Kolmogorov's Theorem

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Our goal is to understand the following theorem.

Kolmogorov's Theorem

Let $\rho, \gamma > 0$ be given, and let $h(\mathbf{q}, \mathbf{p}) = h_0(\mathbf{p}) + h_1(\mathbf{q}, \mathbf{p})$ be a Hamiltonian, with $h_0, h_1 \in \mathcal{A}_\rho$ and $\|h\|_\rho \leq 1$. Suppose the Taylor polynomial of h_0 is

$$h_0(\mathbf{p}) = a + \omega \mathbf{p} + \frac{1}{2} \mathbf{p} \cdot C \mathbf{p} + o(|\mathbf{p}|^2),$$

with $\omega \in \Omega_{\gamma}$ and C is symmetric and invertible. Then for any $\rho_* \leq \rho$, there exists $\epsilon > 0$, which depends on C and γ , but not on the remainder term in $o(|\mathbf{p}|^2)$, such that if $\|h_1\|_{\rho} \leq \epsilon$, there exists a symplectic mapping $\Phi: A_{\rho_*} \to A_{\rho}$ such that if we set $(\mathbf{q}, \mathbf{p}) = \Phi(\mathbf{Q}, \mathbf{P})$ and $H = h \circ \Phi$, we have

$$H(\mathbf{Q}, \mathbf{P}) = A + \omega \mathbf{P} + R(\mathbf{Q}, \mathbf{P}),$$

with $R(\mathbf{Q}, \mathbf{P}) \in O(|\mathbf{P}|^2)$.

Kolmogorov's Theorem

The **KAM** theory is named after

Andreï Kolmogorov,

Vladimir **A**rnold,

Jürgen **M**oser.

Kolmogorov gave an original proof in the 1950s but never published it. Arnold gave a proof of the theorem in 1963 and Moser published a different but related result in 1962.

A Motivating Example

The solar system when planets have zero mass.

Named for William Rowan Hamilton (1805-1865). A reformulation of classical Newtonian mechanics. Relies on a symplectic structure.

Let (X,σ) be a symplectic manifold. That is, X is a differentiable manifold and σ is a nowhere vanishing 2-form such that $d\sigma=0$. A function H on a X has a *symplectic gradient* denoted $\nabla_{\sigma}H$. This gradient is the unique vector field such that for any vector field ξ , we have

$$\sigma(\xi, \nabla_{\sigma} H) = dH(\xi).$$

We will consider a *Hamiltonian differential equation* that makes use of this symplectic gradient:

$$\dot{\mathbf{x}} = (\nabla_{\sigma} H)(\mathbf{x}).$$



For a simple example of a Hamiltonian system, consider $X = \mathbb{R}^{2n}$ with coordinates $(q_1, \ldots, q_n, p_1, \ldots, p_n)$. Take $\sigma = \sum_i dp_i \wedge dq_i$.

Our Hamiltonian differential equation is

$$\dot{q}_{i} = \frac{\partial H}{\partial p_{i}},$$

$$\dot{p}_{i} = -\frac{\partial H}{\partial a_{i}},$$

where these are called the Hamiltonian equations of motion.



How does the Solar System fit this model? Consider the case of a single body of zero mass. (It is sufficient to consider the single body system, since the planets having zero mass means that they do not effect each other). For $\mathbf{x} \in \mathbb{R}^2$ the equation $\ddot{\mathbf{x}} = -\frac{\mathbf{x}}{|\mathbf{x}|^3}$ is

Hamilton's equation for the manifold $X = \mathbb{R}^2 \times \mathbb{R}^2$ with points (\mathbf{q}, \mathbf{p}) , standard symplectic form

$$\sigma = dp_1 \wedge dq_1 + dp_2 \wedge dq_2,$$

and Hamiltonian

$$H(\mathbf{q},\mathbf{p}) = rac{1}{2}(p_1^2 + p_2^2) - rac{1}{\sqrt{q_1^2 + q_2^2}}.$$



After some computation, we can obtain

$$abla_{\sigma}H=\left(\mathbf{p},-rac{1}{\left(q_1^2+q_2^2
ight)^{rac{3}{2}}}\mathbf{q}
ight).$$

The Hamiltonian Differential Equation $\dot{\mathbf{x}} = \nabla_{\sigma} H(\mathbf{x})$ is

$$\begin{aligned} q_1' &= p_1 & p_1' &= -\frac{q_1}{(q_1^2 + q_2^2)^{\frac{3}{2}}}, \\ q_2' &= p_2 & p_2' &= -\frac{q_2}{(q_1^2 + q_2^2)^{\frac{3}{2}}}. \end{aligned}$$
 From here we can recover $\ddot{\mathbf{x}} = -\frac{\mathbf{x}}{|\mathbf{x}|^3}.$

The vector field $\nabla_{\sigma} f$ has a flow ϕ_f^t such that

$$\phi_f^t$$
 preserves f : $f \circ \phi_f^t = f$,

$$\phi_f^t$$
 preserves σ : $(\phi_f^t)^* \sigma = \sigma$.

Our proof of Kolmogorov's Theorem will involve the construction of a symplectic diffeomorphism Φ that is a composition of Hamiltonian flows.

We will need to construct Taylor polynomials of functions of the form $t \mapsto g \circ \phi_f^t$.

To do this we use the Poisson bracket.

Poisson Bracket

Let f and g be two functions on X. The Poisson bracket, denoted $\{f,g\}$, is defined by

$$\{f,g\} = \sigma(\nabla_{\sigma}f,\nabla_{\sigma}g) = df(\nabla_{\sigma}g) = -dg(\nabla_{\sigma}f).$$

We want to make formal the notion of "suffiently irrational." (Later we will see that this will help us to avoid dividing by zero). A number θ is irrational if for all pairs $p,q\in\mathbb{Z}$, we have $\left|\theta-\frac{p}{a}\right|\neq 0$.

So for θ to be "very" irrational, we want $\left|\theta - \frac{p}{q}\right|$ to be "very" different from 0, or "very big."

But that is not quite possible, as we can always approximate an irrational number by rational numbers.



Instead we want to consider what happens with small divisors.

We want
$$\left| \theta - \frac{p}{q} \right|$$
 to be "small" only if the denominator is "big."

We start with diophantine conditions.

Diophantine of Exponent d

A number θ is Diophantine of exponent d if there exists a constant $\gamma>0$ such that for all coprime integers p and q we have

$$\left|\theta - \frac{p}{q}\right| > \frac{\gamma}{|q|^d}.$$

Note that it is a stronger condition to be Diophantine of a smaller exponent.

Diophantine Numbers with Full Measure

For all $\epsilon > 0$, the set of Diophantine numbers of exponent $2 + \epsilon$ is of full measure.

Diophantine Numbers with Full Measure.

Consider numbers in \mathbb{R}/\mathbb{Z} .

For all positive integers q, there are at most q elements of \mathbb{Q}/\mathbb{Z} with denominator q in reduced form. Thus for all $\gamma > 0$,

$$\left|\left\{\theta \in \mathbb{R}/\mathbb{Z} : \left|\theta - \frac{p}{q}\right| < \frac{\gamma}{|q|^{2+\epsilon}}\right\}\right| < 2\gamma \sum_{q=1}^{\infty} \frac{1}{q^{1+\epsilon}}.$$

The intersection of these sets as $\gamma \to 0$ has measure 0, and this intersection is the complement of the set of Diophantine numbers of exponent $2+\epsilon$.

We generalize "sufficiently irrational" from numbers to vectors.

Diophantine Vector in \mathbb{R}^n

Let $\omega = (\omega_1, \dots, \omega_n)$. We say ω is Diophantine if there exists $\gamma > 0$ such that for all vectors with integer coefficients (k_1, \dots, k_n) , we have

$$|k_1\omega_1+\cdots+k_n\omega_n|\geq \frac{\gamma}{(k_1^2+\cdots+k_n^2)^{\frac{n}{2}}}.$$

Let Ω^n_{γ} be the subset of such $\omega \in \mathbb{R}^n$.

We could also rewrite this condition as a dot product

$$|\mathbf{k} \cdot \omega| \le \frac{\gamma}{|\mathbf{k}|^n}.$$

In our example of the solar system, the ω_i represent the frequencies of the planets' orbits.

How common are the vectors ω ? Actually very common, as if we pick components ω_i at random, we will almost surely select such a vector.

Diophantine Vectors are of Full Measure

The union
$$\Omega = \bigcup_{\gamma>0} \Omega_{\gamma}$$
 is of full measure.

Diophantine Vectors are of Full Measure.

Consider $S_{{\bf k},\gamma}$, the region around the hyperplane orthogonal to ${\bf k}$ of thickness $\frac{2\gamma}{|{\bf k}|^{n+1}}.$

The part within the unit cube Q has measure at most $\frac{M\gamma}{|\mathbf{k}|^{n+1}}$, with

M the constant giving the max (n-1)-dim measure of a hyperplane intersected with Q.

The sum $\sum_{\mathbf{k}\in\mathbb{Z}^n-\{0\}}\frac{1}{|\mathbf{k}|^{n+1}}$ is finite, so the volume of $\bigcup_{\mathbf{k}\in\mathbb{Z}^n-\{0\}}S_{\mathbf{k},\gamma}\cap Q$ is

at most a constant times γ .

The intersection of these sets has measure 0, and Ω is the complement, similar to the previous proof.

Torus

Torus

KAM for the Solar System

Let $\mathbf{x}_1(t)$ be a motion of the zero-masses system for which the frequency vector ω is Diophantine. (Then $\mathbf{x}_1(t)$ is dense on the corresponding torus T_1).

Then there exists $\epsilon > 0$ such that, if the planets have masses $m_i < \epsilon$, then there exists a trajectory $\mathbf{x}(t)$ of the system that is dense on a torus T, and there exists a homeomorphism $\Phi: T \to T_1$ such that $\Phi(\mathbf{x}(t)) = \mathbf{x}_1(t)$.

The probability of being on such a trajectory goes to 1 as ϵ goes to 0.



Another Look at the Theorem

Kolmogorov's Theorem

Let $\rho, \gamma > 0$ be given, and let $h(\mathbf{q}, \mathbf{p}) = h_0(\mathbf{p}) + h_1(\mathbf{q}, \mathbf{p})$ be a Hamiltonian, with $h_0, h_1 \in \mathcal{A}_\rho$ and $\|h\|_\rho \leq 1$. Suppose the Taylor polynomial of h_0 is

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with $\omega \in \Omega_{\gamma}$ and C is symmetric and invertible. Then for any $\rho_* \leq \rho$, there exists $\epsilon > 0$, which depends on C and γ , but not on the remainder term in $o(|\mathbf{p}|^2)$, such that if $\|h_1\|_{\rho} \leq \epsilon$, there exists a symplectic mapping $\Phi : A_{\rho_*} \to A_{\rho}$ such that if we set $(\mathbf{q}, \mathbf{p}) = \Phi(\mathbf{Q}, \mathbf{P})$ and $H = h \circ \Phi$, we have

$$H(\mathbf{O}, \mathbf{P}) = A + \omega \mathbf{P} + R(\mathbf{O}, \mathbf{F})$$

$$H(\mathbf{Q}, \mathbf{P}) = A + \omega \mathbf{P} + R(\mathbf{Q}, \mathbf{P}),$$

with $R(\mathbf{Q}, \mathbf{P}) \in O(|\mathbf{P}|^2)$.



We will be solving a system of equations iteratively, using an analog of Newton's method.

Recall that when using Newton's method, we need a way to bound the second derivatives.

We need a way to measure "size," that is, we need to choose a norm for our functions.

Let $X \subseteq \mathbb{C}^k$ be compact. ¹

Let the caligraphic letter \mathcal{X} be the Banach algebra of continuous functions on X that are analytic in the interior and have the sup norm

$$||f||_X = \sum_{\mathbf{x} \in X} |f(\mathbf{x})|.$$

We use the absolute value as the standard Euclidean norm on \mathbb{C}^n .

We consider three regions:

$$\begin{split} & \mathcal{B}_{\rho} = \{\mathbf{p} \in \mathbb{C} : |\mathbf{p}| \leq \rho\}, \\ & \mathcal{C}_{\rho} = \{\mathbf{q} \in \mathbb{C}^{n}/\mathbb{Z}^{n} : |\operatorname{Im}(\mathbf{q})| \leq \rho\}, \\ & \mathcal{A}_{\rho} = \mathcal{C}_{\rho} \times \mathcal{B}_{\rho} = \{(\mathbf{q}, \mathbf{p}) \in \mathbb{C}^{n}/\mathbb{Z}^{n} \times \mathbb{C}^{n} : |\mathbf{p}| \leq \rho, |\operatorname{Im}(\mathbf{q})| \leq \rho\}. \end{split}$$

Denote by \mathcal{B}_{ρ} , \mathcal{C}_{ρ} , and \mathcal{A}_{ρ} the corresponding Banach algebras. We can expand element of \mathcal{B}_{ρ} as power series, and elements of \mathcal{C}_{ρ} as Fourier series:

$$f(\mathbf{z}) = \sum_{\mathbf{k} \in \mathbb{Z}^n} f_{\mathbf{k}} e^{2\pi i \, \mathbf{k} \cdot \mathbf{z}}.$$

We can bound derivatives of analytic functions on balls in terms of the values of the function itself.

Cauchy's Inequalities on Balls

If $f \in \mathcal{B}_{\rho}$, then

$$\|Df\|_{\rho-\delta} \le \frac{1}{\delta} \|f\|_{\rho},$$

$$\|D^2f\|_{\rho-\delta} \le \frac{4}{\delta^2} \|f\|_{\rho}.$$

As a corollary, the case $\delta = \rho$ bounds derivatives at the center of balls.

$$|Df(0)| \le \frac{1}{\rho} \|f\|_{\rho},$$

 $|D^2f(0)| \le \frac{4}{\rho^2} \|f\|_{\rho}.$



Cauchy's Inequalities on Balls.

Take $\mathbf{z} \in B_{\rho-\delta}$ and $\mathbf{u} \in \mathbb{C}^n$. Since $B_{\delta}(\mathbf{z}) \subseteq B_{\rho}$, the function $g: t \mapsto f(\mathbf{z} + t\delta \mathbf{u})$ is defined on the unit disc, and the normal Cauchy inequality implies that

$$\delta |(Df(\mathbf{z}))\mathbf{u}| = |g'(0)| \le ||g||_1 \le ||f||_{\rho}.$$

Applying the argument twice gives the result for the second derivative:

$$|D^2 f(\mathbf{z})(\mathbf{u}, \mathbf{v})| \leq \frac{2}{\delta} \|Df(\mathbf{z})(\mathbf{u})\|_{\rho - \frac{\delta}{2}} |\mathbf{v}| \leq \frac{4}{\delta^2} \|f\|_{\rho} |\mathbf{u}| |\mathbf{v}|.$$

QED



Recall the final equation in Kolmogorov's theorem:

$$H(\mathbf{Q}, \mathbf{P}) = A + \omega \mathbf{P} + R(\mathbf{Q}, \mathbf{P}).$$

This is an equation for a diffeomorphism Φ , where $H = h \circ \Phi$. We solve for this diffeomorphism. To do so we would like to use Newton's method, but using it is not quite sufficient. However we can still do something with a similar flavor.

We obtain Φ as a limit of Φ_i , where

$$\Phi_i = \phi_i \circ \phi_{i-1} \circ \cdots \circ \phi_1.$$

Here, ϕ_i is the Hamiltonian flow for a Hamiltonian function g_i . This g_i is the unknown for which we solve.



At the *i*th step we have a Hamiltonian $\tilde{h} = \Phi_i^* h$.

We expand \tilde{h} up to order 2 in \mathbf{p} , and the coefficients will be Fourier series in \mathbf{q} .

We write $\tilde{h} = \tilde{h}_0 + \tilde{h}_1$:

 \tilde{h}_1 has the terms constant or linear in ${f p}$, except the constant in ${f q}$,

 \tilde{h}_0 is everything else.

We would like to eliminate \tilde{h}_1 , but we won't be able to do so by solving a linear equation.

Instead we will solve a linear equation for a function g such that $\phi_g^* \tilde{h}$ is "better" than \tilde{h} , in some reasonable sense of "better."



We expand $\phi_g^* \tilde{h}$ to first order in g:

$$\phi_{g}^{*}\tilde{h} = \tilde{h} + \{g, \tilde{h}\} + o(|g|)$$

= $\tilde{h}_{0} + \tilde{h}_{1} + \{g, \tilde{h}_{0}\} + \{g, \tilde{h}_{1}\} + o(|g|).$

We want to eliminate the terms, other than the term constant in \mathbf{q} , that are not $O(|\mathbf{p}|)^2$.

Applying the standard Newton's method would necessitate solving the equation

$$\tilde{h}_1 + \{g, \tilde{h}_0\} + \{g, \tilde{h}_1\} \in o(|\mathbf{p}|).$$

But we will do something a bit different.



We can assume that anything we want is small as long as the choice is justified by the resulting inequalities.

So suppose that $\{g, \tilde{h}_1\}$ is of order 2 since g and \tilde{h}_1 are both small.

Then the linear equation that we need to solve is

$$\tilde{h}_1 + \{g, \tilde{h_0}\} \in o(|\mathbf{p}|).$$

This is a *Diophantine differential equation*, and our ability to solve it will rely on the Diophantine vectors that we saw previously.

Diophantine Differential Equations

Let $g \in \mathcal{C}_{
ho}$. We are going to be solving linear equations of the form

$$Df(\omega) = \sum_{i=1}^{n} \omega_i \frac{\partial f}{\partial q_i} = g,$$

with $f \in \mathcal{C}_{\rho'}$ for some $\rho' < \rho$.

Write the functions f and g as Fourier Series:

$$f(\mathbf{q}) = \sum_{\mathbf{k} \in \mathbb{Z}^n} f_{\mathbf{k}} e^{2\pi i \, \mathbf{k} \cdot \mathbf{q}}, \qquad g(\mathbf{q}) = \sum_{\mathbf{k} \in \mathbb{Z}^n} g_{\mathbf{k}} e^{2\pi i \, \mathbf{k} \cdot \mathbf{q}}.$$

The solution is given by

$$f_{\mathbf{k}} = \frac{1}{2\pi i \left(\mathbf{k} \cdot \omega\right)} g_{\mathbf{k}}.$$

We need g_0 to be zero. Then f_0 is arbitrary, and otherwise the series for f is unique.



Diophantine Differential Equations

The convergence properties of f depend on the Diophantine properties of ω .

We cannot divide by zero when computing the Fourier coefficients f_k .

Even though we have convergence for ρ , we might not have boundedness. So we need some $\rho' < \rho$ where we have boundedness, but large enough so that the limit is nonempty.

Diophantine Differential Equations

We will use the following tool for choosing ρ' .

Proposition

If $g \in \mathcal{C}_{\rho}$ and $\omega \in \Omega_{\gamma}$, then for all $\delta \in (0, \rho)$ we have

$$\begin{split} \left\| f \right\|_{\rho-\delta} & \leq \frac{\kappa_n}{\gamma \delta^{2n}} \left\| g \right\|_{\rho}, \\ \left\| Df \right\|_{\rho-\delta} & \leq \frac{\kappa_n}{\gamma \delta^{2n+1}} \left\| g \right\|_{\rho}, \end{split}$$

where κ_n is a constant that depends only on n.

Next we will prove this proposition.



Let's return to $h_1 + \{g, h_0\} \in O(|\mathbf{p}|^2)$ (the Diophantine differential equation that we had before, but without the tildes and with $o(|\mathbf{p}|)$ replaced by $O(|\mathbf{p}|^2)$, which is equivalent since our functions are analytic).

The unknown is g, which we take to be degree 1 in \mathbf{p} :

$$g = \lambda \mathbf{q} + X(\mathbf{q}) + \sum_{i=1}^{n} Y_i(\mathbf{q}) p_i$$

Note that only the linear terms of g contribute to the linear terms of $\{g, h_0\}$.

The unknowns are: λ , X, Y_i .

We can expand X and Y_i as Fourier series, since they are functions of \mathbf{q} only.

Let $a \in \mathbb{R}$, $\omega \in \Omega_{\gamma}$, $R(\mathbf{q}, \mathbf{p}) \in O(|\mathbf{p}|^3)$, and $\overline{A} = 0$ be the average of A on the torus $\mathbf{p} = 0$, we can write

$$h_0(\mathbf{q}, \mathbf{p}) = a + \omega \mathbf{p} + \frac{1}{2} \mathbf{p} \cdot C(\mathbf{q}) \mathbf{p} + R(\mathbf{q}, \mathbf{p}),$$

$$h_1(\mathbf{q}, \mathbf{p}) = A(\mathbf{q}) + B(\mathbf{q}) \mathbf{p}.$$

We then have

$$(h_1 + \{g, h_0\})(\mathbf{q}, \mathbf{p})$$

$$= \omega \cdot \lambda + A(\mathbf{q}) + DX(\mathbf{q})(\omega)$$

$$+ (B(\mathbf{q}) + (\lambda + DX(\mathbf{q}))C(\mathbf{q}) + \omega DY(\mathbf{q})) \cdot \mathbf{p} + O(|\mathbf{p}|^2).$$

We need to solve the equations

$$DX(\mathbf{q})(\omega) = -A(\mathbf{q})$$

$$DY(\mathbf{q})(\omega) = -B(\mathbf{q}) - (\lambda + DX(\mathbf{q}))C(\mathbf{q})$$

for X and Y_i , $i = 1, \ldots, n$.

These are Diophantine differential equations.

Since we have the hypothesis $\overline{A}=0$, we can solve the first equation and find $X\in\mathcal{C}_{\rho'}$ for all $\rho'<\rho$.

We then substitute into the second equation.

We need to solve the equations

$$DX(\mathbf{q})(\omega) = -A(\mathbf{q})$$

$$DY(\mathbf{q})(\omega) = -B(\mathbf{q}) - (\lambda + DX(\mathbf{q}))C(\mathbf{q})$$

for X and Y_i , $i = 1, \ldots, n$.

We can then determine λ since the averages of the right-hand sides in the n are all zero.

Then we can solve the second equation and find $Y_i \in \mathcal{C}_{\rho''}$ for all $\rho'' < \rho'$.