Some topics in ergodic theory

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CHAPTER 1

Introduction

There is no universally accepted concise definition of ergodic theory. I am going to adopt the view that its main purpose is to study statistical patterns in deterministic or random dynamical systems and, specifically, how these patterns depend on the initial state of the system. Let me first try to explain this without going to technical details.

Suppose that we have a system $X$ evolving in a phase space $\mathbb{X}$. This means that at each time $t$ the state of the system denoted by $X_t$ belongs to $\mathbb{X}$. The system may be rather complex, e.g., $X$ can be a turbulent flow of particles in a fluid or gas and at any given time the description of the state of the flow should involve positions and velocities of all these particles. All these details of the flow at any given time have to be encoded as an element of $\mathbb{X}$.

Studying such a system experimentally may be a difficult task because if the behavior of the system is rich, then tracking all the details of the evolution may be hard or impossible. In many cases predicting the trajectory of such a system may be problematic as well, sometimes due to analytic or computational difficulties, sometimes due to intrinsic randomness or lack of stability in the system and due to its sensitivity to the initial data.

However, one may try a statistical approach to describing the properties of the system. Although concrete realizations $(X_t)_{t \geq 0}$ of the system may be erratic, it is still possible that these realizations follow steady statistical patterns.

A usual approach to experimental analysis of the system is to collect statistical information about the system by making repeated observations or measurements. From the mathematical point of view, a measurement is a way to assign to any state $x \in \mathbb{X}$ a number $f(x)$. In other words, it is a function $f : \mathbb{X} \to \mathbb{R}$. In particular, a measurement $f$ of system $X$ at time $t$, produces a number $f(X_t)$. One such measurement hardly tells much about the system, so a natural way to proceed is to conduct the same measurement at different times and find the average of the obtained values. If the measurements are made at times $1, 2, \ldots$, then the result of this procedure is

$$\frac{f(X_1) + \ldots + f(X_n)}{n}.$$ 

The hope is that these time-averaged measurements can serve as approximations to a “true” value one is trying to estimate. Moreover, one hopes that
the more observations we average, the closer the resulting estimate is to the ideal true value. In other words, collecting more and more statistics is useful only if a version of the law of large numbers holds, namely, that there is a number \( \bar{f} \) such that \( \frac{f(X_1) + \ldots + f(X_n)}{n} \to \bar{f} \) as \( n \to \infty \). Moreover, that infinite time horizon average should be independent of the specific initial state of the system or at least be stable with respect to a class of perturbations of the initial state, otherwise it is hard to assign a meaning to this limiting value. So, some questions that naturally arise are: Does the system possess any statistical regularity, i.e., does it make sense to collect statistics about the system? How do the statistical properties of the system depend on the initial condition? Does the system tend to remember the initial condition or does it forget it in the long run? What is remembered and what is forgotten in the long run?

One way to look at these issues is the stability point of view. I would like to illustrate that with a couple of standard mathematical examples.

The first very simple example is a deterministic linear dynamical system with one stable fixed point. A discrete dynamical system is given by a transformation \( \theta \) of a phase space \( \mathbb{X} \). For our example, we take the phase space \( \mathbb{X} \) to be the real line \( \mathbb{R} \) and define the transformation \( \theta \) by \( \theta(x) = ax \), \( x \in \mathbb{R} \), where \( a \) is a real number between 0 and 1. To any point \( x \in \mathbb{X} \) one can associate its forward orbit \( (X_n)_{n=0}^{\infty} \), a sequence of points obtained from \( X_0 = x \) by iterations of the map \( \theta \), i.e., \( X_n = \theta(X_{n-1}) \) for all \( n \in \mathbb{N} \):

\[
X_0 = x = \theta^0(x) \\
X_1 = \theta(X_0) = \theta(x) = \theta^1(x), \\
X_2 = \theta(X_1) = \theta \circ \theta(x) = \theta^2(x), \\
X_3 = \theta(X_2) = \theta \circ \theta \circ \theta(x) = \theta^3(x), \\
\ldots.
\]

We are interested in the behavior of the forward orbit \( (X_n)_{n=0}^{\infty} \) of \( x \) as \( n \to \infty \), where \( n \) plays the role of time. In this simple example, the analysis is straightforward. Namely, zero is a unique fixed point of the transformation: \( \theta(0) = 0 \), and since \( X_n = a^n x \), \( n \in \mathbb{N} \) and \( a \in (0, 1) \) we conclude that as \( n \to \infty \), \( X_n \) converges to that fixed point exponentially fast. Therefore, 0 is a stable fixed point, or a one-point global attractor for the dynamical system defined by \( \theta \), i.e., its domain of attraction coincides with \( \mathbb{R} \). So, due to the contraction and intrinsic stability that is present in the map \( \theta \), there is a fast loss of memory in the system, and no matter what the initial condition is, it gets forgotten in the long run and the points \( X_n = \theta^n(x) \) approach the stable fixed point 0 as \( n \to \infty \).

A completely different example is the following: let \( \mathbb{X} = [0, 1) \) and let \( \theta \) be defined by \( \theta(x) = \{2x\} \). This system exhibits no stability at all. To see that it is convenient to look at this system using binary representations of numbers in \( [0, 1) \): for each \( x \in [0, 1) \) there is a sequence \( (x_i)_{i=1}^{\infty} \) of numbers
$x_i \in \{0, 1\}$ such that $x = 0.x_1x_2x_3\ldots$ in binary notation, i.e.,

$$x = \sum_{i=1}^{\infty} \frac{x_i}{2^i}.$$  

This representation is unique for all $x$ except dyadic numbers. The transformation $\theta$ acts on binary representation as a shift:

$$\theta(0.x_1x_2x_3\ldots) = 0.x_2x_3x_4\ldots.$$  

Therefore,

$$\theta^n(0.x_1x_2x_3\ldots) = 0.x_{1+n}x_{2+n}x_{3+n}\ldots.$$  

To see that there is no stability in this system, let us start with any point $x = 0.x_1x_2x_3\ldots \in [0, 1)$. Let us now take another point $y = 0.y_1y_2y_3\ldots \in [0, 1)$. If $y_n \neq x_n$ for infinitely many values of $n$, then for those values of $n$, the iterates of $\theta^n(x)$ and $\theta^n(y)$ will not be close, one belonging to $(0, 1/2)$ and another to $[1/2, 1)$. (in fact, one can come up with exceptions related to dyadic numbers, but it is clear that they are quite exotic).

So stability at the level of trajectories does not hold in the usual sense. However a statistical version of stability holds true. Namely, one can prove that if $f$ is an integrable Borel function on $[0, 1)$, then for almost every $x \in [0, 1)$,

$$\frac{1}{n} \left( f(x) + f(\theta(x)) + f(\theta^2(x)) + \ldots + f(\theta^{n-1}(x)) \right) \to \int_{[0, 1)} f(x) \, dx. \quad (1.1)$$

So, one can say that in the long run statistical properties of this dynamical system are described by the Lebesgue measure or uniform distribution on $[0, 1)$. In fact, one interpretation of identity (1.1) is that for Lebesgue almost every initial condition $x$ the empirical distribution or empirical measure generated by $\theta$ and assigning mass $n^{-1}$ to each point $x, \theta(x), \ldots, \theta^{n-1}(x)$ converges to the Lebesgue measure. Since the limiting measure is the same for a broad class of initial conditions $x$, we can speak of statistical stability: in the long run the initial value gets forgotten in the statistical properties of the system. We also can interpret the convergence in (1.1) as a law of large numbers.

We see that the Lebesgue measure plays a special role for this system since it describes the limits in (1.1). The reason for this is, as we will see is (a) that the Lebesgue measure is invariant under $\theta$ and (b) it is ergodic, i.e., it is not decomposable into two nontrivial invariant measures.

Similar questions are natural to ask if one considers random maps instead of deterministic ones. One natural way the random maps emerge is via random perturbations of deterministic dynamics. Let us describe one example of this kind. It is a ‘noisy’ modification of our first example above. Recall that in that example we worked with the map $\theta : x \mapsto \theta x$, and 0 was a stable equilibrium point. Let us perturb this dynamical system with noise,
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i.e., a random perturbation that kick the system out of equilibrium. Suppose we have a sequence \((\xi_n)_{n \in \mathbb{Z}}\) of independent Gaussian random variables with mean 0 and variance \(\sigma^2\) defined on some probability space \((\Omega, \mathcal{F}, P)\).

For every \(n \in \mathbb{Z}\) we will now define a random map \(\theta_{n, \omega} : \mathbb{R} \to \mathbb{R}\) by

\[
\theta_{n, \omega}(x) = ax + \xi_n(\omega).
\]

This model is known as an autoregressive-moving-average (ARMA) model of order 1.

A natural analogue of a forward orbit from our first example would be a stochastic process \((X_n)_{n \geq 0}\) emitted from a point \(x\), i.e., satisfying \(X_0 = x\) and, for all \(n \in \mathbb{N}\),

\[
X_n = aX_{n-1} + \xi_n.
\]

However, the stability issue it is not as straightforward here as in the deterministic case. It is clear that there is no fixed equilibrium point that would serve all maps \(\theta_{n, \omega}\) at the same time. The solution of the equation \(\theta_{n, \omega}(x) = x\) for some \(n\) may be irrelevant for all other values of \(n\). Still this system allows for an ergodic result of type (1.1): for any bounded Borel function \(f\), with probability 1,

\[
\frac{f(X_0) + f(X_1) + f(X_2) + \ldots + f(X_{n-1})}{n} \to \int_{\mathbb{R}} f(x) \mu(dx),
\]

where \(\mu\) is the Gaussian measure with mean 0 and variance \(\sigma^2/(1 - a^2)\). The underlying reason for this result is that \(\mu\) is a unique invariant distribution for the Markov semigroup associated with the process. So this example demonstrates a situation where there is no stability in the straightforward deterministic sense, but there is statistical stability.

Roughly speaking the ergodic approach to stochastic stability is to base the study of the system on the description of the set of invariant distributions and their properties, and so this is the main theme of these notes in the context of random dynamics. Nevertheless, we will begin with the basics of ergodic theory of deterministic transformations. The full tentative plan is to study

1. Generalities on ergodic theory of deterministic transformations.
2. Generalities on ergodic theory of random transformations and Markov processes.
3. Finite and countable state space. Markov chains.
5. Continuous time. SDEs.
6. SPDEs: Stochastic Navier-Stokes, Burgers equations.
7. Lyapunov exponents and multiplicative ergodic theory.

In the end of this introduction let us discuss the origin of the word ergodic. In the second half of 19th century, L. Boltzmann worked on foundations of statistical mechanics that was supposed to connect the kinetic theory of gases with thermodynamics. The main idea was that gases in equilibrium
obey a distribution and all macroscopic quantities of interest can be derived from that distribution. Boltzmann introduced the notion of ensemble (that corresponds to Kolmogorov’s notion of probability space) and, among other things, was concerned with the connection with the frequentist notion of probability since computing frequencies is a form of time averaging.

The systems of gas particles Boltzmann was considering preserve the total energy. He called an energy-preserving system an Ergode if there were no other conserved quantities, and he introduced the term Ergodentheorie for the study of such systems. The greek root erg in both energy and ergode relates to work or activity (ergon means work in ancient Greek). It takes no external work to move between states with the same energy, and when studying such systems Boltzmann was led to what may be called a weak form of the ergodic hypothesis: that the path that the system follows explores all available configurations of particles and their velocities with the same total energy. (The Greek word for road or way is odos)

The stronger version of the ergodic hypothesis is usually formulated as “time averages equal ensemble averages”, and can be understood as a form of law of large numbers. So, if $X_1, X_2, \ldots$ are configurations that the system evolves through and $f$ is any observable, then

$$\frac{f(X_1) + \ldots + f(X_n)}{n} \to \int_X f(x)P(dx),$$

where $P$ is the equilibrium distribution, and the right-hand side plays the role of the ensemble average. Of course, no rigorous measure and integration theory or probability theory existed at that time, so the mathematical treatment of these issues had to wait until the 20th century.
CHAPTER 2

Measure-preserving transformations and other basic notions of ergodic theory for deterministic systems

This chapter is devoted to a very sketchy introduction to the ergodic theory of deterministic measure-preserving transformations. Many more topics are studied in introductory texts [Wal82], [Sin76]. A comprehensive modern course on dynamical systems is [KH95].

1. Example: circle rotations

Let us start with an example of a simple dynamical system. Consider a circle $S^1$. It can be understood as the real line modulo 1: $S^1 = \mathbb{R}^1 / \mathbb{Z}^1$, i.e., it is the result of identification of points on the real line $\mathbb{R}^1$ with common fractional part. It can also be parametrized by the segment $[0, 1]$ with identified endpoints.

A rotation of the circle is a transformation $\theta : S^1 \rightarrow S^1$ given by

$$\theta(\omega) = \omega + \alpha \quad \text{(mod 1)}, \quad \omega \in S^1,$$

where $\alpha$ is a real number. Let us look at what happens if we keep watching the dynamics emerging under iterations of the map $\theta$, i.e., the compositions $\theta^1 = \theta$, $\theta^2 = \theta \circ \theta$, $\theta^3 = \theta \circ \theta \circ \theta$, etc. The number of iterations plays the role of time, and we are interested in statistical patterns in these iterations over long time intervals.

If the rotation angle $\alpha$ is rational and can be represented as $\alpha = m/n$ for some coprime numbers $m, n \in \mathbb{N}$, then for all $\omega \in S^1$, all the points $\omega, \theta \omega, \theta^2 \omega, \ldots, \theta^{n-1} \omega$ are distinct, but $\theta^n \omega = \omega$ i.e., after $n$ iterations of $\theta$ the circle returns to its initial position and then the next $n$ iterations $\theta^n \omega, \ldots, \theta^{2n-1} \omega$ coincide with $\omega, \theta \omega, \ldots, \theta^{n-1} \omega$, and then this repeats for $\theta^{2n} \omega, \ldots, \theta^{3n-1} \omega$, and so on. In other words all points $\omega$ are periodic with period $n$.

A more interesting situation occurs if $\alpha$ is irrational. Then there are no periodic points and one can show that the orbit $\{\omega, \theta \omega, \theta^2 \omega \ldots\}$ is dense in $S^1$, and, moreover the following equidistribution theorem established almost simultaneously by P. Bohl [Boh09], H. Weyl [Wey10], and W. Sierpinski [Sie10] (although I have not looked into these papers) holds:
Theorem 2.1. Let $\alpha$ be irrational. Then for any point $\omega \in S^1$, and any interval $I = (a, b) \subset S^1$,

(2.2) \[ \lim_{n \to \infty} \frac{\sum_{j=0}^{n-1} 1_{\theta^j \omega \in I}}{n} = b - a. \]

The sum in the left-hand side computes the total number of visits of the sequence $(\theta^j \omega)_{j=0}^\infty$ to $I$ before time $n$, and the ratio computes the frequency of those visits relative to the total time $n$. Notice that the limiting value on the right-hand side equals to $\text{Leb}(I)$, the Lebesgue measure of $I$. Also notice that the Lebesgue measure plays a special role for this dynamical system, namely it is invariant under rotation $\theta$. It should be intuitively clear what this means. First, imagine mass uniformly spread over the circle and then rotate the circle. The distribution of mass in the circle after the rotation is still uniform, so it coincides with the initial mass distribution.

We will see soon that the fact that the statistics of visits to intervals is described in the long run by an invariant distribution is a general phenomenon explained by ergodic theorems. Here we note that it is easy to extend this result to the following:

Theorem 2.2. Let $\alpha$ be irrational. Then for any point $\omega \in S^1$ and any continuous function $f : S^1 \to \mathbb{R}$,

(2.3) \[ \lim_{n \to \infty} \frac{\sum_{j=0}^{n-1} f(\theta^j \omega)}{n} = \int_{S^1} f(x) dx. \]

Derivation of Theorem 2.2 from Theorem 2.1: Identity (2.2) is a specific case of identity (2.3) for $f = 1_I$. Given (2.2), we can use linearity of sums and integrals to derive (2.3) for functions $f$ representable as

\[ f(\omega) = \sum_{k=1}^{m} c_k 1_{I_k}(\omega), \]

for some system of intervals $I_k \subset S^1$, $k = 1, \ldots, m$. Suppose now $f$ is continuous. Therefore, it is uniformly continuous and for every $\varepsilon > 0$, we can find a step function $f_\varepsilon$ such that $|f(\omega) - f_\varepsilon(\omega)| < \varepsilon$ for all $\omega \in S^1$. Then

\[ \left| \frac{\sum_{j=0}^{n-1} f(\theta^j \omega)}{n} - \int_{S^1} f(x) dx \right| \leq \frac{\sum_{j=0}^{n-1} f_\varepsilon(\theta^j \omega)}{n} - \int_{S^1} f_\varepsilon(x) dx \leq 2\varepsilon, \quad n \in \mathbb{N}, \]

so $\limsup_{n \to \infty}$ of the left-hand side does not exceed $2\varepsilon$. Since $\varepsilon > 0$ is arbitrary, we conclude that the limit of the left-hand side equals zero. \qed

Theorem 2.2 can be interpreted so that for every observable $f$ from a broad class of functions, the time averages approximate the space averages.

Of course, this property fails for rational values of $\alpha = m/n$ despite the fact that the Lebesgue measure is still invariant. The reason is that in that case, for any $\omega \in S^1$, the orbit $(\theta^j \omega)_{j \geq 0}$ explores only a very small part of $S^1$ — a finite set, in fact, — whereas the integral in the right-hand side of (2.3) depends on the behavior of $f$ on the entire $S^1$. This also can be
explained by saying that $S^1$ is decomposable: one can split $S^1$ into smaller massive sets composed of periodic orbits such that the dynamical system can be confined to any of these sets and studied independently on each of them.

Moreover, due to this decomposability of the dynamics, the Lebesgue measure is not a unique invariant measure. For example, we can choose one periodic orbit $(\omega, \theta \omega, \ldots, \theta^{n-1} \omega)$ and obtain an invariant measure by assigning mass $n^{-1}$ to each of its points:

$$\mu = \frac{1}{n} \sum_{j=0}^{n-1} \delta_{\theta^j \omega},$$

where $\delta_x$ denotes the Dirac measure concentrated at $x$, i.e., for any set $A$, \[ \delta_x(A) = 1_{x \in A}. \] Notice that this uniform distribution along the orbit of $\omega$ is not possible for nonperiodic orbits.

We will see later that the Lebesgue measure can be decomposed into a combination of these discrete invariant measures, i.e., can be represented as their (continual) mixture.

Let us now introduce some rigorous definitions in a more abstract framework.

2. Measure-preserving transformations. Formal definitions

Throughout these notes book we will use concepts from measure-theory based probability theory introduced by A.Kolmogorov in 1930’s, see [Kol50]. For a comprehensive introduction to probability, we recommend the book by A.Shiryaev [Shi96], although this is largely a question of habit and taste. There are several good books that can serve this purpose equally well. For example, a more recent textbook [KS07] contains a concise introduction to the mathematical theory of probability and stochastic processes.

We will often work with measurable spaces $(\Omega, \mathcal{F})$, where $\Omega$ is any set and $\mathcal{F}$ is a $\sigma$-algebra on $\Omega$. If $\mathcal{P}$ is a probability measure on $(\Omega, \mathcal{F})$, the triplet $(\Omega, \mathcal{F}, \mathcal{P})$ is called a probability space.

The first two definitions below require only the measurable structure and do not depend on measures.

**Definition 2.1.** Let $(\Omega, \mathcal{F})$ and $(\tilde{\Omega}, \tilde{\mathcal{F}})$ be two measurable spaces. A map $X : \Omega \to \tilde{\Omega}$ is called *measurable* with respect to $(\mathcal{F}, \tilde{\mathcal{F}})$ if for every $A \in \tilde{\mathcal{F}}$, we have $X^{-1} A = \{ \omega : X(\omega) \in A \} \in \mathcal{F}$. Often, such a map is also called $\tilde{\Omega}$-valued random variable. In the case where $(\tilde{\Omega}, \tilde{\mathcal{F}}) = (\mathbb{R}, \mathcal{B}(\mathbb{R}))$, where $\mathcal{B}(\mathbb{R})$ denotes the Borel $\sigma$-algebra on $\mathbb{R}$, $X$ is simply called a random variable.

A specific case of this definition is the following:
DEFINITION 2.2. Let \((\Omega, \mathcal{F})\) be a measurable space. A map or transformation \(\theta : \Omega \to \Omega\) is called measurable if, for every \(A \in \mathcal{F}\), we have \(\theta^{-1}A = \{\omega : \theta \omega \in A\} \in \mathcal{F}\).

Here and often in these notes we follow the tradition to denote the value assigned by the transformation \(\theta\) to argument \(\omega\) by \(\theta \omega\).

A measurable transformation \(\theta\) defines a semigroup of measurable transformations \((\theta^n)_{n \in \mathbb{Z}_+}\). Here \(\theta^n\) is inductively defined by \(\theta^0 = \text{Id}\) (identical transformation, i.e., \(\text{Id}(\omega) \equiv \omega\)) and for \(n \geq 1\), \(\theta^n = \theta \circ \theta^{n-1}\). If a transformation is invertible, i.e., \(\theta\) is a bijection of \(\Omega\) onto itself, then \(\theta^{-1}\) is well-defined and in a similar way \(\theta\) generates a group of transformations \((\theta^n)_{n \in \mathbb{Z}}\). Identity \(\theta^m \circ \theta^n = \theta^{m+n}\) is valid for all \(m, n \in \mathbb{Z}_+\) in the case of the semigroup and for all \(m, n \in \mathbb{Z}\) in the case of the group.

Often in these notes we will omit the composition sign “\(\circ\)” for brevity. For example, this convention allows to rewrite the above group identity written as \(\theta^m \theta^n = \theta^{m+n}\).

DEFINITION 2.3. Let \((\Omega, \mathcal{F}, P)\) be a probability space and \((\tilde{\Omega}, \tilde{\mathcal{F}})\) be a measurable space. Let a map \(X : \Omega \to \tilde{\Omega}\) be \((\mathcal{F}, \tilde{\mathcal{F}})\)-measurable. The pushforward of \(P\) under \(X\) is a measure on \((\Omega, \tilde{\mathcal{F}})\) denoted by \(P_{\theta_X^{-1}}\) and defined by
\[
P_{X^{-1}}(A) = P(X^{-1}A) = P\{\omega : X(\omega) \in A\}, \quad A \in \mathcal{F}.
\]

DEFINITION 2.4. Let \((\Omega, \mathcal{F}, P)\) be a probability space and suppose transformation \(\theta : \Omega \to \Omega\) be measurable. If the pushforward \(P_{\theta^{-1}}\) coincides with \(P\), i.e.,
\[
P(\theta^{-1}A) = P(A), \quad A \in \mathcal{F},
\]
then we say that \(P\) is invariant under \(\theta\), or that \(\theta\) preserves \(P\), or that \(\theta\) is \(P\)-preserving.

It is important to notice that the definition of invariance is based on pre-images under \(\theta\), and not forward images. The reason for that is that the pushforwards of measures are defined in terms of pre-images and there is no natural notion of pullbacks of measures.

PROBLEM 2.1. Suppose transformation \(\theta\) preserves a measure \(P\). Prove that for any \(n \in \mathbb{Z}_+\), \(\theta^n\) also preserves \(P\). Suppose additionally that \(\theta\) is invertible and \(\theta^{-1}\) is measurable. Prove that then a stronger statement holds: for any \(n \in \mathbb{Z}\), \(\theta^n\) also preserves \(P\) (in particular, \(\theta^{-1}\) is \(P\)-preserving).

Let us consider some basic examples of probability spaces with measure-preserving transformations.

Let \(\Omega = \{0, 1, \ldots, n-1\}\), \(\mathcal{F} = 2^\Omega\) (\(\sigma\)-algebra of all subsets of \(\Omega\)), and let \(P\) be the uniform distribution on \(\Omega\). Then a transformation \(\theta\) is \(P\)-preserving if and only if it is a permutation, i.e., a one-to-one map. For instance, the cyclic map \(\theta \omega = \omega + 1 \mod n\) is \(P\)-preserving.
Continuous analogues of this cyclic transformation are circle rotations preserving the Lebesgue measure on $S^1$ equipped with the Borel $\sigma$-algebra, and, more generally, shifts on a $d$-dimensional torus $T^d = \mathbb{R}^d / \mathbb{Z}^d$ given by
\begin{equation}
\theta(\omega_1, \ldots, \omega_d) = (\omega_1 + \alpha_1 \pmod{1}, \ldots, \omega_d + \alpha_d \pmod{1}),
\end{equation}
where $\alpha = (\alpha_1, \ldots, \alpha_d)$ is a fixed vector in $\mathbb{R}^d$.

One more class of examples transformations preserving the Lebesgue measure is provided by interval exchange transformations. Suppose the probability space is $([0, 1], \mathcal{B}([0, 1]), \text{Leb})$. Let us take $m \in \mathbb{N}$ and let intervals $I_k = [a_k, b_k), k = 1, \ldots, m$, form a partitions of $[0, 1)$, i.e., these intervals are mutually disjoint and their union is $[0, 1)$. Let $I'_k = [a'_k, b'_k), k = 1, \ldots, m$, form another partition such that $b'_k - a'_k = b_k - a_k, k = 1, \ldots, m$. Then one can define a transformation $\theta$ of $[0, 1)$ such that for each $I_k$ it coincides with a translation of $I_k$ onto $I'_k$. It is easy to see that thus defined transformation preserves the Lebesgue measure.

It is often useful to check invariance in terms of test functions. From this point on we will often denote expectation (integral) of a random variable $X$ defined on $(\Omega, \mathcal{F}, P)$ by $EX$ or $EX(\omega)$.

**Lemma 2.1.** A measurable map $\theta$ on a probability space $(\Omega, \mathcal{F}, P)$ preserves $P$ if and only if for any bounded measurable function $X : \Omega \rightarrow \mathbb{R}$,
\begin{equation}
EX(\omega) = EX(\theta \omega).
\end{equation}

**Proof:** Definition 2.4 is equivalent to (2.7) for indicator functions $X = 1_A, A \in \mathcal{F}$. So sufficiency of (2.7) is obvious. On the other hand, if (2.7) holds for indicators, then it holds for simple functions (finite linear combinations of indicators). Also, any measurable function $X$ can be approximated by a simple one: for any $\varepsilon > 0$ there is a simple function $X_\varepsilon$ such that $|X(\omega) - X_\varepsilon(\omega)| \leq \varepsilon$. Since (2.7) holds for $X_\varepsilon$, we obtain $|EX(\omega) - EX(\theta \omega)| \leq 2\varepsilon$. Since $\varepsilon$ is arbitrary, (2.7) follows. \(\square\)

**Definition 2.5.** A probability space $(\Omega, \mathcal{F}, P)$ equipped with a measure-preserving transformation $\theta$ is often called a metric dynamical system or a measure-preserving dynamical system. A quadruplet notation $(\Omega, \mathcal{F}, P, \theta)$ is often used.

### 3. Poincaré's recurrence theorem

The first truly contentful result that we are going to prove is the following recurrence theorem proved by H. Poincaré [Poi90], [Poi99].

**Theorem 2.3.** Suppose $(\Omega, \mathcal{F}, P)$ is a probability space, and $\theta$ is a $P$-preserving transformation of $\Omega$. Then for any set $A \in \mathcal{F}$ we have $P(T(A)) = 0$, where $T(A)$ is set of all points in $A$ that never return to $A$ under iterations of $\theta$, i.e.,
\begin{equation}
T(A) = \{\omega \in A : \theta^n \omega \notin A, \ n \in \mathbb{N}\}.
\end{equation}
Proof: Let us notice that the sets $\theta^{-n}T(A)$, $n \in \mathbb{N}$ are mutually disjoint. In fact, if $\omega \in \theta^{-n}T(A) \cap \theta^{-m}T(A)$ and $m < n$, then we denote $\omega' = \theta^m\omega$ and notice that $\omega' \in T(A)$, and at the same time $\theta^{n-m}\omega' = \theta^{n-m}\theta^m\omega = \omega \in T(A)$ which is a contradiction with the definition of $T(A)$.

Due to $\theta$-invariance of $P$, $P(\theta^{-n}T(A)) = P(T(A))$ for all $n$, so

$$\sum_{n \in \mathbb{N}} P(T(A)) = \sum_{n \in \mathbb{N}} P(\theta^{-n}T(A)) = P\left( \bigcup_{n \in \mathbb{N}} \theta^{-n}T(A) \right) \leq P(\Omega),$$

which can hold true only if $P(T(A)) = 0$. \hfill \Box

It is possible to give a quantitative strengthening of this theorem. The following is an adaptation from [Tao08].

Theorem 2.4. Under the conditions of Theorem 2.3,

$$\limsup_{n \to \infty} P(\theta^{-n}A \cap A) \geq P^2(A), \quad A \in \mathcal{F}.$$  

Proof: Invariance of $P$ implies that for any $N \in \mathbb{N}$,

$$E\left[ \sum_{n=1}^{N} 1_{\theta^{-n}A} \right] = NP(A).$$

Using the Cauchy–Schwartz inequality or Lyapunov inequality, we obtain

$$E\left[ \sum_{n=1}^{N} 1_{\theta^{-n}A} \right]^2 \geq \left( E\left[ \sum_{n=1}^{N} 1_{\theta^{-n}A} \right] \right)^2 = N^2P(A)^2. \tag{2.8}$$

The left-hand side of this inequality can be rewritten as

$$E\left[ \sum_{n=1}^{N} 1_{\theta^{-n}A} \right]^2 = \sum_{n,m=1}^{N} P(\theta^{-n}A \cap \theta^{-m}A) = \sum_{n,m=1}^{N} P(A \cap \theta^{-|m-n|}A). \tag{2.9}$$

For any $\varepsilon$, we can find $n_0 = n_0(\varepsilon)$ such that

$$P(A \cap \theta^{-n}A) < L + \varepsilon, \quad n > n_0,$$

where $L = \limsup_{n \to \infty} P(A \cap \theta^{-n}A)$. Therefore, splitting the sum in the right-hand side of (2.9) into two — over $n,m$ satisfying $|n - m| \geq n_0$ and over $n,m$ satisfying $|n - m| < n_0$ — and noticing that there are less than $2n_0N$ terms in the second sum and they all are bounded by 1, we obtain

$$\frac{1}{N^2} \sum_{n,m=1}^{N} P(A \cap \theta^{-|m-n|}A) < L + \varepsilon + \frac{2n_0(\varepsilon)N}{N^2}.$$  

Combining this with (2.8) and (2.9), we obtain

$$P(A)^2 \leq L + \varepsilon + \frac{2n_0(\varepsilon)}{N}.$$  

Taking $N \to \infty$ and then $\varepsilon \to 0$, we obtain the desired result. \hfill \Box
Problem 2.2. Using an appropriate Bernoulli shift, prove that the lower bound in Theorem 2.4 is sharp, i.e., in general $P^2(A)$ cannot be replaced by a larger value.

Theorem 2.4 implies that for any set $A$ of positive measure (no matter how small) there are infinitely many times $n$ such that $P(\theta^{-n}A \cap A) > 0$. In particular, Theorem 2.3 can be viewed as a corollary of Theorem 2.4. 

**Derivation of Theorem 2.3 from Theorem 2.4:** Let us assume that $P(T(A)) > 0$. Then $P(\theta^{-n}T(A) \cap T(A))$ is positive for infinitely many $n \in \mathbb{N}$. For those $n$, there is $\omega \in \theta^{-n}T(A) \cap T(A) \subset \theta^{-n}A \cap A$. So, $\omega \in T(A)$, but $\theta^n \omega \in A$. This contradicts the definition of $T(A)$. □

There is an interesting Zermelo’s paradox related to Poincaré’s recurrence theorem. Consider a gas, i.e., a collection of particles in a box. The evolution of this system of particles can be described by a Hamiltonian dynamical system preserving the Lebesgue measure in the phase space encoding positions and momenta of all particles. If in the beginning all particles are in one half of the box, then as the system evolves it is natural to expect that these particles eventually will be more or less uniformly distributed between this half and the other one. Poincaré’s recurrence theorem allows to conclude though that sooner or later there will be a moment when these particles will all gather within the same half of the box. This seems to contradict both intuition and experience. The explanation is that the time required for such an event to occur in a realistic setting is astronomically large, much larger than the length of any conceivable experiment, and so these events are never observed in practice.

4. Stochastic processes: basic notions

There is a useful interpretation of metric dynamical systems in terms of stationary processes, for example for any random variable $f : X \to \mathbb{R}$, the process $(X_n)_{n \in \mathbb{N}}$ defined by $X_n(\omega) = f(\theta^n \omega)$ is a stationary process. First, let us recall some basic facts concerning stochastic processes in general.

**Definition 2.6.** Let $(\mathbb{X}, \mathcal{X})$ be a measurable space and $T$ be any set. Any collection of $\mathbb{X}$-valued random variables $(X_t)_{t \in T}$ is called a stochastic process.

The set $T$ in this definition plays the role of time axis, and if $T \subset \mathbb{R}$, then for fixed $\omega \in \Omega$, the realization $(X_t(\omega))_{t \in T}$ is also often called the trajectory or sample path corresponding to $\omega$.

The space of these sample paths is denoted by

$$X^T = \{x : T \to \mathbb{X}\}.$$ 

If $T = \mathbb{N}$, then $X^T$ consists of one-sided sequences $(x_1, x_2, \ldots)$, and if $T = \mathbb{Z}$, then $X^T$ consists of two-sided sequences $(\ldots, x_{-2}, x_{-1}, x_0, x_1, x_2, \ldots)$. 

at this point the Bernoulli shift has not been introduced yet
It is often useful to consider an \( (\mathcal{X}, \mathcal{X}) \)-valued process \( X \) as one \( \mathcal{X}_T \)-valued random variable, and so we need to introduce an appropriate \( \sigma \)-algebra on \( \mathcal{X}_T \).

**Definition 2.7.** Suppose \( T \subset \mathbb{R} \). For \( m \in \mathbb{N} \), \( n_1 < \ldots < n_m \), and \( A_1, \ldots, A_m \in \mathcal{X} \), we denote
\[
C_{n_1, \ldots, n_m}(A_1, \ldots, A_m) = \{ x \in \mathcal{X}_T : x_{n_1} \in A_0, \ldots, x_{n_m} \in A_m \}.
\]
Sets of this form are called elementary cylinders.

Let us stress that all cylinders have non-unique representations of the form (2.10). For example,
\[
C_{n_1, \ldots, n_m, n_m+1}(A_1, \ldots, A_m, \mathcal{X}) = C_{n_1, \ldots, n_m}(A_1, \ldots, A_m).
\]

**Definition 2.8.** The cylindric \( \sigma \)-algebra \( \mathcal{X}_T \) is generated by all elementary cylinders.

For any time \( n \in T \), the projection \( \pi_n : \mathcal{X}_T \to \mathcal{X} \) is defined by \( \pi_n(x) = x_n \).

**Theorem 2.5.** The cylindric \( \sigma \)-algebra \( \mathcal{X}_T \) is generated by maps \( \pi_n \).

The family of all elementary cylinders is a \( \pi \)-system, i.e., it is closed under intersection.

**Theorem 2.6 (see [Shi96, Lemma II.2.3]).** If the restrictions of two probability measures on a \( \pi \)-system \( \mathcal{E} \) coincide, then the measures coincide on \( \sigma(\mathcal{E}) \).

The following corollary says that a probability measure on \( (\mathcal{X}_T, \mathcal{X}_T) \) is uniquely defined by its values on elementary cylinders.

**Theorem 2.7.** Let \( P \) and \( Q \) be two measures on \( (\mathcal{X}_T, \mathcal{X}_T) \) such that for any elementary cylinder \( C \), \( P(C) = Q(C) \). Then \( P = Q \).

To formulate an existence result, we need to introduce general cylinders.

**Definition 2.9.** Let \( m \in \mathbb{N} \) and \( n_1, \ldots, n_m \subset T \) satisfy \( n_1 < \ldots < n_m \), we denote by
\[
\mathcal{X}^{n_1, \ldots, n_m} = \sigma(C_{n_1, \ldots, n_m}(A_1, \ldots, A_m), A_1, \ldots, A_m \in \mathcal{X}).
\]
Elements of \( \mathcal{X}^{n_1, \ldots, n_m} \) are called cylinders.

**Theorem 2.8 (Kolmogorov–Daniell extension theorem).** If \( (\mathcal{X}, \mathcal{X}) \) is a Borel space and \( P \) is a nonnegative function on all cylinders such that for any \( m \in \mathbb{N} \) and any \( n_1, \ldots, n_m \subset T \) satisfying \( n_1 < \ldots < n_m \), the restriction of \( P \) to \( \mathcal{X}^{n_1, \ldots, n_m} \) is a probability measure, then there is a unique measure on \( \mathcal{X}_T \) such that its restriction on cylinders coincides with \( P \).

Often in these notes we will work with Borel spaces, as in this theorem. Let us introduce the corresponding definitions.
5. Interpretation of measure-preserving maps via stationary processes

Definition 2.10. We say that measurable spaces \((X, \mathcal{X})\) and \((Y, \mathcal{Y})\) are measurably isomorphic if there is a bijection \(\varphi : X \rightarrow Y\) such that \(\varphi\) is \((X, \mathcal{X})\)-measurable and \(\varphi^{-1}\) is \((Y, \mathcal{Y})\)-measurable.

Definition 2.11. A measurable space \((Y, \mathcal{Y})\) is called a Borel space if it is measurably isomorphic to some space \((X, \mathcal{X})\), where \(X\) is a Borel subset of \([0, 1]\), and \(\mathcal{X}\) is the \(\sigma\)-algebra induced by \(B([0, 1])\) on \(X\).

Remark 2.1. The family of Borel spaces \((X, \mathcal{X})\) is extremely rich and includes, for example, Borel subsets of Polish spaces, see [Kur66]. We recall that a metric set \((X, \rho)\) is called Polish if it is complete and separable.

5. Interpretation of measure-preserving maps via stationary processes

In this section we describe the connection between stationary processes and measure-preserving transformations.

Definition 2.12. Suppose \(\mathbb{T} = \mathbb{N}, \mathbb{Z}_+,\) or \(\mathbb{Z}\) and let \((X_t)_{t \in \mathbb{T}}\) be a stochastic process with values in \(X\). This process is called stationary if for any \(m \in \mathbb{N},\) any \(i_1, \ldots, i_m \in \mathbb{T},\) and any \(r \in \mathbb{N},\) the distributions of \((X_{i_1}, \ldots, X_{i_m})\) and \((X_{i_1+r}, \ldots, X_{i_m+r})\) coincide, i.e., for any \(A_1, \ldots, A_m \in \mathcal{G},\)
\[
P\{X_{i_1} \in A_1, \ldots, X_{i_m} \in A_m\} = P\{X_{i_1+r} \in A_1, \ldots, X_{i_m+r} \in A_m\}. \tag{2.11}
\]

It is sufficient to check this definition for \(r = 1\) since the statement for the general \(r\) follows by induction.

The simplest example of a stationary process is a sequence of independent and identically distributed (i.i.d.) random variables \((X_n)_{n \in \mathbb{N}}\). By definition, this means that there is a probability \(P\) on \((\mathbb{R}, B(\mathbb{R}))\) such that for any \(m\) and any \(i_1, \ldots, i_m \in \mathbb{N},\)
\[
P\{X_{i_1} \in A_1, \ldots, X_{i_m} \in A_m\} = P(A_1)P(A_2)\ldots P(A_m),
\]
so identity (2.11) easily follows.

Lemma 2.2. Suppose \(X\) is an \((\mathbb{X}, \mathcal{X})\)-valued random variable on a probability space \((\Omega, \mathcal{F}, P),\) and \(\theta\) is \(P\)-preserving transformation on \(\Omega.\) Then the stochastic process \((X_n)_{n \in \mathbb{Z}_+}\) defined by
\[
X_n(\omega) = X(\theta^n \omega) \tag{2.12}
\]
is stationary. If \(\theta\) is measurably invertible, then the process \((X_n)_{n \in \mathbb{Z}}\) defined by the same formula is stationary.

Proof: Let us check the definition of stationarity for shift \(r = 1:\)
\[
P\{\omega : X_{i_1+1}(\omega) \in A_1, \ldots, X_{i_m+1}(\omega) \in A_m\} = P\{\omega : X(\theta^{i_1+1} \omega) \in A_1, \ldots, X(\theta^{i_m+1} \omega) \in A_m\}
= P\{\omega : X(\theta^{i_1} \omega) \in A_1, \ldots, X(\theta^{i_m} \omega) \in A_m\}
= P\{\omega : X(\theta^{i_1} \omega) \in A_1, \ldots, X(\theta^{i_m} \omega) \in A_m\}
= P\{\omega : X_{i_1}(\omega) \in A_1, \ldots, X_{i_m}(\omega) \in A_m\},
\]
where the third identity follows from the \( \theta \)-invariance of \( \mathcal{P} \).

One can also make sense of a converse statement: every stationary process can be represented in the form described in Lemma 2.2. Suppose \((X_n)_{n \in \mathbb{Z}^+}\) is an \((\mathbb{X}, \mathcal{A})\)-valued stationary process on a probability space \((\Omega, \mathcal{F}, \mathcal{P})\).

One way to look at the process \(X\) is to view it as a map from \(\Omega\) to the space of trajectories \(\hat{\Omega} = \mathbb{X}^{\mathbb{Z}^+}\), i.e., the space of all functions \(x : \mathbb{Z}^+ \to \mathbb{X}\), i.e., sequences \(x = (x_0, x_1, \ldots)\), equipped with cylindric \(\sigma\)-algebra \(\mathcal{F} = \mathcal{X}^{\mathbb{Z}^+}\). In fact, the process \(X\) is \((\mathcal{F}, \mathcal{\hat{F}})\)-measurable (to see this, it is sufficient to check that \(X\)-pre-images of all cylinders are in \(\mathcal{F}\)). Therefore, we can consider the pushforward \(\hat{\mathcal{P}} = \mathcal{P} X^{-1}\) of \(\mathcal{P}\) to \((\hat{\Omega}, \hat{\mathcal{F}})\) under this map. The result is a probability space \((\hat{\Omega}, \hat{\mathcal{F}}, \hat{\mathcal{P}})\). The process \((\hat{X}_n)_{n \in \mathbb{Z}^+}\) defined by \(\hat{X}_n(x) = \pi_n x = x_n\), \(n \in \mathbb{Z}^+\), is a stationary process under \(\hat{\mathcal{P}}\), because its distribution is the same as that of the original process \(X\).

Let us prove that \(\hat{\mathcal{P}}\) is invariant under the shift map \(\theta : \hat{\Omega} \to \hat{\Omega}\) defined by \(\theta(x_0, x_1, \ldots) = (x_1, x_2, \ldots)\). To check this, let us recall that, due to the Kolmogorov–Daniell extension theorem, measures on \((\hat{\Omega}, \hat{\mathcal{F}})\) are uniquely determined by their values on cylinders

\[
\mathcal{P}(\theta^{-1} C_n(A_0, \ldots, A_{n-1})) = \hat{\mathcal{P}}(C_{n+1}(A_0, \ldots, A_{n-1})) = \mathcal{P}(X_1 \in A_0, \ldots, X_n \in A_{n-1}) = \mathcal{P}(X_0 \in A_0, \ldots, X_{n-1} \in A_{n-1}) = \hat{\mathcal{P}}(C_n(A_0, \ldots, A_{n-1})).
\]

Now we can take the random variable \(\pi_0 : \hat{\Omega} \to \mathbb{X}\) defined by \(\pi_0(x) = x_0\) and notice that the process \(\hat{X}_n\) on \((\hat{\Omega}, \hat{\mathcal{F}}, \hat{\mathcal{P}})\) can be represented as \(\hat{X}_n(x) = \pi_0(\theta^n x)\). This is what we wanted, because this representation is of form (2.12). However, to obtain this representation we had to switch from the original probability space to the canonical space of trajectories. The same procedure applies to processes indexed by \(\mathbb{N}\) and \(\mathbb{Z}\).

It is often convenient to work with stationary processes directly on their trajectory spaces. For example, when working with i.i.d. \((\mathbb{X}, \mathcal{A})\)-valued random variables, it is convenient to work with product measure on \((\mathbb{X}^{\mathbb{Z}^+}, \mathcal{X}^{\mathbb{Z}^+})\). If \(\mathbb{X}\) is finite then \((\mathbb{X}^{\mathbb{Z}^+}, \mathcal{X}^{\mathbb{Z}^+})\) equipped with a product measure and coordinate shift \(\theta\) is called a Bernoulli shift.

6. Ergodicity

In this section we work with a probability space \((\Omega, \mathcal{F}, \mathcal{P})\) equipped with a \(\mathcal{P}\)-preserving transformation \(\theta\). The goal of this section is to define and study the notion of ergodicity or indecomposability of dynamics defined by \(\theta\) with respect to \(\mathcal{P}\).
**Definition 2.13.** A set $A$ is called (backward) invariant if $\theta^{-1}A = A$, forward invariant or positive invariant if $\theta A = A$, and almost invariant or invariant mod 0 if $P(\theta^{-1}A \triangle A) = 0$.

**Lemma 2.3.** (a) If $A \in \mathcal{F}$ is invariant, then $A$ is forward invariant.
(b) If $A \in \mathcal{F}$ is forward invariant then it is almost invariant.
(c) If $A \in \mathcal{F}$ is almost invariant, then there is an invariant set $B$ such that $P(A \triangle B) = 0$.
(d) If $A \in \mathcal{F}$ is forward invariant, then there is an invariant set $B$ such that $P(A \triangle B) = 0$.

**Proof:** To prove part (a), we notice that for any set $A$, $\theta \theta^{-1}A = A$. So, applying $\theta$ to $A = \theta^{-1}A$ we obtain $\theta A = \theta \theta^{-1}A = A$.

To prove part (b), we notice that for any set $A$, $A \subset \theta^{-1}\theta A$. So, applying $\theta^{-1}$ to $A = \theta A$, we obtain $\theta^{-1}A \subset A$. Due to $\theta$-invariance of $P$, we obtain $P(\theta^{-1}A) = P(A)$ and, therefore, $P(\theta^{-1}A \setminus A) = 0$.

Part (d) follows from (b) and (c), so it remains to prove (c). Let $A$ be almost invariant. Let

$$B = \liminf_{n \to \infty} \theta^{-n}A = \bigcup_{n=1}^{\infty} \bigcap_{k=n}^{\infty} \theta^{-k}A.$$  

Then $\theta^{-1}B = \liminf_{n \to \infty} \theta^{-n-1}A = B$, i.e., $B$ is invariant. To prove $P(A \triangle B) = 0$ we first notice that it is easy to check by induction that

$P(\theta^{-n}A \triangle A) = 0, \ n \in \mathbb{N}$.  

(2.13) $P(\theta^{-n}A \triangle A) = 0, \ n \in \mathbb{N}$.

(2.14) $A \setminus B = A \cap \bigcap_{n=1}^{\infty} \bigcup_{k=n}^{\infty} (\theta^{-k}A)^c \subset A \cap \bigcup_{k=1}^{\infty} (\theta^{-k}A)^c = \bigcup_{k=1}^{\infty} (A \setminus \theta^{-k}A)$.

(2.15) $B \setminus A = \bigcup_{n=1}^{\infty} \bigcap_{k=n}^{\infty} \theta^{-k}A \setminus A \subset \bigcup_{n=1}^{\infty} (\theta^{-n}A \setminus A)$.

Now our claim follows from (2.13), (2.14), and (2.15). ⊢

It is easy to check that the families $\mathcal{I}_\theta$ and $\mathcal{I}_\theta^*$ of invariant and almost invariant sets are $\sigma$-algebras. However, the forward invariant sets do not form a $\sigma$-algebra in general and for this reason it is often preferable to work with (inverse)-invariant sets. This is tightly related to the fact that taking pre-image is a more well-behaved operation than taking images, and thus it is easy to define pushforwards of $\sigma$-algebras and measures by taking pre-images, and there is no natural way to define pullbacks of these objects via forward images.

**Problem 2.3.** Give an example of a metric dynamical system such that the collection of forward-invariant sets does not form a $\sigma$-algebra. Hint: complements.
DEFINITION 2.14. A transformation is called **ergodic** if every invariant set has measure 0 or 1.

This definition means that one cannot split our metric dynamical system into two.

**L E M M A 2.4.** A transformation $\theta$ is ergodic if and only if every almost invariant set has measure 0 or 1.

**P R O O F:** The “if” part is obvious since $I \subset I^*$. The converse is a direct consequence of part (c) of Lemma 2.3. □

**D E F I N I T I O N 2.15.** A random variable $X$ is called **invariant** if $X(\omega) = X(\theta \omega)$ for all $\omega \in \Omega$.

**D E F I N I T I O N 2.16.** A random variable $X$ is called **almost invariant** or **invariant mod 0** if $X(\omega) = X(\theta \omega)$ for $P$-almost all $\omega \in \Omega$.

The following obvious statement gives one more reason to define invariant and almost invariant sets as in Definition 2.13.

**L E M M A 2.5.** A set $A$ is invariant if and only if $1_A$ is invariant. A set $A$ is almost invariant if and only if $1_A$ is almost invariant.

**D E F I N I T I O N 2.17.** We say that a (real-valued) random variable $X$ is a.s.-constant if there is a number $c \in \mathbb{R}$ such that $P\{X = c\} = 1$.

**T H E O R E M 2.9.** The following three conditions are equivalent if $\theta$ is a $P$-preserving transformation:

(a) $\theta$ is ergodic;
(b) every almost invariant random variable is a.s.-constant;
(c) every invariant random variable is a.s.-constant;

**P R O O F:** Condition (b) trivially implies (c), and the latter implies (a) as a specific case for indicator random variables. It remains to derive (b) from (c). So we suppose that $X$ is a random variable almost invariant under an ergodic transformation $\theta$. For every real number $t$, the set $A(t) = \{X \leq t\}$ is almost invariant. Ergodicity of $\theta$ and Lemma 2.4 imply $P\{X \leq t\} = 0$ or 1 for any $t \in \mathbb{R}$. Notice that the function $F : \mathbb{R} \to \mathbb{R}$ defined by $F(t) = P\{X \leq t\}$ is the distribution function of $X$. Thus, $\lim_{t \to -\infty} F(t) = 0$, $\lim_{t \to +\infty} F(t)$, $F$ is nondecreasing and left-continuous. Therefore, there is some point $c \in \mathbb{R}$ such that $F(t) = 0$ for all $t \leq c$ and $F(t) = 1$ for all $t > c$. Then $P\{X = c\} = 1$, and the lemma is proven. The following is a proof that does not use the notion of distribution function:

Let us define $c = \sup\{t : P(A(t)) = 0\}$. Then $c$ is finite since $A(t_1) \subset A(t_2)$ for $t_1 < t_2$ and $\bigcap_{t \in \mathbb{R}} A(t) = \emptyset$, $\bigcup_{t \in \mathbb{R}} A(t) = \Omega$. We have

$$P\{X < c\} = P\left(\bigcup_{n=1}^{\infty} A(c - n^{-1})\right) = 0,$$
and
\[
P\{X > c\} = P\left(\bigcup_{n=1}^{\infty} A^c(c + n^{-1})\right) = 0,
\]
so \(P\{X = c\} = 1\).

**Remark 2.2.** It is sufficient to verify conditions (b) and (c) of Theorem 2.9, for bounded random variables. In fact, for a general random variable \(X\) these conditions can be verified by checking them first for its truncations \(X_N = XI_{|X| \leq N}\) and then letting \(N \to \infty\).

**Remark 2.3.** Conditions (b) and (c) of Theorem 2.9 remain necessary and sufficient conditions for ergodicity if stated for complex-valued random variables instead of real-valued ones. To see that, consider the real and imaginary parts of complex valued random variables.

Let us now check ergodicity for some systems.

**Lemma 2.6.** Let \((\Omega, \mathcal{F}, P)\) be the unit circle \(S^1\) equipped with Borel \(\sigma\)-algebra and Lebesgue measure. If \(\alpha\) is irrational, then the circle rotation \(\theta\) by angle \(\alpha\) defined in (2.1) is ergodic.

**Proof:** Let us take an almost invariant bounded random variable \(f\) and prove that it is a.s.-constant. Every bounded measurable function on \(S\) can be decomposed into a Fourier series convergent in \(L^2(S^1)\):
\[
f(\omega) = \sum_{k \in \mathbb{Z}} c_k e^{2\pi i k \omega},
\]
where
\[
c_k = \int_{S^1} e^{-2\pi i k \omega} f(\omega) d\omega, \quad k \in \mathbb{Z}.
\]
Since \(f\) is almost invariant, \(f(\omega)\) and \(f(\theta \omega)\) coincide almost everywhere. So the Fourier series for \(f(\theta \omega)\) coincides with (2.16). Therefore, we can compute \(c_k\) as Fourier coefficients for \(f(\theta \omega)\):
\[
c_k = \int_{S^1} e^{-2\pi i k \omega} f(\omega + \alpha) d\omega = e^{2\pi i k \alpha} \int_{S^1} e^{-2\pi i k \omega} f(\omega) d\omega = e^{2\pi i k c_k}, \quad k \in \mathbb{Z}.
\]
If \(\alpha\) is irrational, then \(e^{2\pi i k c_k} \neq 1\) for all \(k \neq 0\), so \(c_k\) may be nonzero only for \(k = 0\), which means that \(f\) is a.s.-constant. \(\Box\)

If \(\alpha = m/n\) is rational, then \(\theta\) is not ergodic since \(f(\omega) = e^{2\pi i n \omega}\) is a nonconstant invariant function.

**Theorem 2.10.** Let \(\theta\) be a shift on the torus \(T^d\) defined by (2.6). It is ergodic if and only if the identity \(r_1 \alpha_1 + \ldots + r_d \alpha_d = m\) is impossible for any \(r_1, \ldots, r_d, m \in \mathbb{Z}\) such that not all of them are equal to 0.

**Problem 2.4.** Prove Theorem 2.10.
THEOREM 2.11 (A variation on Kolmogorov’s 0-1 Law, REFERENCE????).
Let \((\mathbb{X}, \mathcal{X}, P)\) be a probability space and \((\Omega, \mathcal{F}, P) = (\mathbb{X}^\mathbb{N}, \mathcal{X}^\mathbb{N}(\mathbb{X}), P^\mathbb{N})\), where \(\mathcal{X}^\mathbb{N}\) is the cylindric \(\sigma\)-algebra and \(P^\mathbb{N}\) is the product measure on it. Then the shift transformation \(\theta\) defined by \(\theta(\omega_1, \omega_2, \ldots) = (\omega_2, \omega_3, \ldots)\) is measure preserving and ergodic.

PROOF: The invariance of \(P\) under \(\theta\) follows from the discussion right after Lemma 2.2.

To prove ergodicity, consider an almost invariant set \(A\). It is sufficient to prove that \(P(A) = P^2(A)\) which can be rewritten as \(P(A \cap A) = P(A)P(A)\) or, using the almost invariance of \(A\), as \(P(A \cap \theta^{-n}A) = P(A)P(\theta^{-n}A)\) for all \(n \in \mathbb{N}\). This means that \(A\) is independent of itself or its pre-images under \(\theta^n\).

To prove this, we will approximate \(A\) and \(\theta^{-n}A\) (for large \(n\)) by cylinders that depend on disjoint coordinate sets and thus are independent.

Since \(A \in \mathcal{X}^\mathbb{N}\), it can be approximated by cylinders, i.e., for any \(\varepsilon > 0\) there is a set \(A_\varepsilon \in \mathcal{X}^\mathbb{N}\), a number \(k(\varepsilon) \in \mathbb{N}\) and a set \(B_\varepsilon \in \mathcal{X}^{k(\varepsilon)}\) such that \(A_\varepsilon = \{\omega : (\omega_1, \ldots, \omega_{k(\varepsilon)}) \in B_\varepsilon\}\), and \(P(A_\varepsilon \Delta A) < \varepsilon\).

Let us take any \(n \in \mathbb{N}\) satisfying \(n \geq k(\varepsilon)\). Sets \(A_\varepsilon\) and \(\theta^{-n}A_\varepsilon\) = \(\{\omega : (\omega_{n+1}, \ldots, \omega_{n+k(\varepsilon)}) \in B_\varepsilon\}\) are independent by the definition of the product measure, so

\[
P(A_\varepsilon \cap \theta^{-n}A_\varepsilon) = P(A_\varepsilon)P(\theta^{-n}A_\varepsilon) = P^2(A_\varepsilon).
\]

To deal with the left-hand side, we notice that \(A_\varepsilon\) is very close to an almost invariant set:

\[
|P(A_\varepsilon \cap \theta^{-n}A_\varepsilon) - P(A)| = |P(A_\varepsilon \cap \theta^{-n}A_\varepsilon) - P(A \cap \theta^{-n}A)| \\
\leq |P(A_\varepsilon \cap \theta^{-n}A_\varepsilon) - P(A_\varepsilon \cap \theta^{-n}A)| \\
+ |P(A_\varepsilon \cap \theta^{-n}A) - P(A \cap \theta^{-n}A)| \\
\leq |P(\theta^{-n}(A_\varepsilon \Delta A))| + |P(A_\varepsilon \Delta A)| \leq 2\varepsilon.
\]

For the right-hand side of (2.17), we have

\[
|P^2(A_\varepsilon) - P(A)| = (P(A_\varepsilon) + P(A)) \cdot |P(A_\varepsilon) - P(A)| \leq 2|P(A_\varepsilon \Delta A)| \leq 2\varepsilon,
\]
so \(|P^2(A) - P(A)| \leq 4\varepsilon\). Since \(\varepsilon > 0\) is arbitrary, we have \(P^2(A) = P(A)\), so \(P(A) = 0\) or 1.

\(\square\)
CHAPTER 3

Ergodic theorems for measure-preserving transformations

1. Von Neumann’s ergodic theorem in $L^2$.

Based on [Hal60, pp.13–17]

Suppose we have a probability space $(\Omega, \mathcal{F}, P)$, and let us assume that $P$ is invariant under a measurable transformation $\theta : \Omega \to \Omega$, i.e., for every $A \in \mathcal{F}$, $P(\theta^{-1}A) = P(A)$. In other words, the pushforward $P\theta^{-1}$ of the measure $P$ defined on $A \in \mathcal{F}$ by $P\theta^{-1}(A) = P(\theta^{-1}A)$, coincides with $P$.

We are interested in the convergence of time averages

$$A_n f(\omega) = \frac{1}{n} \sum_{k=0}^{n-1} f(\theta^k \omega)$$

(3.1)

as $n \to \infty$ for measurable functions $f : \Omega \to \mathbb{R}$. There is an elegant proof of this convergence in the sense of $L^2$ due to Riesz.

Let us first introduce an $L^2$ framework. We recall that the space of all measurable functions $f : \mathcal{F} \to \mathbb{C}$ such that $\mathbb{E}|f|^2 < \infty$ with the usual identification of functions that coincide $P$-almost everywhere, is a complex Hilbert space denoted by $L^2(\Omega, \mathcal{F}, P)$ or $L^2$ for brevity, with inner product $\langle f, g \rangle = \mathbb{E}f\bar{g}$ and norm $\|f\| = (\mathbb{E}|f|^2)^{1/2}$.

For any measurable $f$, we will denote by $Uf$ the function defined by $Uf(\omega) = f(\theta \omega)$, $\omega \in \Omega$. The following observation was first made by Koopman in [Koo31] in the context of Hamiltonian systems:

**Lemma 3.1.** The operator $U$ is an isometry in $L^2$, i.e., if $f \in L^2$, then $\|Uf\| = \|f\|$.

**Proof:** For any $A \in \mathcal{F}$, we have $U1_A = 1_{\theta^{-1}A}$. Therefore, the invariance of $P$ under $\theta$ implies

$$\langle U1_A U1_A \rangle = \mathbb{E}1_{\theta^{-1}A}^2 = P(\theta^{-1}A) = P(A).$$

Also, for any disjoint measurable sets $A$ and $B$,

$$\langle U1_A U1_B \rangle = \mathbb{E}1_{\theta^{-1}A}1_{\theta^{-1}B} = 0.$$

Now we can use linearity of $U$ to derive the result for any simple function $f = \sum_{j=1}^{m} c_j 1_{A_j}$, where $m \in \mathbb{N}$, $(A_j)_{j=1}^{m}$ is a collection of mutually disjoint
measurable sets, and $c_j \in \mathbb{C}$ for $j = 1, \ldots, m$:

$$
\|Uf\|^2 = \sum_{i,j=1}^{m} c_i \bar{c}_j \langle U1_{A_i}, U1_{A_j} \rangle 
= \sum_{i,j=1}^{m} c_i \bar{c}_j \langle 1_{\theta^{-1}A_i}, 1_{\theta^{-1}A_j} \rangle = \sum_{i=1}^{m} |c_i|^2 P(A_i) = \|f\|^2 .
$$

For a general $f$, we can find a sequence of simple functions $f_n$ such that $f_n^2(\omega)$ is increasing to $f^2(\omega)$ as $n \to \infty$ for all $\omega \in X$. Then $Uf_n$ also increases to $Uf$ pointwise, and we can use the above identity along with the monotone convergence theorem to write:

$$
\|Uf\|^2 = \lim_{n \to \infty} \|Uf_n\|^2 = \lim_{n \to \infty} \|f_n\|^2 = \|f\|^2 .
$$

\[\Box\]

**Problem 3.1.** This problem is a generalization of Lemma 3.1. Let $(\Omega, F, P)$ be a probability space and $\theta$ be a $P$-preserving transformation of $\Omega$. Let $p \in [1, \infty)$. Prove that if $X$ is a random variable belonging to $L^p(\Omega, F, P)$, then $UX \in L^p(\Omega, F, P)$, and $\|UX\|_{L^p(\Omega, F, P)} = \|X\|_{L^p(\Omega, F, P)}$.

Lemma 3.1 shows that studying the convergence of ergodic averages in $L^2$ is possible via the asymptotic analysis of operators $\sum_{k=0}^{n-1} U^k$, where $U$ is an isometry of a complex Hilbert space.

The first version of the following $L^2$ ergodic theorem appeared in [vN32].

**Theorem 3.1.** Let $U$ be an isometry of a complex Hilbert space $\mathbb{H}$. Let $\pi_I$ be the orthogonal projection onto the space $I = \{ f \in \mathbb{H} : Uf = f \}$. Then, for every $f \in \mathbb{H}$,

$$
A_nf = \frac{1}{n} \sum_{k=0}^{n-1} U^k f
$$

converges to $\pi_I f$ as $n \to \infty$.

**Proof:** Let us denote $G = \{ g - Ug : g \in \mathbb{H} \}$ and prove that $I$ and the closure of $G$ are orthogonal complements of each other. If $\langle f, g - Ug \rangle = 0$ for some $f$ and all $g$, then $\langle f, g \rangle = \langle f, Ug \rangle$ for all $g$. In particular, $\langle f, f \rangle = \langle f, Uf \rangle$. Using this property and the isometry property, we obtain

$$
\langle Uf - f, Uf - f \rangle = \langle Uf, Uf \rangle - \langle Uf, f \rangle - \langle f, Uf \rangle - \langle f, f \rangle
= \|f\|^2 - \|f\|^2 - \|f\|^2 + \|f\|^2 = 0,
$$

so $Uf = f$. Thus $\overline{G} \subseteq I$. Since the isometry property can be rewritten as $\langle Uf, Ug \rangle = \langle f, g \rangle$ for all $f, g \in \mathbb{H}$, we can take any $f \in I$, $g \in \mathbb{H}$ and write

$$
\langle f, g - Ug \rangle = \langle f, g \rangle - \langle f, Ug \rangle = \langle f, g \rangle - \langle Uf, Ug \rangle = \langle f, g \rangle - \langle f, g \rangle = 0 .
$$

Therefore, $I \subseteq \overline{G}$, and thus $I = \overline{G}$. So, now we can decompose any $f \in \mathbb{H}$ as $f = f_I + f_G$, where $f_I \in I$ and $f_G \in G$ are orthogonal projections of $f$. 


onto \( I \) and \( \bar{G} \), and it is sufficient to see how the ergodic averaging acts on each of the spaces \( I \) and \( \bar{G} \).

If \( f \in I \), then \( A_n f = f \).

If \( f \in G \), i.e., if \( f = g - Ug \) for some \( g \in \mathbb{H} \), then

\[
\| A_n f \| = \left\| \frac{1}{n} \sum_{k=0}^{n-1} U^k (g - Ug) \right\| = \left\| \frac{1}{n} g - U^n g \right\| \leq \frac{2}{n} \| g \|
\]

Let now \( f \in \bar{G} \). Then for any \( \varepsilon > 0 \) there is \( f_\varepsilon = g_\varepsilon - Ug_\varepsilon \in G \) such that \( \| f - f_\varepsilon \| \leq \varepsilon \). Therefore,

\[
\| A_n f \| \leq \| A_n (f - f_\varepsilon) \| + \| A_n f_\varepsilon \| \leq \varepsilon + 2\| g_\varepsilon \| / n,
\]

where we used that \( A_n \) is a contraction, i.e., \( \| Ah \| \leq \| h \| \) for all \( h \in \mathbb{H} \). So, \( \limsup_{n \to \infty} \| A_n f \| \leq \varepsilon \), and since the choice of \( \varepsilon \) is arbitrary, we conclude that \( \limsup_{n \to \infty} \| A_n f \| = 0 \), and the proof is complete. \( \Box \)

Notice that in the context of measure-preserving transformations, the space \( I \) always contains all constant functions.

**Problem 3.2.** Prove that if there are no other almost invariant functions, i.e., the transformation is ergodic, then the spaces \( I \) and \( I^\perp \) are, respectively, the space of constants and the space of functions \( f \in L^2 \) with \( Ef = 0 \). Derive that in this case, the limit in the theorem is deterministic and equal \( Ef \).

In general, one can introduce the invariant \( \sigma \)-algebra \( F_I \) generated by all functions from \( I \). Then \( \pi_I f \) can be understood as the conditional expectation \( E(f|F_I) \).

The simplest cases where this theorem applies are rotations of Euclidean spaces \( \mathbb{R}^2 \) and \( \mathbb{R}^3 \) (or their complexifications).

Let us also briefly describe the spectral approach to this theorem based on von Neumann’s original idea. He proved that the group \( (U^n)_{n \in \mathbb{Z}} \) generated by a unitary operator \( U \) admits the following “spectral” representation:

\[
U^n = \int_{[-\pi, \pi)} e^{in\varphi} P(d\varphi), \quad n \in \mathbb{Z},
\]

for some projector-valued measure \( P \) on \( [-\pi, \pi) \). So, splitting the domain of integration into a one point set \( \{0\} \) and its complement, we obtain

\[
A_n = \int_{[-\pi, \pi)} \frac{1}{n} \sum_{k=0}^{n-1} e^{ik\varphi} P(d\varphi) = P(\{0\}) + \int_{[-\pi, \pi) \setminus \{0\}} e^{in\varphi} - \frac{1}{n(e^{i\varphi} - 1)} P(d\varphi),
\]

The integrand in the right-hand side is bounded by 1 and converges to 0, so \( A_n \) converges to \( P(\{0\}) \) which is exactly the projection onto the space of eigenvectors of \( U \) associated to eigenvalue \( e^{i\cdot 1} = 1 \), i.e., the set \( I \) of invariant vectors.

Another instance of this general approach is the law of large numbers for \( L^2 \)-stationary \( \mathbb{C} \)-valued processes. If \( X \) is such a process then there is a
number $a$ and an orthogonal random measure $Z$ on $[-\pi, \pi)$ such that
$$X_n = a + \int_{[-\pi, \pi)} e^{in\varphi} Z(d\varphi), \quad n \in \mathbb{Z}.$$ 
Therefore, we can write
$$\frac{1}{n}(X_0 + \ldots + X_{n-1}) = Z\{0\} + \int_{[-\pi, \pi)\setminus\{0\}} \frac{e^{in\varphi} - 1}{n(e^{i\varphi} - 1)} Z(d\varphi), \quad n \in \mathbb{N},$$
and conclude that
$$\frac{1}{n}(X_0 + \ldots + X_{n-1}) \rightarrow Z\{0\}, \quad n \rightarrow \infty.$$ 
Therefore, the limit is nonrandom and equals zero if the spectral measure has no atom at zero, i.e., $E|Z\{0\}|^2 = 0.$

2. Birkhoff’s pointwise ergodic theorem

The following theorem was first proved by G. Birkhoff in [Bir31].

**Theorem 3.2.** Let $(\Omega, \mathcal{F}, P)$ be a probability space. Suppose $\theta$ is a $P$-preserving measurable transformation, $\mathcal{I}$ is the $\sigma$-algebra of $\theta$-invariant sets, and $f \in L^1(\Omega, \mathcal{F}, P).$ Then, with probability 1,
$$A_nf \rightarrow E[f|\mathcal{I}], \quad n \rightarrow \infty,$$
where the ergodic averages $A_n$ are defined by (3.1) or, equivalently, by (3.2) and $Uf(\omega) = f(\theta\omega).$

If in addition $\theta$ is ergodic then with probability 1,
$$A_nf \rightarrow Ef, \quad n \rightarrow \infty.$$ 

There is a point of view that this is an extremely useful result with no transparent or enlightening proof. Nevertheless, the proof constructed by Y. Katznelson and B. Weiss [KW82] and based on the idea of the nonstandard analysis proof of T. Kamae [Kam82] is simple and conceptual. This is the proof we proceed to explain following closely the original paper [KW82].

**Proof:** Due to linearity of time averages and conditional expectations, it is sufficient to prove the convergence for nonnegative function $f.$ Let us introduce
$$\bar{f}(\omega) = \limsup_{n \rightarrow \infty} A_nf(\omega), \quad \underline{f}(\omega) = \liminf_{n \rightarrow \infty} A_nf(\omega), \quad \omega \in \Omega.$$ 
A priori we do not know if $\bar{f}$ and $\underline{f}$ are finite, but they are almost invariant. To see that, we write
$$|A_nf(\omega) - A_n(\theta\omega)| \leq \frac{1}{n}f(\omega) + \frac{1}{n}f(\theta^n\omega).$$ 
The first term on the right-hand side clearly converges to 0 as $n \rightarrow \infty$ for all $\omega.$ The second term converges to 0 almost surely since due to finiteness
of $Ef$, we obtain for every $\varepsilon > 0$,
\[ \sum_{n=0}^{\infty} P\{f(\theta^n \omega) > \varepsilon n\} < \infty, \]
so the Borel–Cantelli Lemma implies that $f(\theta^n \omega) > \varepsilon n$ at most for finitely many values of $n$.

We need to show that
\[(3.4) \quad E\bar{f}_B \leq Ef_B \leq E\tilde{f}_B, \quad B \in \mathcal{I}. \]
To see this, notice that (3.4) implies $P\{\bar{f} = \tilde{f}\} = 1$, so $P\{A_n f \to \tilde{f}\} = 1$ and $E\bar{f}_B = Ef_B$. Also $\tilde{f}$ is almost invariant and therefore, $T^*$-measurable.

Let us prove the first inequality in (3.4). To that end we will need to use several truncations. Let us take a large number $M > 0$ and define $\bar{f}_M(\omega) = \bar{f}(\omega) \wedge M$, $\omega \in \Omega$.

For all $\omega$, $\bar{f}_M(\omega)$ is finite and does not exceed $\bar{f}(\omega)$. Let us fix $\varepsilon > 0$, define $n(\omega) = \inf \{n : \bar{f}_M(\omega) \leq A_n f(\omega) + \varepsilon\}, \quad \omega \in \Omega,$
and notice that $n(\omega) < \infty$ for all $\omega$. Invariance of $\tilde{f}$ under $\theta$ implies invariance of $\bar{f}_M$, so $\bar{f}_M(\theta^j \omega) = \bar{f}_M(\omega)$ for all $\omega \in \Omega$ and $j \in \mathbb{N}$, and we get
\[(3.5) \quad \sum_{j=0}^{n(\omega)-1} \bar{f}_M(\theta^j \omega) = n(\omega) \bar{f}_M(\omega) \leq \sum_{j=0}^{n(\omega)-1} f(\theta^j \omega) + n(\omega) \varepsilon, \]
where in the second inequality we used the definition of $n(\omega)$. Notice that (3.5) is a lower bound for time averages of $f$ by time averages of $\bar{f}_M$ (an approximation to $\tilde{f}$), and to prove the first inequality in (3.4) we need to convert (3.5) into an integral inequality. If only the upper limit of summation were not random, we would have integrated both sides immediately, but $n$ is random. Nevertheless we shall try to obtain a version of (3.5) with a deterministic number $L$ replacing $n(\omega)$. The idea is to split the summation from 0 to $L - 1$ into random intervals such that on each of them we can apply (3.5). To achieve that, it is convenient though to work with an intermediate version of (3.5) where random variable $n$ is replaced by its bounded truncation.

Let us find $N > 0$ such that $P(C) > 1 - \varepsilon/M$, where $C = \{n(\omega) < N\}$, and introduce $f_{M, \varepsilon} = f1_C + (f \vee M)1_{C^c}$ and $\tilde{n} = n1_C + 1_{C^c}$. Then
\[(3.6) \quad \sum_{j=0}^{\tilde{n}(\omega)-1} \bar{f}_M(\theta^j \omega) = \tilde{n}(\omega) \bar{f}_M(\omega) \leq \sum_{j=0}^{\tilde{n}(\omega)-1} f_{M, \varepsilon}(\theta^j \omega) + \tilde{n}(\omega) \varepsilon, \]
because on $C$ this coincides with (3.5), and on $C^c$ we have $\bar{f}_M \leq M \leq f_{M, \varepsilon}$. 

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Let us now choose \( L > 0 \) large enough to guarantee \( NM/L < \varepsilon \) and for all \( \omega \in \Omega \) define \( n_0(\omega) = 0 \) and, inductively,
\[
    n_k(\omega) = n_{k-1}(\omega) + \tilde{n}(\theta^{n_{k-1}(\omega)}\omega), \quad k \in \mathbb{N}.
\]
The purpose of this is to split the sum from 0 to \( L-1 \) into intervals such that on each of them we can apply (3.6): we define \( k(\omega) = \max\{k : n_k(\omega) \leq L-1\} \) and write
\[
    \sum_{j=0}^{L-1} \bar{f}_M(\theta^j\omega) = \sum_{k=1}^{k(\omega)} \sum_{j=n_{k-1}(\omega)}^{n_k(\omega)-1} \bar{f}_M(\theta^j\omega) + \sum_{j=n_k(\omega)}^{L-1} \bar{f}_M(\theta^j\omega).
\]
Applying (3.6) to each of the \( k(\omega) \) terms and estimating the last special term by \( N\tilde{n}(\omega) \) (here we use the fact that \( \tilde{n}(\omega) \) is uniformly bounded by \( N \), and so is \( L - n_k(\omega) \)), we obtain
\[
    \sum_{j=0}^{L-1} \bar{f}_M(\theta^j\omega) \leq \sum_{j=0}^{L-1} f_{M,\varepsilon}(\theta^j\omega) + L\varepsilon + NM.
\]
Integrating both sides over \( B \) and dividing by \( L \) gives
\[
    \mathbb{E} \bar{f}_M 1_B \leq \frac{1}{L} \sum_{j=0}^{L-1} \mathbb{E} f_{M,\varepsilon}(\theta^j\omega) 1_B + \varepsilon + \frac{NM}{L} \leq \mathbb{E} f_{M,\varepsilon} 1_B + 2\varepsilon,
\]
where we used invariance of \( B \) and measure-preserving property of \( \theta \):
\[
    \mathbb{E} f_{M,\varepsilon}(\theta^j\omega) 1_{\{\omega \in B\}} = \mathbb{E} f_{M,\varepsilon}(\omega) 1_{\{\omega \in \theta^{-j}B\}} = \mathbb{E} f_{M,\varepsilon}(\omega) 1_{\{\omega \in B\}}.
\]
Since \( f_{M,\varepsilon} \leq f + M1_{C^c} \), we can estimate the right-hand side of (3.7) with
\[
    \mathbb{E} f_{M,\varepsilon} 1_B \leq \mathbb{E} f 1_B + MP(C^c) \leq \mathbb{E} f 1_B + \varepsilon,
\]
we obtain
\[
    \mathbb{E} \bar{f}_M 1_B \leq \mathbb{E} f 1_B + 3\varepsilon.
\]
Letting \( \varepsilon \to 0 \) and then \( M \to \infty \), we obtain the first inequality in (3.4).

To prove the second inequality in (3.4), we do not even need to introduce a truncation analogous to \( f_M \). We simply fix \( \varepsilon > 0 \), define
\[
    n(\omega) = \inf \{ n : f(\omega) \geq A_n f(\omega) - \varepsilon \}, \quad \omega \in \Omega,
\]
introduce \( C = \{n(\omega) > N\} \) where \( N \) is chosen so that \( \mathbb{E} f 1_{C^c} < \varepsilon \), define \( f_\varepsilon = f 1_{C} \) and \( \tilde{n} = n 1_{C} + 1_{C^c} \), and proceed in a similar way to the proof of the first inequality in (3.4).

To prove (3.3), it is sufficient to notice that in the ergodic case, \( I^* \) contains only sets of probability 0 and 1, so \( \mathbb{E}[f|I^*] \overset{a.s.}{=} \mathbb{E} f \).

\[\square\]

Problem 3.3. Check the details of the proof of the second inequality in (3.4).
2. BIRKHOFF’S POINTWISE ERGODIC THEOREM

Problem 3.4. Let $(\Omega, \mathcal{F}, P)$ be a probability space and let $\mathcal{G}$ be a $\sigma$-subalgebra of $\mathcal{F}$ such that for each $A \in \mathcal{G}$, $P(A) = 0$ or 1. Prove that $E[X|\mathcal{G}] \overset{a.s.}{=} EX$ for any random variable $X \in L^1(\Omega, \mathcal{F}, P)$.

Theorem 3.3 ($L^1$ ergodic theorem). Under the conditions of theorem 3.2,
\[ \lim_{n \to \infty} E|A_n f - E[f|\mathcal{I}]| = 0. \]

Proof: For $\varepsilon > 0$ we can find a bounded function $f_\varepsilon$ such that $E|f_\varepsilon - f| < \varepsilon$. The pointwise ergodic theorem applies to $f_\varepsilon$ and since the sequence $A_n f_\varepsilon$ is uniformly bounded, we obtain by bounded convergence theorem that
\[ \lim_{n \to \infty} E|A_n f_\varepsilon - E[f_\varepsilon|\mathcal{I}]| = 0, \]

i.e.,
\[ E|A_n f_\varepsilon - E[f_\varepsilon|\mathcal{I}]| < \varepsilon \]
for sufficiently large $n$. On the other hand, due to the measure preserving property,
\[ E|A_n f_\varepsilon - A_n f| \leq \frac{1}{n} \sum_{k=0}^{n-1} E|f_\varepsilon(\theta_k \omega) - f(\theta_k \omega)| \leq \frac{1}{n} \sum_{k=0}^{n-1} E|f_\varepsilon(\omega) - f(\omega)| \leq \varepsilon, \]
and due to the Jensen inequality for conditional expectations,
\[ E|E[f_\varepsilon|\mathcal{I}] - E[f|\mathcal{I}]| \leq EE|f_\varepsilon - f||\mathcal{I}| = E|f_\varepsilon - f| < \varepsilon. \]
Since $\varepsilon$ is arbitrary, the theorem follows from the last three inequalities. □

We can now look at the previous examples from the point of view of the ergodic theorem.

Theorem 3.4 (Kolmogorov’s strong law of large numbers, reference????). Let $(X_n)_{n \in \mathbb{N}}$ be a sequence of independent identically distributed random variables such that $X_1$ has a finite expectation $a$. Then
\[ \frac{1}{n} (X_1 + \ldots + X_n) \overset{a.s.}{\to} a, \quad n \to \infty. \]

Proof: This theorem is a direct consequence of Birkhoff’s ergodic theorem and ergodicity of the standard shift i.i.d. sequences. □

Thus the pointwise ergodic theorem can be viewed as a generalization of the strong law of large numbers.

In fact, the ergodic theorem provides the following description of an invariant measure in the ergodic case: the measure of a set $A$ equals the average occupation time for $A$, i.e., the average fraction of time spent by the trajectory in $A$:
\[ (3.9) \quad \lim_{n \to \infty} \frac{\sum_{j=0}^{n-1} 1_{\theta^j \omega \in A}}{n} = P(A). \]

Theorems 2.1 and 2.2 can also be understood as a corollaries of ergodicity of circle rotations (Theorem 2.10) and the pointwise ergodic theorem.
However, the latter provides convergence only for almost every \( \omega \) whereas the statements of those results, in fact, hold for all \( \omega \in S^1 \) without exception. This gap can be closed easily.

**Problem 3.5.** Follow these lines to construct a complete proof of Theorems 2.1 and 2.2.

One can use the ergodic theorem to give equivalent definition of ergodicity:

**Theorem 3.5.** Let \( \theta \) be a measure preserving transformation on \((\Omega, \mathcal{F}, \mathcal{P})\). The following statements are equivalent:

(a) \( \theta \) is ergodic;
(b) For any \( f \in L^1(\Omega, \mathcal{F}, \mathcal{P}) \), (3.3) holds;
(c) For any \( A \in \mathcal{F} \), (3.9) holds.

**Proof:** Condition (a) implies condition (b) by the ergodic theorem, (c) is a specific case of (b). To derive (a) from (b), we take any invariant set \( A \), use the invariance of \( A \) and of \( A^c \) to write \( S_n 1_A = 1_A \), and since \( A_n 1_A \) converges to \( \mathcal{P}(A) \) a.s., we conclude that \( 1_A \approx \mathcal{P}(A) \), and this can happen only if \( \mathcal{P}(A) = 0 \) or 1.

---

3. Kingman’s subadditive ergodic theorem

**Theorem 3.6 ([Kin73]).** Let \( \theta \) be a measure-preserving and ergodic transformation of a probability space \((\Omega, \mathcal{F}, \mathcal{P})\). Let \((X_n)_{n \in \mathbb{N}}\) be a sequence of \( L^1 \) random variables that satisfy

\[
X_{n+m}(\omega) \leq X_n(\omega) + X_m(\theta^n \omega), \quad n, m, \in \mathbb{N}.
\]

Let \( a = \inf_{n \in \mathbb{N}} (\mathbb{E}X_n/n) \). Then

\[
\frac{X_n}{n} \xrightarrow{a.s.} a, \quad n \to \infty.
\]

**Remark 3.1.** Extension of Birkhoff. Other generalizations are possible. We prove only ergodic case.


**Proof:** Our proof follows closely notes [Lal10] by S. Lalley. Let us first notice that applying property (3.10) inductively, we obtain that for any \( m \in \mathbb{N} \) any finite sequence \((n_k)_{k=1}^m\) of natural numbers,

\[
X_{n_1+\ldots+n_m}(\omega) \leq X_{n_1}(\omega) + X_{n_2}(\theta^{n_1} \omega) + \ldots + X_{n_m}(\theta^{n_1+\ldots+n_{m-1}} \omega).
\]

Defining \( X_0 \equiv 0 \), we see that this inequality remains true also if \((n_k)_{k=1}^m\) is a sequence of nonnegative integers.

Let us now observe that it is sufficient to consider nonpositive random variables. In fact, for a general sequence \((X_n)\) satisfying the conditions of the
theorem, we notice that the sequence of random variables \((Y_n)_{n \in \mathbb{N}}\) defined by

\[
Y_n(\omega) = X_n(\omega) - \sum_{k=0}^{n-1} X_1(\theta^k \omega),
\]

is also subadditive in the sense of (3.10), and all these variables are a.s.-nonpositive due to (3.12). If the theorem holds true for \(Y_n(\omega)\) then it also holds for \(X_n\) since we can apply Birkhoff’s ergodic theorem to the second term in the definition of \(Y_n\). So from now on we may assume that \(X_n \leq 0\).

If \(a > -\infty\), we need to establish the following two inequalities:

\[
\lim_{n \to \infty} \inf \frac{X_n(\omega)}{n} \overset{a.s.}{\geq} a, \quad (3.13)
\]

\[
\lim_{n \to \infty} \sup \frac{X_n(\omega)}{n} \leq a. \quad (3.14)
\]

If \(a = -\infty\), it is sufficient to prove only (3.14).

Let us start with (3.13). We denote its left-hand side by \(b(\omega) \in [-\infty, 0]\) and claim that it is a.s.-constant. In fact, since (3.10) implies \(X_{n+1}(\omega) \leq X_1(\omega) + X_n(\theta \omega)\), by dividing both sides by \(n\) and taking the lim inf we obtain \(b(\omega) \leq b(\theta \omega)\) for all \(\omega\).

**Problem 3.6.** Use the invariance of \(P\) under \(\theta\) to prove that \(b(\omega) \overset{a.s.}{=} b(\theta \omega)\).

So, \(b\) is \(\theta\)-invariant and, due to ergodicity, has to be a.s.-constant with respect to \(P\).

Constants \(a\) and \(b\) are both nonpositive. Let us prove that if \(a > -\infty\), then \(b \geq a\). Suppose that \(b < a - \varepsilon\) for some \(\varepsilon > 0\). If

\[
B_m = \{\omega : \min_{1 \leq n \leq m} X_n(\omega)/n \leq a - \varepsilon\}, \quad m \in \mathbb{N},
\]

then for any \(\delta > 0\), we can find \(m \in \mathbb{N}\) such that \(P(B_m) > 1 - \delta\). Let us fix an \(\omega\) and take an arbitrary \(n \in \mathbb{N}\). Let \(R = \{j \in \{1, \ldots, nm\} : \theta^j \omega \in B_m\}\). By definition of \(R\), for every \(j \in R\) there is \(k = k(j) \in \{1, \ldots, m\}\) such that \(X_k(\theta^j \omega) \leq k(a - \varepsilon)\).

Let us now define two finite sequences \(j_1, \ldots, j_r\) and \(k_1, \ldots, k_r\) depending on \(\omega\). First, we let \(j_1\) be the smallest number in \(R\) and \(k_1 = k(j_1)\). Then we inductively set \(j_i\) to be the smallest number in \(R\) exceeding \(j_{i-1} + k_{i-1} - 1\) and \(k_i = k(j_i)\) until we get to a number \(j_r \leq nm\) such that \(k_r = k(j_r) > \max R\). We have

\[
R \subset \bigcup_{i=1}^{r} \{j_i, j_i + k_i - 1\}, \quad (3.15)
\]

where the intervals in the union on the right-hand side are mutually disjoint.

Applying (3.12) to the sequence

\[
j_1, k_1, j_2 - (j_1 + k_1), k_2, j_3 - (j_2 + k_2), \ldots, k_r, nm + m - (j_r + k_r),
\]

we get

\[
\lim_{n \to \infty} \frac{X_n(\omega)}{n} \overset{a.s.}{\geq} a - \varepsilon > a - \varepsilon > a,
\]

and this contradicts our assumption that \(b < a - \varepsilon\). Therefore, \(b \geq a\) for all \(\omega\) and the proof of (3.13) is complete.

For (3.14), let us consider the set

\[
C_m = \{\omega : \max_{1 \leq n \leq m} X_n(\omega)/n < a\}, \quad m \in \mathbb{N}.
\]

By the ergodic theorem, we can find an \(\omega\) such that \(P(C_m) > 1 - \delta\). Then for any \(n \in \mathbb{N}\), we have

\[
\frac{1}{n} \sum_{k=0}^{n-1} X_1(\theta^k \omega) \leq \frac{1}{n} \sum_{k=0}^{n-1} X_n(\theta^k \omega) = \frac{X_n(\omega)}{n} \leq a - \varepsilon.
\]

Let \(M = \max_{1 \leq n \leq m} X_n(\omega)/n\). Then

\[
M = \frac{1}{n} \sum_{k=0}^{n-1} X_1(\theta^k \omega) \leq a - \varepsilon,
\]

which contradicts our assumption that \(M < a\). Therefore, \(M \geq a\) for all \(\omega\) and the proof of (3.14) is complete.

Hence, we have established that \(b \overset{a.s.}{=} b(\theta \omega)\) for all \(\omega\), and the proof of the theorem is complete.
and throwing out nonpositive terms corresponding to $j_1, j_2 - (j_1 + k_1), \ldots, nm + m - (j_r + k_r)$, we obtain
\[ X_{nm+m}(\omega) \leq X_{k_1}(\theta^{j_1}\omega) + X_{k_2}(\theta^{j_2}\omega) + \ldots + X_{k_r}(\theta^{j_r}\omega). \]
We have $X_{k_i}(\theta^{j_i}\omega) < k_i(a - \varepsilon)$ by definition of $j_i, k_i, i = 1, \ldots, r$, so
\[
(3.16) \quad X_{nm+m}(\omega) \leq (k_1 + \ldots + k_r)(a - \varepsilon) \leq (a - \varepsilon) \sum_{i=1}^{nm} 1_B(\theta^i\omega),
\]
where the last inequality follows from (3.15) and $a - \varepsilon < 0$. Let us now divide both sides of (3.16) by $nm + m$ and take expectations:
\[
E \frac{X_{nm+m}}{nm + m} \leq (a - \varepsilon) \frac{nm}{nm + m} P(B_m).
\]
Recalling that $P(B_m) > 1 - \delta$, letting $n \to \infty$ and using the definition of $a$, we see that
\[
a \leq (a - \varepsilon)(1 - \delta),
\]
and we obtain a contradiction if we choose $\delta$ sufficiently small.

Let us prove (3.14). If all the maps $\theta^m, m \in \mathbb{N}$ are ergodic, then we can use (3.12) to write
\[
X_{nm+k}(\omega) \leq \sum_{j=0}^{n-1} X_m(\theta^{jm}\omega) + X_k(\theta^{nm}\omega)
\]
for $n, m \in \mathbb{N}$ and $k \in \{1, \ldots, m\}$, divide both sides by $nm$, use $X_1 \leq 0$, let $n \to \infty$ and apply Birkhoff’s ergodic theorem to the right-hand side to see that
\[
(3.17) \quad \limsup_{n \to \infty} \frac{X_n(\omega)}{n} \leq \frac{E X_m}{m},
\]
and (3.14) follows. However, $\theta^m$ does not have to be ergodic.

**Problem 3.7.** Give an example of an ergodic transformation $\theta$ such that $\theta^2$ is not ergodic.
So we need to introduce additional averaging to tackle this difficulty. We use (3.12) to write the following $m$ inequalities:

\[ X_{nm+k}(\omega) \leq \sum_{j=0}^{n-2} X_m(\theta^{jm} \omega) + X_{m+k}(\theta^{(n-1)m} \omega) \]

\[ X_{nm+k}(\omega) \leq X_1(\omega) + \sum_{j=0}^{n-2} X_m(\theta^{jm+1} \omega) + X_{m+k-1}(\theta^{(n-1)m+1} \omega), \]

\[ X_{nm+k}(\omega) \leq X_2(\omega) + \sum_{j=0}^{n-2} X_m(\theta^{jm+2} \omega) + X_{m+k-2}(\theta^{(n-1)m+2} \omega), \]

\[ \ldots \]

\[ X_{nm+k}(\omega) \leq X_{m-1}(\omega) + \sum_{j=0}^{n-2} X_m(\theta^{jm+m-1} \omega) + X_{k+1}(\theta^{(n-1)m+k-1} \omega). \]

Let us take the average of these inequalities, use nonpositivity of all the random variables involved:

\[ X_{nm+k}(\omega) \leq \frac{1}{m} \sum_{j=0}^{(n-1)m-1} X_m(\theta^j \omega), \]

Birkhoff’s ergodic theorem now applies to the sum on the right-hand side, so dividing both sides by $nm$ and taking $n \to \infty$, we obtain (3.17). Since $m$ is arbitrary, our claim follows. \qed
CHAPTER 4

Invariant measures

1. Existence of invariant measures

In the last several sections we derived basic dynamical properties of measure-preserving transformations. In particular, we obtained that in the basic framework we have worked with, the statistical properties of dynamics expressed in the form of time averages over long trajectories are described in terms of the invariant measure and the $\sigma$-algebra of almost invariant sets.

In this chapter we change the point of view and notice that a typical situation that often leads to interesting and hard problems, is that a priori we are given only a measurable space $(\Omega, \mathcal{F})$ and a measurable transformation $\theta$ on it. In this case, if we want to study statistical properties of the dynamics with the help of the theory developed above, we first need to find invariant measures, and among them find ergodic ones and only then we may be able to deduce statistical information about trajectories of the system. In general, our conclusions may be different for different initial conditions since there may be several distinct ergodic measures and a function may produce different averages with respect to all these measures.

So we see that it is fundamental to describe the family of all invariant distributions and specifically all ergodic distributions. Notice that here we have made a terminological switch from ergodic transformations to ergodic measures.

First of all we notice that the pushforward of a measure is a linear operator: if $\mu_1$ and $\mu_2$ are finite signed measures (not necessarily probability measures) and $a_1, a_2$ are two numbers, then $a_1\mu_1 + a_2\mu_2$ is also a finite signed measure, and its pushforward under $\theta$ is easily seen to coincide with $a_1\mu_1\theta^{-1} + a_2\mu_2\theta^{-1}$. Therefore, we can speak about a linear operator in the space of finite measures and see straightforwardly that an invariant distribution plays the role of an eigenvector of this operator corresponding to eigenvalue 1. So, we may be interested in studying the structure of the set of all such eigenvectors.

In general, there is a variety of situations that differ from each other in nature.

Even existence of an invariant measure is sometimes hard to establish. However, a compactness argument due to Bogolyubov and Krylov (reference) may often be employed. We give a concrete result here, but we will see developments of this approach throughout these notes.
**Theorem 4.1.** Let $\Omega$ be a metric space equipped with Borel $\sigma$-algebra $\mathcal{B}(\Omega)$ and a measurable transformation $\theta$ on $(\Omega, \mathcal{B}(\Omega))$. Suppose there is a compact set $K$ that is forward invariant under $\theta$, and $\theta$ is continuous on $K$. Then there is an invariant probability measure supported on $K$.

**Proof:** First let us notice that it is sufficient to consider the case where $K = \Omega$. Let now $P$ be any probability measure on $\Omega$. Let $P_k = P^{\theta^{-k}}$, $k \in \mathbb{Z}_+$ and $\bar{P}_n = \frac{1}{n} \sum_{k=0}^{n-1} P_k$. We would like to extract a weakly convergent subsequence from $(\bar{P}_n)_{n \in \mathbb{Z}_+}$. Let us recall that weak convergence $\mu_n \Rightarrow \mu$ of probability measures $(\mu_n)_{n \in \mathbb{Z}_+}$ to a probability measure $\mu$ means that for every bounded continuous function $f: \Omega \to \mathbb{R}$

$$\int_\Omega f(\omega) \mu_n(\omega) \to \int_\Omega f(\omega) \mu(\omega), \quad n \to \infty.$$ 

The Prokhorov theorem (see [Bil68] for a classical introduction to the theory of weak convergence of measures) states that if for any $\varepsilon > 0$ there is a compact set $C$ such that $\mu_n(C) > 1 - \varepsilon$ for all $n$, then one can extract a weakly convergent subsequence $\mu_{n'}$ from $\mu_n$. In our case, for any $\varepsilon > 0$ we can choose $C = \Omega$, so any sequence of measures on $\Omega$ contains a convergent subsequence. So let us choose a sequence $n_i \to \infty$ such that $\bar{P}_{n_i} \Rightarrow \bar{P}$ for some probability $\bar{P}$, $i \to \infty$. Let us prove that $\bar{P}$ is $\theta$-invariant. It is sufficient to check

$$\int_\Omega f(\omega) \bar{P}(d\omega) = \int_\Omega f(\theta \omega) \bar{P}(d\omega),$$

for all continuous bounded functions $f$ since that integrals of functions from that set determines a measure uniquely. For such a function $f$,

$$\int_\Omega f(\omega) \bar{P}(d\omega) = \lim_{i \to \infty} \int_\Omega f(\omega) \bar{P}_{n_i}(d\omega)$$

$$= \lim_{i \to \infty} \frac{1}{n_i} \sum_{k=0}^{n_i-1} \int_\Omega f(\omega) P_k(d\omega)$$

$$= \lim_{i \to \infty} \left[ \frac{1}{n_i} \sum_{k=1}^{n_i} \int_\Omega f(\omega) P_k(d\omega) + \frac{1}{n_i} \int_\Omega f(\omega) \rho(d\omega) - \frac{1}{n_i} \int_\Omega f(\omega) P_{n_i}(d\omega) \right]$$

$$= \lim_{i \to \infty} \frac{1}{n_i} \sum_{k=0}^{n_i-1} \int_\Omega f(\theta \omega) P_k(d\omega) = \int_\Omega f(\theta \omega) \bar{P}(d\omega),$$

which completes the proof. \hfill $\square$

2. Structure of the set of invariant measures.

**Theorem 4.2.** Let $\theta$ be a measurable transformation on $(\Omega, \mathcal{F})$. Suppose $\bar{P}_1$ and $\bar{P}_2$ are two distinct $\theta$-invariant and ergodic measures on $(\Omega, \mathcal{F})$. Then they are singular to each other.
2. Structure of the Set of Invariant Measures.

**Proof:** Let us take a bounded random variable $X$ such that $E_{P_1}X \neq E_{P_2}X$. By Birkhoff's ergodic theorem, there are sets $A_1, A_2 \in \mathcal{F}$ such that $P_1(A_1) = P_2(A_2) = 1$, and that if $i \in \{1, 2\}$, then

$$\frac{1}{n} \sum_{k=0}^{n-1} X(\theta^k \omega) \to E_{P_i}X, \quad \omega \in A_i.$$

Therefore, $A_1 \cap A_2 = 0$, and $P_1$ and $P_2$ are mutually singular. □

For a measurable space with a measurable transformation $(\Omega, \mathcal{F}, \theta)$, the set $I(\Omega, \mathcal{F}, \theta)$ of invariant probability measures is a convex set, i.e., for every $Q_1, Q_2 \in I(\Omega, \mathcal{F}, \theta)$ and every $\alpha \in [0, 1]$, the measure $P = \alpha Q_1 + (1 - \alpha) Q_2$ is also $\theta$-invariant, as a straightforward computations shows. Let us study the extreme points of $I(\Omega, \mathcal{F}, \theta)$. A point $\mu \in I(\Omega, \mathcal{F}, \theta)$ is called an extreme point if conditions $Q_1, Q_2 \in I(\Omega, \mathcal{F}, \theta)$, $\alpha \in (0, 1)$, and

$$(4.1) \quad P = \alpha Q_1 + (1 - \alpha) Q_2,$$

imply $Q_1 = Q_2 = P$.

**Theorem 4.3.** A measure $P$ is an extreme point of $I(\Omega, \mathcal{F}, \theta)$ iff $(\Omega, \mathcal{F}, P, \theta)$ is ergodic.

**Proof:** Let $(\Omega, \mathcal{F}, P, \theta)$ be not ergodic. Then there is an invariant set $B$ with $P(B) \in (0, 1)$. Then $B^c$ is also invariant. Let us define measures $Q_1$ and $Q_2$ by $Q_1(A) = P(A|B)$ and $Q_2(A) = P(A|B^c)$, and set $\alpha = P(B)$. Then $(4.1)$ is satisfied, and both measures $Q_1, Q_2$ are invariant: for any set $A$,

$$Q_1(\theta^{-1} A) = P(\theta^{-1} A|B) = \frac{P(\theta^{-1} A \cap B)}{P(B)} = \frac{P(\theta^{-1} A \cap \theta^{-1} B)}{P(B)} = \frac{P(\theta^{-1} A \cap B)}{P(B)} = \frac{P(A \cap B)}{P(B)} = Q_1(A),$$

and a similar computation applies to $Q_2$. Also, $Q_1 \neq P$ since $Q_1(B) = 1 \neq P(B)$. Therefore, $P$ is not an extreme point.

Next, let $P$ be ergodic and let decomposition $(4.1)$ hold with $\alpha \in (0, 1)$ and invariant measures $Q_1, Q_2$. Let us take any bounded measurable function $f$. Ergodic theorem implies that for a set $A$ such that $P(A) = 0$ and

$$\lim_{n \to \infty} \frac{1}{n} \sum_{k=0}^{n-1} f(\theta^k \omega) = \int_{\Omega} f(\omega)P(d\omega), \quad \omega \in A^c.$$

Since measures $Q_1$ and $Q_2$ are absolutely continuous with respect to $P$, $Q_1(A^c) = Q_2(A^c) = 1$, so we conclude that convergence in $(4.2)$ holds almost surely with respect to $Q_1$ and $Q_2$, so by ergodic theorem,

$$\int_{\Omega} f(\omega)P(d\omega) = E_{Q_1}[f|\mathcal{F}], \quad \text{Q.a.s.}$$
Integrating this identity with respect to $Q_1$ we obtain

$$
\int_{\Omega} f(\omega)P(d\omega) = \int_{\Omega} f(\omega)Q_1(d\omega),
$$

and since $f$ was an arbitrary bounded measurable function, we conclude that $P = Q_1$. Similarly, we obtain $P = Q_2$, so $P$ is an extreme point.

In general, every $\theta$-invariant probability measure can be represented as a mixture of ergodic $\theta$-invariant measures:

$$
P = \int_{\mathcal{E}} P\mu(dP),
$$

where $\mu$ is a measure on $E$, the set of all ergodic measures.

Example with two circles and cylinder rotations.

In the remainder of this section we will make sense of this statement.

One could use an abstract Choquet theorem that says that every point of a compact convex subset of a locally convex topological vector space can be represented as a mixture (integral convex combination) of its extreme points. We will take another, more general and constructive approach. We follow [Sar09] and [EW11].

**Theorem 4.4.**

(1) Let $P$ be a probability measure on a Borel space $(\Omega, \mathcal{F})$ and let $\mathcal{G} \subset \mathcal{F}$ be a $\sigma$-algebra on $\Omega$. Then there is a set $\Omega' \in \mathcal{F}$ satisfying $P(\Omega') = 1$ and a probability kernel $P_\mathcal{G}(\cdot, \cdot)$ from $(\Omega', \mathcal{G})$ to $(\Omega, \mathcal{F})$ with the following property: if $X \in L^1(\Omega, \mathcal{F}, P)$, then

$$
E(X|\mathcal{G})(\omega) \overset{a.s.}{=} \int_{\Omega} X(\sigma)P_\mathcal{G}(\omega, d\sigma).
$$

(2) Suppose additionally $\mathcal{G}$ is countably generated, i.e., $\mathcal{G} = \sigma(E_k, k \in \mathbb{N})$ for some countable family of sets $E_k \in \mathcal{F}, k \in \mathbb{N}$. Then the set $\Omega'$ in the first part of this theorem can be chosen so that for all $\omega \in \Omega'$, $P_\mathcal{G}(\omega, [\omega]_\mathcal{G}) = 1$, where $[\omega]_\mathcal{G} = \bigcap_{A \in \mathcal{G} \in \mathcal{A} \in \mathcal{G}} A$

is the atom of $\mathcal{G}$ containing $\omega$. Moreover, if $\omega_1, \omega_2 \in \Omega'$ and $[\omega_1]_\mathcal{G} = [\omega_2]_\mathcal{G}$, then $P_\mathcal{G}(\omega_1, \cdot) = P_\mathcal{G}(\omega_2, \cdot)$.

(3) If $\mathcal{H} \subset \mathcal{F}$ is another $\sigma$-algebra such that $\mathcal{G} \supseteq \mathcal{H}$ (i.e., for every set $A \in \mathcal{G}$ there is a set $B \in \mathcal{H}$ such that $P(A \triangle B) = 0$, and for every set $A \in \mathcal{H}$ there is a set $B \in \mathcal{G}$ such that $P(A \triangle B) = 0$), then $P_\mathcal{G}(\omega, \cdot) = P_\mathcal{H}(\omega, \cdot)$ for almost all $\omega$.

**Remark 4.1.** The kernel $P$ is called regular conditional probability with respect to $\mathcal{G}$, see [Shi96, Section II.7]

**Proof:** Using the definition of Borel spaces, we can assume that $\Omega$ is a compact segment of $\mathbb{R}$ and $\mathcal{F}$ is a restriction of Borel $\sigma$-algebra onto $\Omega$. Under this assumption we can find a dense countable set in $C(\Omega)$. The
algebra $R$ generated by this set over $\mathbb{Q}$ is also a countable set. For every 
$Y \in R$, we can find a version $G[Y]$ of $E(Y|\mathcal{G})$. Let us now consider the 
following countable set of conditions:

1. $G[pY + qZ](\omega) = pG[Y](\omega) + qG[Z](\omega), \quad p, q \in \mathbb{Q}, Y, Z \in R.$
2. $\min Y \leq G[Y](\omega) \leq \max Y, \quad Y \in R.$

Each of these conditions is violated on an exceptional set of zero measure 
belonging to $\mathcal{G}$. Since there are countably many of these conditions, there is 
a set $A \in \mathcal{G}$ such that $P(A) = 1$ and all the conditions above are satisfied for 
all $\omega \in A$. In particular, for each $\omega \in A$, the functional $Y \mapsto G[Y](\omega)$ is a 
linear functional on $R$ over $\mathbb{Q}$, with norm bounded by 1. So, this functional 
can be extended by continuity to a continuous linear functional on $C(\Omega)$ 
over $\mathbb{R}$, in a unique way. Such functionals can be identified with measures 
by Riesz’s theorem (reference??). Thus for each $\omega \in A$, we obtain a 
measure $P_\omega(\omega, \cdot)$ on $(\Omega, \mathcal{F}, P)$. This measure satisfies

\begin{equation}
\int_{\Omega} Y(\sigma) P(\omega, d\sigma) = G[Y](\omega), \quad \omega \in A, \ Y \in C(\Omega).
\end{equation}

In particular the left-hand side of this identity is a $\mathcal{G}$-measurable function 
for every $Y \in C(\Omega)$.

To see that for every $B \in \mathcal{F}$, $P_\omega(\cdot, B)$ is a $\mathcal{G}$-measurable function it is sufficient 
to represent indicators $1_B$ as pointwise limits of continuous uniformly 
bounded functions $Y_n$, because then by dominated convergence,

$$P_\omega(\omega, B) = \lim_{n \to \infty} \int_{\Omega} Y_n(\sigma) P_\omega(\omega, d\sigma),$$

and a limit of $\mathcal{G}$-measurable functions is a $\mathcal{G}$-measurable function.

Let $\mathcal{A}$ be the collection of sets such that their indicators can be represented 
as pointwise limits of continuous uniformly bounded functions. Then 
$\mathcal{A}$ is a set algebra and a monotone class (see [Shi96, Section II.2] for these notions). Therefore it is a $\sigma$-algebra. Since $\mathcal{A}$ contains open sets, it coincides 
with Borel $\sigma$-algebra. Therefore, our claim that for every $B \in \mathcal{F}$, $P_\omega(\cdot, B)$ 
is a $\mathcal{G}$-measurable function holds true.

**Problem 4.1.** Let $X \in L^1(\Omega, \mathcal{F}, P)$. Prove that there is a sequence 
$X_n \in C(\Omega), n \in \mathbb{N}$ such that $X \overset{a.s.}{=} \sum_{n=1}^{\infty} X_n$ and $\sum_{n=1}^{\infty} \|X_n\|_{L^1(\Omega, \mathcal{F}, P)} < \infty$.

Using this problem, we can write

$$E(X|\mathcal{G})(\omega) \overset{a.s.}{=} \sum_{n=1}^{\infty} E(X_n|\mathcal{G})(\omega) \overset{a.s.}{=} \sum_{n=1}^{\infty} \int_{\Omega} X_n(\sigma) P_\omega(\omega, d\sigma)$$

$$\overset{a.s.}{=} \int_{\Omega} \sum_{n=1}^{\infty} X_n(\sigma) P_\omega(\omega, d\sigma) \overset{a.s.}{=} \int_{\Omega} X(\sigma) P_\omega(\omega, d\sigma).$$

The first identity holds since conditional expectation is a continuous linear 
operator in $L^1$. The second identity follows from (4.4) and continuity of $X_n$. 
The last identity follows from the definition of $X_n$, so it remains to explain
the change of order of integration and summation in the third identity. This
will be justified if we can prove that \( \int_{\Omega} \sum_{n=1}^{\infty} |X_n(\sigma)| P_G(\omega, d\sigma) \) is a.s.-finite. This will follow from

\[
E \int_{\Omega} \sum_{n=1}^{\infty} |X_n(\sigma)| P_G(\omega, d\sigma) < \infty.
\]

Identity

\[
E \int_{\Omega} Y(\sigma) P_G(\omega, d\sigma) = E Y
\]
holds for all \( Y \in \mathbb{R} \) by definition and thus for all \( Y \in C(\Omega) \), so we use it
along with the monotone convergence theorem to see that the left-hand side
of (4.5) equals

\[
\lim_{m \to \infty} E \int_{\Omega} \sum_{n=1}^{m} |X_n(\sigma)| P_G(\omega, d\sigma) \leq \lim_{m \to \infty} m \sum_{n=1}^{m} E|X_n(\sigma)| \leq \sum_{n=1}^{\infty} \|X_n\|_{L^1} < \infty,
\]

which completes the proof of part 1 with \( \Omega' = A \). We will need to adjust \( \Omega' \) in the proof of part 2 that follows. First,

\[
P(E_i|G)(\omega) \overset{a.s.}{=} 1_{E_i}(\omega) \overset{a.s.}{=} P_G(\omega, E_i), \quad i \in \mathbb{N},
\]
where \( P(E_i|G)(\omega) \overset{a.s.}{=} 1_{E_i}(\omega) \) follows by definition of conditional expectation,
and \( P(E_i|G)(\omega) \overset{a.s.}{=} P_G(\omega, E_i) \) is a specific case of (4.3).

Let \( N \supset A^c \) be the union of all exceptional sets in this identity for \( i \in \mathbb{N} \). Then for all \( i \in \mathbb{N} \),

\[
P_G(\omega, E_i) = \begin{cases} 
1, & \omega \in E_i \cap N^c, \\
0, & \omega \in E_i^c \cap N^c, \\
?, & \omega \in N.
\end{cases}
\]

Under the conditions of part 2

\[
[\omega]_G = \bigcap_{i: \omega \in E_i} E_i \cap \bigcap_{i: \omega \notin E_i} E_i^c, \quad \omega \in \Omega.
\]

**Problem 4.2.** Prove representation (4.7). Deduce that \([\omega]_G \) belongs
to \( G \) and, in fact, can be defined as the smallest element of \( G \) containing \( \omega \).

Identities (4.6) and (4.7) now imply \( P_G(\omega, [\omega]_G) = 1, \omega \in N^c \).

Since for any \( B \in \mathcal{F} \) the map \( \omega \mapsto P_G(\omega, B) \) is \( G \)-measurable, \( P_G(\omega_1, B) = P_G(\omega_2, B) \) for any \( \omega_1, \omega_2 \) satisfying \([\omega_1]_G = [\omega_2]_G \). So part 2 holds with
\( \Omega' = N^c \).

To prove part 3, we introduce \( A = \sigma(G, \mathcal{H}) \). Then we take a countable
dense set \( \{X_n\}_{n \in \mathbb{N}} \) in \( C(\Omega) \) and notice that for each \( n \in \mathbb{N} \), \( E(X_n|G) \) and
\( E(X_n|\mathcal{H}) \) are versions of \( E(X_n|A) \), so they coincide almost surely. Therefore,
there is a set \( \Omega' \) of full measure such that

\[
\int P_G(\omega, d\sigma) X_n(\sigma) = \int P_H(\omega, d\sigma) X_n(\sigma), \quad n \in \mathbb{N}, \omega \in \Omega'.
\]
Since measures on $\Omega$ are uniquely defined by the integrals of $X_n$ with respect to them, we conclude that $P_G(\omega, \cdot) = P_H(\omega, \cdot)$ on $\Omega'$.

**Theorem 4.5.** If $P$ is a probability measure on a Borel space $(\Omega, F)$ and $G \subset F$ is a $\sigma$-algebra, then there is a countably generated $\sigma$-algebra $H$ such that $G = H$.

**Proof:** Using the Borel isomorphism we can assume that $\Omega$ is a compact interval on $\mathbb{R}$. Then $L^1(\Omega, F, P)$ is a separable space and thus so is its subset $\{1_A | A \in G\}$. Therefore, there is a family $(A_n)_{n \in \mathbb{N}}$ of sets in $G$ such that all sets from $G$ are approximated by sets $A_n$ with arbitrary precision. Establishing relation (4.8)

$$\sigma(A_n, n \in \mathbb{N}) = G$$

will complete the proof.

**Problem 4.3.** Prove relation (4.8).

**Theorem 4.6.** Let $P$ be a probability measure on a Borel space $(\Omega, F)$, invariant under a measurable transformation $\theta$. Let $P_I(\omega, \cdot)$ be a regular conditional probability with respect to $I$. Then for $P$-almost every $\omega$, the measure $P_I(\omega, \cdot)$ is $\theta$-invariant and ergodic, and $P$ is a mixture or convex combination of those ergodic measures:

$$P = \int \Omega P(\omega) P_I(\omega, \cdot),$$

i.e.,

$$P(A) = \int \Omega P(\omega) P_I(\omega, A), \quad A \in F,$$

and, more generally,

$$EX = \int \Omega P(\omega) \int \Omega P_I(\omega, d\sigma) X(\sigma), \quad X \in L^1(\Omega, F, P).$$

**Proof:** Checking (4.10) is straightforward: according to Theorem 4.4,

$$EX = EE(X|I) = E \int \Omega X(\sigma) P_I(\omega, d\sigma), \quad X \in L^1(\Omega, F, P).$$

From now on we use the Borel isomorphism to assume that $(\Omega, F)$ is a compact segment on $\mathbb{R}$ with Borel $\sigma$-algebra. We choose a dense set $D = \{X_n\}_{n \in \mathbb{N}}$ in $C(\Omega)$.

Let us prove invariance of $P_I(\omega, \cdot)$. By Birkhoff’s ergodic theorem, there is a set $\Omega_1 \in F$ such that $P(\Omega_1) = 1$ and for all $n \in \mathbb{N}$ and $\omega \in \Omega_1$,

$$\int \Omega P_I(\omega, d\sigma) X_n(\sigma) = E(X_n|I)(\omega) = \lim_{N \to \infty} \frac{1}{N} \sum_{k=0}^{N-1} X_n(\theta^k \omega)$$
and
\[ \int_{\Omega} P_{\mathcal{I}}(\omega, d\sigma) X_n(\theta \sigma) d\sigma = E(X_n \circ \theta | \mathcal{I})(\omega) = \lim_{N \to \infty} \frac{1}{N} \sum_{k=0}^{N-1} X_n(\theta^{k+1} \omega). \]

Since \( X_n \) is bounded, the right-hand sides of the last two identities coincide for \( \omega \in \Omega_1 \). Since every \( X \in C(\Omega) \) can be uniformly approximated by functions from \( D \), we obtain that
\[ (4.12) \quad \int_{\Omega} P_{\mathcal{I}}(\omega, d\sigma) X(\sigma) = \int_{\Omega} P_{\mathcal{I}}(\omega, d\sigma) X(\theta \sigma) \]
holds for \( X \in C(\Omega) \) and \( \omega \in \Omega_1 \). For every set \( B \in \mathcal{F} \) there is a sequence of uniformly bounded continuous functions convergent pointwise to \( 1_B \). Bounded convergence implies that (4.12) holds for \( X = 1_B \) which proves that \( \theta \) preserves \( P(\omega, \sigma) \) for \( \omega \in \Omega_1 \).

Let us prove ergodicity of almost all measures \( P_{\mathcal{I}}(\omega, \sigma) \). It is sufficient to prove ergodicity of almost all measures \( P_\mathcal{H}(\omega, \sigma) \) where \( \mathcal{H} \) is a countably generated \( \sigma \)-algebra such that \( \mathcal{I} \subseteq \mathcal{H} \) (the existence of such \( \mathcal{H} \) is guaranteed by Theorem 4.5). The advantage of dealing with \( \mathcal{H} \) is that part 2 of Theorem 4.4 applied to \( \mathcal{H} = \mathcal{G} \) holds on a set \( \Omega' \) of full measure.

**Problem 4.4.** Let \( (\Omega, \mathcal{F}, \mathcal{P}, \theta) \) be a metric dynamical system, where \( (\Omega, \mathcal{F}) \) is a compact subset of \( \mathbb{R} \) with Borel \( \sigma \)-algebra on it. Prove that in this setting ergodicity is equivalent to the following statement: there is a countable dense set \( \{X_n\}_{n \in \mathbb{N}} \) in \( C(\Omega) \) such that
\[ \frac{1}{N} \sum_{i=0}^{N-1} X_n(\theta^i \omega) \xrightarrow{\text{as} N \to \infty} \int_{\Omega} X_n(\omega) \mathcal{P}(d\omega). \]

Using Birkhoff’s ergodic theorem we can find \( \Omega_2 \subseteq \Omega_1 \) such that \( \mathcal{P}(\Omega_2) = 1 \) and for \( \omega \in \Omega_2 \) and any \( n \in \mathbb{N} \)
\[ \frac{1}{N} \sum_{i=0}^{N-1} X_n(\theta^i \omega) \to \int_{\Omega} P_\mathcal{H}(\omega, d\sigma) X_n(\sigma), \quad N \to \infty. \]
Therefore, there is a set \( \Omega_3 \subseteq \Omega_2 \) such that \( \mathcal{P}(\Omega_3) = 1 \) and for all \( \omega \in \Omega_3 \),
\[ P_\mathcal{H}(\omega, B) = 1, \quad \text{where} \]

replace this with an abstract lemma
\[ B = \left\{ \zeta : \frac{1}{N} \sum_{i=0}^{N-1} X_n(\theta^i \zeta) \to \int_{\Omega} P_\mathcal{H}(\zeta, d\sigma) X_n(\sigma), \quad N \to \infty \right\}. \]

Let \( \Omega_4 = \Omega_3 \cap \Omega' \). Then \( \mathcal{P}(\Omega_4) = 1 \). So there is \( \Omega_5 \subseteq \Omega_4 \) with \( \mathcal{P}(\omega, \Omega_4) = 1 \) for all \( \omega \in \Omega_5 \).

Also, \( P_\mathcal{H}(\omega, [\omega]_\mathcal{H}) = 1 \) for \( \omega \in \Omega_4 \). So, \( P_\mathcal{H}(\omega, [\omega]_\mathcal{H} \cap B \cap \Omega_4) = 1 \) for \( \omega \in \Omega_5 \). On the other hand, since \( \omega \mapsto \int \mathcal{P}_\mathcal{H}(\omega, d\sigma) X_n(\sigma) \) is \( \mathcal{H} \)-measurable, we have
\[ \int_{\Omega} P_\mathcal{H}(\zeta, d\sigma) X_n(\sigma) = \int_{\Omega} P_\mathcal{H}(\omega, d\sigma) X_n(\sigma) \]
if \( \omega, \zeta \in \Omega_4 \) and \( \zeta \in [\omega]_\mathcal{H} \).
3. Absolutely continuous invariant measures

Suppose now that $\Omega = (\mathbb{R}^d, \mathcal{B}(\mathbb{R}^d))$ and $\theta$ is a differentiable transforma-

tion. In some cases it is reasonable to expect that there is an absolutely

continuous invariant measure $P$ for $\theta$. Absolute continuity of $P$ with respect

to Lebesgue measure means that there is an $L^1(\mathbb{R}^d, \mathcal{B}(\mathbb{R}^d), \text{Leb})$ function $\rho$ called density such that for any set $B \in \mathcal{B}(\mathbb{R}^d)$,

$$P(B) = \int_B f(\omega) d\omega.$$ 

Densities are not uniquely defined since they can be modified on a zero
measure set.

In what follows, for a differentiable function $f : \mathbb{R}^m \to \mathbb{R}^n$ we denote by

$Df(x)$ the Jacobi matrix of $f$ at point $x$: $(\partial_i f^j(x))_{i=1, \ldots, m, j=1, \ldots, n}$.

**Theorem 4.7.** Let $P(dx) = \rho(x)dx$ be a measure on $(\mathbb{R}^d, \mathcal{B}(\mathbb{R}^d))$. Suppose $\theta : \mathbb{R}^d \to \mathbb{R}^d$ is differentiable and nondegenerate for almost all $x$ (the latter means $\text{Leb}\{x : \det D\theta(x) = 0\} = 0$). Then $P\theta^{-1}$ is also absolutely continuous with respect to Lebesgue measure and the so called transfer operator or Perron–Frobenius operator $L$ applied to $\rho$:

$$L \rho(x) = \sum_{y \in \theta^{-1}x} \frac{\rho(y)}{\det D\theta(y)}, \quad x \in \mathbb{R}^d,$$

gives a density of $P\theta^{-1}$.

Notice that formula (4.13) makes sense for $\text{Leb}$-almost all $x \in \mathbb{R}$ if we follow the usual convention that summation over an empty set is zero.

Since densities that differ only on a set of zero measure define the same measure, it makes sense to consider their equivalence classes and look at the transfer operator as a transformation in $L^1$ understood as a space of equivalence classes.

In particular, we have the following statement.

**Theorem 4.8.** Under the conditions of Theorem 4.7, an absolutely continuous measure $P$ is invariant under transformation $\theta$ if and only if identity

$$(4.14) \quad L \rho(x) = \rho(x)$$

holds for $\text{Leb}$-almost all $x \in \mathbb{R}^d$. In other words, $\rho$ is a fixed point of the transfer operator $L$ or, equivalently, its eigenfunction with eigenvalue 1.
So, the problem of finding invariant densities for smooth maps reduces to solving equation (4.14). Although this equation looks quite innocent, in many concrete cases establishing existence of solutions of (4.14), finding these solutions and exploring their properties poses difficult problems.

Sometimes one can derive existence using conjugation to well-studied dynamical systems. For example, the following is a slightly weakened statement of Denjoy’s theorem (1932):

**Theorem 4.9.** Let \( \theta : S^1 \to S^1 \) be an orientation preserving \( C^2 \)-diffeomorphism with no periodic points. Then there is \( \alpha \in \mathbb{R} \setminus \mathbb{Q} \) such that \( \theta \) is continuously conjugated to \( \theta_{\alpha} \) defined by

\[
\theta_{\alpha}(\omega) = \omega + \alpha.
\]

Continuous conjugation means that there is a homeomorphism \( \phi : S^1 \to S^1 \) such that

\[
\theta = \phi^{-1} \circ \theta_{\alpha} \circ \phi.
\]

M. Herman (1979) and J.-C. Yoccoz (1984) showed that the homeomorphism is actually a diffeomorphism for almost all rotation numbers \( \alpha \) and for Diophantine numbers \( \alpha \), i.e., \( \alpha \) such that \( |q\alpha - p| > cq^{-d} \) for some \( c, d > 1 \) and all nonzero \( p, q \in \mathbb{N} \). Since the Lebesgue measure is invariant for \( \theta_{\alpha} \), its pushforward under \( \phi \) is an absolutely continuous invariant measure for \( \theta \).

Let us look at one concrete example known as the Gauss map. Let \( d = 1 \) and \( \theta x = \{1/x\} \) for \( x \in \mathbb{R}^1 \). Let us check that the function

\[
\rho(x) = \begin{cases} 
\frac{1}{\ln(1+x)}, & x \in [0,1), \\
0, & \text{otherwise}
\end{cases}
\]

is an invariant probability density. Ignoring the normalizing constant \( 1/\ln 2 \), noticing that \( \theta^{-1}x = \{(1/(x+m)) : m \in \mathbb{N}\} \), and computing \( |\theta'(1/(x+m))| = (x+m)^2 \), we see that we need to check

\[
\rho(x) = \sum_{m \in \mathbb{N}} \rho \left( \frac{1}{x+m} \right) \frac{1}{(x+m)^2}, \quad x \in [0,1).
\]

The right-hand side equals

\[
\sum_{m \in \mathbb{N}} \frac{1}{1 + \left( \frac{1}{x+m} \right)^2} \frac{1}{x+m} = \sum_{m \in \mathbb{N}} \frac{1}{(x+m+1)(x+m)} = \sum_{m \in \mathbb{N}} \left( \frac{1}{x+m} - \frac{1}{x+m+1} \right) = \frac{1}{1+x} = \rho(x).
\]

**Theorem 4.10.** Attracting point or \( |\det J| < 1 \) at a fixed point is an obstacle.

**Theorem 4.11.** Suppose \( (\Omega, \mathcal{F}) = ([0,1], \mathcal{B}) \). Let \( m \in \mathbb{N}, m > 1 \) and \( f : [0,1] \to [0,m] \) satisfies \( f(0) = 0 \), \( f(1) = m \) and \( f'(x) > 1 \) for all \( x \in [0,1] \). Then the transformation \( \theta \) defined by \( \theta x = [f(x)] \) has an absolutely
continuous invariant measure. If \( f \in C^r \) for some \( r \geq 2 \), then the invariant density has a \( C^{r-1} \) version.

(A. Rényi, 1953)

**Theorem 4.12 (A. Lasota and J. A. Yorke, 1973).** Let \( \theta : [0, 1] \to [0, 1] \) be a piecewise \( C^2 \) function such that \( \inf |\theta'| > 1 \). Then there is a \( \theta \)-invariant measure absolutely continuous with respect to Lebesgue measure and such that its density has a version with bounded variation.

**Idea of the proof:** It turns out that the following “Lasota–Yorke” inequality holds. There are numbers \( N \in \mathbb{N}, \alpha > 0 \) and \( \beta \in (0, 1) \) such that for every \( f \geq 0 \) with bounded variation \( \nabla f \),

\[
\nabla (\mathcal{L}^N f) \leq \alpha \|f\|_{L^1} + \beta \nabla f.
\]

So, denoting \( f_k = \mathcal{L}^k f \), we obtain

\[
\nabla f_{Nk} \leq \alpha \|f_{N(k-1)}\|_{L^1} + \beta \nabla f_{N(k-1)} \leq \alpha \|f\|_{L^1} + \beta \nabla f_{N(k-1)},
\]

where we used the fact that the set of all probability densities is invariant under the Perron–Frobenius operator \( \mathcal{L} \). Therefore,

\[
\limsup_{k \to \infty} \nabla f_{Nk} \leq \alpha (1 - \beta)^{-1} \|f\|_{L^1},
\]

This along with \( \|f_k\|_{L^1} \leq \|f\|_{L^1} \) guarantees that \( C = \{f_{Nk}\}_{k \in \mathbb{N}} \) is relatively compact in \( L^1 \). Therefore, \( \{f_k\}_{k \in \mathbb{N}} \) is also relatively compact. Mazur’s theorem says that in Banach spaces relative compactness of a set is equivalent to relative compactness of its convex hull, so the set \( \{g_n\}_{n \in \mathbb{N}} \), where

\[
g_n = \frac{1}{n} \sum_{k=0}^{n-1} \mathcal{L}^k f, \quad n \in \mathbb{N}
\]

is also relatively compact. So let us choose a convergent subsequence \( g_{n_k} \overset{L^1}{\to} g \) and notice that

\[
\mathcal{L}g_{n_k} - g_{n_k} = \frac{1}{n} (\mathcal{L}^n_k f - f) \overset{L^1}{\to} 0, \quad k \to \infty.
\]

Since \( \mathcal{L} \) is a continuous operator in \( L^1 \), we conclude that \( \mathcal{L}g = g \).  \( \square \)
CHAPTER 5

Random dynamics

1. Stability in stochastic dynamics

Let us begin with a very simple example, a deterministic linear dynamical system with one stable fixed point.

We take the phase space $X$ to be the real line $\mathbb{R}$ and define the transformation $f: \mathbb{R} \to \mathbb{R}$ by $f(x) = ax$, $x \in \mathbb{R}$, where $a$ is a real number between 0 and 1. To any point $x \in X$ one can associate its forward orbit $(x_n)_{n=0}^{\infty}$, where $x_n = f^n(x)$, $n \geq 0$.

In this simple example, the analysis of the asymptotic behavior of the forward orbit $(x_n)_{n=0}^{\infty}$ as $n \to \infty$, is straightforward and does not require ergodic theory.

Namely, zero is a unique fixed point of the transformation: $f(0) = 0$, and since $x_n = a^n x$, $n \in \mathbb{N}$ and $a \in (0, 1)$ we conclude that as $n \to \infty$, $x_n$ converges to that fixed point exponentially fast. Therefore, 0 is a stable fixed point, or a one-point global attractor for the dynamical system $(\mathbb{R}, f)$. In other words, the domain of attraction coincides with $\mathbb{R}$. So, due to the contraction that is present in the map $f$, there is a fast loss of memory in the system, and no matter what the initial condition is, it gets forgotten in the long run and the points $x_n = f^n(x)$ approach the stable equilibrium 0 as $n \to \infty$.

So far we have been assuming that the evolution begins at time 0, but the picture would not change if we assume that the evolution begins at any other starting time $n_0 \in \mathbb{Z}$. In fact, since the map $f$ is invertible in our example, the full (two-sided) orbit $(x_n)_{n \in \mathbb{Z}} = (f^n x)_{n \in \mathbb{Z}}$ indexed by $\mathbb{Z}$ is well-defined for any $x \in \mathbb{R}$.

Let us now modify the dynamical system of our first example a little by adding noise, i.e., a random perturbation that will kick the system out of equilibrium. The resulting model is often called ARMA(1) (Autoregressive moving average model of order 1). Let us consider some probability space $(\Omega, \mathcal{F}, P)$ rich enough to support a sequence $(\xi_n)_{n \in \mathbb{Z}}$ of independent Gaussian random variables with mean 0 and variance $\sigma^2$. For every $n \in \mathbb{Z}$ we will now define a random map $f_{n, \omega} : \mathbb{R} \to \mathbb{R}$ by

$$f_{n, \omega}(x) = ax + \xi_n(\omega).$$

A natural analogue of a forward orbit from our first example would be a stochastic process $(X_n)_{n \geq n_0}$ emitted at time $n_0$ from point $x$, i.e., satisfying
\( X_{n_0} = x \) and, for all \( n \geq n_0 + 1 \),

\[
X_n = aX_{n-1} + \xi_n.
\]

We also want to describe the long term behavior of the resulting random dynamical system and study its stability properties. However, it is not as straightforward as in the deterministic case. It is clear that there is no fixed point that serves all maps \( f_{n,\omega} \) at the same time. The solution of the equation \( f_{n,\omega}(x) = x \) for some \( n \) may be irrelevant for all other values of \( n \). Still, the system exhibits a pull from infinity towards the origin and contraction. In fact,

\[
|f_{n,\omega}(x) - f_{n,\omega}(y)| = |(ax + \xi_n) - (ay + \xi_n)| = a|x - y|, \quad x, y \in \mathbb{R},
\]

and for the random map \( \Phi_{\omega}^{m,n} \) corresponding to the random dynamics between times \( m \) and \( n \) and defined by

\[
\Phi_{\omega}^{m,n}(x) = f_{n,\omega} \circ f_{n-1,\omega} \circ \cdots \circ f_{m+2,\omega} \circ f_{m+1,\omega},
\]

we obtain

\[
|\Phi_{\omega}^{m,n}(x) - \Phi_{\omega}^{m,n}(y)| = a^{n-m}|x - y|.
\]

The contraction established above should imply some kind of stability. In the random dynamics setting there are several nonequivalent notions of fixed points and their stability.

One way to deal with stability is at the level of distributions. We notice that due to the i.i.d. property of the sequence \( (\xi_n) \), the process \( X_n \) defined above is a homogeneous Markov process with one-step transition probability

\[
P(x, A) = \frac{1}{\sqrt{2\pi} \sigma} \int_{x \in A} e^{-\frac{(y-ax)^2}{2\sigma^2}} \, dy.
\]

If instead of a deterministic initial condition \( x \), say, at time 0 we have a random initial condition \( X_0 \) independent of \( (\xi_n)_{n \geq 1} \) and distributed according to a distribution \( \mu_0 \), then the distribution of \( X_1 \) is given by

\[
\mu_1(A) = \int_{x \in \mathbb{R}} \mu_0(dx)P(x, A),
\]

and a probability measure \( \mu \) on \( \mathbb{R} \) is called invariant or stationary if

\[
\mu(A) = \int_{x \in \mathbb{R}} \mu(dx)P(x, A),
\]

for all Borel sets \( A \). If the dynamics is initiated with initial value distributed according to an invariant distribution, then the resulting process \( (X_n) \) is stationary, i.e., its distribution is invariant under time shifts.

Studying the stability of a Markov dynamical system at the level of distributions involves identification of invariant distributions and establishing convergence of the distribution of \( X_n \) to one of the stationary ones.

In our example, if \( X_0 \) is independent of \( \xi_1 \) and normally distributed with zero mean and variance \( D \), then \( X_1 = aX_0 + \xi_1 \) is also centered Gaussian with variance \( a^2 D + \sigma^2 \). So the distributions at time 0 and time 1 coincide if \( a^2 D + \sigma^2 \)
\( \sigma^2 = D \), i.e., \( D = \sigma^2/(1 - a^2) \). Therefore, the centered Gaussian distribution with variance \( \sigma^2/(1 - a^2) \) is invariant and gives rise to a stationary process. There are several ways to establish uniqueness of this invariant distribution, e.g., the celebrated coupling method introduced by Doeblin in 1930’s which also allows to prove that for any deterministic initial data, the distribution of \( X_n \) converges exponentially fast to the unique invariant distribution in total variation as \( n \to \infty \).

Another way to approach stability is studying random attractors. Let us convince ourselves that in our example, the random attractor contains only one point and that point is a global solution \((X_n)_{n \in \mathbb{Z}}\) defined by

\[
X_n = X(\xi_n, \xi_{n-1}, \xi_{n-2}, \ldots) = \xi_n + a\xi_{n-1} + a^2\xi_{n-2} + \ldots.
\]

Clearly, this series converges with probability 1. Moreover, for any \( n \in \mathbb{Z} \),

\[
aX_{n-1} + \xi_n = \xi_n + a(\xi_{n-1} + a\xi_{n-2} + a^2\xi_{n-3} + \ldots) = \xi_n + a\xi_{n-1} + a^2\xi_{n-2} + a^3\xi_{n-3} + \ldots = X_n,
\]

and \((X_n)_{n \in \mathbb{Z}}\) is indeed a two-sided orbit, i.e., a global (in time) solution of equation (5.1) defined on the entire \( \mathbb{Z} \). Notice that \( X_n \) is a functional of the history of the process \( \xi \) up to time \( n \). Since the process \( \xi \) is stationary, thus constructed \((X_n)_{n \in \mathbb{Z}}\) is also a stationary process. In fact, \( X_n \) is centered Gaussian with variance

\[
\sigma^2 + a^2\sigma^2 + a^4\sigma^2 + \ldots = \sigma^2/(1 - a^2),
\]

which confirms our previous computation of the invariant distribution.

Let us now interpret the global solution \((X_n)_{n \in \mathbb{Z}}\) as a global attractor. We know from the contraction estimate (5.2) that for any \( x \in \mathbb{R} \),

(5.3) \[
|\Phi^{m,n}(x) - X_n| = |\Phi^{m,n}(x) - \Phi^{m,n}(X_m)| = a^{n-m}|x - X_m|.
\]

Using the stationarity of \( X_m \) and applying integration by parts, we see that

\[
\sum_{m<0} P\{|X_m| > |m|\} \leq \sum_{m<0} P\{|X_0| > |m|\} \leq E|X_0| < \infty.
\]

The Borel–Cantelli lemma implies now that \( X_m \) grows at most linearly in \( m \) (in fact, one can prove much better estimates on the growth rate of \( X_m \) since it is a Gaussian stationary process), and (5.3) implies that for any \( x \in \mathbb{R} \),

\[
\lim_{m \to -\infty} |\Phi^{m,n}(x) - X_n| = 0.
\]

In words, if we fix an initial condition \( x \) and run the random dynamics from time \( m \) to time \( n \), then the result of this evolution converges to the special global solution \( X_n \) as we pull the starting time \( m \) back to \(-\infty\). This allows us to call \((X_n)_{n \in \mathbb{Z}}\) a one-point pullback attractor for our random dynamical system. Pullback attractors were introduced in [CF94], [CDF97], [Sch94].
We can adapt the reasoning above to show that there are no other global stationary solutions of (5.1). In fact, if \((Y_n)_{n \in \mathbb{Z}}\) is such a stationary solution then there is a number \(R\) such that for almost all \(\omega\) there is a sequence \(m_k(\omega) \to -\infty\) such that \(|Y_{m_k}| < R\). Using the contraction estimate (5.2), we see that for any \(n\) and \(k\)

\[
|X_n - Y_n| = |\Phi^{m_k,n}X_{m_k} - \Phi^{m_k,n}Y_{m_k}| = a^{n-m_k}|X_{m_k} - Y_{m_k}|.
\]

Using that \(|Y_{m_k}| < R\) and that \(|X_{m_k}|\) grows not faster than linearly in \(m_k\), we take \(m_k\) to \(-\infty\) and conclude that \(X_n = Y_n\) thus proving our uniqueness claim.

We can rephrase the uniqueness and convergence statements as the following One Force — One Solution Principle (1F1S): with probability 1, at any given time there is a unique value of \(X_n\) compatible with the history of the “forcing” \((\xi_k)_{k \leq n}\). One can also say that \(X_n\) is a unique value worked out by the dynamics in the past up to \(n\).

In our example where 1F1S is valid, the global solution plays the role of a one-point random attractor, and the unique invariant distribution can be recovered as the distribution of this random point at, say, time 0. In general, the picture can be more complicated. If a random dynamical system admits an invariant distribution, then at a given time there can be more than one point compatible with the history of random maps. One can consider the union of these points as a random attractor and, moreover, introduce a natural distribution on these points called the sample measure associated to the history of forcing, see [You86], and [Arn98, Chapter 1].

Notice that the invariant distribution in our example results from a balance between two opposing influences. One is the decay or dissipation due to the contractive dynamics. In the absence of randomness the system would simply equilibrate to the stable fixed point. The second mechanism is the “random forcing” that keeps the system from the rest at equilibrium, and the stronger that influence is the more the resulting stationary distribution is spread out. This mechanism is typical for many physical systems.

One of the basic questions of ergodic theory of a dynamical system is the description of the stationary regimes. This is the main subject of the rest of these notes.

2. Markov processes and random transformations

We begin a more formal study of non-deterministic dynamical systems, where the state of the system at time \(n+1\) is not uniquely determined by the state at time \(n\). We will mostly be interested in processes with instantaneous loss of memory. They are usually called Markov processes and defined by the following property: given the history of the process \(X\) up to time \(n\), the state at time \(n+1\) is random and has conditional distribution \(P(X_n, dy)\) that depends only on the value \(X_n\).
DEFINITION 5.1. Let \((\mathcal{X}, \mathcal{X})\) be a measurable space. A function \(P: \mathcal{X} \times \mathcal{X} \to \mathbb{R}\) is called a transition probability, or transition kernel, or probability kernel if for each \(x \in \mathcal{X}\), \(P(x, \cdot)\) is a probability measure, and for each \(B \in \mathcal{X}\), \(P(\cdot, B)\) is an \(\mathcal{X}\)-measurable function.

DEFINITION 5.2. Let \((\mathcal{X}, \mathcal{X})\) be a Borel space. Let \(\rho\) be a probability measure on a \((\mathcal{X}, \mathcal{X})\) and let \(P(\cdot, \cdot)\) be a transition kernel on \((\mathcal{X}, \mathcal{X})\). A \((\mathcal{X}, \mathcal{X})\)-valued process \(X\) defined on a probability space \((\Omega, \mathcal{F}, \mathbb{P})\) with time indexed by \(\{n, n+1, n+2, \ldots\}\) for some \(n \in \mathbb{Z}\) is called a homogeneous Markov process with initial distribution \(\rho\) and one-step transition probability \(P\) if for any \(m \geq 0\) and any sets \(A_0, A_1, \ldots, A_m \in \mathcal{X}\),

\[
(5.4) \quad \mathbb{P}\{X_n \in A_0, \ldots, X_{n+m} \in A_m\} = \int_{A_0} \rho(dx_0) \int_{A_1} P(x_0, dx_1) \cdots \int_{A_{m-2}} P(x_{m-2}, dx_{m-1}) \int_{A_{m-1}} P(x_{m-1}, A_m). 
\]

PROBLEM 5.1. Show that formula (5.4) defines a consistent family of finite-dimensional distributions.

By Kolmogorov’s consistency theorem, this problem implies that there is a unique distribution on \((\mathcal{X}^{n,n+1,\ldots}, \mathcal{X}^{n,n+1,\ldots})\) consistent with the finite-dimensional distributions defined by formula (5.4).

PROBLEM 5.2. Show that Definition 5.2 is equivalent to the following: for any \(m \in \mathbb{N}\) and any bounded and \((\mathcal{X}^n, \mathcal{B}(\mathbb{R}))\)-measurable function \(f: \mathcal{X}^{n+1} \to \mathbb{R}\),

\[
(5.5) \quad \mathbb{E}f(X_0, \ldots, X_{n+m}) = \int_{\mathcal{X}} \rho(dx_0) \int_{\mathcal{X}} P(x_0, dx_1) \cdots \int_{\mathcal{X}} P(x_{m-1}, dx_m) f(x_0, \ldots, x_m). 
\]

Let us convince ourselves that the function \(P(\cdot, \cdot)\) can indeed be interpreted as transition probability.

THEOREM 5.1 (Markov property). Suppose that Definition 5.2 holds and for any \(m\) denote \(\mathcal{F}_m = \sigma(X_k, k \leq m)\). Then

\[
\mathbb{P}(X_{r+1} \in A|\mathcal{F}_r) \overset{\text{a.s.}}{=} \mathbb{P}(X_{r+1} \in A|X_r) \overset{\text{a.s.}}{=} \mathbb{P}(X_r, A), \quad r \geq n, \quad A \in \mathcal{X}. 
\]

PROOF: Without loss of generality we set \(n = 0\). It is sufficient to prove that for any \(m\) and any measurable bounded function \(f: \mathcal{X}^{m+1} \to \mathbb{R}\),

\[
\mathbb{E}P(X_m, A)f(X_0, \ldots, X_m) = \mathbb{E}1_{X_m \in A} f(X_0, \ldots, X_m). 
\]

To prove this formula we can directly compute both sides of this identity using (5.5).

One way to construct a Markov process is to use compositions of i.i.d. maps. More precisely, let \(\mathcal{F}\) be a set of \((\mathcal{X}, \mathcal{X})\)-measurable maps \(f: \mathcal{X} \to \mathcal{X}\) equipped with a \(\sigma\)-algebra \(\mathcal{F}\) such that the map \(F: \mathcal{X} \times \mathcal{F} \to \mathcal{X}\) defined by \(F(x, f) = f(x)\) is \(\mathcal{X} \times \mathcal{F}, \mathcal{X}\)-measurable.
Then suppose \((f_k)_{k \in \mathbb{Z}}\) is an i.i.d. sequence of \((\mathcal{F}, \mathcal{G})\)-valued random variables defined on a probability space \((\Omega, \mathcal{F}, \mathbb{P})\). Let us fix a number \(n \in \mathbb{Z}\) and a distribution \(\rho\) on \((\mathcal{X}, \mathcal{A})\). We let \(X_n\) be distributed according to \(\rho(dx)\) and independent of \(f_k, k > n\), and then define \(X_k, k > n\), inductively by \(X_k = f_k(X_{k-1}) = F(X_{k-1}, f_k)\). For any \(k\), thus constructed functions \(X_k\) are random variables on \((\mathcal{X} \times \Omega, \mathcal{A} \times \mathcal{F})\) which follows from an inductive argument. In fact, \(X_n(x, \omega) = x\) is a random variable, and if \(X_{k-1}(x, \omega)\) is a random variable, then

\[
X_k(x, \omega) = f_{k, \omega}(X_{k-1}(x, \omega)) = F(f_{k, \omega}, X_{k-1}(x, \omega)),
\]

and \(X_k\) is measurable as a composition of measurable maps. It is also convenient to write

\[
X_k(x, \omega) = f_{n,k}(x), \quad \omega \in \Omega, \ x \in \mathcal{X}, \ k \geq n,
\]

where

\[
f_{n,k}(x) = f_{k, \omega} \circ \ldots \circ f_{n+2, \omega} \circ f_{n+1, \omega}(x), \quad x \in \mathcal{X}, \ k \geq n.
\]

Let us take any \(k \in \mathbb{Z}\) and denote

\[
P(x, A) = \mathbb{P}\{\omega : f_{k, \omega}(x) = A\}.
\]

Due to i.i.d. property, this definition does not depend on the choice of \(k\).

**Lemma 5.1.** Thus defined \(P(\cdot, \cdot)\) is a probability kernel.

**Proof:** In this proof we choose \(k = 1\) and work with \(f_1\). The fact that \(P(x, \cdot)\) is a probability measure follows since \(\mathbb{P}\) is a probability measure. If \(A \in \mathcal{A}\), then \(F^{-1}A = \{(x, f) : f(x) \in A\} \in \mathcal{X} \times \mathcal{G}\). Denoting by \(\mathbb{P}f_1^{-1}\) the pushforward of \(\mathbb{P}\) onto \(\mathcal{F}\) by \(f_1\), we see that, for \(x \in \mathcal{X}\), \(P(x, A)\) is the measure assigned by \(\mathbb{P}f_1^{-1}\) to the \(x\)-section of \(F^{-1}A\), i.e., \(\{f : f(x) \in A\}\). (Related to Fubini’s theorem, but is that the right reference? Proof in [Kif86], starting with rectangles) Such functions are measurable.

**Lemma 5.2.** For any bounded measurable function \(g : \mathcal{X} \to \mathbb{R}\),

\[
\int_{\mathcal{X}} P(x, dy)g(y) = \mathbb{E}g(f_{1, \omega}(x)), \quad x \in \mathcal{X}.
\]

**Proof:** The lemma holds for indicator functions \(g = 1_A, A \in \mathcal{A}\), due to (5.8). It holds by linearity for simple functions, and then for general functions by uniform approximation with simple functions.

**Theorem 5.2.** The process \((X_k)_{k \geq n}\) defined in (5.6) is a homogeneous Markov process with initial distribution \(\rho\) and transition kernel \(P\).

**Proof:** Let us prove formula (5.4) by induction assuming \(n = 0\) without loss of generality. For \(m = 0\), the formula is trivially true. Let us now assume that the lemma holds for \(m\) replaced by \(m - 1 \geq 0\). Then

\[
\mathbb{P}\{X_0 \in A_0, \ldots, X_m \in A_m\} = \mathbb{E}1_{X_0 \in A_0} \ldots 1_{X_{m-1} \in A_{m-1}} 1_{f_{m, \omega}(X_{m-1}) \in A_m}.
\]
Denoting $\mathcal{G} = \sigma(X_0, \ldots, X_{m-1})$, we obtain
$$
P\{X_0 \in A_0, \ldots, X_m \in A_m\} = \mathbb{E}[\mathbf{1}_{X_0 \in A_0} \ldots \mathbf{1}_{X_{m-1} \in A_{m-1}} \mathbb{E}[\mathbf{1}_{f_{m,\omega}(X_{m-1}) \in A_m} | \mathcal{G}]] .$$
Since $f_{m,\omega}$ is independent of $\mathcal{G}$ and $X_{m-1}$ is $\mathcal{G}$-measurable we can apply the following lemma:

**Lemma 5.3.** Let $X$ and $Y$ be random variables on a probability space $(\Omega, \mathcal{F}, \mathbb{P})$ taking values in measurable spaces $(\mathcal{X}, \mathcal{X})$ and $(\mathcal{Y}, \mathcal{Y})$, respectively. Let $F : \mathcal{X} \times \mathcal{Y} \to \mathbb{R}$ be a bounded $(\mathcal{X} \times \mathcal{Y}, \mathcal{B}(\mathbb{R}))$-measurable map and suppose that a $\sigma$-algebra $\mathcal{G} \subset \mathcal{F}$ is such that $X$ is $(\mathcal{G}, \mathcal{X})$-measurable and $Y$ is independent of $\mathcal{G}$. Then
$$
\mathbb{E}[F(X,Y) | \mathcal{G}] = a(X),
$$
where $a(x) = \mathbb{E}F(x,Y)$.

To apply this lemma with $X_{m-1}$ for $X$, $f_m$ for $Y$ and $F(X,Y) = 1_{Y(X) \in A_m}$, we find $a(x) = \mathbb{E}\mathbf{1}_{f_{m,\omega}(x) \in A_m} = P(x, A_m)$, and obtain
$$
P\{X_0 \in A_0, \ldots, X_m \in A_m\} = \mathbb{E}[\mathbf{1}_{X_0 \in A_0} \ldots \mathbf{1}_{X_{m-1} \in A_{m-1}} P(X_{m-1}, A_m)].$$
We can now use (5.5) with $m$ replaced by $m-1$ and
$$
f(x_0, \ldots, x_{m-1}) = \mathbf{1}_{x_0 \in A_0} \ldots \mathbf{1}_{x_{m-1} \in A_{m-1}} P(x_{m-1}, A_m),$$
obtaining (5.4).

**Proof of Lemma 5.3:** First, consider the situation where $F(X,Y) = 1_{X \in A, Y \in B} = 1_{X \in A}1_{Y \in B}$ for some sets $A \in \mathcal{X}$, $B \in \mathcal{Y}$. Then
$$
\mathbb{E}[F(X,Y) | \mathcal{G}] = 1_{X \in A} \mathbb{E}[1_{Y \in B} | \mathcal{G}] = 1_{X \in A} P\{Y \in B\},
$$
and, since $1_{X \in A}$ is nonrandom, i.e., constant in $\omega$, we have
$$
a(x) = \mathbb{E}1_{x \in A}1_{Y \in B} = 1_{x \in A} P\{Y \in B\}, \quad x \in \mathcal{X}.
$$
So, the lemma holds for these indicator functions of rectangles. By linearity of expectation, the lemma holds for linear combinations of such indicators. The proof is completed by approximating the generic bounded function with those linear combinations. \hfill \Box

It turns out that the method of obtaining Markov processes described in Theorem 5.2 is, in a sense, universal. Let us state and prove a theorem from [Kif86, Theorem 1.1] based on the idea of R. M. Blumenthal and H. K. Corson [BC72].

**Theorem 5.3.** Let $(\mathcal{X}, \mathcal{X})$ be a Borel space and let $P(\cdot, \cdot)$ be a probability kernel on $(\mathcal{X}, \mathcal{X})$. Then there is a probability space $(\mathcal{F}, \mathcal{G}, \mathbb{Q})$ of measurable maps $f$ on $\mathcal{X}$ such that
$$
\mathbb{Q}\{f : f(x) \in A\} = P(x, A), \quad x \in \mathcal{X}, \ A \in \mathcal{X}.
$$
Proof of Theorem 5.3: This is a version of the original proof of [Kif86]. First, we notice that it is sufficient to consider the case where \( X \) is a Borel subset of \([0, 1]\) and \( \mathcal{X} \) is the restriction of \( B([0, 1]) \) to \( X \).

For each \( x \in X \) we let \( z(x, \cdot) \) be the quantile function of \( P(x, \cdot) \):

\[
z(x, t) = \inf \{ r : P(x, [0, r]) \geq t \}, \quad x \in X, \ t \in [0, 1].
\]

Since for any \( x \in X \),

\[
\text{Leb}\{ t : z(x, t) > r \} = \text{Leb}\{ t : P(x, [0, r]) < t \} = 1 - P(x, [0, r]), \quad r \in [0, 1],
\]

we conclude that \( \text{Leb}\{ t : z(x, t) \leq r \} = P(x, [0, r]) \), and so, \( P(x, \cdot) \) is the pushforward of the Lebesgue measure under \( z(x, \cdot) \):

\[
\text{Leb}\{ t : z(x, t) \in C \} = P(x, C), \quad C \in B([0, 1]).
\]

Therefore, if \( x \) is fixed, then for \( \text{Leb} \)-almost all \( t \in [0, 1] \), \( z(x, t) \in X \).

Let us pick any \( x_0 \in X \) and for all \( t \in [0, 1] \) define a Borel transformation \( f_t : X \to X \) by

\[
f_t(x) = \begin{cases} 
  z(x, t), & x \in X, \ z(x, t) \in X, \\
  x_0, & x \in X, \ z(x, t) \notin X.
\end{cases}
\]

thus obtaining a map \( T \) from \([0, 1]\) to \( \mathbb{F} \). Let us denote by \( \mathcal{F} \) and \( Q \) the pushforwards of \( B([0, 1]) \) and the Lebesgue measure under this map: \( \mathcal{F} = \{ A \subset \mathbb{F} : T^{-1}A \in B([0, 1]) \} \) and \( Q(A) = \text{Leb}(T^{-1}A), \ A \in \mathcal{F} \). For every \( C \in \mathcal{X} \),

\[
Q\{ f : f(x) \in C \} = \text{Leb}\{ t : f_t(x) \in C \} = \text{Leb}\{ t : z(x, t) \in C \} = P(x, C),
\]

which completes the proof. \( \square \)

With this theorem at hand, we can take a product of infinitely many copies of \((\mathbb{F}, \mathcal{F}, Q)\) thus obtaining the following:

**Theorem 5.4.** Let \((X, \mathcal{X})\) be a Borel space, and let \( P(\cdot, \cdot) \) be a probability kernel on \((X, \mathcal{X})\). Then there is a probability space \((\Omega, \mathcal{F}, P)\), a \( \sigma \)-algebra \( \mathcal{F} \) on \( \mathbb{F} \) and a sequence \( (f_k)_{k \in \mathbb{Z}} \) of i.i.d. \((\mathbb{F}, \mathcal{F})\)-valued random variables on \((\Omega, \mathcal{F}, P)\) such that for any \( n \in \mathbb{Z} \) and any distribution \( \rho \) on \((X, \mathcal{X})\), the construction described before Theorem 5.2 produces a Markov process on \((X_k)_{k \geq n}\) with initial distribution \( \rho \) and transition probability \( P(\cdot, \cdot) \).

It is clear that there are many ways to realize a Markov process with a given transition probability. The dynamical properties of resulting system of maps may drastically depend on the details of such representation. For example, let \( (\xi_k)_{k \in \mathbb{N}} \) be an i.i.d. sequence of uniformly distributed random variables on \([0, 1]\) defined on \((\Omega, \mathcal{F}, P)\). Then for any measurable \( g : [0, 1] \to [0, 1] \) the random dynamical system on \([0, 1]\) defined by

\[
f_{k, \omega}(x) = g(x) + \xi_k(\omega) \pmod{1}, \quad x \in [0, 1], \ \omega \in \Omega,
\]

generates a Markov process with transition function \( P(x, \cdot) \equiv \text{Leb} \). However, if \( g(x) \equiv x \), then for any \( \omega \in \Omega \) and any \( x, y \in [0, 1] \),

\[
d(f^0_\omega(x), f^m_\omega(y)) \equiv d(x, y),
\]
where transformations $f_{\omega}^{n,m}$ have been defined in (5.7). However, this property does not hold true, say, for $g(x) \equiv 2x \pmod{1}$.

Let us now summarize the construction introduced in this section. At different times the evolution in $X$ is governed by different transformations chosen from a “common pool”, a transformation space $\mathcal{F}$. These transformations form an $\mathcal{F}$-valued stationary process and, when studying the evolution in $X$, it is convenient also to keep track of that stationary process.

We have a state space $X$ equipped with a $\sigma$-algebra $\mathcal{X}$ and a set $\mathcal{F}$ of $X$-measurable transformations equipped with a $\sigma$-algebra $\mathcal{F}$ and a measure $Q$ on $\mathcal{F}$. The main canonical probability space $(\Omega, \mathcal{F}, P)$ will be then $(\mathcal{F}, \mathcal{F}), (\mathcal{Q}, \mathcal{Q})$ or $(\mathcal{F}, \mathcal{F}, Q)$. The shift operator $\theta : \Omega \to \Omega$ given by $(\theta \omega)_i = \omega_{i+1}$, $i \in \mathbb{Z}$, or, respectively, $i \in \mathbb{N}$, preserves $P$.

It is convenient to identify random transformations with coordinates of $\omega$, so we can write $f_\omega = \omega_1$. We also introduce $(\mathcal{F}, \mathcal{F})$-valued random variables $(f_n)_{n \in \mathbb{N}}$ defined by $f_{n,\omega} = \omega_n = f_{\theta^n \omega}$. They are i.i.d. with common distribution $Q$.

The joint evolution in $X$ and $\Omega$ is given by the following skew product transformation $\Theta : X \times \Omega \to X \times \Omega$ defined by

$$\Theta(x, \omega) = (f_\omega x, \theta \omega) = (\omega_1 x, \theta \omega).$$

Thus the space $(X \times \Omega, \mathcal{X} \times \mathcal{F})$ equipped with transformation $\Theta$ is a usual dynamical system. Iterations of $\Theta$ are given by

$$\Theta^n(x, \omega) = (f_{\omega}^n x, \theta^n \omega) = (\omega_1^n x, \theta^n \omega),$$

where transformations $f_{\omega}^{n,m}$ have been defined in (5.7) and

$$\omega^{m,n} = \omega^n \circ \ldots \circ \omega^{m+1}.$$

Skew products satisfy the following cocycle property:

$$f_{\omega}^{0,m_1+m_2}(x) = f_{\theta^{m_1} \omega}^{0,m_2} \circ f_{\omega}^{0,m_1},$$

or

$$\omega^{0,m_1+m_2}(x) = (\theta^{m_1} \omega)^{0,m_2} \circ \omega^{0,m_1}.$$

In general, one can study random dynamical systems as skew-products satisfying the cocycle property without introducing product measures as above or any kind of measures. For example, linearization of differentiable maps.

3. Invariant measures

Suppose we have an $(\mathcal{X}, \mathcal{X})$-valued Markov process $(X_k)_{k \geq 0}$ with one-step transition probability function $P(\cdot, \cdot)$. If the distribution of $X_0$ is given by a measure $\rho$, then

$$P\{X_1 \in A\} = \rho P(A), \quad A \in \mathcal{X},$$

explore this example
where the convolution of a measure $\rho$ and a kernel $P$ is defined by

$$\rho P(A) = \int_X \rho(dx) P(x, A), \quad A \in \mathcal{X}.$$ 

Moreover, the transition kernel $P$ defines the following semigroup on $\mathcal{P}(\mathbb{X}, \mathcal{X})$ (the set of probability measures on $(\mathbb{X}, \mathcal{X})$): $\rho P^0 = \rho$ and for $n \in \mathbb{N}$,

$$\rho P^n = \rho P^{n-1} P.$$ 

In analogy with the deterministic case that corresponds to transition kernels given by delta functions, we are interested in the situation where the distribution after the transformation is the same as the initial distribution, i.e., it is invariant under the action of semigroup $(P^n)_{n \geq 0}$. In other words, we want to have $P\{X_1 \in A\} = P\{X_0 \in A\}$, or

$$(5.12) \quad \rho(A) = \int_X \rho(dx) P(x, A), \quad A \in \mathcal{X}.$$ 

**Definition 5.3.** A measure $\rho$ on $(\mathbb{X}, \mathcal{X})$ is called invariant for transition probability $P(\cdot, \cdot)$ if (5.12) holds.

As in deterministic case, we can define invariance of distributions in terms of test functions:

**Lemma 5.4.** A measure $\rho$ on $(\mathbb{X}, \mathcal{X})$ is invariant for transition probability $P(\cdot, \cdot)$ if and only if for any bounded measurable function $g : \mathbb{X} \to \mathbb{R}$,

$$(5.13) \quad \int_{\mathbb{X}} \rho(dx) g(x) = \int_{\mathbb{X}} \rho(dx) \int_{\mathbb{X}} P(x, dy) g(y).$$

**Proof:** Definition 5.3 is equivalent to (5.13) for indicator functions $g = 1_A$, $A \in \mathcal{X}$. By linearity it is also equivalent to (5.13) for simple functions and the general statement is obtained by uniform approximation of bounded functions by simple ones. \qed

**Problem 5.3.** Prove that if $\rho$ is invariant for $P$, then the Markov process with finite-dimensional distributions defined by (5.4) is stationary.

For example, let $\mathbb{X} = \{1, 2, 3\}$ with $\mathcal{X} = 2^\mathbb{X}$, and $P(x, A) = |A \setminus \{x\}|/2$ for any $A \subset \mathbb{X}$ which means that, given that at any time step the system changes its current state to a new one chosen uniformly from the remaining two states. Denoting

$$p_{ij} = P(i, \{j\}) = \begin{cases} \frac{1}{2}, & j \neq i, \\ 0, & j = i, \end{cases}$$

we obtain the following system of equations for an invariant measure $\rho$:

$$\rho_j = \sum_{i=1}^3 \rho_i p_{ij}, \quad j \in \mathbb{X},$$

where $\rho_j = \rho\{j\}$. This system of linear equations is easily seen to have a line of solutions $\rho_1 = \rho_2 = \rho_3$. Since we are interested in probability measures,
we have to set \( \rho_1 = \rho_2 = \rho_3 = 1/3 \), and so this transition kernel has a unique invariant distribution.

Another example. Let us check that the uniform distribution on \([0,1]\) is invariant for the following probability kernel on the real line.

\[
P(x, A) = \frac{1}{2} \delta_x(A) + \frac{1}{2} \delta_{1+x}(A), \quad x \in \mathbb{R}.
\]

Let us take any Borel \( A \subset [0,1] \) and compute

\[
\int_{[0,1]} P(x, A) dx = \frac{1}{2} \int_{[0,1]} 1_A(x/2) dx + \frac{1}{2} \int_{[0,1]} 1_A((1+x)/2) dx
\]

\[
= \frac{1}{2} \text{Leb}\left\{ x \in [0,1] : \frac{x}{2} \in A \right\} + \frac{1}{2} \text{Leb}\left\{ x \in [0,1] : \frac{1+x}{2} \in A \right\}
\]

\[
= \frac{1}{2} \text{Leb}((2A) \cap [0,1]) + \frac{1}{2} \text{Leb}((2A - 1) \cap [0,1])
\]

\[
= \text{Leb}(A \cap [0,1/2]) + \text{Leb}(A \cap [1/2,1]) = \text{Leb}(A).
\]

In fact, there are no other invariant distributions, but checking this is not so easy as in the first example.

**Problem 5.4.** Use binary decompositions of numbers in \([0,1]\) to give another proof of invariance of the uniform distribution in this example.

**Problem 5.5.** Find an invariant distribution for the following probability kernel on the real line:

\[
P(x, A) = \frac{1}{2} \delta_{x/2}(A) + \frac{1}{2} \delta_{1+x/2}(A), \quad x \in \mathbb{R}.
\]

Let us consider another example on the real line. Let us fix two parameters \( a \in (0,1) \) and \( \sigma^2 > 0 \) and suppose that \( P(x, \cdot) \) is a Gaussian measure with mean \( ax \) and variance \( \sigma^2 \). Let us find a number \( r^2 > 0 \) such that a centered Gaussian distribution with variance \( r^2 \) is invariant. We need to make sure that for any Borel \( A \subset \mathbb{R} \),

\[
\int_A dx \frac{e^{-x^2}}{\sqrt{2\pi}} = \int_{\mathbb{R}} dx \frac{e^{-x^2}}{\sqrt{2\pi}} \int_A dy \frac{e^{-|y-ax|^2}}{\sqrt{2\pi\sigma}}.
\]

After a change of variables \( ax = z \), the right-hand side becomes

\[
\int_{\mathbb{R}} dx \frac{e^{-z^2/(2\sigma^2 a^2)}}{\sqrt{2\pi a}} \int_A dy \frac{e^{-|y-z|^2/\sigma^2}}{\sqrt{2\pi\sigma}} = \int_A p(y) dy,
\]

where \( p \) is the convolution of densities of two centered Gaussian distributions with variances \( r^2 a^2 \) and \( \sigma^2 \), i.e., \( p \) is a centered Gaussian density with variance \( r^2 a^2 + \sigma^2 \). So we will have an invariant Gaussian distribution if and only if \( r^2 = r^2 a^2 + \sigma^2 \), i.e., \( r^2 = \sigma^2/(1 - a^2) \). We will prove later that there are no other invariant distributions for this Markov semigroup.

Let us suppose now that the Markov process with transition kernel \( P(\cdot, \cdot) \) is, in fact, generated by a skew-product transformation \( \Theta \) on the canonical space \((X \times \Omega, \mathcal{X} \times \mathcal{F})\), where \((\Omega, \mathcal{F}, P) = (\mathbb{F}^N, \mathcal{F}^N, Q^N)\). The notion of an
5. RANDOM DYNAMICS

An invariant measure for deterministic transformations applies to this situation, so we can use either Definition 2.4 or Lemma 2.1. However, we are interested not just in any invariant measure on \((\mathbb{X} \times \Omega, \mathcal{X} \times \mathcal{F})\), but only in those whose marginal distribution on \((\Omega, \mathcal{F})\) is \(\mathbb{P}\).

**Definition 5.4.** We say that a measure \(\mu\) on \((\mathbb{X} \times \Omega, \mathcal{X} \times \mathcal{F})\) is invariant under \(\Theta\) and agrees with \(\mathbb{P}\) if (i) for any bounded \(\mathcal{X} \times \mathcal{F}\)-measurable function \(g : \mathbb{X} \times \Omega \to \mathbb{R}\),

\[
\int_{\mathbb{X} \times \Omega} \mu(dx, d\omega) g(x, \omega) = \int_{\mathbb{X} \times \Omega} \mu(dx, d\omega) g(\Theta(x, \omega)),
\]

and (ii) \(\mu\) projects onto \(\mathbb{P}\) under the canonical projection \(\mathbb{X} \times \Omega \to \Omega\), i.e.

\[
\mu(\mathbb{X} \times A) = \mathbb{P}(A), \quad A \in \mathcal{F}.
\]

**Theorem 5.5 ([Ohn83]).** A probability measure \(\mathbb{\rho}\) is invariant for \(\mathbb{P}(\cdot, \cdot)\) if and only if \(\mathbb{\rho} \times \mathbb{P}\) is invariant for \(\Theta\).

**Proof:** Suppose \(\mathbb{\rho} \times \mathbb{P}\) is \(\Theta\)-invariant. Let us take any bounded measurable function \(g : \mathbb{X} \to \mathbb{R}\) and check (5.13). In the following calculation we introduce a function \(G(x, \omega) = g(x):\)

\[
\int_{\mathbb{X}} \mathbb{\rho}(dx) \int_{\mathbb{X}} \mathbb{P}(x, dy) g(y) = \int_{\mathbb{X}} \mathbb{\rho}(dx) \mathbb{E} g(\omega_1 x)
\]

\[
= \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) g(\omega_1 x) = \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) G(\omega_1 x, \theta \omega)
\]

\[
= \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) G(\Theta(x, \omega)) = \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) G(x, \omega)
\]

\[
= \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) g(x) = \int_{\mathbb{X}} \mathbb{\rho}(dx) g(x).
\]

Suppose now that \(\mathbb{\rho}\) is \(\mathbb{P}(\cdot, \cdot)\)-invariant. It is sufficient to check (5.14) for indicators of rectangles, i.e., \(g(x, \omega) = 1_{x \in A} 1_{\omega \in B}\). The left-hand side of (5.14) equals \(\mathbb{\rho}(A)\mathbb{P}(B)\), so let us compute the right-hand side:

\[
\int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) g(\Theta(x, \omega)) = \int_{\mathbb{X} \times \Omega} (\mathbb{\rho} \times \mathbb{P})(dx, d\omega) g(f_\omega, \theta \omega)
\]

\[
= \int_{\mathbb{X}} \mathbb{\rho}(dx) \int_{\Omega} \mathbb{P}(dw) 1_{\omega_1(x) \in A} 1_{\theta \omega \in B}
\]

Since \(\omega_1\) and \(\theta \omega\) are independent under \(\mathbb{P} = \mathbb{Q}^N\), the r.h.s. equals

\[
= \int_{\mathbb{X}} \mathbb{\rho}(dx) \int_{\mathbb{X}} \mathbb{Q}(d\omega_1) 1_{\omega_1(x) \in A} \int_{\mathbb{X}(2,3,...)} \mathbb{Q}^{(2,3,...)}(d\omega_2, d\omega_3 \ldots) 1_{(\omega_2, \omega_3 \ldots) \in B}
\]

\[
= \int_{\mathbb{X}} \mathbb{\rho}(dx) P(x, A) \mathbb{P}\{\theta \omega \in B\} = \mathbb{\rho}(A) \mathbb{P}(B),
\]

and the lemma is proved. \(\square\)
In the case of \((\Omega, \mathcal{F}, \mathcal{P}) = (\mathbb{R}^\mathbb{Z}, \mathcal{F}^\mathbb{Z}, Q^\mathbb{Z})\), relation (5.14) can be derived in the same way for \(\mu = \rho \times \mathcal{P}\) from invariance of \(\rho\), but only for functions \(g\) that do not depend on the past, i.e.,

\[
g(x, (\ldots, \omega_{-1}, \omega_0, \omega_1, \omega_2 \ldots)) = g(x, (\omega_1, \omega_2, \ldots)).
\]

However, \(\rho \times \mathcal{P}\) does not have to be invariant in this case, and it is actually natural to expect invariant distributions not to have the product structure since the state of the system in \(x\) is largely determined by the dynamics in the past defined by \((\ldots, \omega_{-2}, \omega_{-1}, \omega_0)\) whereas product structure would mean independence of the state and the noise input.

In general, any measure on \((X \times \Omega, \mathcal{X} \times \mathcal{F})\) that projects on \(\mathcal{P}\) can be disintegrated into conditional measures and represented as \(\mu(dx, d\omega) = \mu_\omega(dx)\mathcal{P}(d\omega) = \mu(\omega, dx)\mathcal{P}(d\omega)\) for a probability kernel \(\mu(\cdot, \cdot)\) from \((\Omega, \mathcal{F})\) to \((X, \mathcal{X})\), i.e., the following formula of complete probability holds:

\[
\int_{X \times \Omega} f \, d\mu = \int_\Omega \left( \int_X f(x, \omega) \mu_\omega(dx) \right) \mathcal{P}(d\omega).
\]

Here and further we follow [Arn98] where more general random dynamical systems are considered.

**Theorem 5.6.** Let \(\Theta\) be a measurable random dynamical system on a Borel space \((X, \mathcal{X})\) over \((\Omega, \mathcal{F}, \mathcal{P}) = (\mathbb{R}^\mathbb{Z}, \mathcal{F}^\mathbb{Z}, Q^\mathbb{Z})\) defined by (5.9). A measure \(\mu\) on \((X \times \Omega, \mathcal{X} \times \mathcal{F})\) agreeing with \(\mathcal{P}\) is invariant iff \(\mathcal{P}\)-a.e. \(\omega\) its disintegration \(\mu_\omega\) satisfies

\[
(5.16) \quad \mu_\omega f_\omega^{-1} = \mu_\theta \omega.
\]

**Proof:** For any set \(B \times F \in \mathcal{X} \times \Omega,\)

\[
\mu_\Theta^{-1}(B \times F) = \int_{\Theta^{-1}F} \mu_\omega f_\omega^{-1}(B)\mathcal{P}(d\omega)
\]

and

\[
\mu(B \times F) = \int_F \mu_\omega(B)\mathcal{P}(d\omega) = \int_F \mu_\omega(B)\mathcal{P} \theta^{-1}(d\omega) = \int_{\theta^{-1}F} \mu_\theta \omega(B)\mathcal{P}(dw).
\]

If (5.16) holds, then these identities imply \(\mu(B \times F) = \mu \Theta^{-1}(B \times F)\), and it follows that \(\mu = \mu \Theta^{-1}\).

On the other hand, if \(\mu\) is invariant, then since one can represent any set in \(\mathcal{F}\) as \(\theta^{-1}F\) for some \(F\), we conclude that for almost all \(\omega\), \(\mu_\theta \omega(B) = \mu_\omega f_\omega^{-1}(B)\). The exceptional set depends on \(B\), but since \((X, \mathcal{X})\) is a Borel space, any measure \((X, \mathcal{X})\) is uniquely defined by its values on a countable collection of sets, so we can choose a universal set \(A\) of full measure such the measures \(\mu_\theta \omega\) and \(\mu_\omega f_\omega^{-1}\) coincide for \(\omega \in A\).

The sample measures \(\mu_\omega\) are called Markov or non-anticipating if for any \(B \in \mathcal{X}\), the random variable \(\mu_\omega(B)\) is measurable with respect to \(\mathcal{F}_- = \sigma(\omega_k, k \leq 0)\).

Let us establish the existence of Markov or non-anticipating sample measures in the compact case. This is a version of Proposition 2.1 from [You86].
THEOREM 5.7. Let \((X, \mathcal{X})\) be a compact space with Borel \(\sigma\)-algebra and let \((\Omega, \mathcal{F}, \mathbb{P}) = (\mathbb{F}^\mathbb{Z}, \mathbb{F}^\mathbb{Z}, Q^\mathbb{Z})\) where the measure \(Q\) is concentrated on continuous transormations of \(\mathbb{F}\). Suppose a probability measure \(\rho\) is invariant for the corresponding Markov transfer operator \(P(\cdot, \cdot)\) defined by (5.8). Then there is an invariant measure \(\mu\) of the random dynamical system with non-anticipating sample measures such that

\[
\rho(B) = \mu(B \times \Omega) = \int_{\Omega} \mu_\omega(B) \mathbb{P}(d\omega).
\]

**Proof:** Let us take any bounded measurable function \(g : X \rightarrow \mathbb{R}\) and define

\[
P_{n,\omega}(g) = \int_X \rho(dx) g(\omega_{-n,0}x).
\]

In particular,

\[(5.17) \quad P_{n,\omega}(B) = P_{n,\omega}(1_B)\]

defines a probability measure \(P_{n,\omega}\) on \((X, \mathcal{X})\). The sequence of these measures is tight, so it always has limit points. Let us identify them.

Let us prove that \(P_{n,\omega}(g)\) is a martingale with respect to \((G_n)_{n \in \mathbb{N}}\), where \(G_n = \sigma(\omega_{-n+1}, \ldots, \omega_0)\). We need to check that for every \(\mathcal{F}_n\)-measurable function \(h\),

\[
\mathbb{E}P_{n,\omega}(g)h(\omega_{-n+1}, \ldots, \omega_0) = \mathbb{E}P_{n+1,\omega}(g)h(\omega_{-n+1}, \ldots, \omega_0).
\]

The right-hand side equals

\[(5.18) \quad \int_X \rho(dx) \mathbb{E}g(\omega_{-n-1,0}x) h(\omega_{-n+1}, \ldots, \omega_0) = \int_X \rho(dx) \mathbb{E}F((\omega_{-n+1}, \ldots, \omega_0), \omega_{-n}x),
\]

where

\[F((\omega_{-n+1}, \ldots, \omega_0), y) = g(\omega_0 \ldots \omega_{-n+1}y) h(\omega_{-n+1}, \ldots, \omega_0).
\]

Since \(\omega_{-n}x\) is independent of \((\omega_{-n+1}, \ldots, \omega_0)\),

\[
\mathbb{E}F((\omega_{-n+1}, \ldots, \omega_0), \omega_{-n}x) = \mathbb{E}a(\omega_{-n}x),
\]

where

\[a(y) = \mathbb{E}F((\omega_{-n+1}, \ldots, \omega_0), y).
\]

So, the right-hand side of (5.18) equals

\[
\mathbb{E} \int_X \rho(dx) a(\omega_{-n}x) = \int_X \rho(dx) a(x)
\]

\[
= \int_X \rho(dx) \mathbb{E}F((\omega_{-n+1}, \ldots, \omega_0), x)
\]

\[
= \int_X \rho(dx) \mathbb{E}g(\omega_0 \ldots \omega_{-n+1}x) h(\omega_{-n+1}, \ldots, \omega_0)
\]

\[= \mathbb{E}P_{n,\omega}(g)h(\omega_{-n+1}, \ldots, \omega_0),
\]

and our claim is established.
Notice that $P_{n,\omega}(g)$ is a bounded martingale. Therefore, it a.s. converges to a random variable $G[g]$, measurable w.r.t. $\mathcal{F}_-$. Let us now take a dense set $\{g_n\}_{n\in\mathbb{N}}$ in $C(\mathbb{X})$. The algebra $R$ generated by this set over $\mathbb{Q}$ is also a countable set. For every $Y \in R$, $G[Y]$ is an a.s.-limit of $P_{n,\omega}(Y)$. Let us now consider the following countable set of conditions:

1. $G[pY + qZ](\omega) = pG[Y](\omega) + qG[Z](\omega)$, $p, q \in \mathbb{Q}$, $Y, Z \in R$.
2. $\min Y \leq G[Y](\omega) \leq \max Y$, $Y \in R$.

Each of these conditions is violated on an exceptional set of zero measure belonging to $\mathcal{F}_-$. Since there are countably many of these conditions, there is a set $\Omega' \in \mathcal{F}_-$ such that $P(\Omega') = 1$ and all the conditions above are satisfied for all $\omega \in \Omega'$. In particular, for each $\omega \in \Omega'$, the functional $Y \mapsto G[Y](\omega)$ is a linear functional on $R$ over $\mathbb{Q}$, with norm bounded by 1. So, this functional can be extended by continuity to a continuous linear functional on $C(\mathbb{X})$ over $\mathbb{R}$ in a unique way. Such functionals can be identified with measures by Riesz’s theorem. Thus for each $\omega \in \Omega'$, we obtain a measure $\mu_\omega(\cdot)$ on $(\mathbb{X}, \mathcal{X})$ that is $\mathcal{F}_-$-measurable w.r.t. $\omega$.

Approximating a generic function $g \in C(\mathbb{X})$ uniformly by functions from $R$, it is easy to see then that if $\omega \in \Omega'$, then $P_{n,\omega}(g) \to \int_{\mathbb{X}} \mu_\omega(dx)g(x)$ as $n \to \infty$. Therefore, measures $P_{n,\omega}$ defined for $\omega \in \Omega'$ by (5.17) converge weakly to $\mu_\omega$. It follows that $P_{n,\theta\omega}$ converges weakly to $\mu_{\theta\omega}$ for $\omega \in \theta^{-1}\Omega'$. On the other hand the construction of $\mu_\omega$ implies that $P_{n,\theta\omega} = P_{n-1,\omega}f_\omega^{-1}$ and thus $P_{n,\theta\omega}$ weakly converges to $\mu_\omega f_\omega^{-1}$ for $\omega \in \Omega'$. We conclude that (5.16) holds on the set of full measure $\Omega' \cap \theta^{-1}(\Omega')$. 

In fact, one does not have to require compactness as the following combination of Theorem 1.7.2 and Corollary 1.7.6 from [Arn98] shows:

**Theorem 5.8.** If $f_\omega$ is a system of i.i.d. continuous maps on a Polish space $(\mathbb{X}, \mathcal{X})$, then a bijection between invariant non-anticipating Markov measures on $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{N}$ and product invariant measures $\rho \times \mathcal{P}$ on $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{N}$ is given by the following two statements:

1. If $\mu$ is an invariant non-anticipating invariant measure on $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{Z}$, then $\mu^+ = \rho \times Q^\mathbb{N}$, where

\[
\rho(A) = \int_{\mathbb{Z}^-} Q^\mathbb{Z}^-(d\omega_-)\mu_{\omega_-}(A),
\]

is invariant for $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{N}$.

2. If $\mu^+ = \rho \times Q^\mathbb{N}$ is invariant on $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{N}$, then the non-anticipating measure $\mu$ on $(\mathbb{X}, \mathcal{X}) \times (\mathcal{F}, \mathcal{F})^\mathbb{Z}$ whose conditional measures w.r.t. $\mathcal{F}^{\mathbb{Z}^-}$ are given by

\[
\mu_\omega(A) = \lim_{n \to \infty} \mu^+(f_\omega^{-n,\theta})^{-1}(A)
\]

is invariant.
5. RANDOM DYNAMICS

4. Ergodicity for Markov processes and systems of i.i.d. transformations

**Definition 5.5.** If $P(\cdot, \cdot)$ is a probability kernel on $(\mathcal{X}, \mathcal{X})$, then for any bounded $\mathcal{X}$-measurable function $g$, $Pg$ is a function defined by

$$Pg(x) = \int_{\mathcal{X}} P(x, dy)g(y), \quad x \in \mathcal{X}.$$  

**Definition 5.6.** Let $P(\cdot, \cdot)$ be a probability kernel on $(\mathcal{X}, \mathcal{X})$ and let $\rho$ be a probability measure on $(\mathcal{X}, \mathcal{X})$. A bounded $\mathcal{X}$-measurable function $g$ on $\mathcal{X}$ is called $(P, \rho)$-invariant if $Pg$ is $\rho$-a.s. constant.

**Definition 5.7.** Let $P(\cdot, \cdot)$ be a probability kernel on $(\mathcal{X}, \mathcal{X})$. A $P$-invariant probability measure $\rho$ on $(\mathcal{X}, \mathcal{X})$ is called ergodic if any $(P, \rho)$-invariant function is $\rho$-a.s. constant.

Ergodicity of measures on $(\mathcal{X} \times \Omega, \mathcal{X} \times \mathcal{F})$ under deterministic transformation $\Theta$ has been defined in the first part of these notes. In fact it was defined in terms of nonexistence of nontrivial invariant sets.

**Definition 5.8.** Let $P(\cdot, \cdot)$ be a probability kernel on $(\mathcal{X}, \mathcal{X})$ and let $\rho$ be a probability measure on $(\mathcal{X}, \mathcal{X})$. A set $A$ is called $(P, \rho)$-invariant if $1_A$ is a $(P, \rho)$-invariant function.

**Theorem 5.9.** Let $P(\cdot, \cdot)$ be a probability kernel on $(\mathcal{X}, \mathcal{X})$. A $P$-invariant probability measure $\rho$ on $(\mathcal{X}, \mathcal{X})$ is ergodic iff any $(P, \rho)$-invariant set $A$ satisfies $\rho(A) \in \{0, 1\}$.

**Proof:** One of the implications is obvious. Let us prove the other one. Suppose that $g$ is a $(P, \rho)$-invariant function. Let us prove that for any $a$ the set $\{x : g(x) > a\}$ is $(P, \rho)$-invariant.

To that end first we prove that if $g$ is $(P, \rho)$-invariant then $|g|$ is $(P, \rho)$-invariant. To see that, we can write $|g| = |Pg| \leq P|g|$. Since, in addition $P$-invariance of $\rho$ implies $\int (P|g| - |g|)d\rho = 0$, we conclude that $P|g| = |g|$ $\rho$-a.s.

Then $g^+ = (g + |g|)/2$ is also $(P, \rho)$-invariant.

Also, if $g_1$ and $g_2$ are $(P, \rho)$-invariant, then $g_1 - g_2$ is $(P, \rho)$-invariant and so are $g_1 \lor g_2 = (g_1 - g_2) \lor 0 + g_2$ and $g_1 \land g_2 = -((-g_1) \lor (-g_2))$.

Since 1 is $(P, \rho)$-invariant, we conclude from all of the above properties that $(n(g - a)^+) \land 1$, $n \in \mathbb{N}$ is a sequence of $(P, \rho)$-invariant functions. Therefore, the indicator of $\{x : g(x) > a\}$ being the pointwise limit of this sequence is also $(P, \rho)$-invariant.

Since indicators of sets $\{x : g(x) > a\}$ for all $a \in \mathbb{Q}$ are invariant, all these sets have $\rho$-measure 0 or 1. Therefore, $g$ is $\rho$-a.s. constant \( \square \)

**Theorem 5.10.** A $P$-invariant measure $\rho$ is ergodic iff $\rho \times P$ is ergodic on $(\mathcal{X} \times \Omega, \mathcal{X} \times \mathcal{F})$, where $(\Omega, \mathcal{F}, P) = (\mathbb{P}^\mathbb{N}, \mathbb{B}^\mathbb{N}, Q^\mathbb{N})$.

**Proof:** Let us first assume that $\rho \times P$ is ergodic. Suppose then that $Pg = g$ $\rho$-a.s. and is not a constant $\rho$-a.s. Then there is a number $c$ such
that $\rho(B) \notin \{0, 1\}$ where $B = \{x : g(x) \geq c\}$. This set is $(P, \rho)$-invariant as follows from the proof of Theorem 5.9. So,

$$1_B(x) = P1_B(x) = \int_Q 1_B(f(x))Q(df).$$

Therefore, if $x \in B$, then $f(x) \in B$ for $Q$-a.e. $f$, and if $x \notin B$, then $f(x) \notin B$ for $Q$-a.e. $f$. So,

$$1_B(\Theta(x, \omega)) = 1_B(\omega_1 x) = 1_B(x), \quad \rho \times P - \text{a.s.}$$

Since $\rho \times P$ is ergodic, $1_B$ is $\rho$-a.s. constant, and we obtain a contradiction with $\rho(B) \notin \{0, 1\}$. Therefore, $g$ is constant $\rho$-a.s., which proves that $\rho$ is ergodic.

Let us prove the converse implication. Suppose that $\rho$ is ergodic and $h$ is a bounded measurable function on $\mathbb{X} \times \Omega$ satisfying $h(\Theta(x, \omega)) = h(x, \omega)$, for $\rho \times P$-a.e. $(x, \omega)$. We must prove that $h$ is constant $\rho \times P$-a.s. Invariance of $h$ implies that for $\rho \times P$-a.e. $(x, \omega)$,

$$h(x, \omega) = h(f_0^m x, (\omega_{m+1}, \omega_{m+2}, \ldots)), \quad m \in \mathbb{N}.$$ 

Let $h_0(x) = \mathbb{E} h(x, \omega)$. Then, for $\rho$-a.e. $x$,

$$h_0(x) = \mathbb{E}[h(x, (\omega_1, \omega_2, \ldots))]_{\omega_1} = \mathbb{E}[\mathbb{E}[h(\omega_1 x, (\omega_2, \ldots))]_{\omega_1}$$

$$= \mathbb{E}h_0(\omega_1 x) = P h_0(x).$$

Due to ergodicity of $\rho$, there is a constant $c$ such that $h_0 = c$.

Also, if

$$h_m(x, (\omega_1, \ldots, \omega_m)) = \mathbb{E}[h(x, (\omega_1, \omega_2, \ldots))]_{\omega_1, \ldots, \omega_m]},$$

then for $\rho \times P$-a.e. $(x, \omega)$,

$$h_m(x, (\omega_1, \ldots, \omega_m)) = \mathbb{E}[h(\omega_0^m x, (\omega_{m+1}, \omega_{m+2}, \ldots)) \mid (\omega_1, \ldots, \omega_m)]$$

$$= h_0(\omega_0^m x) = c.$$ 

For $\rho$-a.e. $x$ we can apply P. Lévy’s theorem on continuity of conditional expectation and obtain that for $\rho \times P$-a.e. $(x, \omega)$,

$$c = \mathbb{E}[h(x, (\omega_1, \omega_2, \ldots)) \mid (\omega_1, \ldots, \omega_m)]$$

$$\rightarrow \mathbb{E}[h(x, (\omega_1, \omega_2, \ldots)) \mid (\omega_1, \omega_2, \ldots)] = h(x, (\omega_1, \omega_2, \ldots)),$$

i.e., $h(x, (\omega_1, \omega_2, \ldots)) = c$. 

\[ \square \]

4.1. Structure of the set of invariant distributions. Similarly to the deterministic case, the following statements hold true for the set $I_P$ of invariant measures under a kernel $P(\cdot, \cdot)$ on a Borel space $(\mathbb{X}, \mathcal{X})$:

**Theorem 5.11.**

1. $I_P$ is a convex set.
2. $P(\cdot, \cdot)$-ergodic measures are extreme points of $I_P$.
3. Two different ergodic measures are mutually singular.
4. Every measure $\rho \in I_P$ can be represented as

$$\rho = \int_X \nu(dx) \rho_x,$$

$$\rho(A) = \int_X \nu(dx) \rho_x(A), \quad A \in \mathcal{X},$$

for some kernel $\rho(\cdot)$ on $(\mathbb{X}, \mathcal{X})$. For $\rho$-almost all $x$, the measure $\rho_x(\cdot)$ is ergodic.

Similar statements hold true for random dynamical systems

4.2. Ergodicity of a measures with respect to a transition kernel vs. ergodicity of the associated Markov measure on the trajectory space.

**Theorem 5.12.** A $P$-invariant measure $\rho$ on $(\mathbb{X}, \mathcal{X})$ is ergodic if and only if the Markov measure $P_\rho$ on the space $(\mathbb{X}^{\mathbb{Z}_+}, \mathcal{X}^{\mathbb{Z}_+})$ generated by initial distribution $\rho$ and transition probability $P(\cdot, \cdot)$ according to (5.4) is ergodic with respect to shift $\theta$ on $(\mathbb{X}^{\mathbb{Z}_+}, \mathcal{X}^{\mathbb{Z}_+})$ defined by

$$\theta(x_0, x_1, x_2, \ldots) = (x_1, x_2, \ldots), \quad (x_0, x_1, x_2, \ldots) \in \mathbb{X}^{\mathbb{Z}_+}.$$

**Proof:** Suppose $(\mathbb{X}^{\mathbb{Z}_+}, \mathcal{X}^{\mathbb{Z}_+}, P, \theta)$ is ergodic. Let $A \in \mathcal{X}$ be $(P, \rho)$ invariant. The set $A^{\mathbb{Z}_+} \in \mathcal{X}^{\mathbb{Z}_+}$ is invariant with respect to $\theta$, so $P_\rho(A^{\mathbb{Z}_+}) \in \{0, 1\}$ due to ergodicity. Also, $(P, \rho)$-invariance of $A$ implies that $\rho(A) = P_\rho(A^{\mathbb{Z}_+})$. Hence $\rho(A) \in \{0, 1\}$, which proves ergodicity of $\rho$.

Let us prove the converse implication. Suppose $\rho$ is ergodic with respect to $P$. Let us take any $\theta$-invariant set $A \in \mathcal{X}$ and prove that $P_\rho(A) \in \{0, 1\}$. Let $\delta_x$ be the Dirac measure concentrated at $x \in \mathbb{X}$. Using $A = \theta^{-1} A$ and the Markov property, for any $x \in \mathbb{X}$ we can write

$$P_{\delta_x}(A) = P_{\delta_x}(\theta^{-1} A)$$

$$= P_{\delta_x}((X_1, X_2, \ldots) \in A)$$

$$= \int_{\mathbb{X}} P(x, dy) P_{\delta_y}((X_0, X_1, \ldots) \in A)$$

$$= \int_{\mathbb{X}} P(x, dy) P_{\delta_y}(A).$$

So, we obtain that the function $f : \mathbb{X} \to [0, 1]$ defined by $f(x) = P_{\delta_x}(A)$ is constant. In particular, it is $(P, \rho)$-invariant, so by ergodicity it is $\rho$-a.s. constant. Let us denote that constant by $p$.

Let us denote $F_k = \sigma\{X_0, \ldots, X_k\}, \ k \in \mathbb{Z}_+$. Then

$$1_A \overset{P-\text{a.s.}}{\lim}_{k \to \infty} P(A|F_k) = \lim_{k \to \infty} P(\theta^{-k} A|F_k)$$

$$= \lim_{k \to \infty} P(\theta^{-k} A|X_k) = \lim_{k \to \infty} p = p,$$

which can hold only if $p \in \{0, 1\}$ so that $P(A) \in \{0, 1\}$. \qed
CHAPTER 6

Invariant distributions for Markov chains in discrete spaces

1. Perron–Frobenius Theorem

There are several statements that pass under this title. Here we follow the exposition in [KH95].

**Theorem 6.1.** Let \( d \in \mathbb{N} \) and suppose \( A \) is a \( d \times d \) matrix with nonnegative entries such that for some \( n \in \mathbb{N} \) all entries of \( A^n \) are positive. Then there is an eigenvector \( x_0 \) with all positive components. If \( x \) is an eigenvector with nonnegative components, then \( x = cx_0 \) for some \( c > 0 \). The eigenvalue \( \lambda \) associated with \( x_0 \) is simple and all other eigenvalues are less than \( \lambda \) in absolute value.

**Proof:** For \( x \in \mathbb{R}^d \) we will write \( |x|_1 = \sum_{i=1}^{d} |x_i| \). Let us introduce \( C = \{ x \in \mathbb{R}^N : x_i \geq 0, i = 1, \ldots, d \} \) and \( \Delta = \{ x \in C : \sum_{i=1}^{d} x_i = 1 \} \). The simplex \( \Delta \) is the convex hull of its extreme points \( e_1 = (1, 0, 0, \ldots), e_2 = (0, 1, 0, \ldots), \ldots, e_d = (0, 0, \ldots, 1) \).

For every \( x \in \Delta \) we introduce \( Tx = Ax/|Ax|_1 \in \Delta \).

Let us study some properties of \( \Delta_0 = \bigcap_{k=1}^{\infty} T^k \Delta \). Images of convex closed sets under \( T \) are convex closed sets, so all sets \( T^k \Delta, k \in \mathbb{N} \) are convex and closed, and so is their intersection \( \Delta_0 \). Since \( T^{k+1} \Delta \subset T^k \Delta \) for all \( k \in \mathbb{N} \), we obtain \( \Delta_0 \neq \emptyset \).

The conditions of the theorem imply that

\[
T^m \Delta \subset \text{Int} \Delta, \quad m \geq n,
\]

where \( \text{Int} \Delta \) denotes the relative interior of \( \Delta \). Therefore, \( \Delta_0 \subset \text{Int} \Delta \).

Let us show that \( \Delta_0 \) has at most \( d \) extreme points. Namely, let us choose an increasing sequence \((n_k)_{k \in \mathbb{N}}\) such that \( \lim_{k \to \infty} T^{n_k} e_i = x_i \) for some \( x_i \in \Delta_0 \), \( i = 1, \ldots, d \), and prove that all ext \( \Delta_0 \subset \{x_1, \ldots, x_d\} \).

If \( x \in \Delta_0 \) then \( x \in T^{n_k} \Delta \) for all \( k \in \mathbb{N} \) and therefore \( x \) is a convex combination of extreme points of \( T^{n_k} \Delta \). All these are images of extreme points of \( \Delta \) under \( T^{n_k} \), so we obtain a representation \( x = \sum_{i=1}^{d} \alpha_i^{(k)} T^{n_k} e_i \) for some \( \alpha_i^{(k)} \geq 0 \) satisfying \( \sum_{i=1}^{d} \alpha_i^{(k)} = 1, k \in \mathbb{N} \). Choosing an increasing sequence \((k(m))_{m \in \mathbb{N}}\) such that \( \lim_{m \to \infty} \alpha_i^{(n_k(m))} = \alpha_i \) for some \( \alpha_i \), we obtain \( x = \sum_{i=1}^{d} \alpha_i x_i \), so if \( x \) is an extreme point it has to coincide with one of \( x_i \), \( i = 1, \ldots, d \).
Since $\Delta_0 = T\Delta_0$ and $\text{ext} \Delta_0$ is finite, we obtain $\text{ext} \Delta_0 = T\text{ext} \Delta_0$, in other words $T$ acts a permutation on $\text{ext} \Delta_0$. Therefore there is a number $m \in \mathbb{N}$ such that $T^m x = x$ for every $x \in \text{ext} \Delta_0$. So, every $x \in \text{ext} \Delta_0$ is an eigenvector of $A^n$ with a positive eigenvalue.

Let us prove that, in fact, there cannot be two distinct eigenvectors with those properties. There are two cases to exclude: (i) there are $x, y \in \Delta_0$ and $\nu > 0$ such that $x \neq y$ and $A^m x = \nu x$ and $A^m y = \nu y$; (ii) there are $x, y \in \Delta_0$ and $\nu > \mu > 0$ such that $A^m x = \nu x$ and $A^m y = \mu y$.

In case (i), there is a number $\nu > \mu > 0$ such that $z = x + \mu (y - x) \in \partial \Delta$. Therefore $A^{mn} z = \nu^n z \in \partial \mathcal{C}$, so $T^{mn} z \in \partial \Delta$ which contradicts $T^{mn} \Delta \subset \text{Int} \Delta$.

In case (ii), we can use $y \in \text{Int} \Delta$ to choose $\varepsilon > 0$ small enough to ensure that $z = y - \varepsilon x \in \text{Int} \mathcal{C}$. Then $A^{km} z = \mu^k y - \varepsilon \nu^k x$, so for sufficiently large $k$, $A^{km} z \notin \mathcal{C}$ which is a contradiction.

We conclude that $\Delta_0$ contains a unique point $x_0$. Recalling that $\Delta_0$ is $T$-invariant, we obtain that $Tx_0 = x_0$, i.e., $Ax_0 = \lambda x_0$ for some $\lambda > 0$.

Let us show that if $Ax = \mu x$ for any $x \in \mathcal{C}$ that is not a multiple of $x_0$ and satisfies $x \neq 0$, then $|\mu| < \lambda$.

If $\mu = \pm \lambda$, then we can find $\alpha \in \mathbb{R}$ such that $x = x_0 + \alpha x \in \partial \mathcal{C} \setminus \{0\}$. Then $A^{2k} z = \lambda^{2k} z$ for all $k \in \mathbb{N}$ which contradicts (6.1).

If $\mu \in \mathbb{R}$ and $|\mu| > \lambda$, then we have already proved that $x \notin \mathcal{C}$. Also we can find $\varepsilon > 0$ small enough to guarantee that $z = x_0 + \varepsilon x \in \text{Int} \mathcal{C}$. Then for large values of $m$, the direction of $A^{2m+1} z$ will be close to the direction of $\pm x$, so we will have $A^{2m+1} z \notin \mathcal{C}$ which contradicts $A$-invariance of $\mathcal{C}$.

In general, if $\mu = \rho e^{i2\pi \phi}$ for some $\rho > 0$ and $\phi \in \mathbb{R}$, then there is a plane $L$ such that the action of $A$ on $L$ is multiplication by $\rho$ and rotation by angle $\phi$.

If $\phi = k/l$ for some $k \in \mathbb{Z}$ and $l \in \mathbb{N}$, then $A^l$ acts on $L$ as multiplication by $\rho^l$, and we can apply the above analysis to $A^l$ and conclude that $\rho < \mu$.

If $\phi$ is irrational, then for any $\delta > 0$, the direction of $A^k x$ will get $\delta$-close to that of $x$ for an infinitely many values of $k$. Let us denote the set of those values of $k$ by $K$. So we choose $x \in L \setminus \mathcal{C}$ and $\varepsilon$ small enough to ensure $z = x_0 + \varepsilon x \in \text{Int} \mathcal{C}$. If $\rho > \lambda$, then for sufficiently large $k \in K$, the direction of $A^k z$ will be close to that of $x$, and so $A^k z \notin \mathcal{C}$ for those values of $k$, a contradiction. If $\rho = \lambda$, then we choose $y \in L$ so that $z = \frac{x_0 + \alpha y}{|x_0 + \alpha y|} \in \partial \Delta$.

Since $\phi$ is irrational,

$$T^k z = \frac{x_0 + e^{i2\pi \phi} \alpha y}{|x_0 + e^{i2\pi \phi} \alpha y|}$$

will get arbitrarily close to $\partial \Delta$ for infinitely many values of $k$ which contradicts (6.1).

To finish the proof that $\lambda$ is simple, we must prove that there are no nontrivial Jordan blocks for matrix $A$. Suppose there is such a block. Then there is a vector $x_1$ such that $Ax_1 = \lambda x_1 + x_0$. Let $y = \lambda x_1 + x_0$. Then $Ay = \lambda (y + x_0)$. Now we can choose $\varepsilon$ such that $z = x_0 - \varepsilon y \in \text{Int} \mathcal{C}$, and
compute by induction
\[ A^n z = \lambda^n((1 - n\varepsilon)x_0 - \varepsilon y), \quad n \geq 0. \]
We see that for sufficiently large \( n \), \( A^n z \notin C \) which is a contradiction. \( \square \)

2. Hilbert projective metric approach.

Let us give another proof of the Perron–Frobenius theorem based on contraction in so called Hilbert projective metric. We follow exposition in [Bal00] that, in turn, follows [Fur60]

For two points \( x, y \in \mathbb{R}^d \) we will write \( x \leq y \) and \( y \geq x \) if \( y - x \in C \).

We write \( x \sim y \) if \( x = ry \) for some \( r > 0 \). Let us denote \( C^* = C \setminus \{0\} \) and introduce
\[
\alpha(x,y) = \sup\{r \in \mathbb{R}_+ : rx \leq y\}, \quad x,y \in C^*.
\]

We note that \( \alpha(x,y) \in [0, \infty) \) for all \( x,y \in C \) and \( \alpha(x,x) = 1 \) for all \( x \in C \). Also,
\[
(6.2) \quad \alpha(x,z) \geq \alpha(x,y)\alpha(y,z), \quad x,y,z \in C^*,
\]

since \( z \geq \alpha(y,z)y \) and \( y \geq \alpha(x,y)x \). Next we define \( \Gamma(x,y) = \alpha(x,y)\alpha(y,x) \).

Inequality (6.2) implies
\[
(6.3) \quad \Gamma(x,z) \geq \Gamma(x,y)\Gamma(y,z), \quad x,y,z \in C^*.
\]

Applying this to \( z = x \), and noticing that \( \Gamma(x,x) = 1 \), we obtain
\[
0 \leq \Gamma(x,y) \leq 1, \quad x,y \in C^*.
\]

Moreover, \( \Gamma(x,y) = 1 \) iff \( x \sim y \). One part of this statement is obvious since for every \( r > 0 \), \( \alpha(x,rx) = r \). To see the converse implication, we notice that \( \alpha(x,y)\alpha(y,x) = 1 \) implies that there is \( r > 0 \) such that \( x \geq ry \) and \( y \geq r^{-1}x \). The latter is equivalent to \( ry \geq x \), so these two inequalities can hold together only if \( x = ry \).

We can now define the Hilbert projective metric by
\[
\Theta(x,y) = -\ln \Gamma(x,y) = -\ln \alpha(x,y) - \ln \alpha(y,x) \in [0, \infty].
\]

Notice that \( \Theta(x,y) \) is infinite if \( x,y \in \partial C \) and \( x \not\sim y \). The following triangle inequality follows from (6.3):
\[
\Theta(x,z) \leq \Theta(x,y) + \Theta(y,z), \quad x,y,z \in C^*.
\]

From the properties of \( \Gamma \), we derive that \( \Theta(x,y) = \Theta(x,ry) \) for any \( x,y \in C^* \) and \( r > 0 \), and \( \Theta(x,y) = 0 \) iff \( x \sim y \), so \( \Theta \) is a pseudo-metric on \( C^* \) and a metric on \( \Delta = C^*/\sim \).

Other ways to represent \( \Theta \) are
\[
\Theta(x,y) = -\ln \left( \min_i \frac{x_i}{y_i} \cdot \min_j \frac{y_j}{x_j} \right) = \ln \left( \max_i \frac{y_i}{x_i} \cdot \max_j \frac{x_j}{y_j} \right) = \ln \max_{ij} \frac{x_jy_i}{y_jx_i}.
\]

Let us prove that the map \( T : \Delta \to \Delta \) is nonexpanding. First, we notice that \( x - ry \in C \) implies \( A(x - ry) = Ax - rAy \in C \). Therefore \( \alpha(Ax, Ay) \geq \alpha(x,y) \), so \( \Theta(Ax, Ay) \leq \Theta(x,y) \).
Assuming now that all the entries of \( A \) are positive, let us prove that the contraction coefficient is actually strictly less than 1. First we recall that in this case \( T\Delta \) is a compact subset of \( \text{Int} \Delta \). Therefore, expressions \( \ln(x_jy_i/(y_jx_i)) \) are uniformly bounded with respect to \( x, y \in \Delta \), and
\[
D = \text{diam}(T\Delta) = \sup\{\Theta(x, y) : x, y \in \Delta\} < \infty,
\]
or, equivalently
\[
\delta = \inf\{\Gamma(x, y) : x, y \in \text{C}^*\} > 0.
\]
These numbers are related by \( \delta = e^{-D} \). The following theorem establishes the strict contraction property of \( \theta \).

**Theorem 6.2.** If \( A \) has all positive entries, then for all \( x, y \in \Delta \),
\[
(6.4) \quad \Theta(Tx, Ty) \leq \frac{1 - \sqrt{\delta}}{1 + \sqrt{\delta}} \Theta(x, y).
\]

**Proof:** Let us take two distinct points \( x, y \in \Delta \). We can assume that \( \Theta(x, y) < \infty \), i.e., both numbers \( \alpha_1 = \alpha(x, y) \) and \( \alpha_2 = \alpha(y, x) \) are nonzero. Then \( y - \alpha_1 x \in \text{C}^* \) and \( x/\alpha_2 - y \in \text{C}^* \). Applying the definition of \( \delta \) to those vectors, we obtain that there are two numbers \( \lambda_1, \lambda_2 \geq 0 \) such that \( \lambda_1 \lambda_2 \geq \delta \) and
\[
A(x/\alpha_2 - y) \geq \lambda_1 A(y - \alpha_1 x),
\]
\[
A(y - \alpha_1 x) \geq \lambda_2 A(x/\alpha_2 - y),
\]
i.e.,
\[
Ax \geq \frac{1 + \lambda_1}{\alpha_2 + \lambda_1 \alpha_1} Ay, \quad Ay \geq \frac{\alpha_1 + \lambda_2}{1 + \lambda_2} Ax.
\]
So,
\[
\Theta(Tx, Ty) = \Theta(Ax, Ay) \leq -\ln \left( \frac{1 + \lambda_1}{\frac{1 + \lambda_2}{(\frac{1}{\alpha_2 + \lambda_1 \alpha_1})(1 + \lambda_2)} \left( \frac{1}{\lambda_2} - \lambda_1 \right) e^s}{(e^s + \frac{1}{\lambda_2})(e^s + \lambda_1)} ds \right)
\]
\[
= \frac{1}{\lambda_1 \lambda_2} \left( \int_0^{1/(\alpha_1 \alpha_2)} \frac{1}{\lambda_2} - \lambda_1 dt \right) = \left( \int_0^{1/(\alpha_1 \alpha_2)} e^s \frac{\lambda_1 e^s}{(e^s + \frac{1}{\lambda_2})(e^s + \lambda_1)} ds \right)
\]
Elementary analysis shows that the maximum of the expression \( \lambda_1 t/((t + \frac{1}{\lambda_2})(t + \lambda_1)) \) over \( t \geq 0 \) is attained at \( t = \sqrt{\lambda_1/\lambda_2} \) and equals \((1 + 1/\sqrt{\lambda_1 \lambda_2})^{-2}\). Therefore,
\[
\Theta(Tx, Ty) \leq \ln \frac{1}{\alpha_1 \alpha_2} \left( \frac{1}{\lambda_1 \lambda_2} - \frac{1}{(\frac{1}{\sqrt{\lambda_1 \lambda_2}} + 1)^2} \right) = \frac{1}{\sqrt{\lambda_1 \lambda_2}} - \frac{1}{1 + \sqrt{\lambda_1 \lambda_2}} \Theta(x, y) = \frac{1 - \sqrt{\delta}}{1 + \sqrt{\delta}} \Theta(x, y).
\]
The function $t \mapsto (1 - t) / (1 + t)$ decays for $t \geq 0$, so the last expression is bounded by the right-hand side of (6.4).

This strict contraction property implies existence and uniqueness of a fixed point of $T$ on $\Delta$, along with exponential convergence of iterations of $T^n \Delta$ to that fixed point. If the requirement of positivity of all entries is satisfied not for $A$ but for $A^n$ for some $m \in \mathbb{N}$, then the above reasoning still can be applied to $A^n$, since $A$ itself is non-expanding.

3. Another proof of ergodic theorem for Markov chain

We follow Bulinskii & Shiryaev. I think their proof is borrowed from [Chu67].

**Theorem 6.3.** Let $P$ be a transition probability on $X$ where $X = \mathbb{N}$ or $\{1, \ldots, N\}$ for some $N \in \mathbb{N}$. Suppose that there is $j_0 \in X$, $n_0 \in \mathbb{N}$ and $\delta > 0$ such that $P_{ij}^{n_0} \geq \delta$ for all $i \in \mathbb{N}$. Then there is a vector $\pi : X \to \mathbb{R}_+$ with $\sum_{i \in X} \pi(i) = 1$ such that

$$\pi(i) = \sum_{j \in X} \pi(j) P_{ji}, \quad i \in X,$$

and

$$|P_{ij}^n - \pi(j)| \leq (1 - \delta)^{\lfloor n/n_0 \rfloor}, \quad n \in \mathbb{N}.$$

**Proof:** Let

$$m_n(j) = \inf_{i \in X} P_{ij}^n, \quad M_n(j) = \sup_{i \in X} P_{ij}^n, \quad n \in \mathbb{N}.$$

Clearly, $m_n(j) \leq P_{ij}^n \leq M_n(j)$ for all $n, i, j$. We notice that $m_n(j)$ is nondecreasing in $n$, $M_n(j)$ is nonincreasing, since for any $n, n' \in \mathbb{N},$

$$m_{n' + n}(j) = \inf_{i \in X} \sum_k P_{ik}^n P_{k'j}^{n'}(k, j) \geq m_n(j) \inf_{i \in X} \sum_k P_{ik}^{n'} = m_n(j),$$

$$M_{n' + n}(j) = \sup_{i \in X} \sum_k P_{ik}^n P_{k'j}^{n'}(k, j) \leq m_n(j) \sup_{i \in X} \sum_k P_{ik}^{n'} = M_n(j),$$

Let us prove that $\lim_{n \to \infty} (M_n(j) - m_n(j)) = 0$.

$$M_n(j) - m_n(j) = \sup_{a, b} P_{aj}^n + \sup_{-b} (-P_{bj}^n) = \sup_{a, b} (P_{aj}^n - P_{bj}^n)$$

$$= \sup_{a, b} \sum_k (P_{ak}^{n_0} - P_{bk}^{n_0}) P_{kj}^{n-n_0}$$

$$= \sup_{a, b} \left( \sum_k^+ (P_{ak}^{n_0} - P_{bk}^{n_0}) P_{kj}^{n-n_0} + \sum_k^- (P_{ak}^{n_0} - P_{bk}^{n_0}) P_{kj}^{n-n_0} \right)$$

$$\leq \sup_{a, b} \left( M_{n-n_0}(j) \sum_k^+ (P_{ak}^{n_0} - P_{bk}^{n_0}) + m_{n-n_0}(j) \sum_k^- (P_{ak}^{n_0} - P_{bk}^{n_0}) \right),$$
where $\sum_k^+(\text{respectively}, \sum_k^−)$ denotes the sum over nonnegative (respectively, negative) values of $P_n(a, k) - P_n(b, k)$. Since $\sum_k P_n^0 = \sum_k P_n^0 = 1$, we have

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) + \sum_k^−(P_n^0 a_k - P_n^0 b_k) = 0,$$

so,

$$\sum_k^−(P_n^0 a_k - P_n^0 b_k) = -\sum_k^+(P_n^0 a_k - P_n^0 b_k),$$

and

$$M_n(j) - m_n(j) \leq (M_{n−n_0}(j) - m_{n−n_0}(j)) \sup_{a,b} \sum_k^+(P_n^0 a_k - P_n^0 b_k).$$

We claim that $\sum_k^+(P_n^0 a_k - P_n^0 b_k) < 1 - \delta$ for all $a, b$. In fact, if $j_0$ is not included in the summation $\sum_k^+$, then

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) \leq \sum_k P_n^0 a_k - P_n^0 a_{j_0} \leq 1 - \delta.$$  

If $j_0$ is included in the summation $\sum_k^+$, then

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) \leq \sum_k P_n^0 a_k - P_n^0 b_{j_0} \leq 1 - \delta,$$

so our claim is established and from (6.7) and monotonicity of $M$ and $m$, we conclude by induction that $M_n(j) - m_n(j) \leq (1 - \delta)^{[n/n_0]}$. This, along with monotonicity and boundedness of $M_n(j)$ and $m_n(j)$, implies now that there is a common limit $\pi(j)$ for $M_n(j)$ and $m_n(j)$, and, moreover, $|M_n(j) - \pi(j)| \leq (1 - \delta)^{[n/n_0]}$ and $|m_n(j) - \pi(j)| \leq (1 - \delta)^{[n/n_0]}$.

Invariance of $\pi$ under the action of $P$ follows by taking $n \to \infty$ on both sides of $P_{ij}^{n+1} = P_i^k P_{kj}$.

4. Entropy approach

This proof is based on the fact that relative entropy of a Markov chain is monotone in time. We follow [Lig05] closely.

Let $\nu$ and $\mu$ be two probability distributions on $X = \{1, \ldots, N\}$. We define the relative entropy (also known as Kullback–Leibler divergence) of $\mu$ with respect to $\nu$ by

$$H(\mu | \nu) = \sum_{i=1}^N \mu(i) \ln \frac{\mu(i)}{\nu(i)}.$$

Introducing

$$\varphi(x) = \begin{cases} x \ln x, & x > 0, \\ 0, & x = 0, \end{cases}$$

where $\sum_k^+(\text{respectively}, \sum_k^−)$ denotes the sum over nonnegative (respectively, negative) values of $P_n^0(a, k) - P_n^0(b, k)$. Since $\sum_k P_n^0 = \sum_k P_n^0 = 1$, we have

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) + \sum_k^−(P_n^0 a_k - P_n^0 b_k) = 0,$$

so,

$$\sum_k^−(P_n^0 a_k - P_n^0 b_k) = -\sum_k^+(P_n^0 a_k - P_n^0 b_k),$$

and

$$M_n(j) - m_n(j) \leq (M_{n−n_0}(j) - m_{n−n_0}(j)) \sup_{a,b} \sum_k^+(P_n^0 a_k - P_n^0 b_k).$$

We claim that $\sum_k^+(P_n^0 a_k - P_n^0 b_k) < 1 - \delta$ for all $a, b$. In fact, if $j_0$ is not included in the summation $\sum_k^+$, then

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) \leq \sum_k P_n^0 a_k - P_n^0 a_{j_0} \leq 1 - \delta.$$  

If $j_0$ is included in the summation $\sum_k^+$, then

$$\sum_k^+(P_n^0 a_k - P_n^0 b_k) \leq \sum_k P_n^0 a_k - P_n^0 b_{j_0} \leq 1 - \delta,$$

so our claim is established and from (6.7) and monotonicity of $M$ and $m$, we conclude by induction that $M_n(j) - m_n(j) \leq (1 - \delta)^{[n/n_0]}$. This, along with monotonicity and boundedness of $M_n(j)$ and $m_n(j)$, implies now that there is a common limit $\pi(j)$ for $M_n(j)$ and $m_n(j)$, and, moreover, $|M_n(j) - \pi(j)| \leq (1 - \delta)^{[n/n_0]}$ and $|m_n(j) - \pi(j)| \leq (1 - \delta)^{[n/n_0]}$.

Invariance of $\pi$ under the action of $P$ follows by taking $n \to \infty$ on both sides of $P_{ij}^{n+1} = P_i^k P_{kj}$.

4. Entropy approach

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$$H(\mu | \nu) = \sum_{i=1}^N \mu(i) \ln \frac{\mu(i)}{\nu(i)}.$$

Introducing

$$\varphi(x) = \begin{cases} x \ln x, & x > 0, \\ 0, & x = 0, \end{cases}$$

we can rewrite relative entropy as
\[ H(\mu|\nu) = \sum_{i=1}^{N} \nu(i) \varphi \left( \frac{\mu(i)}{\nu(i)} \right). \]

**Lemma 6.1.** Suppose \( \nu \) is a probability measure on \( X \) such that \( \nu(i) > 0 \) for all \( i \in X \), and assume that \( \nu \) is invariant under \( P \). Then for any probability measure \( \mu \) on \( X \) we have
\[ H(\mu P|\nu) \leq H(\mu|\nu), \]
and identity holds if and only if
\[ \frac{\mu(i)}{\nu(i)} = \frac{\mu(j)}{\nu(j)} \]
holds for all \( i \) and \( j \) such that there is \( l \) such that \( P(i,l) > 0 \) and \( P(j,l) > 0 \).

**Proof:**
\[
H(\mu P|\nu) = \sum_{i=1}^{N} \nu(i) \varphi \left( \sum_{j=1}^{N} \frac{\mu(j)P(j,i)}{\nu(i)} \right) \\
= \sum_{i=1}^{N} \nu(i) \varphi \left( \sum_{j=1}^{N} \frac{\mu(j)}{\nu(j)} \cdot \frac{\nu(j)P(j,i)}{\nu(i)} \right).
\]
Since \( \varphi \) is convex and
\[ \sum_{j=1}^{N} \frac{\nu(j)P(j,i)}{\nu(i)} = 1, \]
the right-hand side of (6.8) we can continue
\[
H(\mu P|\nu) \leq \sum_{i=1}^{N} \nu(i) \sum_{j=1}^{N} \varphi \left( \frac{\mu(j)}{\nu(j)} \right) \cdot \frac{\nu(j)P(j,i)}{\nu(i)} \\
= \sum_{j=1}^{N} \nu(j) \varphi \left( \frac{\mu(j)}{\nu(j)} \right) \sum_{i=1}^{N} P(j,i) = H(\mu|\nu).
\]
Also, since \( \nu(i) > 0 \) for all \( \nu \), the inequality becomes strict if and only if at least one inner summation over \( j \) produces a strict inequality. Due to strict convexity of \( \varphi \), this happens if there are nonzero weights \( \frac{\nu(j)P(j,i)}{\nu(i)} \) corresponding to different values of \( \frac{\mu(j)}{\nu(j)} \).

**Proof of convergence:** Following the Krylov–Bogolyubov approach, we conclude that for every \( i \in X \), every limit point of the sequence
\[
\frac{1}{n} \sum_{k=1}^{n} P(i,j)
\]
is an invariant distribution. Let \( \nu \) be one of these limit points. Then for every \( i \in \mathbb{X} \), \( \nu(i) > 0 \) since \( \nu(i) \geq \sum \nu(j)P^m(j,i) > 0 \). So we can apply Lemma 6.1 and conclude that for any distribution \( \mu \) on \( \mathbb{X} \), \( H(\mu P^n|\nu) \) is a nonincreasing function. Continuity of \( H \) implies that if there is a sequence \( n' \to \infty \) such that \( \mu P^n \) converges to a distribution \( \lambda \), \( H(\lambda|\nu) = \lim_{n \to \infty} H(\mu P^n|\nu) \). On the other hand, for any \( k \in \mathbb{N} \), \( \lambda P^k = \lim \mu P^{n+k} \), so \( H(\lambda P^k|\nu) = \lim_{n \to \infty} H(\mu P^n|\nu) = H(\lambda|\nu) \). Applying Lemma 6.1 and the condition of the theorem we conclude that \( \nu = \lambda \).

So we just proved that every partial limit of \( \mu P^n \) equals \( \nu \), and the theorem follows. \( \square \)

5. Coupling

Let us prove a slightly weaker version of Theorem 6.3 using a method called coupling that was introduced by W. Doeblin [Doe38].

Let us introduce a Markov process \((X_n, Y_n)\) on \( \mathbb{X} \times \mathbb{X} \) with the following properties: (i) \( X_n \) and \( Y_n \) evolve independently and according to transition probabilities \( P_{ij} \) until the moment when \( X_n = Y_n \); (ii) from that point on, the processes \( X_n \) and \( Y_n \) evolve together, staying equal to each other for all future times, and the evolution still happens according to transition probabilities \( P_{ij} \).

To describe this more formally, we need to introduce a transition probability \( Q_{(i_1,i_2),(j_1,j_2)} \) on \( \mathbb{X} \times \mathbb{X} \) by

\[
Q_{(i_1,i_2),(j_1,j_2)} = \begin{cases} 
P_{i_1j_1}P_{i_2j_2}, & i_1 \neq i_2, \\
P_{i_1j_1}, & i_1 = i_2, j_1 = j_2, \\
0, & i_1 = i_2, j_1 \neq j_2.
\end{cases}
\]

Notice that the components \( X_n \) and \( Y_n \) of the resulting Markov process are Markov processes with transition probabilities \( P \): for any \( i_1, j_1 \in \mathbb{X} \),

\[
P\{X_{n+1} = j_1, X_n = i_1\} = \sum_{i_2,j_2} P\{X_{n+1} = j_1, Y_{n+1} = j_2, X_n = i_1, Y_n = i_2\}
\]

\[
= \sum_{i_2,j_2} P\{X_n = i_1, Y_n = i_2\}Q_{(i_1,i_2),(j_1,j_2)}
\]

\[
= \sum_{i_2 \neq i_1} \sum_{j_2} P\{X_n = i_1, Y_n = i_2\}P_{i_1j_1}P_{i_2j_2} + P\{X_n = i_1, Y_n = i_1\}P_{i_1j_1}
\]

\[
= \sum_{i_2 \neq i_1} P\{X_n = i_1, Y_n = i_2\}P_{i_1j_1} + P\{X_n = i_1, Y_n = i_1\}P_{i_1j_1}
\]

\[
= P\{X_n = i_1\}P_{i_1j_1},
\]

and a similar computation is valid for \( P(Y_{n+1} = j_2, Y_n = i_2) = P_{i_2,j_2} \) for all \( i_2, j_2 \in \mathbb{X} \).
Let us denote the distribution of \((X_n, Y_n)_{n \in \mathbb{Z}_+}\) on \(\mathbb{X}^{\mathbb{Z}_+}\) given \((X_0, Y_0) = (i, j)\) by \(Q_{i,j}\) and prove that 
\[
\lim_{n \to \infty} Q_{i,j}\{X_n = Y_n\} = 1, \quad i, j \in \mathbb{X}.
\]
Let us denote 
\[
\tau = \min\{n \in \mathbb{Z}_+: X_n = Y_n\}.
\]
Then 
\[
Q_{i,j}\{\tau \leq n_0\} \geq \sum_{k \in \mathbb{X}} P_{i,k}^{n_0} P_{j,k}^{n_0} \geq \delta^2
\]
or, equivalently, 
\[
Q_{i,j}\{\tau \geq n_0\} \leq 1 - \delta^2.
\]
Iterating this reasoning and using the Markov property gives then that 
\[
Q_{i,j}\{X_{rn_0} \neq Y_{rn_0}\} = Q_{i,j}\{\tau \geq rn_0\} \leq (1 - \delta^2)^r
\]
and 
\[
Q_{i,j}\{X_n \neq Y_n\} \leq (1 - \delta^2)^{[n/n_0]},
\]
so our claim holds. Now 
\[
|P_{i,j}^{n} - P_{i,j}^{n}| = |Q_{i_1,i_2}\{X_n = j\} - Q_{i_1,i_2}\{Y_n = j\}| \\
\leq Q_{i_1,i_2}\{X_n \neq Y_n\} \leq (1 - \delta^2)^{[n/n_0]}.
\]
Combining this with (6.5) and (6.6), we complete the proof. \(\square\)

The coupling we used in this proof is not optimal, so the rate of convergence we established is not as strong as in Theorem 6.3 — in the final estimate \((1 - \delta)^{[n/n_0]}\) got replaced with \((1 - \delta^2)^{[n/n_0]}\). However one can use a better way to couple two processes called maximal coupling and recover the optimal rate. More on this in forthcoming chapters.
6. Classification of states of a Markov chain


**Definition 6.1. Filtration**

General Markov property, see [Shi96]. Let $P(\cdot, \cdot)$ be a transition probability on $(X, \mathcal{X})$. We denote by $P_\rho$ the Markov distribution on $(\Omega, \mathcal{F}) = (X^\mathbb{Z}_+, \mathcal{X}^\mathbb{Z}_+)$ generated by initial distribution $\rho$ and transition probability $P(\cdot, \cdot)$. We denote $P_x = P_{\delta_x}, x \in X$. We denote the shift on $(X^\mathbb{Z}_+, \mathcal{X}^\mathbb{Z}_+)$ by $\theta$. We also denote $F_n = \sigma(X_0, \ldots, X_n)$.

**Theorem 6.4.** For any bounded measurable function $H : \Omega \to \mathbb{R}$ and for any initial distribution $\rho$,

$$E_\rho(H \circ \theta^n|F_n)(\omega) = E_{X_n}(\omega)H, \quad P_\rho - a.s.$$

**Definition 6.2.** Stopping time $\tau$ w.r.t. $(F_n)_{n \geq 0}$ iff $\{\tau \leq n\} \in F_n$ for every $n \geq 0$ (or, equivalently, $\{\tau = n\} \in F_n$ for every $n \geq 0$).

**Definition 6.3.** Sigma-algebra $F_\tau = \{A \in F : A \cap \{\tau = n\} \in F_n\}$ is interpreted as the $\sigma$-algebra of events observed on a random interval $[0, \tau]$.

**Theorem 6.5.** If $\tau$ is a finite stopping time, then the following strong Markov property holds true:

$$E_\rho(H \circ \theta^\tau|F_\tau) = E_{X_\tau}(H), \quad P_\rho - a.s.$$

6.2. Classification based on algebraic properties of transition probabilities.

**Definition 6.4.** Inessential states, essential states.

**Definition 6.5.** Irreducible classes of communicating states

**Definition 6.6.** Cyclic subclasses.

6.3. Classification based on recurrence. Let $\tau_i = \min\{k \geq 1 : X_k = i\}$

**Definition 6.7.** The state $i$ is transient if $P_i\{\tau_i < \infty\} < 1$.

**Definition 6.8.** The state $i$ is recurrent if $P_i\{\tau_i < \infty\} = 1$.

**Definition 6.9.** The state $i$ is positive recurrent if $E_i\tau_i < \infty$.

**Theorem 6.6.** In every communication class, all states are of the same type.

**Theorem 6.7.** If $h$ is recurrent, then there is an invariant measure (maybe unbounded) on the communication class of $h$. If $h$ is positively recurrent, there is an invariant probability distribution.
6. CLASSIFICATION OF STATES OF A MARKOV CHAIN

**Proof:** Let \( \pi_h = 1 \) and let \( \pi_i, i \neq h \) be the average time spent in \( i \) by the process started at \( h \) before coming back to \( h \).

\[
\pi_i = \mathbb{E}_h \sum_{k=1}^{\infty} 1_{\{X_k = i, k < \tau_h\}} = \sum_{k=1}^{\infty} \mathbb{P}_h\{X_k = i, k < \tau\}
\]

So, for \( m \neq h \),

\[
(\pi \mathbf{P})_m = 1 \cdot \mathbb{P}_{hm} + \sum_{i \neq h} \sum_{k=1}^{\infty} \mathbb{P}_h\{X_k = i, k < \tau_h\} \mathbb{P}_{hm}
\]

\[
= \mathbb{P}_h\{X_1 = m, 1 < \tau_h\} + \sum_{k=2}^{\infty} \mathbb{P}_h\{X_k = m, k < \tau_h\}
\]

\[
= \sum_{k=1}^{\infty} \mathbb{P}_h\{X_k = m, k < \tau_h\} = \pi_m,
\]

and

\[
(\pi \mathbf{P})_h = \mathbb{P}_{hh} + \sum_{k=1}^{\infty} \sum_{j \neq h} \mathbb{P}_h\{X_k = i, k < \tau_h\} \mathbb{P}_{dh}
\]

\[
= \mathbb{P}_h\{\tau_h = 1\} + \sum_{k=2}^{\infty} \mathbb{P}_h\{\tau_h = k\} = \sum_{k=1}^{\infty} \mathbb{P}_h\{\tau_h = k\} = 1 = \pi_h.
\]

When is this measure \( \pi \) finite?

\[
\sum_{i \neq h} \pi_i = \mathbb{E}_h \sum_{k=1}^{\infty} \sum_{i \neq h} 1_{\{X_k = i, k < \tau_h\}} = \mathbb{E}_h (\tau_h - 1) = \mathbb{E}_h \tau_h - 1,
\]

so the measure \( \pi \) is finite iff \( h \) is positive recurrent. \( \square \)

Relation (6.9) also implies that if \( \mu \) is the probability invariant measure obtained from \( \pi \) by normalization, then

\[
\mu_i = \frac{1}{1 + \sum_{i \neq h} \pi_i} = \frac{1}{\mathbb{E}_h \tau_h}.
\]

In fact, this invariant distribution is unique. We will study a more general approach to uniqueness later, but there is a simple explanation based on the renewal theorem.

Let \( (T_k)_{k=1}^{\infty} \) be an i.i.d. sequence of positive random variables with \( \mathbb{E}T_1 = a \in (0, \infty) \). We set \( S_0 = 0 \),

\[
S_n = T_1 + \ldots + T_n, \quad n \in \mathbb{N},
\]

and define

\[
N(t) = \max\{n \geq 0 : S_n \leq t\}, \quad t \geq 0.
\]

**Theorem 6.8.** With probability 1, \( \mathbb{E}N(t)/t \to 1/a \).
IDEA OF PROOF: We have
\[ S_{N(t)} \leq t < S_{N(t)+1}, \quad t \geq 0. \]
Therefore,
\[ \frac{S_{N(t)}}{N(t)} \leq \frac{t}{N(t)} < \frac{S_{N(t)+1}}{N(t)}, \quad t \geq 0. \]
By the strong law of large numbers, both l.h.s. and r.h.s. converge to \( a \) a.s., and so does \( t/N(t) \). So we have \( N(t)/t \to 1/a \), and to finish the proof one has to convert this into convergence of integrals, we omit that part. \( \square \)

Coming back to Markov processes, suppose \( i \) is a recurrent state and \( \rho \) is an invariant distribution for \( P \). Let \( N(m) \) be the number of returns to \( i \) by time \( m \). Under \( P_\rho \), the process \((X_k)\) is stationary, so
\[
(6.10) \quad \mathbb{E}N(m) = \mathbb{E}_\rho \sum_{k=1}^m 1_{\{X_k = i\}} = \sum_{k=1}^m P_\rho \{X_k = i\} = mp_i.
\]
On the other hand, the renewal theorem gives \( \mathbb{E}N(m)/m \to 1/\mathbb{E}_i \tau_i \), so we conclude that \( p_i = 1/\mathbb{E}_i \tau_i \). (In fact, we cannot apply the renewal theorem straightforwardly, because the first hitting time has distribution different from all the next return times, but this can easily be taken into account).

In the opposite direction, suppose \( \rho \) is an invariant distribution and \( i \) is a state with \( \rho(i) > 0 \). Then, due to (6.10) there will be infinitely many returns to \( i \) (so \( i \) is recurrent) and \( \mathbb{E}N(m)/m = \rho_i \), so according to renewal theorem \( \mathbb{E}_i \tau_i = 1/\rho_i \).

6.4. Lyapunov–Foster criterion for positive recurrence.

THEOREM 6.9. Let \( P(\cdot, \cdot) \) be a transition kernel on a countable space \( \mathbb{X} \). Let \( Lf = Pf - f \) for any function \( f \) such that \( Pf \) is well-defined. The process is positive recurrent iff there is a function \( V : \mathbb{X} \to \mathbb{R}_+ \) such that \( LV(x) \leq -1 \) for all but finitely many \( x \in \mathbb{X} \).

REMARK 6.1. \( V \) is called a Lyapunov function

PROOF: The set \( A = \{x \in \mathbb{X} : LV(x) > -1\} \) is finite. For \( x \notin A \) we can write
\[
V(x) \geq 1 + PV(x) \geq 1 + \int_{\mathbb{X}} P(x, dx_1)V(x_1) \geq 1 + \int_{A^c} P(x, dx_1)V(x_1)
\]
Applying the same estimate to \( V(x_1) \) we obtain iteratively

\[
V(x) \geq 1 + \int_{A^c} P(x, dx_1) \left( 1 + \int_{A^c} P(x_1, dx_2) V(x_2) \right)
\]

\[
\geq 1 + \int_{A^c} P(x, dx_1) + \int_{A^c} \int_{A^c} P(x, dx_1) P(x_1, dx_2) V(x_2)
\]

\[
\geq \ldots
\]

\[
\geq 1 + \int_{A^c} P(x, dx_1) + \int_{A^c} \int_{A^c} P(x, dx_1) P(x_1, dx_2)
\]

\[
+ \ldots + \int_{A^c} \ldots \int_{A^c} P(x, dx_1) P(x_1, dx_2) \ldots P(x_{n-1}, x_n)
\]

\[
+ \int_{A^c} \ldots \int_{A^c} P(x, dx_1) P(x_1, dx_2) \ldots P(x_{n}, x_{n+1}) V(x_{n+1}).
\]

Omitting the last nonnegative term, we obtain

\[
V(x) \geq 1 + P_x \{ \tau_A > 1 \} + P_x \{ \tau_A > 2 \} + \ldots + P_x \{ \tau_A > n \}, \quad n \in \mathbb{N},
\]

where \( \tau_A = \inf \{ n \in \mathbb{N} : X_n \in A \} \). Therefore

\[
V(x) \geq \sum_{n=0}^{\infty} P_x \{ \tau_A > n \} = \sum_{n=0}^{\infty} \sum_{k=n+1}^{\infty} P_x \{ \tau_A = k \} = \sum_{k=1}^{\infty} \sum_{n=0}^{k-1} P_x \{ \tau_A = k \}
\]

\[
= \sum_{k=1}^{\infty} k P_x \{ \tau_A = k \} = E_x \tau_A.
\]

So the set \( A \) is positive recurrent, i.e. \( E_x \tau_A < \infty \). Therefore, there is an invariant distribution assigning positive weight to all states in \( A \) (see the paragraph below this proof). All states assigned positive weight according to the invariant distribution are recurrent.

In the opposite direction, if the process is recurrent, then for any state \( x \in X \) we can take an arbitrary non-empty finite set \( A \subset X \) and define \( V(x) = E_x \tau_A \). Then for any \( x \notin A \)

\[
V(x) = E_x \tau_A = \int_X P(x, dy)(1 + E_y \tau_A) = 1 + \int_X P(x, dy) E_y \tau_A
\]

\[
= 1 + \int_X P(x, dy)V(y) = 1 + PV(x),
\]

so \( LV(x) = -1 \). \( \square \)

Suppose the set \( A \) is positive recurrent, i.e. \( E_x \tau_A < \infty \) for all \( x \in X \). Suppose all states in \( A \) are in the same communication class. Then there is an invariant measure defined in the following way. Consider a new transition probability \( Q_{ij} \) on \( A \) defined by

\[
Q_{ij} = P_i \{ X_{\tau_A} = j \}, \quad i, j \in A.
\]
This transition matrix defines an irreducible Markov chain on $A$. Let us denote its unique invariant probability by $\nu$. For all $i \in X$ we define
\[
\pi_i = \mathbb{E}_\nu \sum_{k=0}^{\infty} 1\{X_k = i, k < \tau\} = \sum_{h \in A} \nu_h \sum_{k=0}^{\infty} P_h\{X_k = i, k < \tau\}.
\]
We claim that this is an invariant measure for $P$. For $j \notin A$,
\[
(\pi P)_j = \sum_{i \in X} \sum_{h \in A} \nu_h \sum_{k=0}^{\infty} P_h\{X_k = i, k < \tau\} P_{ij}
= \sum_{h \in A} \nu_h \sum_{k=0}^{\infty} P_h\{X_k+1 = j, k+1 < \tau\} = \pi_j.
\]
For $j \in A$,
\[
(\pi P)_j = \sum_{i \in X} \sum_{h \in A} \nu_h \sum_{k=0}^{\infty} P_h\{X_k = i, k < \tau\} P_{ij}
= \sum_{h \in A} \nu_h \sum_{k=0}^{\infty} P_h\{X_{k+1} = j, \tau = k+1\}
= \sum_{h \in A} \nu_h P_h\{X_{\tau} = j\} = \sum_{h \in A} \nu_h Q_{hj} = \nu_j.
\]
This finishes the proof of the claim since $\nu_j = \pi_j$ for $j \in A$. 
CHAPTER 7

Invariant distributions for Markov processes in general spaces

1. Existence via Krylov–Bogolyubov approach

**Definition 7.1.** A transition probability $P(\cdot, \cdot)$ on a metric space $X$ with Borel $\sigma$-algebra $\mathcal{X}$ is Feller if for every bounded continuous $f : X \to \mathbb{R}$, $Pf$ is also continuous.

Feller property means that $P(x, \cdot)$ is continuous in $x$ in the topology of weak convergence.

Krylov–Bogolyubov theorem for Markov processes:

**Theorem 7.1.** Let $\rho$ be a probability measure on $(X, \mathcal{X})$ such that the sequence $(\mu_n)_{n \in \mathbb{N}}$ defined by

\[
\mu_n = \frac{1}{n} \sum_{j=0}^{n-1} \rho P^j, \quad n \in \mathbb{N}
\]

is tight. Then there is a $P$-invariant measure.

**Proof:** By Prokhorov’s theorem, tightness implies relative compactness, so there is a sequence $n_k \uparrow \infty$ and a probability measure $\nu$ such that $\mu_{n_k}$ converges to $\nu$ weakly. Let us take any bounded continuous function $f$ and compute

\[
|\nu Pf - \nu f| = \lim_{k \to \infty} \left| \frac{1}{n_k} \sum_{j=0}^{n_k-1} \rho P^j + 1 - \frac{1}{n_k} \sum_{j=0}^{n_k-1} \rho P^j f \right|
\]

\[
= \lim_{k \to \infty} \left\| \frac{1}{n_k} (P^{n_k} f - f) \right\|_{\infty} \leq \lim_{k \to \infty} \frac{2\|f\|_{\infty}}{n_k} = 0.
\]

□

**Theorem 7.2.** Suppose $(X, \mathcal{X})$ is a metric space. Let $V : X \to \mathbb{R}$ be a measurable function such that $U_c = \{x \in X : V(x) \leq c\}$ is a pre-compact set for any $c > 0$. If a transition probability $P(\cdot, \cdot)$ on $(X, \mathcal{X})$ satisfies

\[
PV(x) \leq \alpha V(x) + \beta, \quad x \in X,
\]

for some $\alpha \in (0, 1)$ and $\beta \geq 0$, then there is an invariant distribution for $P$. 

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Proof: By iterating (7.2), we obtain

\[ P^2V(x) \leq \alpha(\alpha V(x) + \beta) + \beta \leq \alpha^2 V + \alpha \beta + \beta, \]

\[ P^n V(x) \leq \alpha^n V(x) + \beta(\alpha^{n-1} + \ldots + \alpha + 1) \leq \alpha^n V(x) + \frac{\beta}{1 - \alpha}. \]

Therefore, defining for any \( x \in X \), \( \mu_n \) by (7.1) with \( \rho = \delta_x \)

\[ \mu_n V = \frac{1}{n} \sum_{j=0}^{n-1} P^j V(x) \leq V(x) + \frac{\beta}{1 - \alpha}. \]

We claim that \( \mu_n \) is a tight family since due to Markov’s inequality

\[ \mu_n \{ x : V > c \} \leq \frac{\mu_n V}{c} \leq \frac{1}{c} \left( V(x) + \frac{\beta}{1 - \alpha} \right). \]

This estimate on the measure \( \mu_n \) of the pre-compact set \( \{ x : V > c \} \) does not depend on \( n \), and can be made arbitrarily small by choosing sufficiently large values of \( c \).

□

Example: ARMA with \( V(x) = x^2 \).

2. Doeblin condition

We say that a transition probability \( P(\cdot, \cdot) \) on \( (X, \mathcal{X}) \) satisfies the Doeblin condition if there is a probability measure \( \nu \) on \( (X, \mathcal{X}) \) and a number \( \rho \in (0, 1) \) such that \( P(x, A) \geq \rho \nu(A) \) for all \( x \in X \) and all \( A \in \mathcal{X} \).

The measure \( \nu \) is called a minorizing measure.

Theorem 7.3. If Doeblin’s condition holds true, then there is an invariant measure.

Proof: We introduce a kernel \( Q(x, \cdot) = \nu(\cdot) \) that is identically equal to \( \nu \). We define

(7.3) \[ \rho = p \sum_{n=0}^{\infty} \nu(P - pQ)^n. \]

This series converges and defines a probability measure since

\[ \rho(X) = p \sum_{n=0}^{\infty} (1 - p)^n = 1. \]

Let us prove that \( \rho \) is invariant:

\[ \rho P = p p Q + \rho(P - pQ) = p \nu + p \sum_{n=0}^{\infty} \nu(P - pQ)^{n+1} \]

\[ = p \nu + p \sum_{n=1}^{\infty} \nu(P - pQ)^n = p \sum_{n=0}^{\infty} \nu(P - pQ)^n = \rho. \]

□
Let us try to understand the “resolvent” formula (7.3). Since
\[ P(x, dy) = p\nu(dy) + (1 - p)R(x, dy), \]
where \( R(x, dy) = (P(x, dy) - \nu(dy))/(1 - p) \) is a transition probability, one can represent the Markov evolution in the following way: one considers an i.i.d. family of Bernoulli variables \((\kappa_i)\) taking values 1 with probability \( p \) and 0 with probability \( 1 - p \). To obtain \( X_n \) given \( X_{n-1} \) we apply the following procedure: Conditioned on \( \kappa_n = 1 \), the distribution of \( X_n \) is \( \nu \), Conditioned on \( \kappa_n = 0 \), the distribution of \( X_n \) is \( R(X_{n-1}, \cdot) \).

So let us now find the distribution at time 0. First, let us condition on the event
\[ B_n = \{\kappa_{-n} = 1, \kappa_{-n+1} = 0, \ldots, \kappa_0 = 0\}. \]
Conditioned on this event, \( X_{-n} \sim \nu, X_{-n+1} \sim \nu R, \ldots, X_0 \sim \nu R^n \). Probability of \( B_n \) equals \( p(1 - p)^n \). Using the complete probability formula gives (7.3).

**Theorem 7.4.** If Doeblin’s condition is satisfied, then there is at most one invariant distribution for \( P \).

**Proof:** Let \( \rho_1 \) and \( \rho_2 \) be two different invariant distributions. Let us prove that they are equal to each other using the coupling method. Let us organize a Markov process on \((\mathcal{X} \times \mathcal{X}, \mathcal{X} \times \mathcal{X})\) with initial distribution \( \rho_1 \times \rho_2 \) and special transition probabilities \( P((x_1, x_2), \cdot) \) that we proceed to describe. So, we are going to describe the distribution of \((X^1_{n+1}, X^2_{n+1})\) conditioned on \((X^1_n, X^2_n) = (x_1, x_2)\).

If \( x_1 \neq x_2 \) then we will need a Bernoulli r.v. \( \kappa \) taking value 1 with probability \( p \) and value 0 with probability \( 1 - p \). Let \( \xi, \eta_1, \eta_2 \) be a r.v.’s independent of \( \kappa \) with distributions \( \nu, (P(x_1, \cdot) - p\nu)/(1 - p), \) and \( (P(x_2, \cdot) - p\nu)/(1 - p), \) respectively. Now we define \( P((x_1, x_2), \cdot) \) to be the joint distribution of \( \mathcal{X} \)-valued random variables \( Y_1 \) and \( Y_2 \) defined by:
\[
\begin{align*}
Y_1 &= \kappa \xi + (1 - \kappa) \eta_1, \\
Y_2 &= \kappa \xi + (1 - \kappa) \eta_2.
\end{align*}
\]

If \( x_1 = x_2 \) we let \( Y \) be distributed according to \( P(x_1, \cdot) \) and set \( P((x_1, x_2), \cdot) \) to be the joint distribution of \((Y, Y)\).

Notice that in both cases the transition probabilities \( P((x_1, x_2), \cdot) \) project onto \( P(x_1, \cdot) \) and \( P(x_2, \cdot) \). This is obvious for \( x_1 = x_2 \), whereas if \( x_1 \neq x_2 \) then for any \( A \in \mathcal{X} \),
\[
P\{Y_1 \in A\} = P\{Y_1 \in A; \kappa = 1\} + P\{Y_1 \in A; \kappa = 0\}
= pP\{\xi \in A\} + (1 - p)P\{\eta_1 \in A\}
= p\nu(A) + (1 - p)(P(x_1, A) - p\nu(A))/(1 - p) = P(x_1, A),
\]
and a similar computation shows \( P\{Y_2 \in A\} = P(x_2, A) \).

Therefore, the Markov process \((X^1_n, X^2_n)\) defined by the initial distribution \( \rho_1 \times \rho_2 \) and transition probabilities \( P((x_1, x_2), \cdot) \) satisfies the following
property: the distributions of $X_1$ and $X_2$ coincide with distributions of Markov process with initial distribution $\rho_1$ and $\rho_2$, respectively, and transition probabilities $P(\cdot, \cdot)$.

Moreover, if $X_1^n = X_2^n$ for some $n$, then $X_m^1 = X_m^2$ holds for all $m \geq n$. Therefore,

$$ P_{\rho_1 \times \rho_2} \{ X_n^1 \neq X_2 \} \leq P_{\rho_1 \times \rho_2} \{ \tau > n \}, \quad n \in \mathbb{N}, $$

where $\tau = \min \{ k \in \mathbb{N} : X_k^1 = X_k^2 \}$.

In particular, for any $A \in \mathcal{X}$,

$$ |\rho_1(A) - \rho_2(A)| = \left| P_{\rho_1} \{ X_n \in A \} - P_{\rho_2} \{ X_n \in A \} \right| $$

$$ = \left| P_{\rho_1 \times \rho_2} \{ X_1^n \in A \} - P_{\rho_1 \times \rho_2} \{ X_2^n \in A \} \right| $$

$$ \leq P_{\rho_1 \times \rho_2}(\{ X_1^n \in A \} \Delta \{ X_2^n \in A \}) $$

$$ \leq P_{\rho_1 \times \rho_2} \{ X_n^1 \neq X_2^n \} \leq P_{\rho_1 \times \rho_2} \{ \tau > n \}. $$

However,

$$ P_{\rho_1 \times \rho_2} \{ \tau > n \} \leq (1 - p)^n, $$

so taking $n \to \infty$ we obtain $\rho_1(A) = \rho_2(A)$ which completes the proof. □

The stopping time $\tau$ in the above proof is called the coupling time.

The proof can easily be modified to provide the proof of the following strengthening of Theorem 7.4:

**Theorem 7.5.** Let Doeblin’s condition be satisfied. If $\rho$ is the invariant distribution (the uniqueness of which is guaranteed by Theorem 7.4), then for any other initial distribution $\nu$,

$$ \| \nu P^n - \rho \|_{TV} \leq (1 - p)^n. $$

□

3. Harris positive recurrence condition

We will say that a transition probability $P(\cdot, \cdot)$ on $(\mathcal{X}, \mathcal{X})$ defines a Harris (positive) recurrent process if there is a set $K \in \mathcal{X}$, a probability measure $\nu$ on $(\mathcal{X}, \mathcal{X})$ and a number $p \in (0, 1)$ such that the following conditions hold true:

(A) $\sup_{x \in A} E_x \tau_A < \infty$

(B) $P(x, B) > p \nu(B)$ for all $x \in A$.

**Theorem 7.6.** If $P(\cdot, \cdot)$ satisfies the Harris recurrence condition, then there is a unique invariant distribution.

**Proof:** We represent the distribution of $X_n$ given $X_{n-1} \in A$ as follows: let $(\kappa_n)$ be an i.i.d. sequence Bernoulli r.v. taking value 1 with probability $p$ and value 0 with probability $1 - p$. Conditioned on $\{\kappa_n = 1\}$, $X_n \sim \nu$. Conditioned on $\{\kappa_n = 0\}$, $X_n \sim R(x, \cdot)$, where $R(x, dy) = (P(x, dy) - \nu(dy))/(1 - p)$.

We will observe an extended process $(X_n, \kappa_n)$. 
We create the following sequence of stopping times:
\[
\tau_1 = \min \{ n \geq 1 : X_{n-1} \in A, \kappa_{n-1} = 1 \},
\]
\[
\tau_k = \min \{ n \geq \tau_{k-1} + 1 : X_{n-1} \in A, \kappa_{n-1} = 1 \}.
\]
Then \((X_{\tau_k})\) forms a Markov process with invariant distribution \(\nu\).

Denoting the length of one excursion by \(\sigma = \tau_k - \tau_{k-1}\)
One can also use condition (A) to estimate
\[
\tilde{\mathbb{E}} \sigma \leq \sum_{m=1}^{\infty} p(1 - p)^{m-1} m \mathbb{E} \tau_A < \infty.
\]

We denote by \(\tilde{P}\) the probability on the extended probability space that includes the information about \(\kappa\)’s, and by \(\tilde{\mathbb{E}}\) the corresponding expectation.

Therefore, the average occupation measure \(\rho\) defined by
\[
\rho(B) = \tilde{\mathbb{E}}_{\nu} \sum_{m=0}^{\infty} 1_{\{X_m \in B, m < \sigma\}} = \sum_{m=0}^{\infty} \tilde{P}_{\nu}\{X_m \in B, m < \sigma\}, \quad B \in \mathcal{X},
\]
is finite.

Let us prove that \(\rho\) is invariant. Let us first notice that
\[
\rho P(B) = \int_{\mathcal{X}} \rho(dx) P(x, B)
\]
\[
= \int_{A^c} \rho(dx) P(x, B) + (1 - p) \int_A \rho(dx) R(x, B) + p \int_A \rho(dx) \nu(B)
\]
\[
= \sum_{m=0}^{\infty} \tilde{P}_{\nu}\{X_m \in A^c, X_{m+1} \in B, m + 1 < \sigma\}
\]
\[
+ \sum_{m=0}^{\infty} \tilde{P}_{\nu}\{X_m \in A, X_{m+1} \in B, m + 1 < \sigma\}
\]
\[
+ \sum_{m=0}^{\infty} \tilde{P}_{\nu}\{X_m \in A, X_{m+1} \in B, \sigma = m + 1\}.
\]
Now we (i) combine the first two sums into one and (ii) use the invariance of \(\nu\) under the embedded Markov chain to derive
\[
\sum_{m=0}^{\infty} \tilde{P}_{\nu}\{X_m \in A, X_{m+1} \in B, \sigma = m + 1\} = \nu(B).
\]
The result is
\[ \rho P(B) = \sum_{m=0}^{\infty} \tilde{P}_\nu \{ X_{m+1} \in B, m + 1 < \sigma \} + \nu(B) \]
\[ = \sum_{m=1}^{\infty} \tilde{P}_\nu \{ X_m \in B, m < \sigma \} + \tilde{P}_\nu \{ X_0 \in B, 0 < \sigma \} \]
\[ = \sum_{m=0}^{\infty} \tilde{P}_\nu \{ X_m \in B, m < \sigma \} = \rho(B), \]
where we used \( \sigma \geq 1 \) in the second identity.

One can base a uniqueness proof on this representation or use a coupling argument.

In terms of Lyapunov function, suppose we are in conditions of Theorem 7.2. If (7.2) holds, then
\[ PV(x) - V(x) < (-1 - \beta)/(1 - \alpha). \]
Therefore, \( PV(x) - V(x) < -1 \) for all \( x \) such that \( V(x) > (-1 - \beta)/(1 - \alpha) \). Let us denote by \( A \) the complement to that set. The reasoning in the proof of Theorem 6.9 implies that the set \( A \) is positively recurrent, and the conditions of Theorem 7.2 imply that \( A \) is compact. In many situations (e.g. in SDE’s under some nondegeneracy conditions) we know that condition (B) is satisfied. In particular, this reasoning holds true for our ARMA example.
Bibliography


