + \int_{\Omega \setminus A} \left| \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} \right| d\eta \\ &\geq \left| \int_A \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} d\eta - \int_{\Omega \setminus A} \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} d\eta \right| \\ &= |\mu(A) - \nu(A) - \mu(\Omega \setminus A) + \nu(\Omega \setminus A)| \\ &= 2|\mu(A) - \nu(A)|. \end{aligned}$$

Therefore, $\|\mu - \nu\|_{\text{TV}} \geq 2 \sup(\{|\mu(A) - \nu(A)| : A \subseteq \Omega \text{ measurable}\})$. On the other hand, let $A = \left\{ \frac{d\mu}{d\eta} \geq \frac{d\nu}{d\eta} \right\}$.

Then as

$$\int_A \frac{d\mu}{d\eta} d\eta = \mu(A) = 1 - \mu(\Omega \setminus A) = 1 - \int_{\Omega \setminus A} \frac{d\mu}{d\eta} d\eta,$$

it follows that

$$\int_A \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} d\eta = \int_{\Omega \setminus A} \frac{d\nu}{d\eta} - \frac{d\mu}{d\eta} d\eta = \frac{1}{2} \|\mu - \nu\|_{\text{TV}}.$$

Therefore, as

$$0 \leq \int_A \frac{d\mu}{d\eta} - \frac{d\nu}{d\eta} d\eta = \mu(A) - \nu(A) = |\mu(A) - \nu(A)|$$

it follows that $\|\mu - \nu\|_{\text{TV}} \leq 2 \sup(\{|\mu(A) - \nu(A)| : A \subseteq \Omega \text{ measurable}\})$. □

Exercise 7.5.12

Solution. Clearly $\eta \geq 0$, and

$$\eta(\Omega) = \sum_{n=1}^{\infty} \frac{1}{2^n} \mu_n(\Omega) = \sum_{n=1}^{\infty} \frac{1}{2^n} = 1.$$

Moreover, $\eta(A) \leq \eta(B)$ for $A \subseteq B \subseteq \Omega$. If $A_1, A_2, \dots \subseteq \Omega$ are disjoint sets then

$$\eta\left(\bigcup_{i=1}^{\infty} A_i\right) = \sum_{n=1}^{\infty} \frac{1}{2^n} \mu_n\left(\bigcup_{i=1}^{\infty} A_i\right) = \sum_{n=1}^{\infty} \sum_{i=1}^{\infty} \frac{1}{2^n} \mu_n(A_i).$$

We note that the sum is absolutely convergent, as the terms are non-negative and it is bounded above by one, so we can exchange the order of summation to deduce that

$$\eta\left(\bigcup_{i=1}^{\infty} A_i\right) = \sum_{i=1}^{\infty} \sum_{n=1}^{\infty} \frac{1}{2^n} \mu_n(A_i) = \sum_{i=1}^{\infty} \eta(A_i).$$

Therefore, η is countably additive and hence a probability measure. □

Exercise 7.6.2

Solution. Proceed by induction on n .

- When $n = 1$ we have

$$\begin{aligned} \pi(A) &= \mathbb{P}(X_0 \in A, X_1 \in \mathcal{X}) \\ &= \mathbb{P}(X_0 \in A, X_1 \in A) + \mathbb{P}(X_0 \in A, X_1 \in \mathcal{X} \setminus A) \\ &\stackrel{(1)}{=} \mathbb{P}(X_0 \in A, X_1 \in A) + 0 \\ &= \mathbb{P}(X_0 \in A, X_1 \in A), \end{aligned}$$

where (1) follows by the invariance of A .

- Assuming that

$$\mathbb{P}(X_0 \in A, \dots, X_n \in A) = \pi(A)$$

it follows that

$$\begin{aligned}\pi(A) &= \mathbb{P}(X_0 \in A, \dots, X_n \in A, X_{n+1} \in \mathcal{X}) \\ &= \mathbb{P}(X_0 \in A, \dots, X_n \in A, X_{n+1} \in A) + \mathbb{P}(X_0 \in A, \dots, X_n \in A, X_{n+1} \in \mathcal{X} \setminus A) \\ &\stackrel{(1)}{=} \mathbb{P}(X_0 \in A, \dots, X_n \in A, X_{n+1} \in A) + 0 \\ &= \mathbb{P}(X_0 \in A, \dots, X_n \in A, X_{n+1} \in A),\end{aligned}$$

where (1) follows by the invariance of A .

Therefore, by induction we conclude that

$$\mathbb{P}(X_0 \in A_0, \dots, X_n \in A) = \pi(A)$$

for every $n \in \mathbb{N}$.

□

8 Ergodic Theory

8.1 Birkhoff's Ergodic Theorem for Dynamical Systems

Definition 8.1.1. A dynamical system consists of a probability space $(\Omega, \mathcal{F}, \mathbb{P})$ and a measure-preserving map $\theta : \Omega \rightarrow \Omega$. That is, $\mathbb{P}(\theta^{-1}(A)) = \mathbb{P}(A)$ for every $A \in \mathcal{F}$.

Definition 8.1.2. Given a measurable transformation θ on $(\Omega, \mathcal{F}, \mathbb{P})$ a set A such that $\theta^{-1}(A) = A$ is said to be θ -invariant. In particular, the θ -invariant σ -algebra $\mathcal{I} \subset \mathcal{F}$ is given by

$$\mathcal{I} = \{A \in \mathcal{F} : \theta^{-1}(A) = A\}. \quad (8.1.1)$$

Exercise 8.1.3. Show that \mathcal{I} , as given by (8.1.1), is indeed a σ -algebra.

Definition 8.1.4. A measurable function $f : \Omega \rightarrow \mathbb{R}$ is θ -invariant if $f \circ \theta = f$.

Exercise 8.1.5. Let $f : \Omega \rightarrow \mathbb{R}$ be a \mathcal{F} -measurable function. Then f is invariant if and only if f is measurable with respect to \mathcal{I} , as given by (8.1.1).

Similarly, a \mathcal{F} -measurable function $f : \Omega \rightarrow \mathbb{R}$ is invariant if and only if f is measurable with respect to \mathcal{I} as defined above.

Definition 8.1.6. Consider a dynamical system $(\Omega, \mathcal{F}, \mathbb{P})$ with θ a measure-preserving map. Then θ is ergodic if for any $A \in \Omega$ a θ -invariant set, we have $\mathbb{P}(A) \in \{0, 1\}$. Ergodicity is also a property of \mathbb{P} , and so \mathbb{P} is said to be ergodic with respect to θ .

Proposition 8.1.7. Let $(\Omega, \mathcal{F}, \mathbb{P})$ with θ be a dynamical system. Then the following are equivalent.

1. \mathbb{P} is ergodic with respect to θ .
2. Every θ -invariant integrable function $f : \Omega \rightarrow \mathbb{R}$ is almost surely constant.
3. Every θ -invariant bounded function $f : \Omega \rightarrow \mathbb{R}$ is almost surely constant.

Proof. (2) \Rightarrow (3). This is clear as every bounded function is integrable with respect to a probability measure.

(3) \Rightarrow (1). Let $f = \mathbf{1}_A$, where A is an invariant set. Then f is invariant and bounded and so is almost surely constant, that is $f \in \{0, 1\}$. Which implies that $\mathbb{P}(A) = f \in \{0, 1\}$ meaning \mathbb{P} is ergodic.

(1) \Rightarrow (2). Let f be an integrable and invariant function. Then by Exercise 8.1.5 it follows that f is \mathcal{I} -measurable. Consider the sets

- $A_+ := \{\omega \in \Omega : f(\omega) > \mathbb{E}(f)\}$,
- $A_- := \{\omega \in \Omega : f(\omega) < \mathbb{E}(f)\}$, and
- $A_0 := \{\omega \in \Omega : f(\omega) = \mathbb{E}(f)\}$.

These sets form a disjoint partition of Ω , and so by the ergodic property of \mathbb{P} it follows that exactly one has full measure whilst the others have zero-measure. If $\mathbb{P}(A_+) = 1$ then

$$0 = \int_{\Omega} f - \mathbb{E}(f) \, d\mathbb{P} = \int_{A_+} f - \mathbb{E}(f) \, d\mathbb{P},$$

which implies that $f - \mathbb{E}(f) = 0$ almost surely on A_+ which is a contradiction. Similarly, we have that $\mathbb{P}(A_-) \neq 1$. Therefore, $\mathbb{P}(A_0) = 1$ which implies that f is almost surely constant. \square

Theorem 8.1.8 (Maximal Ergodic Theorem). Let $(\Omega, \mathcal{I}, \mathbb{P})$ with θ be a dynamical system, where \mathcal{I} is as given by (8.1.1). Let $f : \Omega \rightarrow \mathbb{R}$ be such that $\mathbb{E}(|f|) < \infty$. Let

$$S_N(\omega) = \sum_{n=0}^{N-1} f(\theta^n \omega)$$

and

$$M_N(\omega) = \max \{S_0(\omega), \dots, S_N(\omega)\}$$

where $S_0 = 0$. Then

$$\int_{\{M_N > 0\}} f(\omega) \mathbb{P}(d\omega) \geq 0$$

for every $N \geq 1$.

Proof. Observe that for $0 \leq k \leq N$ and every $\omega \in \Omega$ we have

$$f(\omega) + S_k(\theta\omega) - S_{k+1}(\omega)$$

and

$$S_k(\theta\omega) \leq M_N(\theta\omega).$$

This implies that,

$$f(\omega) + M_N(\theta\omega) \geq f(\omega) + S_k(\theta\omega) = S_{k+1}(\omega).$$

Therefore,

$$f(\omega) \geq \max(S_1(\omega), \dots, S_N(\omega)) - M_N(\theta\omega). \quad (8.1.2)$$

Furthermore, on the set $\{M_N > 0\}$ we have

$$M_N(\omega) = \max(S_1(\omega), \dots, S_N(\omega)).$$

Combined with (8.1.2) it follows that $f(\omega) \geq M_N(\omega) - M_N(\theta\omega)$ on $\{M_N > 0\}$. Note that as $M_N \geq 0$ we have

$$\begin{aligned} \mathbb{E}(M_N) &= \int_{\Omega} M_N(\omega) d\mathbb{P}(\omega) \\ &= \int_{\{M_N=0\}} M_N(\omega) d\mathbb{P}(\omega) + \int_{\{M_N>0\}} M_N(\omega) d\mathbb{P}(\omega) \\ &= 0 (\mathbb{P}(\{M_N = 0\})) + \int_{\{M_N>0\}} M_N(\omega) d\mathbb{P}(\omega) \\ &= \int_{\{M_N>0\}} M_N(\omega) d\mathbb{P}(\omega). \end{aligned}$$

Therefore,

$$\begin{aligned} \int_{\{M_N>0\}} f(\omega) \mathbb{P}(d\omega) &\geq \int_{\{M_N>0\}} M_N(\omega) - M_N(\theta\omega) \mathbb{P}(d\omega) \\ &\geq \mathbb{E}(M_N) - \int_{A_N} M_N(\omega) \mathbb{P}(d\omega), \end{aligned}$$

where $A_N := \{\theta\omega : M_N(\omega) > 0\} \subseteq \Omega$. Hence,

$$\int_{\{M_N>0\}} f(\omega) \mathbb{P}(d\omega) \geq 0.$$

□

Theorem 8.1.9 (Birkhoff's Ergodic Theorem). Let $(\Omega, \mathcal{F}, \mathbb{P})$ with θ be a dynamical system and let \mathcal{I} be as in (8.1.1). For $f : \Omega \rightarrow \mathbb{R}$ with $\mathbb{E}(|f|) < \infty$ it follows that

$$\lim_{N \rightarrow \infty} \left(\frac{1}{N} \sum_{n=0}^{N-1} f(\theta^n \omega) \right) = \mathbb{E}(f|\mathcal{I})$$

almost surely.

Proof. By replacing f with $f - \mathbb{E}(f|\mathcal{I})$, we can assume without loss of generality that $\mathbb{E}(f|\mathcal{I}) = 0$. Let

$$\bar{\eta} = \limsup_{n \rightarrow \infty} \frac{S_n}{n}$$

and

$$\underline{\eta} = \liminf_{n \rightarrow \infty} \frac{S_n}{n}.$$

Note that $\bar{\eta}(\theta\omega) = \bar{\eta}(\omega)$, so that for $\epsilon > 0$ it follows that

$$A^\epsilon = \{\bar{\eta}(\omega) > \epsilon\} \in \mathcal{I}.$$

Let

$$f^\epsilon(\omega) = (f(\omega) - \epsilon)\mathbf{1}_{A^\epsilon}(\omega)$$

to make analogous definition of S_N^ϵ and M_N^ϵ as those made in Theorem 8.1.8. Then using Theorem 8.1.8 it follows that

$$\int_{\{M_N^\epsilon > 0\}} f^\epsilon(\omega) \mathbb{P}(d\omega) \geq 0$$

for $N \geq 1$. Next, observe that the sequence of sets $\{M_N^\epsilon > 0\}$ indexed by N is increasing to

$$B^\epsilon := \left\{ \sup_N S_N^\epsilon > 0 \right\} = \left\{ \sup_N \frac{S_N^\epsilon}{N} > 0 \right\}.$$

Noting that

$$\frac{S_N^\epsilon(\omega)}{N} = \begin{cases} 0 & \bar{\eta}(\omega) \leq \epsilon \\ \frac{S_N(\omega)}{N} - \epsilon & \bar{\eta}(\omega) > \epsilon, \end{cases}$$

it follows that

$$B^\epsilon = \{\bar{\eta} > \epsilon\} \cap \left\{ \sup_N \frac{S_N}{N} > \epsilon \right\} = \{\bar{\eta} > \epsilon\} = A^\epsilon.$$

As $\mathbb{E}(|f^\epsilon|) \leq \mathbb{E}(|f|) + \epsilon < \infty$, using the dominated convergence theorem we deduce that

$$0 \leq \lim_{N \rightarrow \infty} \int_{\{M_N^\epsilon > 0\}} f^\epsilon(\omega) \mathbb{P}(d\omega) = \int_{A^\epsilon} f^\epsilon(\omega) \mathbb{P}(d\omega).$$

Therefore,

$$\begin{aligned} 0 &\leq \int_{A^\epsilon} f^\epsilon(\omega) \mathbb{P}(d\omega) \\ &= \int_{A^\epsilon} f(\omega) - \epsilon \mathbb{P}(d\omega) \\ &= \int_{A^\epsilon} f(\omega) \mathbb{P}(d\omega) - \epsilon \mathbb{P}(A^\epsilon) \\ &= \int_{A^\epsilon} \mathbb{E}(f|\mathcal{I})(\omega) \mathbb{P}(d\omega) - \epsilon \mathbb{P}(A^\epsilon) \\ &= 0 - \epsilon \mathbb{P}(A^\epsilon). \end{aligned}$$

Therefore, $\mathbb{P}(A^\epsilon) = 0$ for every $\epsilon > 0$, implying that $\bar{\eta} \leq 0$ almost surely. Repeating a similar argument with $-f$ we see that $\underline{\eta} \geq 0$. Hence, $\bar{\eta} = \underline{\eta} = 0$ which completes the proof. \square

Corollary 8.1.10. *If a dynamical system $(\Omega, \mathcal{F}, \mathbb{P})$ with θ is ergodic, then*

$$\lim_{N \rightarrow \infty} \left(\frac{1}{N} \sum_{n=0}^{N-1} f(\theta^n \omega) \right) = \mathbb{E}(f)$$

almost surely.

Proof. As $\mathbb{E}(f|\mathcal{I})$ is \mathcal{I} -measurable, if θ is ergodic it follows by statement 2. of Proposition 8.1.7 that $\mathbb{E}(f|\mathcal{I})$ is almost surely constant. More specifically, $\mathbb{E}(f|\mathcal{I}) = \mathbb{E}(f)$. Therefore,

$$\lim_{N \rightarrow \infty} \left(\frac{1}{N} \sum_{n=0}^{N-1} f(\theta^n \omega) \right) \stackrel{\text{Thm 8.1.9}}{=} \mathbb{E}(f|\mathcal{I}) = \mathbb{E}(f).$$

□

8.2 Birkhoff's Ergodic Theorem for Markov Chains

Having seen how ergodic theorems can be derived for dynamical systems, it will be useful to investigate how to arrive at a dynamical system from a Markov chain. For this we investigate the space of sequences $\mathcal{X}^{\mathbb{N}}$, and more generally $\mathcal{X}^{\mathbb{Z}}$. More specifically, we consider the shift operator θ on these spaces, which is such that for $x := (x_0, x_1, \dots) \in \mathcal{X}^{\mathbb{N}}$ we have

$$\theta(x) = (x_1, x_2, \dots)$$

and similarly for $x := (\dots, x_{-1}, x_0, x_1, \dots)$. In particular, for $x \in \mathcal{X}^{\mathbb{N}}$ we let

$$(\theta_n x)(m) = x_{n+m},$$

and similarly for $x \in \mathcal{X}^{\mathbb{Z}}$. Note that on $\mathcal{X}^{\mathbb{Z}}$ we have $\theta = \theta_1$ and $\theta^{-1} = \theta_{-1}$. Therefore, as before we can consider

$$\mathcal{I} = \{C \in \mathcal{B}(\mathcal{X}^{\mathbb{Z}}) : \theta^{-1}C = C\}. \quad (8.2.1)$$

Thus to understand how the theory of dynamical systems can be used on Markov chains it will be important to understand how to construct two-side Markov processes on $\mathcal{X}^{\mathbb{Z}}$. To do so it will be necessary to work with a family of transition probabilities $P = (P(x, \cdot), x \in \mathcal{X})$ and a P -invariant probability measure π , that is $\pi = \int_{\mathcal{X}} P(x, \cdot) \pi(dx)$.

8.2.1 Constructing Two-Sided Markov Processes

One approach to constructing a Markov chain $(X_n)_{n \in \mathbb{Z}}$ is referred to as the finite-dimensional approach. In this case, a probability measure \mathbb{P}_π measure is constructed using finite-dimensional distributions and Kolmogorov's extension theorem. More specifically, $\mu_{n,m}$ denotes the distribution of $(X_{-n}, \dots, X_0, \dots, X_m)$ which is given by

$$P(z_{m-1}, dz_m) \dots P(z_0, dz_1) \dots P(z_{-n}, dz_{-n+1}) \pi(dz_{-n}) = \prod_{k=-n}^{m-1} P(z_k, dz_{k+1}) \pi(dz_{-n}).$$

By the invariance of π it follows that $(\mu_{n,m})_{n,m \in \mathbb{Z}}$ is a consistent family of probability measures. Therefore, using Theorem 3.2.7 it follows that \mathbb{P}_π on $\mathcal{X}^{\mathbb{Z}}$ defines a stationary Markov chain with transition probabilities P and $\text{Law}(X_n) = \pi$ for $n \in \mathbb{Z}$.

Another approach to constructing a Markov chain $(X_n)_{n \in \mathbb{Z}}$ is referred to as the time shift approach. Here we start with a Markov chain $(Y_n)_{n \in \mathbb{N}}$ which has π as an invariant initial distribution and let $(X_n^{(m)})_{n \geq -m}$ be such that

$$\left(X_{-m}^{(m)}, X_{-(m-1)}^{(m)}, X_{-(m-2)}^{(m)}, \dots \right) = \left(X_{-(m-1)}^{(m-1)}, X_{-(m-2)}^{(m-1)}, X_{-(m-3)}^{(m-1)}, \dots \right)$$

with

$$\left(X_{-1}^{(1)}, X_1^{(1)}, X_0^{(1)}, \dots \right) = (Y_0, Y_1, Y_2, \dots).$$

In the limit the process $(X_n^{(m)})$ gives the required Markov process on $\mathcal{X}^{\mathbb{Z}}$.

8.2.2 Birkhoff's Ergodic Theorem for Stationary Markov Chains as Dynamical Systems

Lemma 8.2.1. *The probability space $(\mathcal{X}^{\mathbb{Z}}, \mathcal{B}(\mathcal{X}^{\mathbb{Z}}), \mathbb{P}_{\pi})$ with the shift operator θ is a dynamical system.*

Proof. Let $(X_n)_{n \in \mathbb{Z}}$ be a time-homogeneous Markov chain with law \mathbb{P}_{π} . By the finite-dimensional construction, it follows that

$$\mathbb{P}(X_{-n} \in A_{-n}, \dots, X_m \in A_m) = \int_{A_{-n}} \cdots \int_{A_m} P(x_{m-1}, dx_m) \cdots P(x_{-n}, dx_{-(n-1)}) \pi(dx_{-n}), \quad (8.2.2)$$

similarly,

$$\mathbb{P}(X_{-(n-1)} \in A_{-n}, \dots, X_{m+1} \in A_m) = \int_{A_{-(n-1)}} \cdots \int_{A_{m+1}} P(x_m, dx_{m+1}) \cdots P(x_{-(n-1)}, dx_{-(n-2)}) \pi(dx_{-(n-1)}). \quad (8.2.3)$$

Clearly, the right-hand sides of (8.2.2) and (8.2.3) are identical, and so

$$\mathbb{P}(X_{-n} \in A_{-n}, \dots, X_m \in A_m) = \mathbb{P}(X_{-(n-1)} \in A_{-n}, \dots, X_{m+1} \in A_m),$$

which is to say that θ preserving map and so defined a dynamical system. \square

Remark 8.2.2. *Under the product topology, it is apparent that θ is continuous. Thus, the dynamical system of Lemma 8.2.1 is referred to as the continuous dynamical system.*

Definition 8.2.3.

1. A measure \mathbb{P} is ergodic for θ if for every $A \in \mathcal{I}$, where \mathcal{I} is as given by (8.2.1), we have $\mathbb{P}(A) \in \{0, 1\}$.
2. An invariant measure π that induces an ergodic probability measure \mathbb{P}_{π} for θ is said to be ergodic with respect θ .

Applying Theorem 8.1.9 to $(\theta_n a)_{n \in \mathbb{Z}}$ gives Corollary 8.2.4.

Corollary 8.2.4. *Let $f : \mathcal{X} \rightarrow \mathbb{R}$ be integrable and let $\tilde{f} : \mathcal{X}^{\mathbb{Z}} \rightarrow \mathbb{R}$ be given by*

$$\tilde{f}(\dots, a_{-1}, a_0, a_1, \dots) = f(a_0).$$

Then $\tilde{f}(\theta_n a) = f(a_n)$, so that

$$\frac{1}{n} \sum_{k=1}^n f(a_k) \xrightarrow{n \rightarrow \infty} \mathbb{E}_{\pi}(\tilde{f} | \mathcal{I})$$

almost surely with respect to \mathbb{P}_{π} . Moreover, if π is ergodic then

$$\frac{1}{n} \sum_{k=1}^n f(a_k) \xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}} f d\pi$$

almost surely with respect to \mathbb{P}_{π} .

Remark 8.2.5. *An analogous statement holds over $\mathcal{X}^{\mathbb{N}}$.*

8.2.3 Birkhoff's Ergodic Theorem for Markov Chains

Consider $(X_n)_{n \in \mathbb{Z}}$ a stationary time-homogeneous Markov chain on a probability space $(\Omega, \mathcal{F}, \mathbb{P})$ with transition probability P and a P -invariant initial distribution π . Currently, our ergodic theory makes statements about

$(X_n)_{n \in \mathbb{Z}}$ as a dynamical system on $(\mathcal{X}^{\mathbb{Z}}, \mathcal{B}(\mathcal{X}^{\mathbb{Z}}), \mathbb{P}_\pi)$ with θ being the shift operator. Here we expand our ergodic theory to make statements about $(X_n)_{n \in \mathbb{Z}}$ in terms of \mathbb{P} instead of \mathbb{P}_π . For this, let

$$I_P = \{\pi \in \mathcal{P}(\mathcal{X}) : T\pi = \pi\}.$$

Theorem 8.2.6. *Let $(X_n)_{n \in \mathbb{Z}}$ be a stationary Markov process with $X_0 \sim \pi$, where π is an invariant probability measure. Then the following hold.*

1. Let $f : \mathcal{X}^{\mathbb{Z}} \rightarrow \mathbb{R}$ be an integrable function. Let $\bar{f} = \mathbb{E}_\pi(f|\mathcal{I})$, then

$$\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \xrightarrow{n \rightarrow \infty} \bar{f}(X(\omega))$$

for almost every ω with respect to \mathbb{P} .

2. Moreover, if π is ergodic then

$$\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}^{\mathbb{Z}}} f d\mathbb{P}_\pi$$

for almost every ω with respect to \mathbb{P} .

Proof. For $f : \mathcal{X}^{\mathbb{Z}} \rightarrow \mathbb{R}$ in $L^1(\mathbb{P}_\pi)$ consider

$$E := \left\{ a \in \mathcal{X}^{\mathbb{Z}} : \frac{1}{n} \sum_{k=1}^n f(\theta^k a) \xrightarrow{n \rightarrow \infty} \mathbb{E}_\pi(\bar{f}|\mathcal{I}) \right\}.$$

By Corollary 8.2.4 it follows that $\mathbb{P}_\pi(E) = 1$. In particular, we have that

$$\mathbb{P}(\{\omega : X(\omega) \in E\}) = \mathbb{P}_\pi(E) = 1.$$

Therefore, for almost every ω with respect to \mathbb{P} we have that $\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega))$ converges. \square

Theorem 8.2.7. *Let $P = P(x, \cdot)$ be a transition probability with an invariant measure π . Let $(X_n)_{n \in \mathbb{Z}}$ be a time-homogeneous Markov process with transition probabilities P and initial position $X_0 = x$. Then for π -almost every $x \in \mathcal{X}$ the following hold.*

1. For $f : \mathcal{X}^{\mathbb{Z}} \rightarrow \mathbb{R}$ an integrable function $\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega))$ converges for almost every ω with respect to \mathbb{P} .

2. Moreover, if π is ergodic, then

$$\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}} f d\mathbb{P}_\pi$$

for almost every ω with respect to \mathbb{P} .

Proof. Suppose that $X_0 \sim \pi$. Then by Theorem 8.2.6 it follows that

$$\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \xrightarrow{n \rightarrow \infty} \bar{f}(X(\omega))$$

for almost every ω with respect to \mathbb{P} . Therefore, by the dominated convergence theorem it follows that

$$\mathbb{E} \left(\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \mid \sigma(X_0) \right) \xrightarrow{n \rightarrow \infty} \mathbb{E}(\bar{f}(X(\omega)) \mid \sigma(X_0))$$

for almost every ω with respect to \mathbb{P} . Therefore, for almost every $x \in \mathcal{X}$ with respect to π we have

$$\mathbb{E} \left(\frac{1}{n} \sum_{k=1}^n f(\theta^k X(\omega)) \mid X_0 = x \right) \xrightarrow{n \rightarrow \infty} \mathbb{E}(\bar{f}(X(\omega)) \mid X_0 = x)$$

for almost every ω with respect to \mathbb{P} . □

Corollary 8.2.8. *Let P be a transition probability with an ergodic invariant probability measure π . Then for $g : \mathcal{X} \rightarrow \mathbb{R}$ in $L^1(\pi)$ it follows that*

$$\frac{1}{n} \sum_{k=1}^n g(X_k(\omega)) \xrightarrow{n \rightarrow \infty} \int_{\mathcal{X}} g \, d\pi$$

for almost every ω with respect to \mathbb{P} .

Proof. Let $\tilde{g} : \mathcal{X}^{\mathbb{Z}} \rightarrow \mathbb{R}$ be given by $\tilde{g}(y) := g(y_0)$. Then using Theorem 8.2.7 with \tilde{g} and noting that

$$\int_{\mathcal{X}^{\mathbb{Z}}} \tilde{g} \, d\mathbb{P}_{\pi} = \int_{\mathcal{X}} g \, d\pi$$

the result follows. □

Proposition 8.2.9. *Ergodic invariant probability measures for a time-homogeneous Markov chain are equal or mutually singular.*

Proof. Let π_1 and π_2 be distinct invariant probability measures. Let $f : \mathcal{X} \rightarrow \mathbb{R}$ be a bounded measurable function such that

$$\int_{\mathcal{X}} f \, d\pi_1 = \int_{\mathcal{X}} f \, d\pi_2, \tag{8.2.4}$$

which exists as the measures are distinct. Let $(X_n)_{n \in \mathbb{Z}}$ be a Markov chain with $X_0 = x$. Then for $i = 1, 2$ let

$$E_i := \left\{ x : X_0 = x, \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n f(X_k) = \int_{\mathcal{X}} f \, d\pi_i \, \mathbb{P} \text{ almost everywhere} \right\}.$$

By Corollary 8.2.8 the set E_i is well-defined and is such that $\pi_i(E_i) = 1$. By (8.2.4) we have that $E_1 \cap E_2 = \emptyset$ and so $\pi_1(E_2) = 0$ which means that π_1 and π_2 are mutually singular if they are not equal. □

8.3 The Structure Theorem

The structure theorem relates to the invariant probability measures of a time-homogeneous Markov chain. More specifically, we consider a chain on \mathcal{X} with transition probability P , and with T being the corresponding transition operator. Moreover, let

$$I_P := \{ \pi \in \mathcal{P}(\mathcal{X}) : T\pi = \pi \} \tag{8.3.1}$$

consist of the probability measures that are invariant under T .

Exercise 8.3.1. *Let I_P be as given in (8.3.1). Show that the following statements hold.*

1. *If $\pi_1, \pi_2 \in I_P$ then $t\pi_1 + (1-t)\pi_2 \in I_P$ for $t \in [0, 1]$.*
2. *If T is Feller, then I_P is closed.*

Definition 8.3.2. A probability measure $\pi \in I_P$ is an extremal of I_P if π cannot be decomposed as $\pi = t\pi_1 + (1-t)\pi_2$ for $t \in (0, 1)$ and $\pi_1, \pi_2 \in I_P$ distinct.

To work towards the structure theorem we need to understand the measurability of set-theoretic constructions.

Definition 8.3.3. The set-theoretic difference of sets A and B is given by $A \triangle B := A \cup B \setminus (A \cap B)$.

Proposition 8.3.4. Let $(\Omega, \mathcal{F}, \mathbb{P})$ be a probability space. Let $A, B, (A_\alpha)_{\alpha \in \mathcal{A}}$ and $(B_\alpha)_{\alpha \in \mathcal{A}}$ be elements of \mathcal{F} . Then the following statements hold.

1. $A^c \triangle B^c = A \triangle B$.
2. $(\bigcup_{\alpha \in \mathcal{A}} A_\alpha) \triangle (\bigcup_{\alpha \in \mathcal{A}} B_\alpha) \subset \bigcup_{\alpha \in \mathcal{A}} (A_\alpha \triangle B_\alpha)$.
3. For a measurable function $f : \Omega \rightarrow \Omega$ we have $f^{-1}(A \triangle B) = f^{-1}(A) \triangle f^{-1}(B)$.
4. $(A \triangle B) \triangle (B \triangle C) = A \triangle C$.
5. $\mathbb{P}(A \triangle B) = 0$ implies that $\mathbb{P}(A) = \mathbb{P}(B)$.

Proof.

1. Proceeding directly,

$$\begin{aligned} A^c \triangle B^c &= ((\Omega \setminus A) \setminus (\Omega \setminus B)) \cup ((\Omega \setminus B) \setminus (\Omega \setminus A)) \\ &= (A \setminus B) \cup (B \setminus A) \\ &= A \triangle B. \end{aligned}$$

2. Let $\omega \in (\bigcup_{\alpha \in \mathcal{A}} A_\alpha) \triangle (\bigcup_{\alpha \in \mathcal{A}} B_\alpha)$ then

$$\omega \in \left(\bigcup_{\alpha \in \mathcal{A}} A_\alpha \right) \setminus \left(\bigcup_{\alpha \in \mathcal{A}} B_\alpha \right) \cup \left(\bigcup_{\alpha \in \mathcal{A}} B_\alpha \right) \setminus \left(\bigcup_{\alpha \in \mathcal{A}} A_\alpha \right).$$

Suppose without loss of generality that $\omega \in (\bigcup_{\alpha \in \mathcal{A}} A_\alpha) \setminus (\bigcup_{\alpha \in \mathcal{A}} B_\alpha)$, with $\omega \in A_{\bar{\alpha}}$. It follows that $\omega \in A_{\bar{\alpha}} \setminus B_{\bar{\alpha}}$ and so $\omega \in A_{\bar{\alpha}} \triangle B_{\bar{\alpha}}$. Therefore,

$$\left(\bigcup_{\alpha \in \mathcal{A}} A_\alpha \right) \triangle \left(\bigcup_{\alpha \in \mathcal{A}} B_\alpha \right) \subset \bigcup_{\alpha \in \mathcal{A}} (A_\alpha \triangle B_\alpha).$$

3. This follows from the fact that the pre-image is distributed over unions and intersections.
4. Figure 12a is a visual representation of

$$((A \setminus B) \cup (B \setminus A)) \setminus ((B \setminus C) \cup (C \setminus B)) \tag{8.3.2}$$

and Figure 12b is a visual representation of

$$((B \setminus C) \cup (C \setminus B)) \setminus ((A \setminus B) \cup (B \setminus A)). \tag{8.3.3}$$

Noting that $(A \triangle B) \triangle (B \triangle C)$ is given by the union of (8.3.2) and (8.3.3) we use Figure 12c to conclude that

$$(A \triangle B) \triangle (B \triangle C) = A \triangle C.$$

5. As $A \setminus B$ and $B \setminus A$ are disjoint sets it follows that

$$\mathbb{P}(A \triangle B) = \mathbb{P}(A \setminus B) + \mathbb{P}(B \setminus A).$$

Therefore, if $\mathbb{P}(A \triangle B) = 0$ it must be the case that $\mathbb{P}(A \setminus B) = \mathbb{P}(B \setminus A) = 0$. Noting that

$$\mathbb{P}(A) = \mathbb{P}(A \cap B) + \mathbb{P}(A \setminus B)$$

and

$$\mathbb{P}(B) = \mathbb{P}(A \cap B) + \mathbb{P}(B \setminus A).$$

It follows that

$$\mathbb{P}(A) = \mathbb{P}(A \cap B) = \mathbb{P}(B).$$

□

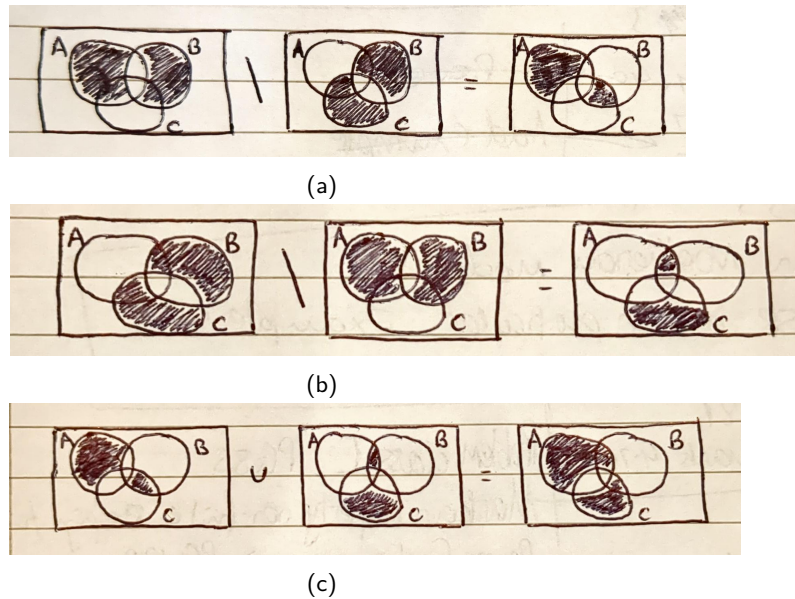


Figure 12: Visual proof for statement 3. of Proposition 8.3.4.

Definition 8.3.5. Let $(\Omega, \mathcal{F}, \mathbb{P})$ be a probability space. Then for $A, B \in \mathcal{F}$ we say $A \sim B$ if and only if $\mathbb{P}(A \triangle B) = 0$.

Definition 8.3.6. Let $(\Omega, \mathcal{F}, \mathbb{P})$ be a probability space.

1. The σ -algebra \mathcal{F} is complete with respect to \mathbb{P} if whenever $B \in \mathcal{F}$ and $\mathbb{P}(B) = 0$, then for any $A \subset B$ it follows that $A \in \mathcal{F}$.
2. The completion $\bar{\mathcal{F}}$ of \mathcal{F} with respect to \mathbb{P} is the smallest σ -algebra containing \mathcal{F} that is complete with respect to the measure \mathbb{P} .

Recall, that we can view our time-homogeneous Markov process as the canonical stochastic process on the probability space $(\mathcal{X}^{\mathbb{Z}}, \mathcal{B}(\mathcal{X}^{\mathbb{Z}}), \mathbb{P}_{\pi})$, where \mathbb{P}_{π} is the law of the chain with transition probability P and initial distribution π . We can then consider θ and θ^{-1} measure preserving transformations on $\mathcal{X}^{\mathbb{Z}}$ and

$$\mathcal{F}_n^m := \bigvee_{k=-n}^m \sigma(X_k) \subset \mathcal{B}(\mathcal{X}^{\mathbb{Z}}).$$

Lemma 8.3.7. Let $A \in \mathcal{B}(\mathcal{X}^{\mathbb{Z}})$, then for any $\epsilon > 0$ there exists an $N > 0$ and $A_\epsilon \in \mathcal{F}_{-N}^N$ such that

$$\mathbb{P}(A \Delta A_\epsilon) < \epsilon.$$

Proof. Consider

$$\mathcal{B}_0 := \{A \in \mathcal{B}(\mathcal{X}^{\mathbb{Z}}) : \text{for all } \epsilon > 0, \text{ there exists } N \in \mathbb{N} \text{ and } A_\epsilon \in \mathcal{F}_{-N}^N \text{ such that } \mathbb{P}(A \Delta A_\epsilon) < \epsilon\}.$$

Clearly, $\emptyset \in \mathcal{B}_0$. By statement 1. of Proposition 8.3.4 it follows that if $A \in \mathcal{B}_0$ then $A^c \in \mathcal{B}_0$. Now consider $(A_j)_{j \in \mathbb{N}} \subset \mathcal{B}_0$. By construction there exists a sequence $(N_j)_{j \in \mathbb{N}} \subset \mathbb{N}$ with corresponding events $A'_j \in \mathcal{F}_{-N_j}^{N_j}$ such that $\mathbb{P}(A_j \Delta A'_j) \leq \epsilon 2^{-j}$. Since \mathbb{P} is a finite measure, there exists a J such that with $A := \bigcup_{j \in \mathbb{N}} A_j$ we have $\mathbb{P}(A \Delta \bigcup_{j \leq J} A_j) \leq \epsilon$. Therefore,

$$\begin{aligned} \mathbb{P}\left(A \Delta \bigcup_{j \leq J} A'_j\right) &= \mathbb{P}\left(\left(A \Delta \bigcup_{j \leq J} A_j\right) \Delta \left(\bigcup_{j \leq J} A_j \Delta \bigcup_{j \leq J} A'_j\right)\right) \\ &\leq \mathbb{P}\left(A \Delta \bigcup_{j \leq J} A_j\right) + \mathbb{P}\left(\bigcup_{j \leq J} A_j \Delta \bigcup_{j \leq J} A'_j\right) \\ &\leq \epsilon + \mathbb{P}\left(\bigcup_{j \leq J} (A_j \Delta A'_j)\right) \\ &\leq \sum_{j \leq J} 2\epsilon. \end{aligned}$$

Hence, as $\bigcup_{j \leq J} A'_j \in \mathcal{F}_{-N}^N$ where $N = \max_{j \leq J} (N_j)$ we have that $\bigcup_{j \in \mathbb{N}} A_j \in \mathcal{B}_0$ which makes \mathcal{B}_0 a σ -algebra. It follows that $\mathcal{B}_0 = \mathcal{B}(\mathcal{X}^{\mathbb{Z}})$ which completes the proof. \square

Lemma 8.3.8. For any $A \in \mathcal{I} := \{C \in \mathcal{B}(\mathcal{X}^{\mathbb{Z}}) : \theta^{-1}C = C\}$ and $l \in \mathbb{Z}$, there exists $\hat{A}_l \in \sigma(X_l)$ such that $A \sim \hat{A}_l$.

Proof. Let $A \in \mathcal{I}$, then for any $\epsilon > 0$ we can use Lemma 8.3.7 to find a $N \in \mathbb{N}$ and $A_\epsilon \in \mathcal{F}_{-N}^N$ such that $\mathbb{P}(A \Delta A_\epsilon) < \epsilon$. Since,

$$\theta_{-1}(A \Delta A_\epsilon) = \theta_{-1}(A) \Delta \theta_{-1}(A_\epsilon) = A \Delta \theta_{-1}(A_\epsilon),$$

and \mathbb{P} is θ -invariant, it follows that

$$\mathbb{P}(A \Delta \theta_{-k} A_\epsilon) < \epsilon \tag{8.3.4}$$

for all $k \geq 0$. In particular, observe that $\theta^{-(N+k)} A_\epsilon \in \mathcal{F}_k^{2N+k} \subset \mathcal{F}_k^\infty$ for fixed $k \geq 0$ and arbitrary $\epsilon > 0$. Therefore, with ϵ and N as previously given, fix k and let $\epsilon_m = \frac{\epsilon}{2^m}$. Furthermore, let

$$D_n^\epsilon := \theta^{-(N+k)} A_{\frac{\epsilon}{2^n}} \in \mathcal{F}_k^\infty$$

and

$$D := \bigcap_{m \geq 1} \bigcup_{n \geq 1} D_n^{\epsilon_m} \in \mathcal{F}_k^\infty.$$

Note that using (8.3.4) we have $\mathbb{P}(A \Delta D_n^\epsilon) < \frac{\epsilon}{2^n}$ and so

$$\mathbb{P}\left(\bigcap_{n \geq 1} (A \setminus D_n^{\epsilon_m})\right) = \lim_{n \rightarrow \infty} \mathbb{P}(A \setminus D_n^{\epsilon_m}) \leq \lim_{n \rightarrow \infty} \frac{\epsilon_m}{2^n} = 0. \tag{8.3.5}$$

On the other hand,

$$\begin{aligned}
\mathbb{P}(D \setminus A) &\leq \mathbb{P}\left(\bigcup_{n=1}^{\infty} D_n^{\epsilon_m} \setminus A\right) \\
&= \mathbb{P}\left(\bigcap_{n=1}^{\infty} (A \setminus D_n^{\epsilon_m})\right) \\
&\leq \frac{\epsilon}{2^n} \\
&\leq \frac{1}{m} \\
&\xrightarrow{m \rightarrow \infty} 0.
\end{aligned}$$

Therefore, $\mathbb{P}(D \setminus A) = 0$. Furthermore,

$$\begin{aligned}
\mathbb{P}(A \setminus D) &= \mathbb{P}\left(A \setminus \bigcap_{m \geq 1} \bigcup_{n \geq 1} D_n^{\epsilon_m}\right) \\
&= \mathbb{P}\left(\bigcup_{m \geq 1} \bigcap_{n \geq 1} (A \setminus D_n^{\epsilon_m})\right) \\
&\stackrel{(8.3.5)}{=} 0
\end{aligned}$$

and so $A \sim D$. Therefore, for any k with $D^{(k)} := D$ we have $D^{(k)} \in \mathcal{F}_k^{\infty}$ and $\mathbb{P}(A \triangle D^{(k)}) = 0$. Similarly using θ_{-1} for any k there is a $D^{(-k)} \in \mathcal{F}_{-\infty}^{(-k)}$ such that $\mathbb{P}(A \triangle D^{(-k)}) = 0$. Let $l \in (-k, k)$ then

$$\begin{aligned}
\mathbb{E}(\mathbf{1}_A | \sigma(X_l)) &= \mathbb{E}(\mathbf{1}_A^2 | \sigma(X_l)) \\
&= \mathbb{E}(\mathbf{1}_{D^{(-k)}} \mathbf{1}_{D^{(k)}} | \sigma(X_l)) \\
&= \mathbb{E}(\mathbf{1}_{D^{(-k)}} | \sigma(X_l)) \mathbb{E}(\mathbf{1}_{D^{(k)}} | \sigma(X_l)) \\
&= (\mathbb{E}(\mathbf{1}_A | \sigma(X_l)))^2,
\end{aligned}$$

which implies that $\mathbb{E}(\mathbf{1}_A | \sigma(X_l))(\omega) \in \{0, 1\}$ almost surely. Let

$$\hat{A} := \{\omega \in \mathcal{X}^{\mathbb{Z}} : \mathbb{E}(\mathbf{1}_A | \sigma(X_l))(\omega) = 1\} \in \sigma(X_l).$$

Then $\mathbb{E}(\mathbf{1}_A | \sigma(X_l)) = \mathbf{1}_{\hat{A}}$ and $\mathbb{E}(\mathbf{1}_{A^c} | \sigma(X_l)) = \mathbf{1}_{\hat{A}^c}$. Then for any $E \in \sigma(X_l)$ we have

$$\mathbb{P}(A \cap E) = \mathbb{E}(\mathbb{E}(\mathbf{1}_A | \sigma(X_l))) = \mathbb{P}(\hat{A} \cap E).$$

In particular, $\mathbb{P}(A \cap \hat{A}^c) = 0$ and $\mathbb{P}(\hat{A} \cap A^c) = 0$. Hence, $\mathbb{P}(A \triangle \hat{A}) = 0$. Thus letting $\hat{A}_l := \hat{A}$ we have $A \sim \hat{A}_l$. \square

Proposition 8.3.9. For any $A \in \mathcal{I} := \{C \in \mathcal{B}(\mathcal{X}^{\mathbb{Z}}) : \theta^{-1}C = C\}$, there exists $\bar{A} \in \mathcal{B}(\mathcal{X})$ such that

$$A \sim \prod_{i \in \mathbb{Z}} \bar{A}.$$

Proof. By Lemma 8.3.8, there exists an $\hat{A} \in \sigma(X_0)$ with $A \sim \hat{A}$. Let $\bar{A} \in \mathcal{B}(\mathcal{X})$ be such that

$$\hat{A} := \{\omega \in \mathcal{X}^{\mathbb{Z}} : X_0(\omega) \in \bar{A}\}.$$

Then by the invariance of A with follows that

$$\mathbb{P} \left(A \triangle \theta_{-n} \hat{A} \right) = \mathbb{P} \left(\theta_{-n} \left(A \triangle \hat{A} \right) \right) = 0,$$

which implies that

$$\mathbb{P} \left(\bigcup_{n \in \mathbb{Z}} A \triangle \theta^{-n} \hat{A} \right) = 0.$$

As $\theta_{-n} \hat{A} = \{\omega : X_n(\omega) \in \bar{A}\}$ we have that

$$\bigcap_{k=0}^n \theta_{-k} \hat{A} = \{\omega : X_0(\omega) \in \bar{A}, \dots, X_n(\omega) \in \bar{A}\} \sim A.$$

Thus,

$$\{X_i \in \bar{A} : i \in \mathbb{Z}\} = \prod_{i \in \mathbb{Z}} \bar{A} \sim A.$$

□

Corollary 8.3.10. *Let π be an invariant probability measure for P . Then π is ergodic if and only if every π -invariant set \bar{A} is such that $\pi(\bar{A}) \in \{0, 1\}$.*

Proof. (\Rightarrow). If \bar{A} is π -invariant, then using Exercise 7.6.2 it follows that

$$\mathbb{P}_\pi \left(\prod_{i \in \mathbb{Z}} \bar{A} \right) = \pi(\bar{A}).$$

As we assume \mathbb{P}_π is ergodic and $\prod_{i \in \mathbb{Z}} \bar{A}$ is a θ -invariant set, it follows that $\pi(\bar{A}) \in \{0, 1\}$.

(\Leftarrow). For $A \in \mathcal{I}$ we have from Proposition 8.3.9 that there exists a $\bar{A} \in \mathcal{B}(\mathcal{X})$ such that $A \sim \prod_{i \in \mathbb{Z}} \bar{A}$. □

Proposition 8.3.11. *Let π be an invariant probability measure for P . Then π is ergodic if and only if π is an extremal of I_P , where I_P is as given in (8.3.1).*

Proof. (\Rightarrow). Suppose that π is not extremal, with $\pi = t\pi_1 + (1-t)\pi_2$ for $t \in (0, 1)$ and $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$ distinct. Moreover, suppose that π is ergodic. Then for any θ -invariant set A we have

$$t\mathbb{P}_{\pi_1}(A) + (1-t)\mathbb{P}_{\pi_2}(A) \in \{0, 1\},$$

thus $\mathbb{P}_{\pi_1}(A) = \mathbb{P}_{\pi_2} = 0$ or $\mathbb{P}_{\pi_1}(A) = \mathbb{P}_{\pi_2} = 1$ making π_1 and π_2 ergodic. Therefore, by Proposition 8.2.9 π_1 and π_2 are mutually singular. Let $E \in \mathcal{B}(\mathcal{X})$ be such that $\pi_1(E) = 1$, and $\pi_2(E) = 0$, so that $\mathbb{P}_{\pi_1} \left(\prod_{i \in \mathbb{Z}} E \right) = 1$ and $\mathbb{P}_{\pi_2} \left(\prod_{i \in \mathbb{Z}} E \right) = 0$. It follows that

$$\mathbb{P}_\pi \left(\prod_{i \in \mathbb{Z}} E \right) = t\mathbb{P}_{\pi_1} \left(\prod_{i \in \mathbb{Z}} E \right) + (1-t)\mathbb{P}_{\pi_2} \left(\prod_{i \in \mathbb{Z}} E \right) = t < 1,$$

which contradicts π being ergodic.

(\Leftarrow). Suppose π is not ergodic. Then by Corollary 8.3.10 there exists a π -invariant set F with $\pi(F) := t \in (0, 1)$. Let $\pi_1, \pi_2 \in \mathcal{P}(\mathcal{X})$ be given by

$$\pi_1(B) = \frac{1}{t}\pi(B \cap F),$$

and

$$\pi_2(B) = \frac{1}{1-t}\pi(B \cap F^c).$$

By the π -invariance of F we have $P(x, F) = 1$ for almost every $x \in F$ with respect to π . Using similar arguments to those made in Lemma 7.6.3, we have that π restricted to F is invariant. That is, π_1 is invariant. On the other hand,

$$\pi(F^c) = \int_F P(x, F^c) d\pi + \int_{F^c} P(x, F^c) d\pi = \int_{F^c} P(x, F^c) d\pi,$$

which implies that $P(x, F^c) = 1$ for almost every $x \in F^c$ with respect to π . Therefore, π_2 is invariant, and thus $\pi_1, \pi_2 \in I_P$ which contradicts π being extremal. \square

Theorem 8.3.12. *Given a time-homogeneous transition probability P , with transition operator T let I_P be as given by (8.3.1). Then for*

$$\mathcal{E} = \{\pi \in \mathcal{P}(\mathcal{X}) : \pi \text{ ergodic}\} \cap I_P$$

the following statements hold.

1. I_P is convex with being the set \mathcal{E} of its extremal points.
2. For $\pi_1, \pi_2 \in \mathcal{E}$, either π_1 and π_2 are equal or they are mutually singular.
3. Every $\pi \in I_P$ can be written as $\pi = t\pi_1 + (1-t)\pi_2$ for some $\pi_1, \pi_2 \in \mathcal{E}$ and $t \in [0, 1]$.

Proof.

1. I_P being convex follows from statement 1. of Exercise 8.3.1. Proposition 8.3.11 shows that the set of extremal points of I_P coincides with \mathcal{E} .
2. This is shown in Proposition 8.2.9.
3. This follows from statement 1.

\square

Corollary 8.3.13. *If a time-homogeneous Markov process admits more than one invariant measure, it admits at least two ergodic invariant measures.*

Proof. Let $\pi_1, \pi_2 \in I_P$ be distinct. We are done if $\pi_1, \pi_2 \in \mathcal{E}$. In any other case, we must have that either π_1 or π_2 is not ergodic. Suppose without loss of generality that $\pi_1 \in I_P \setminus \mathcal{E}$. By statement 3. of Theorem 8.3.12 we can write $\pi_1 = t\mu_1 + (1-t)\mu_2$ where $\mu_1, \mu_2 \in \mathcal{E}$. It is clear that μ_1 is not equal to μ_2 , otherwise $\pi_1 = \mu_1$ which contradicts π_1 not being ergodic. \square

Corollary 8.3.14. *If a time-homogeneous Markov process has a unique invariant measure π , then π is ergodic.*

Proof. In this case $I_P = \{\pi\}$ and so π is an extremal point. Therefore, by statement 1. of Theorem 8.3.12 we have that π is ergodic. \square

Proposition 8.3.15. *Let $A \subset \mathcal{X}$ be a P -invariant set. Let $A_0 = A$ and inductively let*

$$A_n = \{x \in \mathcal{X} : P(x, A_{n-1}) > 0\}.$$

Suppose $\mathcal{X} = \bigcup_{n=1}^{\infty} A_n$ and $A = \bigcup_{k=1}^m B_k$ for $(B_k)_{k=1, \dots, m}$ disjoint closed P -invariant sets. If the time-homogeneous Markov chain restricted to B_k has a unique invariant measure π_k , then π_k is ergodic. Moreover, the $(\pi_k)_{k=1, \dots, m}$ are the only ergodic invariant probability measures of the chain.

Proof. On B_k , the ergodicity follows from Corollary 8.3.14. If $\pi \in I_P$, then using Proposition 7.6.7 it follows that $\pi(A) = 1$. Using arguments made in the proof of Proposition 8.3.11 we have that the restriction on π onto the B_k is an invariant probability measure for P . By the uniqueness of invariant probability measures on B_k we can uniquely write

$$\pi = \sum_{k=1}^m \pi(B_k) \pi_k.$$

□

8.4 Solution to Exercises

Exercise 8.1.3

Solution. As $\emptyset \in \mathcal{F}$, and $\theta^{-1}(\emptyset) = \emptyset$ we have $\emptyset \in \mathcal{I}$. Suppose that $A \in \mathcal{I}$, then

$$\theta^{-1}(A^c) = (\theta^{-1}(A))^c = A^c,$$

and so $A^c \in \mathcal{I}$. Now consider $(A_n)_{n \in \mathbb{N}} \subset \mathcal{I}$, then

$$\theta^{-1}\left(\bigcup_{n \in \mathbb{N}} A_n\right) = \bigcup_{n \in \mathbb{N}} \theta^{-1}(A_n) = \bigcup_{n \in \mathbb{N}} A_n,$$

and so $\bigcup_{n \in \mathbb{N}} A_n \in \mathcal{I}$. Therefore, \mathcal{I} forms a σ -algebra. □

Exercise 8.1.5

Solution. Let $f = \mathbf{1}_A$. Note that

$$f = \mathbf{1}_A = \mathbf{1}_{\omega: \omega \in A}$$

and

$$f \circ \theta = \mathbf{1}_{\omega: \theta\omega \in A} = \mathbf{1}_{\theta^{-1}(A)}.$$

Therefore, $f \circ \theta = f$ if and only if $A = \theta^{-1}(A)$ which is to say that $A \in \mathcal{I}$. Therefore, $f \circ \theta = f$ if and only if f is \mathcal{I} -measurable. Extending the above argument by linearity it follows that $f = \sum_{i=1}^n a_i \mathbf{1}_{A_i}$ is θ -invariant if and only if f is measurable with respect to θ . As measurability is preserved under limits, and any non-negative function is the limit of a sequence of simple functions it follows that $f : \Omega \rightarrow [0, \infty]$ is θ -invariant if and only if f is \mathcal{I} -measurable. Then for general $f : \Omega \rightarrow [0, \infty]$, using the decomposition $f = f^+ - f^-$, we conclude that $f : \Omega \rightarrow \mathbb{R}$ is θ -invariant if and only if f is \mathcal{I} -measurable. □

Exercise 8.3.1

Solution.

1. This follows directly from the linearity of the transition operator T .
2. T being Feller means that it is a continuous map from $\mathcal{P}(\mathcal{X})$ to $\mathcal{P}(\mathcal{X})$ under the weak topology. Therefore, for $(\pi_n)_{n \in \mathbb{N}} \subset I_P$ a sequence converging weakly to a limit π we have that

$$T\pi = T \lim_{n \rightarrow \infty} \pi_n = \lim_{n \rightarrow \infty} T\pi_n = \lim_{n \rightarrow \infty} \pi_n = \pi.$$

Therefore, $\pi \in I_P$ which means that I_P is closed. □

9 Appendix

9.1 Gaussian Measures

Definition 9.1.1. A measure μ on \mathbb{R}^n is Gaussian if there exists a non-negative symmetric matrix K and vector $m \in \mathbb{R}^n$ such that

$$\int_{\mathbb{R}^n} e^{i\langle \lambda, x \rangle} \mu(dx) = \exp\left(i\langle \lambda, m \rangle - \frac{1}{2}\langle K\lambda, \lambda \rangle\right).$$

If K is non-degenerate then the density with respect to the Lebesgue measure is given by

$$\frac{1}{\sqrt{(2\pi)^n \det(K)}} \exp\left(-\frac{1}{2}\langle K^{-1}(x - m), x - m \rangle\right).$$

In this case, m is called the mean, and K the covariance operator.

Remark 9.1.2.

1. We are using the notation

$$\langle u, v \rangle = u^\top v.$$

2. A Gaussian measure is specified entirely by its mean and covariance operator.

Theorem 9.1.3. Let X be a Gaussian random variable on \mathbb{R}^n with mean m , covariance operator K and let $A : \mathbb{R}^d \rightarrow \mathbb{R}^d$ be a linear map. Then AX is a Gaussian random variable with covariance operator AKA^\top .

Proof. For $\lambda \in \mathbb{R}^n$ we have

$$\begin{aligned} \mathbb{E}(\exp(i\langle \lambda, AX \rangle)) &= \mathbb{E}(\exp(i\langle A^\top \lambda, X \rangle)) \\ &\stackrel{(1)}{=} \exp\left(i\langle \lambda, Am \rangle - \frac{1}{2}\langle KA^\top \lambda, A^\top \lambda \rangle\right) \\ &= \exp\left(i\langle \lambda, Am \rangle - \frac{1}{2}\langle AKA^\top \lambda, \lambda \rangle\right) \end{aligned}$$

where (1) follows as X is a Gaussian random variable with covariance operator K and mean m . Therefore, as a Gaussian is determined by its mean and covariance, we deduce that AX is a Gaussian random variable with mean Am and covariance AKA^\top . \square

Proposition 9.1.4. If $(X_k)_{k=1}^n$ are independent Gaussian random variables on \mathbb{R}^d then for $a_k \in \mathbb{R}$ the random variable $\sum_{k=1}^n a_k X_k$ is also Gaussian.

Proof. Suppose that X_k has mean m_k and covariance operator K_k . Then for $\lambda \in \mathbb{R}^n$ observe that

$$\begin{aligned}
\mathbb{E} \left(\exp \left(k \left\langle \lambda, \sum_{k=1}^n a_k X_k \right\rangle \right) \right) &= \mathbb{E} \left(\exp \left(k \sum_{k=1}^n \langle a_k \lambda, X_k \rangle \right) \right) \\
&= \mathbb{E} \left(\prod_{k=1}^n \exp (k \langle a_k \lambda, X_k \rangle) \right) \\
&\stackrel{(1)}{=} \prod_{k=1}^n \mathbb{E} (\exp (k \langle a_k \lambda, X_k \rangle)) \\
&= \prod_{k=1}^n \exp \left(k \langle a_k \lambda, m_k \rangle - \frac{1}{2} \langle K_k (a_k \lambda), a_k \lambda \rangle \right) \\
&= \exp \left(\sum_{k=1}^n k \langle \lambda, a_k m_k \rangle - \frac{1}{2} \langle (a_k^2 K_k) \lambda, \lambda \rangle \right) \\
&= \exp \left(k \left\langle \lambda, \sum_{k=1}^n a_k m_k \right\rangle - \frac{1}{2} \left\langle \sum_{k=1}^n a_k^2 K_k \lambda, \lambda \right\rangle \right).
\end{aligned}$$

Where (1) is justified by the independence of the X_k . Hence, $\sum_{i=k=1}^n a_k X_k$ is a Gaussian measure with mean $\sum_{k=1}^n a_k m_k$ and covariance operator $\sum_{k=1}^n a_k^2 K_k$. \square

9.2 The Doeblin Coupling

Definition 9.2.1. For random variables X and Y with state space \mathcal{X} , a coupling is a random variable $Z = (X', Y')$ with state space $\mathcal{X} \times \mathcal{X}$ such that

1. $\text{Law}(X) = \text{Law}(X')$, and
2. $\text{Law}(Y) = \text{Law}(Y')$.

Remark 9.2.2. Note that for a given X and Y there can be lots of different couplings. This is because $\text{Law}(X)$ and $\text{Law}(Y)$ do not determine $\text{Law}(Z)$, but only determine its marginals.

Definition 9.2.3. Consider independent Markov chains $(X_n)_{n \in \mathbb{N}}$ and $(X'_n)_{n \in \mathbb{N}}$ on a discrete state space \mathcal{X} , with transition probabilities P , and initial distributions μ and ν respectively. The process $(Z_n)_{n \in \mathbb{N}}$ given by $Z_n = (X_n, X'_n)$ is known as the Doeblin coupling. Moreover, $T := \inf \{n \in \mathbb{N} : X_n = X'_n\}$ is known as the coalescing time of the chains $(X_n)_{n \in \mathbb{N}}$ and $(X'_n)_{n \in \mathbb{N}}$.

Exercise 9.2.4. Let $(\mathcal{G}_n)_{n \in \mathbb{N}}$ and $(\mathcal{G}'_n)_{n \in \mathbb{N}}$ be filtrations of independent σ -algebras. Let $(X_n)_{n \in \mathbb{N}}$ be a time-homogeneous Markov process with respect to the filtration $(\mathcal{G}_n)_{n \in \mathbb{N}}$. Then, $(X_n)_{n \in \mathbb{N}}$ is a time-homogeneous Markov process with respect to the filtration $(\mathcal{G}_n \vee \mathcal{G}'_n)_{n \in \mathbb{N}}$, where $\mathcal{G}_n \vee \mathcal{G}'_n$ is the σ -algebra generated by $\mathcal{G}_n \cup \mathcal{G}'_n$.

Lemma 9.2.5. *Assume the setup of Definition 9.2.3 and let*

$$Y_n = \begin{cases} X_n & n < T \\ X'_n & n \geq T. \end{cases}$$

Then $(Y_n)_{n \in \mathbb{N}}$ is a Markov process with initial distribution μ and transition probabilities P . In particular,

$\text{Law}(Y) = \text{Law}(X)$.

Proof. Let $(\mathcal{F}_n)_{n \in \mathbb{N}}$ and $(\mathcal{G}_n)_{n \in \mathbb{N}}$ be the natural filtrations of $(X_n)_{n \in \mathbb{N}}$ and $(X'_n)_{n \in \mathbb{N}}$ respectively. Then let $\mathcal{F}_n = \mathcal{F}_n^0 \vee \mathcal{G}_n^0$. For $f \in \mathcal{B}_b(\mathcal{X})$ it follows that

$$\begin{aligned} \mathbb{E}(f(Y_{n+1})|\mathcal{F}_n) &= \mathbb{E}(f(Y_{n+1})\mathbf{1}_{\{T \leq n\}}|\mathcal{F}_n) + \mathbb{E}(f(Y_{n+1})\mathbf{1}_{\{T > n\}}|\mathcal{F}_n) \\ &= \mathbf{1}_{\{T \leq n\}}\mathbb{E}(f(X'_{n+1})|\mathcal{F}_n) + \mathbf{1}_{\{T > n\}}\mathbb{E}(f(X_{n+1})|\mathcal{F}_n) \\ &= \mathbf{1}_{\{T \leq n\}}Pf(X'_n) + \mathbf{1}_{\{T > n\}}Pf(X_n) \\ &= \mathbf{1}_{\{T \leq n\}}Pf(Y_n) + \mathbf{1}_{\{T > n\}}Pf(Y_n) \\ &= Pf(Y_n). \end{aligned}$$

Where we have implicitly used the result of Exercise 9.2.4. Furthermore, we have used the fact that T is a $(\mathcal{F}_n)_{n \in \mathbb{N}}$ -stopping time, and hence $\{T \leq n\}$ and $\{T > n\}$ are \mathcal{F}_n -measurable. Hence, $(Y_n)_{n \in \mathbb{N}}$ is a Markov process with transition probability P . Moreover, it is clear that $Y_0 = X_0$, as if $T = 0$ then $Y_0 = X'_0 = X_0$ by the definition of T . So that $\text{Law}(Y_0) = \text{Law}(X_0) = \mu$, which implies that $\text{Law}(Y) = \text{Law}(X)$ as $(Y_n)_{n \in \mathbb{N}}$ and $(X_n)_{n \in \mathbb{N}}$ have the same transition probability. \square

Lemma 9.2.6. Assume the setup of Definition 9.2.3, then

$$\sum_{j \in \mathcal{X}} |\mathbb{P}(X_n = j) - \mathbb{P}(X'_n = j)| \leq 2\mathbb{P}(T > n).$$

Proof. Let $j \in \mathcal{X}$ then

$$\begin{aligned} |\mathbb{P}(X_n = j) - \mathbb{P}(X'_n = j)| &\stackrel{(1)}{=} |\mathbb{P}(Y_n = j) - \mathbb{P}(X'_n = j)| \\ &= |\mathbb{P}(Y_n = j) - \mathbb{P}(X'_n = j, T > n) - \mathbb{P}(Y_n = j, T \leq n)| \\ &\stackrel{(2)}{=} \mathbb{P}(Y_n = j, T > n) - \mathbb{P}(Y'_n = j, T > n). \end{aligned}$$

Where (1) follows by the result of Lemma 9.2.5 and (2) is an application of the law of total probability. Therefore,

$$\begin{aligned} \sum_{j \in \mathcal{X}} |\mathbb{P}(X_n = j) - \mathbb{P}(X'_n = j)| &\leq \sum_{j \in \mathcal{X}} \mathbb{P}(Y_n = j, T > n) + \mathbb{P}(X'_n = j, T > n) \\ &\leq 2\mathbb{P}(T > n). \end{aligned}$$

\square

Exercise 9.2.7. Let $(X_n)_{n \in \mathbb{N}}$ and $(X'_n)_{n \in \mathbb{N}}$ be independent Markov chains on a discrete state space \mathcal{X} and transition probabilities P . Let $\mathcal{G}_n = \sigma(X_k : k \leq n)$, $\mathcal{G}'_n = \sigma(X'_k : k \leq n)$ and $\mathcal{F}_n = \mathcal{G}_n \vee \mathcal{G}'_n$. Show that

$$\mathbb{P}(X_{n+1} = j, X'_{n+1} = j' | \mathcal{F}_n) = \mathbb{P}(X_{n+1} = j | X_n) \cdot \mathbb{P}(X'_{n+1} = j' | X'_n). \quad (9.2.1)$$

Lemma 9.2.8. Assume the setup of Definition 9.2.3, then the Doeblin coupling Z_n is a time-homogeneous Markov chain on $\mathcal{X} \times \mathcal{X}$ with transition probabilities Q and initial distribution $\mu \otimes \nu$, where

$$Q_{(i,i'),(j,j')} = P_{ij}P_{i'j'}$$

for all $i, i', j, j' \in \mathcal{X}$.

Proof. By the independence of X_0 and X'_0 it follows immediately that $\text{Law}(X_0, X'_0) = \mu \otimes \nu$. With $(\mathcal{F}_n)_{n \in \mathbb{N}}$ as defined in the proof of Lemma 9.2.5, for $j, j' \in \mathcal{X}$ it follows that

$$\begin{aligned} \mathbb{P}(Z_{n+1} = (j, j') | \mathcal{F}_n) &= \mathbb{P}(X_{n+1} = j, X'_{n+1} = j' | \mathcal{F}_n) \\ &\stackrel{\text{Ex 9.2.7}}{=} \mathbb{P}(X_{n+1} = j | X_n) \cdot \mathbb{P}(X'_{n+1} = j' | X'_n) \\ &= P_{X_n, j} P_{X'_n, j'} \\ &=: Q_{Z_n, (j, j')}. \end{aligned}$$

Therefore, $\mathbb{P}(Z_{n+1} = (j, j') | \mathcal{F}_n) = \mathbb{P}(Z_{n+1} = (j, j') | Z_n)$ and so $(Z_n)_{n \in \mathbb{N}}$ is a time-homogeneous Markov process with transition probability Q . \square

Lemma 9.2.9. *Suppose that $S \subseteq \mathbb{N}$ is a non-empty set with the property that for $s, s' \in S$ it follows that $s + s' \in S$. Then letting $d = \gcd(S)$, there must exist a $K > 0$ such that for any $k \geq K$ one has $kd \in S$.*

Proof. By considering $S' = \{\frac{s}{d} : s \in S\}$ we can assume without loss of generality that $d = 1$. Let $\{d_1, \dots, d_n\} \subseteq S$ be such that $\gcd(\{d_1, \dots, d_n\}) = 1$. By Bezout's identity we know that there exist integers a_1, \dots, a_n such that

$$\sum_{i=1}^n a_i d_i = 1.$$

Let $M = \sum_{i=1}^n d_i$. Then for $l = 0, \dots, M-1$ it follows that

$$NM + l = \sum_{i=1}^n (N + la_i) d_i.$$

As $k \leq M$ we can choose N_0 such that $N_0 + la_i \geq 0$. Therefore, the sum on the right-hand side can be thought of as summing $N + la_i$ copies of d_i for each i . Hence, by the additive property of S it follows that $NM + l \in S$ for all $l = 0, \dots, M-1$ and $N \geq N_0$. Letting $l = 0$ we deduce that $kd = k \in S$ for every $k \geq N_0M$. \square

Lemma 9.2.10. *Suppose i is aperiodic and recurrent, then there exists an $N \in \mathbb{N}$ such that $P_{ii}^n > 0$ for every $n > N$.*

Proof. As i is recurrent we know that $R(i) \neq \emptyset$. Moreover, if $n_1 \in R(i)$ and $n_2 \in R(i)$ then $P_{ii}^{n_1} > 0$ and $P_{ii}^{n_2} > 0$ and so $P_{ii}^{n_1+n_2} > 0$ by the Chapman-Kolmogorov equation. This implies that $n_1 + n_2 \in R(i)$ and so $R(i)$ has the additive property. Consequently, as i is aperiodic we have that $\gcd(R(i)) = 1$ and so by Lemma 9.2.9 that there exists an $N \in \mathbb{N}$ such that $n \in R(i)$ for all $n \geq N$. In other words, $P_{ii}^n > 0$ for $n \geq N$. \square

Lemma 9.2.11. *If P is irreducible, aperiodic, and positive recurrent, then Q , as defined in Lemma 9.2.8, is irreducible and positive recurrent.*

Proof. For any $i \in \mathcal{X}$ we know by Lemma 9.2.10 that there exists an N such that $P_{ii}^n > 0$ for any $n \geq N$. Thus by the irreducibility of P we know that for any $j \in \mathcal{X}$ there exists an m such that $P_{ij}^m > 0$. Consequently,

$$P_{ij}^{n+m} \geq P_{ii}^n P_{ij}^m > 0$$

for all $n \geq N$. More succinctly, we can say that $P_{ij}^n > 0$ for sufficiently large n . Therefore, for $(i, j), (i', j') \in \mathcal{X} \times \mathcal{X}$ there exists some $N \in \mathbb{N}$ such that $P_{ij}^n > 0$ and $P_{i'j'}^n > 0$ which implies that

$$Q_{(i, i'), (j, j')}^n = P_{ij}^n P_{i'j'}^n > 0$$

for all $n \geq N$. Which proves that Q is irreducible. Consequently, with π being the invariant measure of P we know that $\pi \otimes \pi$ is an invariant measure of Q . Having established that Q is irreducible this implies that Q is positive recurrent. \square

Lemma 9.2.12. Assume the setup of Definition 9.2.3 and in particular that P is irreducible, aperiodic and positive recurrent. Then,

$$\mathbb{P}(T < \infty) = 1.$$

Proof. Let

$$T_{(i,i')} := \inf \{n \geq 1 : Z_n(X_n, X'_n) = (i, i')\}.$$

As Q is irreducible we know that

$$\mathbb{P}(T_{(i,i')} < \infty) = 1$$

for all $(i, i') \in \mathcal{X} \times \mathcal{X}$. In particular, letting $i = i'$ it is clear that $T \leq T_{(i,i)}$, hence, $\mathbb{P}(T < \infty) \geq \mathbb{P}(T_{(i,i)} < \infty) = 1$. \square

9.3 Solution to Exercises

Exercise 9.2.4

Solution. Consider the set $\mathcal{D} = \{A \cap B : A \in \mathcal{G}_n, B \in \mathcal{G}'_n\}$. Note \mathcal{D} is a π -system. Moreover, $\mathcal{G} := \mathcal{G}_n \vee \mathcal{G}'_n$ forms a λ -system with $\mathcal{D} \subseteq \mathcal{G}$. So by Theorem 2.2.8 we know that $\sigma(\mathcal{D}) \subseteq \mathcal{G}$. However, as $\mathcal{G}_n \subseteq \mathcal{D}$ and $\mathcal{G}'_n \subseteq \mathcal{D}$, by taking $A = \mathcal{X}$ and $B = \mathcal{X}$ respectively, we know that $\mathcal{G}_n \cup \mathcal{G}'_n \subseteq \sigma(\mathcal{D})$ which implies that $\mathcal{G} = \sigma(\mathcal{G}_n \cup \mathcal{G}'_n) \subseteq \sigma(\mathcal{D})$. Therefore, $\sigma(\mathcal{D}) = \mathcal{G}$. Now for $f \in \mathcal{B}_b(\mathcal{X})$ and $n \in \mathbb{N}$ we know that $\mathbb{E}(f(X_{n+1})|\mathcal{G}_n)$ is \mathcal{G} measurable. Moreover, for $A \cap B \in \mathcal{D}$ it follows that

$$\begin{aligned} \int_{A \cap B} \mathbb{E}(f(X_{n+1})|\mathcal{G}_n) d\mathbb{P} &= \int \mathbb{E}(f(X_{n+1})\mathbf{1}_{A \cap B}|\mathcal{G}_n) d\mathbb{P} \\ &\stackrel{(1)}{=} \int \mathbb{E}(f(X_{n+1})\mathbf{1}_{A \cap B}) d\mathbb{P} \\ &= \int_{A \cap B} f(X_{n+1}) d\mathbb{P}, \end{aligned}$$

where in (1) we have used the fact that $f(X_{n+1})\mathbf{1}_{A \cap B}$ is independent of \mathcal{G}_n . Therefore, as $\sigma(\mathcal{D}) = \mathcal{G}$ we deduce that

$$\int_C \mathbb{E}(f(X_{n+1})|\mathcal{G}_n) d\mathbb{P} = \int_C f(X_{n+1}) d\mathbb{P}$$

for all $C \in \mathcal{G}$. Hence, $\mathbb{E}(f(X_{n+1})|\mathcal{G}_n \vee \mathcal{G}'_n) = \mathbb{E}(f(X_{n+1})|\mathcal{G}_n)$. Therefore, as $(X_n)_{n \in \mathbb{N}}$ is Markov with respect to \mathcal{G}_n we conclude that

$$\begin{aligned} \mathbb{E}(f(X_{n+1})|\mathcal{G}_n \vee \mathcal{G}'_n) &= \mathbb{E}(f(X_{n+1})|\mathcal{G}_n) \\ &= \mathbb{E}(f(X_{n+1})|X_n) \end{aligned}$$

and so $(X_n)_{n \in \mathbb{N}}$ is Markov with respect to $\mathcal{G}_n \vee \mathcal{G}'_n$. \square

Exercise 9.2.7

Solution. Let $\mathcal{G}_n = \sigma(X_k : k \leq n)$, $\mathcal{G}'_n = \sigma(X'_k : k \leq n)$ so that $\mathcal{F}_n = \mathcal{G}_n \vee \mathcal{G}'_n$. The right-hand side of (9.2.1) is equal to $\mathbb{P}(X_{n+1} = j|\mathcal{G}_n) \cdot \mathbb{P}(X'_{n+1} = j'|\mathcal{G}'_n)$, by the Markov property, and so is \mathcal{F}_n -measurable. Recall, that \mathcal{F}_n is generated by sets of the form $G_1 \cap G_2$ where $G_1 \in \mathcal{G}_n$ and $G_2 \in \mathcal{G}'_n$. Using the independence of \mathcal{G}_n and \mathcal{G}'_n it follows that,

$$\begin{aligned} \mathbb{E}(\mathbb{P}(X_{n+1} = j|\mathcal{G}_n) \cdot \mathbb{P}(X'_{n+1} = j'|\mathcal{G}'_n) \mathbf{1}_{G_1 \cap G_2}) &= \mathbb{E}\left(\mathbb{E}(\mathbf{1}_{\{X_{n+1}=j\}}\mathbf{1}_{G_1}|\mathcal{G}_n) \mathbb{E}(\mathbf{1}_{\{X'_{n+1}=j'\}}\mathbf{1}_{G_2}|\mathcal{G}'_n)\right) \\ &= \mathbb{E}\left(\mathbb{E}(\mathbf{1}_{\{X_{n+1}=j\}}\mathbf{1}_{G_1}|\mathcal{G}_n)\right) \mathbb{E}\left(\mathbb{E}(\mathbf{1}_{\{X'_{n+1}=j'\}}\mathbf{1}_{G_2}|\mathcal{G}'_n)\right) \\ &= \mathbb{E}(\mathbf{1}_{\{X_{n+1}=j\}}\mathbf{1}_{G_1}) \mathbb{E}(\mathbf{1}_{\{X'_{n+1}=j'\}}\mathbf{1}_{G_2}) \\ &= \mathbb{E}(\mathbf{1}_{\{X_{n+1}=j\} \cap \{X'_{n+1}=j'\}}\mathbf{1}_{G_1 \cap G_2}) \\ &= \mathbb{E}(\mathbb{P}(X_{n+1} = j, X'_{n+1} = j') \mathbf{1}_{G_1 \cap G_2}). \end{aligned}$$

It follows that

$$\begin{aligned}\mathbb{P}(X_{n+1} = j, X'_{n+1} = j') &= \mathbb{P}(X_{n+1} = j | \mathcal{G}_n) \cdot \mathbb{P}(X'_{n+1} = j' | \mathcal{G}'_n) \\ &= \mathbb{P}(X_{n+1} = j | X_n) \cdot \mathbb{P}(X'_{n+1} = j' | X'_n).\end{aligned}$$

□