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## CHAPTER 0

## Overview

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## 1. Motivations/Applications

A few sample results that turn out to be related to homogeneous dynamics are listed.
1.1. Horocycles on constant negative curvature surfaces. Equip $\mathbb{H}^{2}:=\{x+i y \in \mathbb{C}, y>0\}$ with the metric $\frac{\mathrm{dx}^{2}+\mathrm{dy}^{2}}{y^{2}}$. Let $\Gamma \leq \operatorname{Isom}\left(\mathbb{H}^{2}\right)$ be a discrete (torsion free) subgroup such that $\mathbb{H}^{2} / \Gamma$ is compact (such a subgroup is called a uniform lattice). Then $\mathbb{H}^{2} / \Gamma$ is a compact surface of constant negative curvature. Conversely, every surface with constant negative curvature arises this way. Let $\pi: \mathbb{H}^{2} \rightarrow \mathbb{H}^{2} / \Gamma=M$ be the quotient map.

Consider a horocycle $\mathscr{H}$ in $\mathbb{H}^{2}$. Explicitly, for each $v \in\{x+i y, y=0\}$, a horocycle based at $v$ is a circle (with respect to the Euclidean metric) in $\mathbb{H}^{2}$ tangent to $\{y=0\}$ at $v$. For $v=\infty$, a horocycle based at $v$ is a horizontal line above $\{y=0\}$.


Now we take the image of $\mathscr{H}$ under the projection $\pi$.
Theorem 1.1 ([Hed36]). For every $\mathscr{H}, \pi(\mathscr{H})$ is dense in $M$.
If $M=\mathbb{H}^{2} / \Gamma\left(\Gamma \leq \operatorname{Isom}\left(\mathbb{W}^{2}\right)\right.$ still discrete $)$ is just of finite volume, then
THEOREM 1.2. $\quad$ 1. $\pi(\mathscr{H})$ is either closed or dense in $M$.
2. Let $\pi\left(\mathscr{H}_{i}\right)$ be a sequence of closed horocycles, then as the length goes to infinity, $\pi\left(\mathscr{H}_{i}\right)$ becomes dense in $M$.

Remark 1.3. Assume $M=\mathbb{H}^{2} / \Gamma$ has finite volume. Then there exists closed $\pi(\mathscr{H})$ iff $M$ is non-compact.

By comparison, the image under $\pi$ of a geodesic is very different. The image could be closed, dense, or in between. And closed geodesics do not necessarily equidistribute towards the volume measure (though on average they do equidistribute).
1.2. Isometric immersion of hyperbolic spaces. Let $\mathbb{M}^{3}$ be the three dimensional hyperbolic space $\{(x+i y, z) \in \mathbb{C} \times \mathbb{R}, z>0\}$ equipped with the metric $\frac{1}{z^{2}}\left(\mathrm{dx}^{2}+\mathrm{dy}^{2}+\mathrm{dz}{ }^{2}\right)$. Let $\Gamma \leq \mathbb{H}^{3}$ be a discrete (torsion free) subgroup, such that $\mathbb{H}^{3}$ is compact (finite volume suffices). Consider an isometric embedding $\iota: \mathbb{H}^{2} \rightarrow \mathbb{H}^{3}$. The image of $\iota$ can be explicitly described. There are two cases:

1. given a circle on $\{z=0\}$, then there exists a unique half-sphere in $\mathbb{H}^{3}$ whose boundary is this given circle;
2. given a line on $\{z=0\}$, then there exists a unique half-plane in $\mathbb{H}^{3}$ whose boundary is this given line.
Then $\iota\left(\Vdash^{2}\right)$ is either a half-sphere or a half-plane described above. Similarly, we consider the image of $\iota\left(\mathbb{H}^{2}\right)$ under $\pi: \mathbb{H}^{3} \rightarrow \mathbb{H}^{3} / \Gamma=: M$,

Theorem 1.4. 1. $\pi\left(\iota\left(\mathbb{H}^{2}\right)\right)$ is either closed or dense in $M$;
2. Given an infinite sequence of distinct closed $\pi\left(\iota_{i}\left(\mathbb{W}^{2}\right)\right)$, then $\lim _{i} \pi\left(t_{i}\left(\mathbb{H}^{2}\right)\right)$ is dense in $M$.

REMARK 1.5. That the volume of $\pi\left(t_{i}\left(\mathbb{H}^{2}\right)\right)$ would go to infinity is automatic.
1.3. Oppenheim conjecture/Margulis theorem. Consider a non-degenerate real quadratic form in three (larger than 3 also ok) variables, viewed as a function $Q: \mathbb{R}^{3} \rightarrow \mathbb{R}$. Assume it is indefinite. Note that if $Q$ is a quadratic form with rational coefficients or proportional to such a form, then $Q\left(\mathbb{Z}^{3}\right)$ is discrete in $\mathbb{R}$.

THEOREM 1.6. IfQ is NOT proportional to a quadratic form with rational coefficients, then $Q\left(\mathbb{Z}^{3}\right)$ is dense in $\mathbb{R}$.

REMARK 1.7. It is also true replacing $\mathbb{Z}^{3}$ by primitive vectors. The proof is a bit harder.
Remark 1.8. This is false if $Q$ has two variables.
The above theorem admits a quantitative version in certain cases. Let $Q$ be a (nondegenerate) quadratic form in 4 variables of signature $(3,1)$ (what follows does not apply to signature $(2,2),(1,2))$. Assume $Q$ is irrational as above.

Theorem 1.9. There exists $\lambda_{Q}>0$ such that for every $a<b \in \mathbb{R}$,

$$
\#\left\{x \in \mathbb{Z}^{4} \mid Q(x) \in(a, b),\|x\| \leq T\right\} \sim \operatorname{Vol}\left\{x \in \mathbb{R}^{4} \mid Q(x) \in(a, b),\|x\| \leq T\right\} \sim \lambda_{Q}(b-a) T^{2}
$$

1.4. Littlewood conjecture. Let $\alpha \in \mathbb{R}$ (assume everything is irrational just in case of some trivialities). By pigeon-hole principle(?), one can show that

$$
\inf _{\left(m_{\neq 0}, n\right) \in \mathbb{Z}^{2}}|m| \cdot|m \alpha+n| \leq 1 .
$$

On the other hand there exists $\alpha$ ("badly approximable numbers") such that

$$
\inf _{\left(m_{\neq 0}, n\right) \in \mathbb{Z}^{2}}|m| \cdot|m \alpha+n|>0 .
$$

The Littlewood conjecture is
Conjecture 1.10. For every pair $(\alpha, \beta) \in \mathbb{R}^{2}$ irrational,

$$
\inf _{\left(m_{\neq 0}, n_{1}, n_{2}\right) \in \mathbb{Z}^{3}}|m| \cdot\left|m \alpha+n_{1}\right| \cdot\left|m \beta+n_{2}\right|=0 .
$$

To make it look closer to the Oppenheim conjecture, you may write $l_{\alpha}(x, y, z):=\alpha x+y$, $l_{\beta}(x, y):=\beta x+z$ and $\varphi(x, y, z):=x$. Let $L(x, y, z):=\varphi \cdot l_{\alpha} \cdot l_{\beta}$. Then the conjecture asserts that when $L$ is "irrational", then $\inf _{(x, y, z) \in \mathbb{Z}^{3}}|L(x, y, z)|$ is dense at 0 . By comparison, the Oppenheim conjecture is equivalent to $Q\left(\mathbb{Z}^{3}\right)$ being dense at 0 .

Our current knowledge is
Theorem 1.11. The set

$$
\left\{(\alpha, \beta) \in \mathbb{R}^{2}, \text { that fails this conjecture }\right\}
$$

has Hausdorff dimension 0.

Remark 1.12. For every $\delta>0$, there exists $(\alpha, \beta)$ such that

$$
\liminf _{n \in \mathbb{Z}} n^{1+\delta}\|n \alpha\|\|n \beta\|>0
$$

According to [Gal62], this was done in [Spe42]. So the exponent on $n$ is the best one can hope for. On the other hand, maybe one can improve $n$ by $\log n$ (see [Gal62] for some restrictions though).
1.5. Quantum unique ergodicity. Let $(M, d)$ be a closed hyperbolic surface of constant negative curvature. Let $\Delta$ be the Laplacian operator $-y^{2}\left(\frac{\partial^{2}}{\partial x^{2}}+\frac{\partial^{2}}{\partial y^{2}}\right)$.
Fact 1. Eigenvalues of $\Delta$ are non-negative and discrete in $\mathbb{R}$, say enumerated as

$$
0=\lambda_{1}<\lambda_{2}<\ldots
$$

Fact 2. For each $\lambda_{i}$, the eigenspace $E_{\lambda_{i}}$ consists of smooth functions and has finite dimension;
Fact 3. Different eigenspaces are mutually orthogonal and $L^{2}(M)$ is spanned by them. (see e.g. Thm 3.2.1, Jost, Riemannian Geometry; Thm 4.43, Gallot, Hulin, Lafontaine.)

Now take $f_{i} \in E_{\lambda_{i}}$. We are interested in the limiting behavior of the sequence of measures $\left\{\left|f_{i}\right|^{2} \mathrm{Vol}\right\}$, normalized to be probability measures.

A theorem (quantum ergodicity) of Snirelman says that there exists a density one subsequence $n_{i}$ such that $\lim \left|f_{n_{i}}\right|^{2} \mathrm{Vol}=\mathrm{Vol}$ (suitably normalized) in the weak* topology (this theorem holds for more general compact Riemannian manifold, as long as the geodesic flow is ergodic, a property that holds for every negatively curved surface).

Conjecture 1.13 (Quantum unique ergodicity). $\lim \left|f_{n}\right|^{2} \mathrm{Vol}=$ Vol holds without passing to any subsequence.

This is still open. Progress is made when the fundamental group is a "congruence subgroup" where there is an additional supply of operators, called Hecke operators, that commute with the Laplacian.

Theorem 1.14. Assume $\left\{f_{i}\right\}$ is a sequence of Hecke-Laplacian eigenfunctions. Then

$$
\lim \left|f_{n}\right|^{2} \mathrm{Vol}=\mathrm{Vol}
$$

in the weak* topology.
In the non-compact congruence case, this also holds for Hecke-Maass forms whose proof requires one more step to guarantee non-divergence.

## 2. Measure rigidity

2.1. Unipotent flows. Consider $\mathrm{SL}_{2}(\mathbb{R})$ and a discrete subgroup $\Gamma$. Equip $\mathrm{SL}_{2}(\mathbb{R})$ with a right invariant Riemannian metric. Then the volume measure $\mathrm{m}_{X}$ on $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ is left invariant under $\mathrm{SL}_{2}(\mathbb{R})$. We normalize it to be a probability measure.

Consider the subgroup

$$
\mathrm{U}:=\left\{u_{s}:=\left[\begin{array}{ll}
1 & s \\
0 & 1
\end{array}\right], s \in \mathbb{R}\right\}
$$

Theorem 2.1. Assume $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ is compact. Then $\mathrm{m}_{X}$ is the unique U -invariant probability measure.

This would immediately imply the denseness result above.
Theorem 2.2. Assume $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ has finite volume. Then each U -invariant probability measure is a convex combination (possible in the form of an integral) of the following

1. $\mathrm{m}_{X}$;
2. the U -invariant measure supported on a closed (and compact) orbit of U .

The implication to orbit closure requires an analysis on this convex combination.
In general, Ratner's measure classification on ergodic invariant measures for Ad-unipotent flows roughly reads as follows.

THEOREM 2.3 (Measure rigidity theorem). Assume the following

- a connected Lie group (2nd countable) $G$ together with a discrete subgroup $\Gamma$;
- a one-parameter Ad-unipotent subgroup $\mathrm{U}=\left\{\mathbf{u}_{s}, s \in \mathbb{R}\right\}$ of $G$.

Then every $U$-invariant ergodic probability measure $\mu$ on $G / \Gamma$ is homogeneous.
Ad-unipotent means that the image of $U$ under the Adjoint representation in $G L(\mathfrak{g})$ consist of unipotent matrices.

For a measure $\mu$ on $G / \Gamma$ define the closed subgroup of $G$ by

$$
H:=G_{\mu}:=\left\{g \in G, g_{*} \mu=\mu\right\}
$$

We say that a probability measure $\mu$ is homogeneous if there exists $x \in X=G / \Gamma$ such that $\mu(H x)=1$.

REMARK 2.4. When $G$ is a semisimple closed subgroup of $\mathrm{SL}_{n}$, Ad-unipotent is the same as being unipotent in $\mathrm{SL}_{n}$.

REMARK 2.5. Various "connected" assumptions may be dropped with similar conclusions. E.g. one may consider $\mathbf{u}_{s \in \mathbb{Z}}$.

REMARK 2.6. Let $H, x$ be as in the theorem and the definition above. Then $H x$ is closed in $G / \Gamma$. This is proved in [Rag72, Sec.1.13] assuming G/Г admits a finite G-invariant measure (i.e., $\Gamma$ is a lattice in $G$ ), but the proof carries through without this assumption.

REMARK 2.7. Let $H, x$ be as in the theorem and the definition above. Then by modifying $x$, one can show that $H x=H^{\circ} x$ and U is contained in $H^{\circ}$.

THEOREM 2.8 (Equidistribution and topological rigidity I). Further assume that $\Gamma$ is a lattice in $G$. Then for every $x$, there exists $\mathrm{U} \leq H \leq G$ closed connected subgroup such that

1. $H x$ is closed and supports an $H$-invariant probability measure $\mu_{H}$;
2. for every bounded continuous function $f: G / \Gamma \rightarrow \mathbb{R}$,

$$
\lim _{T \rightarrow \infty} \int_{0}^{T} f\left(u_{t} \cdot x\right) \mathrm{dt} \text { exists and is equal to } \int f(x) \mu_{H}(x)
$$

3. $\mathrm{U} \curvearrowright \mu_{H}$ is ergodic;
4. $\mathrm{U} \cdot x$ is dense in $H \cdot x$.

The logic of Ratner is

$$
\text { Measure rigidity } \Longrightarrow \text { Equidistribution } \Longrightarrow \text { Topological rigidity. }
$$

Nevertheless, there is a different (potential) approach by deducing the topological rigidity bypassing ergodic theory.

The topological rigidity is the original Raghunathan's conjecture.
THEOREM 2.9 (Topological rigidity II). Let $G, \Gamma$ be as in the last theorem. Let $L \leq G$ be a Lie subgroup generated by one-parameter Ad-unipotent subgroups. Then for every $x \in G / \Gamma$, there exists $L \leq H \leq G$ and $V \leq L$ some one-parameter Ad-unipotent subgroup (of $G$ ) such that

1. Hx is closed and supports an H-invariant probability measure;
2. $\overline{L x}=\overline{V x}=H x$;
3. $V \curvearrowright \mu_{H}$ is ergodic.
2.2. Higher rank diagonalizable action. Fact: Let $a_{t}:=\left[\begin{array}{ll}e^{t} & \\ & e^{-t}\end{array}\right]$. Then the $\left\{a_{t}\right\}_{t \in \mathbb{R}}$ action on $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ admits many invariant probability measures/closed sets and they are not easy to classify. The conjecture is that the situation would become better in higher rank.

Let

$$
A:=\left\{\left.\left[\begin{array}{lll}
e^{t_{1}} & & \\
& e^{t_{2}} & \\
& & e^{t_{3}}
\end{array}\right] \right\rvert\, t_{i} \in \mathbb{R}, \sum t_{i}=0\right\} \cong \mathbb{R}^{2} .
$$

Consider the $A \frown \mathrm{SL}_{3}(\mathbb{R}) / \mathrm{SL}_{3}(\mathbb{Z})$.
Conjecture 2.10. - Every ergodic invariant probability measure is homogeneous;

- Every bounded (in the unbounded case, statements need to be modified) orbit of A is homogeneous.

Of course one can propose similar (but necessarily more complicated) conjectures for other (semisimple) Lie groups $G$ and other $A$ 's.

Theorem 2.11. Let $G:=\mathrm{SL}_{3}(\mathbb{R}), \Gamma=\mathrm{SL}_{3}(\mathbb{Z})$ and $A$ same as above. Let $\mu$ be an $A$-invariant ergodic probability measure on $G / \Gamma$. Assume for some $a \in A, h_{\mu}(a)>0$. Then $\mu$ is the $G$ invariant probability measure on $G / \Gamma$.

The topological implication is that
Theorem 2.12. The Hausdorff dimension of

$$
\{x \in G / \Gamma, A x \text { is bounded }\}
$$

is 2 .
Note that the union of compact $A$-orbits is a countable union, hence also has Hausdorff dimension 2.

A theorem of slightly different flavor, related to the AQUE theorem above, is
Theorem 2.13. Let $G:=\mathrm{SL}_{2}(\mathbb{R}) \times \mathrm{SL}_{2}(\mathbb{R}), \Gamma \leq G$ irreducible (e.g. $\Gamma=\mathrm{SL}_{2}(\mathbb{Z}[\sqrt{2}])$ ). Let $H:=$ $\{e\} \times \mathrm{SL}_{2}(\mathbb{R})$. And

$$
A:=\left\{\left(a_{t}, \mathrm{id}\right), t \in \mathbb{R}\right\}
$$

Let $\mu$ be an A-invariant probability measure such that

- $h(a, v)>0$ for every ergodic component $v$ of $\mu$;
- $\mu$ is $H$-recurrent (some assumption weaker than $H$-invariant), then $\mu$ is the $G$-invariant probability measure.

Remark 2.14. Same conclusion holds replacing the 2nd factor $\mathrm{SL}_{2}(\mathbb{R})$ by $\mathrm{SL}_{2}\left(\mathbb{Q}_{p}\right)$. This $p$-adic version is what is required for the AQUE theorem.

REMARK 2.15. This theorem is not easily reduced to the ergodic case due to the recurrence condition.

REMARK 2.16. Whether one can eliminate the entropy assumption remains an open problem.

## 3. Further reading

Here are some general references.
[BM00] is a nice introduction to homogeneous dynamics including a proof of Oppenheim conjecture in the last chapter.

Einsiedler and Ward have a (ongoing) book project on homogeneous dynamics available on the authors' homepages.

What we plan to cover in this course (and almost everything I write here) can be found in the monograph $\left[\mathrm{EEE}^{+} 10\right]$.

## CHAPTER 1

## Denseness of horocycles

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## 1. Summary

In [Hed36], the author considers curves in the hyperbolic disk with constant "geodesic curvature" (let cur denote this number), measuring how far a curve is away from being a geodesic. Under this constancy condition, curves are divided into 4 types:

1. $\operatorname{cur}=0$;
2. $\operatorname{cur} \in(0,1)$;
3. cur $=1$;
4. cur $>1$.

Geodesics belong to type 1. Type 2 are those equidistant to some geodesic. Type 3 are horocycles. Type 4 are Euclidean circles in the interior of the disk model. The paper is about their "transitivity modulo $\Gamma$ ", namely their image in certain hyperbolic surfaces (orbifolds).

This chapter is about horocycles, that is, curves of type 3 assuming the relevant surface is compact. Actually our discussion applies to the unit-tangent bundle, not just to the surface itself. Type 4 are compact. It is claimed that type 2 behave like type 1 . One can also show that as the curvature tends to 1 , type 2 and type 4 asymptotically behave like type 3 .

Firstly, let us introduce some notations

- $\mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{cc}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}, \mathrm{A}:=\left\{\mathbf{a}_{t}: \left.=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\} ;$
- Let $\Gamma$ be a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{R})$ and let $\mathrm{X}:=\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$, equipped with the quotient topology;
- for $g \in \mathrm{SL}_{2}(\mathbb{R})$, write $[g]_{\Gamma}$ for its image in X.

One can identify X (at least when $\{ \pm 1\} \subset \Gamma$ and $\Gamma / \pm 1$ is torsion free) with the unit tangent bundle of some hyperbolic surface. Moreover, orbits of A are geodesics and orbits of $U$ are horocyles.

Theorem 1.1. Assume in addition that $\Gamma$ is cocompact in $\mathrm{SL}_{2}(\mathbb{R})$, namely, X is compact. Then the action of U on X is minimal, that is to say, for every $x \in \mathrm{X}, \mathrm{U} . x$ is dense in X .

In a dual formulation, this says that for every nonzero vector $v \in \mathbb{R}^{2}, \Gamma . v$ is dense in $\mathbb{R}^{2}$. This fails for $\Gamma=\mathrm{SL}_{2}(\mathbb{Z})$, but $\mathrm{SL}_{2}(\mathbb{Z})$ is not cocompact.

REMARK 1.2. The proof below applies equally well to the discrete case $\left\{\mathbf{u}_{s}\right\}_{s \in \mathbb{Z}}$ with the same conclusion. Namely, for every $x \in \mathrm{X},\left\{\mathbf{u}_{s} . x\right\}_{s \in \mathbb{Z}, s \geq 0}$ is dense in X . However, whether $\left\{u_{s^{2}} . x\right\}_{s \in \mathbb{Z}}$ is dense in X seems unknown (see [Zhe21] and references therein).

This chapter roughly corresponds to [BM00, Chapter IV, Section 2].

## 2. Injectivity radius

We fix some right invariant metric $d(\cdot, \cdot)$ on $\mathrm{SL}_{2}(\mathbb{R})$, compatible with the topology. We will not be bothered about the explicit form of the metric. So just take its existence as a fact. Assuming this, define the quotient metric on $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ by

$$
d\left([g]_{\Gamma},[h]_{\Gamma}\right):=\inf _{\gamma \in \Gamma} d(g \gamma, h)=\inf _{\gamma_{1}, \gamma_{2} \in \Gamma} d\left(g \gamma_{1}, h \gamma_{2}\right)
$$

Fix such a metric, we can define the injectivity radius at a point $x \in \mathrm{X}$ by

$$
\operatorname{Inj} \operatorname{Rad}(x):=\inf \{\delta>0 \mid g \mapsto g \cdot x \text { is injective on } d(g, \mathrm{id})<\delta\}
$$

Since $\Gamma$ is discrete, $\operatorname{InjRad}(x)>0$ for all $x \in X$. Also note that $\operatorname{InjRad}$ is continuous. Hence if $\Gamma$ is cocompact, there exists (and we fix such an) $r_{\mathrm{X}}>0$ such that $\operatorname{InjRad}(x) \geq r_{\mathrm{X}}$ for all $x \in \mathrm{X}$.

Also, one can check that for $d\left(g_{i}, \mathrm{id}\right)<\frac{r_{x}}{4}$ for $i=1,2$, we have $d\left(g_{1} \cdot x, g_{2} \cdot x\right)=d\left(g_{1}, g_{2}\right)$.
LEMmA 2.1. Assume $\Gamma$ is cocompact. Then there are no compact U -orbits in X , that is, for every $s \neq 0$ and $x \in \mathrm{X}, \mathbf{u}_{s} . x \neq x$. As every unipotent matrix in $\mathrm{SL}_{2}(\mathbb{R})$ is conjugate to an element of U , this implies that $\Gamma$ contains no (non-identity) unipotent matrices.

Proof. Assume otherwise, then we can find $g_{0} \in \mathrm{SL}_{2}(\mathbb{R})$ such that

$$
s_{0}:=\inf \left\{s>0 \mid \mathbf{u}_{s} . g_{0} \Gamma=g_{0} \Gamma\right\}>0
$$

In the current case inf is actually achieved at $s_{0}$ and $\mathbf{u}_{s_{0}} . g_{0} \Gamma=g_{0} \Gamma$. Consider

$$
\begin{aligned}
& \mathbf{a}_{-t} \mathbf{u}_{s_{0}} g_{0} \Gamma=\mathbf{a}_{-t} g_{0} \Gamma \\
\Longrightarrow & {\left[\begin{array}{cc}
1 & e^{-2 t} s_{0} \\
0 & 1
\end{array}\right]\left[\begin{array}{cc}
e^{-t} & 0 \\
0 & e^{t}
\end{array}\right] g_{0} \Gamma=\left[\begin{array}{cc}
e^{-t} & 0 \\
0 & e^{t}
\end{array}\right] g_{0} \Gamma . }
\end{aligned}
$$

As $t \rightarrow+\infty$, this implies the existence of compact orbit of $U$ of arbitrarily small period, which is impossible due to the fact $r_{\mathrm{X}}>0$. More explicitly, for $t$ large enough such that

$$
d\left(\mathrm{id},\left[\begin{array}{cc}
1 & e^{-2 t} s_{0} \\
0 & 1
\end{array}\right]\right)<r_{\mathrm{X}}
$$

One has, by the definition of $r_{\mathrm{X}}$, that $\mathbf{u}_{e^{-2 t} s_{0}}=\mathrm{id}$, or in other words, $s_{0}=0$. Here is a picture


Existence of closed horocydes
$\Rightarrow$ Existence of short closed horocycles

Corollary 2.2. Assume $\Gamma$ is cocompact and take $x \in \mathrm{X}$. There exist $t_{n}, s_{n} \rightarrow+\infty$ with $\left|t_{n}-s_{n}\right| \rightarrow \infty$ such that $d\left(x_{n}, y_{n}\right) \rightarrow 0$, where $x_{n}:=\mathbf{u}_{t_{n}} \cdot x$ and $y_{n}:=\mathbf{u}_{s_{n}} \cdot x$

Proof. The map $t \mapsto \mathbf{u}_{t} . x$ from $\mathbb{R}_{\geq 0}$ to X is injective. Since X is compact, we may apply pigeon-hole principle.

## 3. Additional invariance

Now we start to prove the theorem. The crucial notion here is
DEFINITION 3.1. Let a (semi)group G act on a topological space $W$ by homeomorphisms. A nonempty subset $V$ of $W$ is said to be $G$-minimal iff it is closed, $G$-stable and contains no proper non-empty closed $G$-stable subset.

Let $Y$ be a U-minimal set in the orbit closure $\overline{U . x_{0}}$. The existence of $Y$ is guaranteed by the compactness of $X$ and Zorn's lemma. Our strategy is to find some $y \in Y$ and a larger group whose orbit based at $y$ is contained in $Y$.

Proof of Theorem 1.4. Keep notations as Coro.2.2 above. When $n$ is large, we find some $A_{n} \in \mathrm{SL}_{2}(\mathbb{R})$ with $d\left(A_{n}, \mathrm{id}\right) \leq r_{\mathrm{X}} / 4$ such that $y_{n}=A_{n} x_{n}$. Write

$$
A_{n}=\left[\begin{array}{cc}
1+a_{n} & b_{n} \\
c_{n} & 1+d_{n}
\end{array}\right] \quad \text { with } a_{n}, b_{n}, c_{n}, d_{n} \rightarrow 0
$$

Coro.2.2 ensures that $A_{n}$ is not an upper triangular unipotent matrix.
The key calculation is:

$$
\begin{align*}
\mathbf{u}_{s} A_{n} \mathbf{u}_{s}^{-1} & =\left[\begin{array}{ll}
1 & s \\
0 & 1
\end{array}\right]\left[\begin{array}{cc}
1+a_{n} & b_{n} \\
c_{n} & 1+d_{n}
\end{array}\right]\left[\begin{array}{cc}
1 & -s \\
0 & 1
\end{array}\right] \\
& =\left[\begin{array}{ll}
1 & s \\
0 & 1
\end{array}\right]\left[\begin{array}{cc}
1+a_{n} & b_{n}-s\left(1+a_{n}\right) \\
c_{n} & 1+d_{n}-s c_{n}
\end{array}\right]  \tag{1}\\
& =\left[\begin{array}{cc}
1+a_{n}+s c_{n} & b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n} \\
c_{n} & 1+d_{n}-s c_{n}
\end{array}\right] .
\end{align*}
$$

Case I, $c_{n}=0$ for infinitely many $n$.
This case is left to you as an exercise.
Case II. $c_{n} \neq 0$ for $n$ large enough.
Equa.(1) above suggests that the upper right corner dominates when $s$ is large (this is called "shearing phenomenon", we will return to this point later).


We will kill the upper right corner according to the following computation

$$
\mathbf{u}_{t}\left(\mathbf{u}_{s} A_{n} \mathbf{u}_{s}^{-1}\right)=\left[\begin{array}{cc}
1+a_{n}+(s+t) c_{n} & b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n}+t\left(1+d_{n}-s c_{n}\right)  \tag{2}\\
c_{n} & 1+d_{n}-s c_{n}
\end{array}\right] .
$$

Define $t=t(s)$ by imposing the following equality

$$
\begin{gather*}
b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n}+t\left(1+d_{n}-s c_{n}\right)=0 \\
\Longleftrightarrow  \tag{3}\\
t=-\frac{b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n}}{d_{n}-s c_{n}}=-\frac{b_{n}-s a_{n}-s}{1+d_{n}-s c_{n}}-s
\end{gather*}
$$

The range of $s$ for which the $t(s)$ is ill-defined will be excluded from the discussion (see $s=s_{n, \delta}$ below, where one has $1+d_{n}-s c_{n}=1 \pm \delta$ with $\delta$ small). With this choice of $t=t(s)$,

$$
\mathbf{u}_{t}\left(\mathbf{u}_{s} A_{n} \mathbf{u}_{s}^{-1}\right)=\left[\begin{array}{cc}
\left(1+d_{n}-s c_{n}\right)^{-1} & 0  \tag{4}\\
c_{n} & 1+d_{n}-s c_{n}
\end{array}\right] .
$$

Now for $\delta>0$ (we will let $\delta \rightarrow 0$ in a moment), choose $s=s_{n, \delta} \geq 0$ such that either $1+d_{n}-s c_{n}=$ $1+\delta$ or $1-\delta$, depending on the signature of $c_{n}$. So

$$
s_{n, \delta}=\frac{d_{n}-\delta}{c_{n}} \text { or } \frac{d_{n}+\delta}{c_{n}},
$$

whichever is positive.
Define

$$
y_{n, \delta}^{\prime}:=\mathbf{u}_{t(s)} \mathbf{u}_{s} \cdot y_{n}, \quad x_{n, \delta}^{\prime}:=\mathbf{u}_{s} . x_{n}, \quad \text { where } s=s_{n, \delta}
$$

Then by definition

$$
\begin{align*}
y_{n, \delta}^{\prime} & =\mathbf{u}_{t(s)} \mathbf{u}_{s} A_{n} \mathbf{u}_{s}^{-1} \mathbf{u}_{s} \cdot x_{n}=\mathbf{u}_{t(s)} \mathbf{u}_{s} A_{n} \mathbf{u}_{s}^{-1} \cdot x_{n, \delta}^{\prime} \\
& =\left[\begin{array}{cc}
(1 \pm \delta)^{-1} & 0 \\
0 & (1 \pm \delta)
\end{array}\right] \cdot x_{n, \delta}^{\prime} \tag{5}
\end{align*}
$$

Fix $\delta$, let $n$ vary. By passing to a subsequence $n_{k}$, assume that $y_{n, \delta}^{\prime}$ and $x_{n, \delta}^{\prime}$ converge to $y_{\infty, \delta}$ and $x_{\infty, \delta}$ respectively. Hence

$$
y_{\infty, \delta}^{\prime}=\left[\begin{array}{cc}
(1+\delta)^{-1} & 0 \\
0 & (1+\delta)
\end{array}\right] \cdot x_{\infty, \delta}^{\prime} \text { or }\left[\begin{array}{cc}
(1-\delta)^{-1} & 0 \\
0 & (1-\delta)
\end{array}\right] \cdot x_{\infty, \delta}^{\prime}
$$

Without loss of generality, assume that the first case happens for infinitely many $\delta>0$ converging to 0 . It looks like we are not making any progress except that the "transverse difference" is now in the direction of the diagonal, which normalizes $U$. So it is time to invoke the following general fact, which is why we introduced the notion of minimal set.

Lemma 3.2. Let $\Gamma \curvearrowright Z$ by homeomorphisms. $\Gamma$ is a semi-group and $Z$ a topological space. Assume that $V$ is a $\Gamma$-minimal set and $W$ is a $\Gamma$-invariant closed set. If $\phi \in \operatorname{Homeo}(X)$ normalizes (the image of) $\Gamma$ and there exist $v_{0} \in V$ and $w_{0} \in W$ with $\phi\left(v_{0}\right)=w_{0}$. Then $\phi(V)$ is contained in $W$.

Proof of the Lemma.

$$
\phi(V)=\phi\left(\overline{\Gamma \cdot v_{0}}\right)=\overline{\phi\left(\Gamma \cdot v_{0}\right)}=\overline{\Gamma \cdot w_{0}} \subset W
$$

From the lemma (applied to $V=W=Y, Z=\mathrm{X}$ ), we see that for a set of $\delta$ converging to 0 and for every $y \in Y$,

$$
\left[\begin{array}{cc}
(1+\delta)^{-1} & 0 \\
0 & (1+\delta)
\end{array}\right] . Y \subset Y
$$

Since $\left\{g \in \mathrm{SL}_{2}(\mathbb{R}) \mid g Y \subset Y\right\}$ is a closed sub-semigroup,

$$
\left\{\mathbf{a}_{t} \mid t \geq 0\right\} . Y \subset Y
$$

is contained in $Y$ for every $y$. By definition $\left\{\mathbf{u}_{s} \mid s \geq 0\right\} . Y \subset Y$. By minimality, actually $\left\{\mathbf{u}_{s} \mid s \geq 0\right\} . Y=$ $Y$.

Fix some $y \in Y$, and take a limit point $y^{\prime}$ of $\mathbf{a}_{t} y$ as $t \rightarrow+\infty$. Then the orbit of the full group A. $y^{\prime} \subset Y \Longrightarrow$ AU. $y^{\prime} \subset Y$. Thus we are done modulo Lem.4.1.

## 4. A duality argument

Let $B^{+}:=\left\{\mathbf{a}_{t} \mathbf{u}_{s}\right\}_{t, s \in \mathbb{R}}=\mathrm{A} \cdot \mathrm{U}$ and $B:=\left\{( \pm 1) \mathbf{a}_{t} \mathbf{u}_{s}\right\}_{t, s \in \mathbb{R}} . B^{+}$is the identity component of $B$ and $B=B^{+} \sqcup(-1) B^{+}$where we have abbreviated the matrix $\left[\begin{array}{cc}-1 & 0 \\ 0 & -1\end{array}\right]$ as " -1 ".

Lemma 4.1. The action of $B^{+}$on X is minimal.
Remark 4.2. The lemma also holds when only assuming $\Gamma$ to be discrete and of finite covolume (referred to as a lattice). Actually, the lemma follows iff the limit set for $\Gamma$ is the full boundary, which includes some infinitely generated, infinite covolume examples.

Proof. We are going to show that the $B$-action is minimal first and then explain why this is sufficient.

An equivalent formulation is that the $\Gamma$-action on $\mathrm{SL}_{2}(\mathbb{R}) / B$ is minimal. To prove this, we will take a geometric point of view.

Recall that $\mathrm{SL}_{2}(\mathbb{R})$ acts on the upper half space $\mathscr{H}^{2}:=\{z=x+i y \mid x \in \mathbb{R}, y>0\}$ by

$$
\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right] \cdot z:=\frac{a z+b}{c z+d} .
$$

This action preserves the Riemannian metric (referred to as the hyperbolic metric)

$$
\left(\mathrm{dx}^{2}+\mathrm{dy} \mathrm{y}^{2}\right) / y^{2}
$$

Geodesics under the hyperbolic metrics are (Euclidean) circles perpendicular to the x -axis together with all the vertical lines.

Another important point is that as the $y$-coordinate approaches 0 , the (hyperbolic) distance between two points of (Euclidean) distance $=1$ actually goes to $\infty$. The $\mathrm{SL}_{2}(\mathbb{R})$-action extends continuously to the "boundary" defined by

$$
\partial \mathscr{H}^{2}:=\{(x, y), y=0\} \sqcup\{\infty\} .
$$

where the topology near $\infty$ is defined as the "one-point compactification". Thus topologically the boundary is a circle. The action at $\infty$ is given as follows

$$
\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right] \cdot \infty=\frac{a \infty+b}{c \infty+d}=\left\{\begin{array}{lc}
a / c, & \text { if } c \neq 0 \\
\infty, & \text { otherwise }
\end{array} .\right.
$$

Why care? Note that the stabilizer of $\infty$ is exactly $B$ and the action is transitive on $\partial \mathscr{H}^{2}$ (Exercise: convince yourself that this gives a topological homeomorphism $\mathrm{SL}_{2}(\mathbb{R}) / B \cong \mathscr{H}^{2}$ ) Thus it suffices to show that the action of $\Gamma$ on $\partial \mathscr{H}^{2}$ is minimal.

Claim 4.1. For every $z \in \mathscr{H}^{2}$, the orbit closure $\overline{\Gamma \cdot z} \supset \partial \mathscr{H}^{2}$.
Assuming the claim, let $W$ be a closed $\Gamma$-invariant set on $\partial \mathscr{H}^{2} \cong S^{1}$. Thus its complement consists of disjoint union of open intervals (labelled as $I_{i}$ 's). Take such an interval $I_{0}$ with endpoints $w_{1}, w_{2}$. We argue that $\Gamma \cdot \widehat{w_{1} w_{2}}$ (the unique geodesic connecting $w_{1}$ and $w_{2}$ ) never contains $I_{0}$ in its closure, which contradicts against the above claim. Indeed, $\Gamma$ translates of $\overline{w_{1} w_{2}}$ are just geodesics with endpoints outside the region between $\overline{w_{1} w_{2}}$ and $I_{0}$.


Hence we are done.
Proof of Claim 4.1. By co-compactness, we can find a bounded region $\mathscr{B} \subset \mathscr{H}^{2}$ (whose diameter under hyperbolic distance is denoted by $\operatorname{diam}(\mathscr{B})$ ) such that $\Gamma . \mathscr{B}=\mathscr{H}^{2}$. For every $z_{\neq \infty} \in \partial \mathscr{H}^{2}$ (the case $z=\infty$ is left to you) and a neighborhood $\mathscr{N}_{z, r_{0}}$ of radius $r_{0}$ (in the Euclidean metric) of $z$, we are going to show that some $\gamma \cdot \mathscr{B}$ is contained in $\mathscr{N}_{z, r_{0}}$. Indeed we can find $\gamma \in \Gamma, b \in \mathscr{B}$ such that $\gamma b . z \in \mathscr{N}_{z, r_{0} / 2}$. When $r_{0}$ is sufficiently small one can show that for $z^{\prime} \in \mathscr{H}^{2}$

$$
d_{\text {Hyperbolic }}\left(z^{\prime}, \gamma b . z\right) \leq \operatorname{diam}(\mathscr{B}) \Longrightarrow d_{\text {Euclidean }}\left(z^{\prime}, \gamma b . z\right) \leq r_{0} / 2 .
$$

Applying this to $z^{\prime}=\gamma . z$ finishes the proof.
Finally, as promised, we explain how to get the minimality of $B^{+}$-action from that of $B$. So take $x_{0} \in \mathrm{X}$ and we know that

$$
\overline{B \cdot x_{0}}=\overline{B^{+} \cdot x_{0} \cup B^{+}(-1) \cdot x_{0}}=\mathrm{X} .
$$

As $\overline{B^{+} x_{0}} \cup \overline{B^{+}(-1) x_{0}}$ is $B$-invariant and closed, it is also equal to X . As X is connected (well, the group $\mathrm{SL}_{2}(\mathbb{R})$ is connected), their intersection $\overline{B^{+} x_{0}} \cap \overline{B^{+}(-1) x_{0}}$ is non-empty. But this again, is a $B$-invariant closed set, so has to be the full X . In particular $\overline{B^{+} x_{0}}=X$. And the proof completes.

## 5. Exercises

EXERCISE 5.1. Let $G:=\mathrm{SL}_{2}(\mathbb{R})$ act continuously on a locally compact Hausdorff topological space $X$. Take $x \in X$, let $G_{x}:=\{g \in g, g . x=x\}$. Assume G. $x$ is closed, or is open in its closure, show that the bijection induced from $g \mapsto g . x$

$$
G / G_{x} \rightarrow G . x
$$

is a homeomorphism where $G / G_{x}$ is equipped with the quotient topology and $G . x$ is equipped with the subspace topology.

Hint: Baire's Category theorem.
EXERCISE 5.2. Assume $\Gamma$ is cocompact in $\mathrm{SL}_{2}(\mathbb{R})$. Consider the standard action of $\mathrm{SL}_{2}(\mathbb{R})$ on $\mathbb{R}^{2}$. Show that for every $v_{\neq 0} \in \mathbb{R}^{2}, \Gamma . v$ is dense in $\mathbb{R}^{2}$.

Exercise 5.3. Show that $\mathrm{SL}_{2}(\mathbb{R})$ does not admit any bi-invariant Riemannian metric.
5.1. A more geometric take on horocycles. I assume you have some familiarity with geometry on the upper half space in this section.

Notations:

- $\mathbb{H}^{2}:=\left\{(x, y) \in \mathbb{R}^{2}, y>0\right\}$ equipped with the metric $\frac{\mathrm{dx}^{2}+\mathrm{dy}^{2}}{y^{2}}$ and the left action of $\mathrm{SL}_{2}(\mathbb{R})$ via fractional linear transformations;
- $T^{1}\left(\mathbb{H}^{2}\right)$ is the unit tangent bundle of $\mathbb{H}^{2}$;
- $\partial \mathbb{H}^{2}:=\left\{(x, 0) \in \mathbb{R}^{2}, x \in \mathbb{R}\right\} \sqcup\{\infty\}$ be the boundary of $\mathbb{H}^{2}$; The topology on $\{(x, y), x \in$ $\mathbb{R}, y \geq 0\}$ is the natural topology and the topology on $\overline{\mathbb{H}^{2}}:=\mathbb{H}^{2} \sqcup \partial \mathbb{H}^{2}$ is the one-point compactification topology. The action of $\mathrm{SL}_{2}(\mathbb{R})$ extends continuously to $\overline{\mathbb{H}^{2}}$;
- Let $\Gamma_{0}$ be a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{Z})$ such that $\Gamma_{0} \backslash \mathbb{H}^{2}$ is a closed surface of genus $g \geq 2$;
- Let $\Gamma_{0}^{\prime}:=\left[\Gamma_{0}, \Gamma_{0}\right]$, recall that $\Gamma_{0}^{\prime}$ is a normal subgroup of $\Gamma_{0}$ and $\Gamma_{0} / \Gamma_{0}^{\prime} \cong \mathbb{Z}^{2 g}$;
- For $x \in \mathbb{H}^{2}$ and a discrete subgroup $\Gamma$ of $\mathrm{SL}_{2}(\mathbb{R})$, define the limit set $\operatorname{Limit}_{x}(\Gamma):=\overline{\Gamma \cdot x} \backslash$ $\Gamma . x$ in $\overline{\mathbb{H}^{2}}$.
EXERCISE 5.4. $\operatorname{Limit}_{x}(\Gamma) \subset \partial \mathbb{H}^{2}$ for every discrete subgroup $\Gamma$ of $\mathrm{SL}_{2}(\mathbb{R})$ and every $x \in \mathbb{H}^{2}$.
EXERCISE 5.5. For every $x, y \in \mathbb{H}^{2}$ and discrete subgroup $\Gamma$ of $\operatorname{SL}_{2}(\mathbb{R}), \operatorname{Limit}_{x}(\Gamma)=\operatorname{Limit}_{y}(\Gamma)$.
Thus the limit set is independent of the choice of base point and we henceforth denote it by $\operatorname{Limit}(\Gamma)$.

EXERCISE 5.6. Let $\Gamma$ be a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{R})$. Show that $\operatorname{Limit}(\Gamma)$ is a $\Gamma$-minimal set.
(A $\Gamma$-set is said to be $\Gamma$-minimal iff either it is empty or for every $x$ in this set, $\Gamma . x$ is dense in this set. Actually Limit( $\Gamma$ ), if infinite, is the unique nonempty $\Gamma$-minimal set)

Recall that for every geodesic $Y$ (or closed convex subset) on $\mathbb{H}^{2}$ and every $x \in \mathbb{H}^{2}$, there is a unique point, denoted as $\pi_{Y}(x)$, in $Y$ such that

$$
\operatorname{dist}(x, Y)=\operatorname{dist}\left(x, \pi_{Y}(x)\right)
$$

For every $x \in T^{1}\left(\mathbb{H}^{2}\right)$, let $x^{+}:=\lim _{t \rightarrow+\infty} g_{t} \cdot x$ and $x^{-}:=\lim _{t \rightarrow-\infty} g_{t} \cdot x$ where $g_{t}$ denotes the geodesic flow. Let $\overline{x^{-} x^{+}}$be the unique geodesic in $T^{1} \mathbb{W}^{2}$ connecting $x^{-}$and $x^{+}$. By abuse of notation we also let $\overline{x^{-} x^{+}}$denote its projection to $\mathbb{H}^{2}$. Fix some point $o \in \mathbb{H}^{2}$ (say, take $o=(0,1)$ ), and $x \in T^{1} \mathbb{W}^{2}$, let $t=t_{o}(x)$ be the unique real number such that

$$
x=g_{t} \cdot \pi \widetilde{x^{-x^{+}}}(o) .
$$

(a priori, $\pi_{\overline{x^{-} x^{+}}}(o)$ is just an element in $\mathbb{H}^{2}$ but we identify it with the unique element on $\overline{x^{-} x^{+}} \subset T^{1} \mathbb{H}^{2}$ whose projection to $\mathbb{H}^{2}$ is $\pi \overline{x^{-} x^{+}}(o)$ )

EXERCISE 5.7. The map $\Phi_{o}: T^{1} \sharp^{2} \rightarrow\left(\partial \Vdash^{2} \times \partial \Vdash^{2} \backslash \Delta \partial \Vdash^{2}\right) \times \mathbb{R}$ defined by

$$
x \mapsto \Phi_{o}(x):=\left(x^{-}, x^{+}, t_{o}(x)\right)
$$

is a homeomorphism.
This is the so-called Hopf coordinate.
EXERCISE 5.8. Check that $\Phi_{o}\left(g_{t} \cdot x\right)=\left(x^{-}, x^{+}, t_{o}(x)+t\right)$.
EXERCISE 5.9. Check that for $\gamma \in \mathrm{SL}_{2}(\mathbb{R}), \Phi_{o}(\gamma \cdot x)=\left(\gamma \cdot x^{-}, \gamma \cdot x^{+}, *\right)$ for some real number $*$.
Thus the orbits of $\Gamma$ on $T^{1} \mathbb{H}^{2} /\left\{g_{t}\right\}_{t \in \mathbb{R}}$ corresponds to the orbits of $\Gamma$ on $\partial \Vdash^{2} \times \partial \Vdash^{2} \backslash \Delta \partial \Vdash^{2}$.
Exercise 5.10. Let $\Gamma$ be a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{R})$. Using the fact that $g_{t}$-action on $\Gamma \backslash T^{1} \Vdash^{2}$ is not minimal, show that the action of $\Gamma$ on $\partial \Vdash^{2} \times \partial \Vdash^{2} \backslash \Delta \partial \Vdash^{2}$ is not minimal.

This action is still quite chaotic, at least when $\Gamma$ is a lattice, but if we take one step further, it becomes totally discontinuous.

Let FAT $\Delta$ be the "fat diagonal" in $\partial \mathbb{H}^{2} \times \partial \Vdash^{2} \times \partial \Vdash^{2}$, i.e.

$$
\text { FAT } \Delta:=\left\{\left(x_{1}, x_{2}, x_{3}\right) \in\left(\partial \Vdash^{2}\right)^{3}, x_{i}=x_{j}, \exists i \neq j\right\} .
$$

EXERCISE 5.11. Let $\Gamma$ be a subgroup of $\mathrm{SL}_{2}(\mathbb{R})$. Show that the diagonal $\Gamma$-action on $\left(\partial \mathbb{H}^{2}\right)^{3}$ । FATA is conjugate to the $\Gamma$-action on $\mathbb{H}^{2}$.

Now turn to the special $\Gamma_{0}, \Gamma_{0}^{\prime}$ we defined. Recall in Lec 2 we have shown that $\operatorname{Limit}\left(\Gamma_{0}\right)$ is the full $\partial \Vdash^{2}$. Show that also

EXERCISE 5.12. $\operatorname{Limit}\left(\Gamma_{0}^{\prime}\right)=\partial \Vdash^{2}$.
(Hint: use Exer 5.5 and the fact that $\Gamma_{0}^{\prime}$ is a normal subgroup)
EXercise 5.13. Use this and the "thin" property of hyperbolic space to show that closed geodesics are dense in $\Gamma_{0}^{\prime} \backslash T^{1} \sharp^{2}$.
(In Lec. 3 we established denseness of closed geodesics on $\mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$ by constructing an explicit one and then considering commensurable lattices)

For a point $v$ on $\partial \Vdash^{2}$ and $x \in \mathbb{H}^{2}$, let $\mathscr{H}_{\nu}(x)$ be the unique horocycle - the unique Euclidean circle tangent to $\partial \Vdash^{2}$ at $v(\neq \infty)$ and passing through $x$ (when $v=\infty, \mathscr{H}_{v}(x)$ is a horizontal line passing through $x$ ). We shall think of $\mathscr{H}_{\nu}(x)$ as a subset of $T^{1} \mathbb{W}^{2}$ by equipping every point $\mathscr{H}_{\nu}(x)$ with the unique unit tangent vector that is orthogonal to $\mathscr{H}_{\nu}(x)$ and pointing towards $\nu$.

In Lec. 2 we have shown that the projection of every horocycle is dense in $\Gamma_{0} \backslash T^{1} \mathbb{W}^{2}$. Here is a more geometric approach following Hedlund's paper.

EXERCISE 5.14. Show that for every nonempty open interval $I \subset \partial \Vdash^{2}$ and $x \in \Vdash^{2}$, the set

$$
\bigcup_{\nu \in I} \mathscr{H}_{\nu}(x)
$$

is dense in $\Gamma_{0}^{\prime} \backslash T^{1} \mathbb{H}^{2}$.
(Hint: use Exer.5.13)
EXERCISE 5.15. Let $v \in \partial \mathbb{H}^{2}$, show that if there exists $x \in \mathbb{H}^{2}$ such that $\mathscr{H}_{\nu}(x)$ is dense in $\Gamma_{0}^{\prime} \backslash T^{1} \mathbb{H}^{2}$, then $\mathscr{H}_{\nu}(y)$ is dense in $\Gamma_{0}^{\prime} \backslash T^{1} \mathbb{H}^{2}$ for every $y \in \mathbb{H}^{2}$.

EXercise 5.16. The set of $v$ such that $\mathscr{H}_{\nu}(x)$ is dense in $\Gamma_{0}^{\prime} \backslash T^{1} \sharp^{2}$ is dense in $\partial \sharp^{2}$.
Let $\mathscr{D}$ be a Dirichlet fundamental domain for $\Gamma_{0}^{\prime}$. Accept the fact that if $\Gamma_{0}^{\prime}$ were finitely generated, then $\mathscr{D}$ would have only finitely many sides.

EXercise 5.17. Show that $\Gamma_{0}^{\prime}$ is not finitely generated.
Exercise 5.18. Let $v \in \partial \Vdash^{2} \cap \overline{\mathscr{D}}$, then $\mathscr{H}_{\nu}(x)$ is not dense in $\Gamma_{0}^{\prime} \backslash T^{1} \mathbb{H}^{2}$.
(Hint: without loss of generality assume $v=\infty$, argue that, fixing a base point $o$, there is an upper bound for the $y$-coordinate of $\gamma . o$ as $\gamma$ varies in $\Gamma_{0}^{\prime}$.)

Since $\mathscr{H}_{\nu}(x)$ is not compact in $\Gamma_{0}^{\prime} \backslash T^{1} \mathbb{H}^{2}$, we have demonstrated an orbit of the horocycle flow that is neither dense nor compact.

EXercise 5.19. Take some $y \in T^{1} \mathbb{W}^{2}$ such that $\left\{g_{t} . y\right\}$ is compact in $\Gamma_{0} \backslash T^{1} \mathbb{H}^{2}$. Show that $\mathscr{H}_{y^{+}}(x)$ is dense in $\Gamma_{0} \backslash T^{1} \mathbb{H}^{2}$.
(Hint: approximate some dense horocycle in $T^{1} \mathbb{H}^{2}$ )

EXERCISE 5.20. Let $v \in \partial \mathbb{H}^{2}$ and fix some $x \in \mathbb{H}^{2}$. Suppose the Euclidean radius of $\gamma . \mathcal{H}_{v}(x)$ can be arbitrarily large as $\gamma$ varies in $\Gamma_{0}$. Then $\mathscr{H}_{\nu}(x)$ is dense in $\Gamma_{0} \backslash T^{1} \mathbb{H}^{2}$.
(When the horocycle is based at infinity, by saying the Euclidean radius is large, we mean that the horocycle could be very low) (Hint: show that you can approximate every periodic geodesic)

EXERCISE 5.21. Show that indeed, since $\Gamma_{0} \backslash T^{1} \sharp^{2}$ is compact, that the Euclidean radius of $\gamma \cdot \mathscr{H}_{\nu}(x)$ can be arbitrarily large as $\gamma$ varies in $\Gamma_{0}$ for every pair $v \in \partial \uplus^{2}$ and $x \in \mathbb{H}^{2}$.
(Hint: use the fact that the some (well, in the current case, every) geodesic stemming from $v$ is bounded in $\Gamma_{0} \backslash T^{1} \mathbb{H}^{2}$ )

## A dynamical reformulation of Oppenheim conjecture

Back to the Top.
We recommend the last chapter of [BM00] for an elementary account of the proof of Oppenheim conjecture. See [Mar97, LM14, BGHM10] for history and more recent stories.

## 1. The statement

The goal of this and the next lecture is to prove a weak Oppenheim conjecture. In this lecture we will reduce the proof to a dynamical statement whose proof is delegated to the next lecture. A stronger form will be treated later with the help of non-divergence of unipotent flows.

Theorem 1.1. Let $Q$ be a non-degenerate indefinite quadratic from with real coefficients in $N \geq 3$ variables. Assume that $Q$ is not a scalar multiple of some quadratic form with rational coefficients. Then the closure of $Q\left(\mathbb{Z}^{N} \backslash \mathbf{0}\right)$ contains 0 .

REMARK 1.2. This theorem says nothing nontrivial to the quadratic form $Q_{1}=x y-\sqrt{2} z^{2}$ since $Q_{1}(1,0,0)=0$. However, it is nontrivial for $Q_{2}=x^{2}+y^{2}-\sqrt{2} z^{2}$ since the value of $Q_{2}$ at integral points can never be 0 unless $(x, y, z)=(0,0,0)$.

Later we will specialize to the case when $N=3$, from which the general case would follow. Details are left to the reader.

REMARK 1.3. Counter examples exist when $N=2$. For instance consider the quadratic form $Q\left(x_{1}, x_{2}\right):=\left(x_{1}-\sqrt{2} x_{2}\right) x_{2}$. Note that $\sqrt{2}$ is badly approximable which means that there exists $c>0$ such that $\{\sqrt{2} m\}|m| \geq c$ for all non-zero integer $m$ where $\{\cdot\}$ stands for the distance to the nearest integer. We will sketch a dynamical explanation below.

## 2. The space of lattices

For a quadratic form $Q$ in $N$ variables, define for $k=\mathbb{R}, \mathbb{Q}, \mathbb{Z}$,

$$
\begin{equation*}
\mathrm{SO}_{Q}(k):=\left\{g \in \mathrm{SL}_{N}(k) \mid Q \circ g=Q\right\} . \tag{6}
\end{equation*}
$$

The definition makes sense for $Q$ irrational. It might happen that $\mathrm{SO}_{Q}(\mathbb{Z})$ is trivial. If $M_{Q}$ is the symmetric matrix representing of $Q$, i.e. $Q(v)=v^{t r} M_{Q} v$ ( $v$ written as a column vector), then

$$
\begin{equation*}
\mathrm{SO}_{Q}(k):=\left\{g \in \mathrm{SL}_{N}(k) \mid g^{t r} M_{Q} g=M_{Q}\right\} . \tag{7}
\end{equation*}
$$

One can compute the Lie algebra of $\mathrm{SO}_{Q}(\mathbb{R})$ as

$$
\mathfrak{s o}_{Q}=\left\{X \in \mathfrak{s l}_{n}(\mathbb{R}) \mid M_{Q} X+X^{t r} M_{Q}=0\right\} .
$$

Where does it act on?
Definition 2.1. A subgroup $\Lambda$ of $\mathbb{R}^{N}$ is said to be a (unimodular) lattice if $\Lambda$ is discrete and cocompact in $\mathbb{R}^{N}\left(\right.$ with $\left.\operatorname{Vol}\left(\mathbb{R}^{N} / \Lambda\right)=1\right)$.

Here Vol is taken with respect to the standard Euclidean metric on $\mathbb{R}^{N}$.

EXAMPLE 2.2. $\mathbb{Z}^{N}$ is a unimodular lattice in $\mathbb{R}^{N}$.
EXAMPLE 2.3. $\mathbb{Z}[\sqrt{2}]$ may be viewed as a lattice in $\mathbb{R}^{4}$ by the geometric embedding, i.e.

$$
\Lambda:=\{(x, y) \mid x, y \in \mathbb{Z}[\sqrt{2}], y=\sigma(x)\}
$$

where $\sigma$ is the nontrivial element in $\operatorname{Gal}(\mathbb{Q}(\sqrt{2}) / \mathbb{Q})$.
EXAMPLE 2.4. $\mathbb{Z}[\sqrt{-2}]$ may be viewed as a lattice in $\mathbb{R}^{2}$ by identifying it with $\mathbb{C}$, explicitly,

$$
\Lambda=\{(x, \sqrt{2} y) \mid x, y \in \mathbb{Z}\}
$$

EXAMPLE 2.5. You can get a unimodular lattice starting from a lattice by multiplying a scalar.


Explicitly, for every discrete subgroup $\Lambda$ of $\mathbb{R}^{N}$, one can find $\nu_{1}, \ldots, \nu_{n}$ in $\mathbb{R}^{N}$ that are $\mathbb{R}$ linearly independent and $\Lambda=\mathbb{Z} \nu_{1} \oplus \mathbb{Z} \nu_{2} \oplus \ldots \oplus \mathbb{Z} \nu_{n}$. Such a set $\left\{v_{1}, \ldots, v_{n}\right\}$ will be called a basis of $\Lambda$. And $n$ is called the rank of $\Lambda$. $\Lambda$ is a lattice iff $n=N$. Conversely, given $n$ vectors $\nu_{1}, \ldots, v_{n}$ that are $\mathbb{R}$-linearly independent, the subgroup $\mathbb{Z} \nu_{1}+\ldots+\mathbb{Z} v_{n}$ is a discrete subgroup of $\mathbb{R}^{N}$.

Assume a discrete subgroup $\Lambda$ has rank $N$ with basis $\left(v_{1}, \ldots, v_{N}\right)$. Then $\operatorname{Vol}\left(\mathbb{R}^{N} / \Lambda\right)=\left|\operatorname{det}\left(v_{1}, . ., v_{N}\right)\right|=$ $\left\|\nu_{1} \wedge \ldots \wedge \nu_{N}\right\|$. This is because

$$
\left\{a_{1} v_{1}+\ldots+a_{N} v_{N} \mid a_{i} \in[0,1)\right\}
$$

forms a strict fundamental domain for $\mathbb{R}^{N} / \Lambda$, namely, it is in bijection with $\mathbb{R}^{N} / \Lambda$ under the quotient map. Also, let us recall that

$$
\operatorname{Vol}\left(\mathbb{R}^{N} / \Lambda\right)=\left\|\nu_{1}\right\| \cdot \operatorname{dist}\left(\nu_{2}, \mathbb{R} \nu_{1}\right) \cdot \operatorname{dist}\left(\nu_{3}, \mathbb{R} \nu_{1}+\mathbb{R} \nu_{2}\right) \cdot \ldots \cdot \operatorname{dist}\left(\nu_{N}, \mathbb{R} \nu_{1}+\ldots+\mathbb{R} \nu_{N-1}\right)
$$

Thus $\Lambda$ is a unimodular lattice iff $\operatorname{det}\left(v_{1}, \ldots, v_{N}\right)= \pm 1$.
It is useful to be familiar with quotient constructions in Euclidean spaces. More precisely, an Euclidean space is a finite-dimensional $\mathbb{R}$-linear space together with a non-degenerate positive definite quadratic form (or an "inner product", if you prefer). The "standard" $\mathbb{R}^{N}$ is nothing but the vector space $\mathbb{R}^{N}$ together with the form $Q_{s t d}\left(x_{1}, \ldots, x_{N}\right):=x_{1}^{2}+\ldots+x_{N}^{2}$. Once an Euclidean space is given, we can talk about distance, volume...

If $W$ is an $\mathbb{R}$-subspace of $\mathbb{R}^{N}$, then we think of $W$ as an Euclidean space by restricting the quadratic form to $W$. Since $Q_{s t d}$ is positive definite, its restriction to every subspace is also positive definite. Also, the quotient $\mathbb{R}^{N} / W$ is equipped with a natural Euclidean structure by identifying it with the orthogonal complement of $W$ in $\mathbb{R}^{N}$. Alternatively, you can define the quotient metric on $\mathbb{R}^{N} / W$ and then argue that it comes from a quadratic form. These two methods give the same Euclidean structure on $\mathbb{R}^{N} / W$.

Definition 2.6. Let $\mathrm{X}_{N}$ be the set of unimodular lattices in $\mathbb{R}^{N}$ equipped with the Chabauty topology.

Alternatively, one may think of $\mathrm{X}_{N}$ as the set of all lattices of $\mathbb{R}^{N}$ up to $\mathbb{R}^{*}$-action.
A detailed treatment of Chabauty topology may be found in [BP92, Chapter E, Section 1]. For us, it suffices to know that under the Chabauty topology, a sequence $\left(\Lambda_{n}\right) \subset \mathrm{X}_{N}$ converges to $\Lambda \in \mathrm{X}_{N}$ iff one can find a basis $v_{1}^{n}, \ldots, v_{N}^{n}$ of $\Lambda_{n}$ such that as $n \rightarrow \infty,\left(v_{i}^{n}\right)_{n}$ converges to some $v_{i}^{\infty} \in \mathbb{R}^{N}$ for every $i=1, \ldots, N$ and $\Lambda=\oplus_{i} \mathbb{Z} v_{i}^{\infty}$.

Note that for a sequence $\left(\Lambda_{n}\right) \subset \mathrm{X}_{N}$, if there are bases $\left(v_{1}^{n}, \ldots, v_{N}^{n}\right)$ with $\left(v_{i}^{n}\right)_{n}$ converging to some $v_{i}^{\infty} \in \mathbb{R}^{N}$ for every $i$, then $\left(v_{1}^{\infty}, \ldots, v_{N}^{\infty}\right)$ are automatically $\mathbb{R}$-linearly independent and they span a lattice $\Lambda$ with covolume $\operatorname{Vol}\left(\mathbb{R}^{N} / \Lambda\right)=1$.

The space $\mathrm{X}_{\mathrm{N}}$ admits a natural action of $\mathrm{SL}_{N}(\mathbb{R})$ and
LEMMA 2.7. The mapg $\mapsto g \cdot \mathbb{Z}^{N}$ from $\mathrm{SL}_{N}(\mathbb{R})$ to $\mathrm{X}_{\mathrm{N}}$ descends to a homeomorphism $\mathrm{SL}_{N}(\mathbb{R}) / \mathrm{SL}_{N}(\mathbb{Z}) \cong$ $\mathrm{X}_{\mathrm{N}}$.

Proof. $\mathrm{SL}_{N}(\mathbb{Z})$ is equal to the stabilizer of $\mathbb{Z}^{N}$ in $\mathrm{SL}_{N}(\mathbb{R})$, this proves the injectivity.
For every $\Lambda \in \mathrm{X}_{N}$, find a basis $v_{1}, \ldots, v_{N}$. Replacing $v_{1}$ by $-v_{1}$ if necessary, assume $M:=$ $\left(v_{1}, \ldots, v_{N}\right)\left(v_{i}\right.$ written as column vectors) has determinant 1 . Then $M . \mathbb{Z}^{N}=\Lambda$. This proves the surjectivity.

We leave it to the reader to convince himself/herself that the map is open and continuous.

Definition 2.8. For a discrete subgroup $\Lambda \leq \mathbb{R}^{N}$ we define

$$
\begin{equation*}
\operatorname{sys}(\Lambda):=\inf _{\nu \neq 0} \in \Lambda \nu v \| \tag{8}
\end{equation*}
$$

where $\|\cdot\|$ is the standard Euclidean norm.
Clearly sys $(\Lambda)>0$.
You may interpret sys( $\Lambda$ ) as the length of the smallest geodesic in the quotient flat torus $\mathbb{R}^{N} / \Lambda$.

One can check that sys : $\mathrm{X}_{N} \rightarrow \mathbb{R}_{>0}$ is continuous.
The following is sometimes referred to as Mahler's criterion.
Lemma 2.9. $\quad$ 1. A set $\mathscr{B} \subset \mathrm{X}_{\mathrm{N}}$ does not have compact closure (we will simply write unbounded later) iffor every $\varepsilon>0$ there exists $\Lambda \in \mathscr{B}$ with $\operatorname{sys}(\Lambda) \leq \varepsilon$.
2. For every $\varepsilon>0$, the set

$$
\{\Lambda \mid \operatorname{sys}(\Lambda) \geq \varepsilon\}
$$

is compact in $\mathrm{X}_{N}$.
Definition 2.10. For a discrete subgroup $\Lambda$ of $\mathbb{R}^{N}$, we let $\|\Lambda\|:=\operatorname{Vol}(V / \Lambda)$ where $V$ is the $\mathbb{R}$-linear span of $\Lambda$. For a lattice $\Lambda$ of some Euclidean space $V$, we let $\|\Lambda\|_{V}:=\operatorname{Vol}(V / \Lambda)$.

As we have discussed, if $v_{1}, \ldots, v_{n}$ is a basis of $\Lambda$, then

$$
\|\Lambda\|=\left\|\nu_{1}\right\| \cdot \operatorname{dist}\left(v_{2}, \mathbb{R} . \nu_{1}\right) \cdot \ldots \cdot \operatorname{dist}\left(\nu_{n}, \mathbb{R} \nu_{1}+\ldots+\mathbb{R} v_{n-1}\right)
$$

Let us also remark that $\operatorname{dist}\left(\nu_{2}, \mathbb{R} \nu_{1}\right)=\left\|\nu_{2}\right\|_{\mathbb{R}^{N} / \mathbb{R} \nu_{1}}$ and more generally

$$
\operatorname{dist}\left(v_{k}, \mathbb{R} v_{1}+\ldots+\mathbb{R} v_{k-1}\right)=\left\|v_{k}\right\|_{\mathbb{R}^{N} /\left(\mathbb{R} \nu_{1}+\ldots+\mathbb{R} v_{k-1}\right)} .
$$

Proof of Lem.2.9. 1. follows from the continuity of sys. Let us prove 2.
Fix some $\varepsilon>0$ and take $\Lambda \in \mathrm{X}_{N}$ satisfying sys $(\Lambda) \geq \varepsilon$. It suffices to construct a basis of $\Lambda$ with bounded distance to the origin.

Consider the projection $p: \mathbb{R}^{N} \rightarrow \mathbb{R}^{N} / \Lambda$. As $\operatorname{Vol}\left(\mathbb{R}^{N} / \Lambda\right)=1, p$ restricted to the subset $[-1,1]^{N}$ is not injective. This shows that there exists $v_{\neq 0} \in \Lambda,\|v\| \leq C_{1}(N)$ for some positive constant $C_{1}(N)$ depending only on $N$. In particular, if we choose $\nu_{1} \in \Lambda$ such that

$$
\left\|\nu_{1}\right\|=\operatorname{sys}(\Lambda),
$$

then $\left\|\nu_{1}\right\| \leq C_{1}(N)$. Note that $\nu_{1}$ is primitive in the sense that $\nu_{1}$ is not an integral multiple of any vector in $\Lambda$ other than $\pm \nu_{1}$.

Let $\pi_{1}$ be the projection from $\mathbb{R}^{N}$ to $V_{1}:=\mathbb{R}^{N} / \mathbb{R} \nu_{1}$. Since $\Lambda_{1}:=\pi_{1}(\Lambda)$ has rank $N-1$ and spans $V_{1}$, we have that $\Lambda_{1}$ is discrete and actually a lattice in $V_{1}$.

Note that

$$
\begin{aligned}
1= & \|\Lambda\|=\left\|\nu_{1}\right\| \cdot \operatorname{dist}\left(\nu_{2}, \mathbb{R} \nu_{1}\right) \cdot \ldots \cdot \operatorname{dist}\left(\nu_{N}, \mathbb{R} \nu_{1}+\ldots+\mathbb{R} v_{N-1}\right) \\
= & \left\|\nu_{1}\right\| \cdot\left\|\pi_{1}\left(\nu_{2}\right)\right\|_{V_{1}} \cdot \operatorname{dist}\left(\pi_{1}\left(\nu_{3}\right), \mathbb{R} \pi_{1}\left(\nu_{2}\right)\right) \cdot \ldots \cdot \operatorname{dist}\left(\pi_{1}\left(\nu_{N}\right), \mathbb{R} \pi_{1}\left(\nu_{2}\right)+\ldots+\mathbb{R} \pi_{1}\left(v_{N-1}\right)\right) \\
= & \left\|\nu_{1}\right\| \cdot\left\|\Lambda_{1}\right\|_{V_{1}} \geq \varepsilon \cdot\left\|\Lambda_{1}\right\|_{V_{1}} \\
& \Longrightarrow\left\|\Lambda_{1}\right\|_{V_{1}} \leq \varepsilon^{-1}=: C_{2}(\varepsilon) .
\end{aligned}
$$

Now choose $\nu_{2} \in \Lambda \backslash \mathbb{R} \nu_{1}$ such that

$$
\left\|\pi_{1}\left(\nu_{2}\right)\right\|=\operatorname{sys}_{V_{1}}\left(\Lambda_{1}\right) .
$$

A similar argument as above shows that $\left\|\pi_{1}\left(\nu_{2}\right)\right\|<C_{3}(N, \varepsilon)$. By modifying $\nu_{2}$ by some integral multiple of $\nu_{1}$, we assume that $\left\|\nu_{2}\right\|<C_{3}(N, \varepsilon)=C_{3}$ with a possibly larger $C_{3}$.

Next we want to argue that $\operatorname{sys}_{V_{1}}\left(\Lambda_{1}\right)>c_{1}(N, \varepsilon)$ for some constant $c_{1}(N, \varepsilon)>0$ (we will soon see that can take $c_{1}=0.4 \varepsilon$ ) depending only on $N, \varepsilon$.

Say we have a nonzero vector in $\Lambda_{1}$ of length smaller than $\lambda$. Then its lift $v \in \Lambda$ has the property $0<\operatorname{dist}\left(\nu, \mathbb{R} \nu_{1}\right)<\lambda$. So if we write $v=x . \nu_{1}+w$ for some $w$ orthogonal to $\nu_{1}$ then $\|w\| \leq \lambda$. Let $n_{x}$ be the nearest integer to $x$, then $\nu^{\prime}:=\left(x-n_{x}\right) \nu_{1}+w \in \Lambda$ has norm

$$
\left\|v^{\prime}\right\| \leq\left|x-n_{x}\right|\left\|\nu_{1}\right\|+\lambda \leq 0.5\left\|\nu_{1}\right\|+\lambda
$$

So if we had chosen $\lambda=0.4 \varepsilon \leq 0.4 \operatorname{sys}(\Lambda)$, then $\left\|v^{\prime}\right\| \leq 0.9$ sys $(\Lambda)$, which is a contradiction. Thus every non-zero vector in $\Lambda_{1}$ has length greater than $0.4 \varepsilon$.

Let $\pi_{2}$ be the natural projection $\mathbb{R}^{N} \rightarrow \mathbb{R}^{N} /\left(\mathbb{R} \nu_{1}+\mathbb{R} \nu_{2}\right)=: V_{2}$. By abuse of notation, also denote the natural projection $V_{1} \rightarrow V_{2}$ by $\pi_{2}$.

With similar arguments, $\Lambda_{2}:=\pi_{2}\left(\Lambda_{1}\right)$ is a lattice in $V_{2}$ and

$$
\left\|\Lambda_{1}\right\|_{V_{1}}=\left\|\pi_{1}\left(v_{2}\right)\right\| \cdot\left\|\Lambda_{2}\right\|_{V_{2}} \Longrightarrow\left\|\Lambda_{2}\right\|_{V_{2}} \leq c_{1}^{-1} \cdot C_{2}=: C_{4}(N, \varepsilon)=: C_{4} .
$$

With similar arguments, $\operatorname{sys}_{V_{2}}\left(\Lambda_{2}\right)>c_{2}(N, \varepsilon)$. So we can find $\nu_{3}, \ldots$ up to $v_{N}$ with bounded norms. And one can check that each step you get a primitive subgroup of $\Lambda$ and $\left\{v_{1}, \ldots, v_{N}\right\}$ forms a basis of $\Lambda$. So we are done.

Definition 2.11. A primitive subgroup of $\Lambda$ is a subgroup $\Delta$ such that the $\mathbb{Q}$-span (or equivalently, the $\mathbb{R}$-span) of $\Delta$ intersecting with $\Lambda$ gives back $\Delta$.

The $\mathbb{Z}$-span of two primitive subgroups may not be primitive. e.g., consider $(1,1),(1,-1)$ in $\mathbb{Z}^{2}$, each of which is primitive, but they span a index 2 subgroup of $\mathbb{Z}^{2}$, hence not primitive.

## 3. Values of a quadratic form and orbits of its symmetric group

Now comes the equivalent formulation of weak Oppenheim. For a rational quadratic form $Q$, this would imply that $\mathrm{SO}_{Q}(\mathbb{Z})$ is not cocompact in $\mathrm{SO}_{Q}(\mathbb{R})$ if $Q(\nu)=0$ admits a solution in $v_{\neq 0} \in \mathbb{Z}^{N}$ (in which case we say $Q$ is isotropic over $\mathbb{Q}$ ). When $N \geq 5$, a rational indefinite quadratic form is always isotropic over $\mathbb{Q}$ (see [0'M63, 63:19, 66:1])

Lemma 3.1. For a non-degenerate quadratic form $Q$ in $N$ variables with real coefficients, the following two are equivalent:

1. the closure of $Q\left(\mathbb{Z}^{N} \backslash 0\right)$ contains 0 ;
2. the orbit closure of $\mathrm{SO}_{Q}(\mathbb{R})$ based at the identity coset is unbounded in $\mathrm{X}_{\mathrm{N}}$. In other words, $\mathrm{SO}_{Q}(\mathbb{R}) . \mathbb{Z}^{N}$ contains non-zero vectors of arbitrarily small length.

Proof of $2 \Longrightarrow 1$. By assumption and Mahler's criterion, there exists $g_{n} \in \mathrm{SO}_{Q}(\mathbb{R})$ and $u_{n}(\neq 0) \in \mathbb{Z}^{N}$ such that $g_{n} \cdot u_{n}$ tends to $\mathbf{0}$. Hence

$$
Q\left(u_{n}\right)=Q\left(g_{n} \cdot u_{n}\right) \rightarrow 0 .
$$

And we are done.
Remark 3.2. For the proof of Thm.1.1 this direction is sufficient. However we feel that it is conceptually better to do the converse, too. Actually, this provides a different way of understanding why Thm.1.1 fails when $N=2$ - it suffices to find a bounded, yet non-closed orbit of the diagonal group $A$ on $\mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$. And one can do this by constructing two closed orbits of A and a third orbit Ay such that in the forward direction, Ay approximates one closed orbit and in the backward direction Ax approximates the other. This relies on the fact that closed A-orbits are dense (for instance, one can find one by explicit construction and then consider all lattices commensurable to it) and an argument with local coordinates in stable/unstable/flow direction.

Why is this sufficient? Note that if $Q$ is an indefinite rational quadratic form in two variable, then either $Q$ is $\mathbb{Q}$-equivalent to $Q_{0}=x y$ or $Q_{1}=x^{2}-b y^{2}$ for some $b>0$ and $\sqrt{b} \notin \mathbb{Q}$. In the former case, the orbit of $\mathrm{SO}_{Q}(\mathbb{R})$ based at the identity coset diverges (that is, the orbit map is proper) and in the second case the orbit is compact, stabilizer of which comes from certain elements in $\mathbb{Q}(\sqrt{b})$.

Now go back to the proof of $1 \Longrightarrow 2$ of Lem.3.1. We need the following fact.
Lemma 3.3. For every $r_{\neq 0} \in \mathbb{R}, \mathrm{SO}_{Q}(\mathbb{R})$ acts transitively on the level set

$$
V_{r}:=\left\{v \in \mathbb{R}^{N} \mid Q(v)=r\right\} .
$$

And for $r=0$, there are at most 2 -orbits of $\mathrm{SO}_{Q}(\mathbb{R})$ on $V_{0} \backslash\{\mathbf{0}\}$.
Proof. By linear algebra, up to change of $\mathbb{R}$-coordinate (i.e. up to $\mathrm{GL}_{N}(\mathbb{R})$ ), we may and do assume that $Q$ takes the form

$$
Q\left(x_{1}, \ldots, x_{N}\right)=\left(x_{1}^{2}+\ldots+x_{s}^{2}\right)-\left(x_{s+1}^{2}+\ldots+x_{s+t}^{2}\right)=: Q_{1}\left(x_{1}, \ldots, x_{s}\right)-Q_{2}\left(x_{s+1}, \ldots, x_{s+t}\right)
$$

for some $s+t=N$ and $s, t \in \mathbb{Z}_{\geq 0}$. When one of $s, t$ is equal to 0 , the form is definite and we assume that we already know how to handle this case.

For $\boldsymbol{x} \in \mathbb{R}^{N}$, we write $v_{x}:=\left(x_{1}, \ldots, x_{s}\right)$ and $w_{x}:=\left(x_{s+1}, \ldots, x_{N}\right)$.
Now we fix $r_{0}$ and if $V_{r_{0}}$ is empty there is nothing to prove. So assume otherwise and take $\boldsymbol{x}_{0} \in V_{r_{0}}$. Let $r_{1}:=Q_{1}\left(v_{x_{0}}\right)$ and $r_{2}:=Q_{2}\left(w_{x_{0}}\right)$. Thus by transitivity in the (positive) definite case, we can find $k_{i} \in \mathrm{SO}_{Q_{i}}(\mathbb{R})(\mathrm{i}=1,2)$ such that

$$
\begin{aligned}
k_{1} \cdot v_{x_{0}} & =\left(\sqrt{r_{1}}, 0, \ldots, 0\right) \\
k_{2} \cdot w_{x_{0}} & =\left(\sqrt{r_{2}}, 0, \ldots, 0\right) .
\end{aligned}
$$

Let $\mathrm{SO}_{\left(x_{1}^{2}-x_{s+1}^{2}\right)}(\mathbb{R})$ be embedded in $\mathrm{SO}_{Q}(\mathbb{R})$ by leaving the rest of the coordinates unchanged. When $r=r_{1}-r_{2} \neq 0$, it acts on $x_{1}^{2}-x_{s+1}^{2}=r$ transitively. The level sets are not connected, but the group $\mathrm{SO}_{(x y)}(\mathbb{R})$ is also not! Both have 2 components.

As for the case $r=0$, one can show that $\mathrm{SO}_{\left(x_{1}^{2}-x_{s+1}^{2}\right)}(\mathbb{R})$ has two orbits on $x_{1}^{2}-x_{s+1}^{2}=0$.
Here is an illustration of the proof by pictures


Remark 3.4. Assume $Q$ is indefinite non-degenerate. One can further show that when $N \geq 3, \mathrm{SO}_{Q}(\mathbb{R})$ acts on $V_{0} \backslash \mathbf{0}$ transitively. When $N=2, \mathrm{SO}_{Q}(\mathbb{R})$ acts on $V_{0} \backslash \mathbf{0}$ with exact two orbits.

Proof of $1 \Longrightarrow 2$. By assumption for every $\varepsilon>0$ there exists $u_{\varepsilon \neq 0} \in \mathbb{Z}^{N}$ such that $\left|Q\left(u_{\varepsilon}\right)\right| \leq$ $\varepsilon$. On the other hand, By the (proof of) Lem.3.3, there exists non-zero $u_{\varepsilon}^{1}, u_{\varepsilon}^{2} \in \mathrm{SO}_{Q}(\mathbb{R}) . u_{\varepsilon}$ such that

$$
\left\|u_{\varepsilon}^{1}\right\|,\left\|u_{\varepsilon}^{2}\right\| \leq \theta(Q, \varepsilon)=\theta
$$

where $\theta$ tends to 0 (for a fix $Q$ ) as $\varepsilon$ does so. Hence $\operatorname{sys}\left(g_{\varepsilon} \mathbb{Z}^{N}\right) \leq \theta$ and we see that $\mathrm{SO}_{Q}(\mathbb{R}) \cdot \mathbb{Z}^{N}$ is unbounded as $\varepsilon \rightarrow 0$ by Lem.2.9.

Now we specialize to $N=3$.
In light of Lem.3.1, to prove Thm.1.1, it is sufficient to show that $\mathrm{SO}_{Q}(\mathbb{R}) \cdot \mathbb{Z}^{3}$ is unbounded. Find $g_{0} \in \mathrm{SL}_{3}(\mathbb{R})$ such that $Q \circ g_{0}^{-1}$ is a scalar multiple of $Q_{0}=2 x_{1} x_{3}-x_{2}^{2}$. Then

$$
\mathrm{SO}_{Q_{0}}=g_{0} \mathrm{SO}_{Q} g_{0}^{-1} .
$$

So sufficient to show that $\mathrm{SO}_{Q_{0}}(\mathbb{R}) g_{0} \cdot \mathbb{Z}^{3}$ is unbounded in $\mathrm{X}_{3}$, which will follow from
Theorem 3.5. Let $\Lambda \in \mathrm{X}_{3}$ be such that $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ is bounded, then $\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R}) . \Lambda$ is closed, and hence compact.

In some sense we cheated a little bit. Because we are going to use a trick on quadratic forms. And the true dynamical result we are going to prove is (to be proved in the next lecture):

Theorem 3.6. Let $\Lambda \in \mathrm{X}_{3}$ be such that $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ is bounded, then either $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ is closed and hence compact, or the closure of $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ contains a $\left\{\mathbf{v}_{s}\right\}_{s \geq 0}$-orbit or a $\left\{\mathbf{v}_{s}\right\}_{s \leq 0}$ orbit. where

$$
\mathbf{v}_{s}:=\exp \left(s \cdot\left[\begin{array}{lll}
0 & 0 & 1 \\
& 0 & 0 \\
& & 0
\end{array}\right]\right)=\left[\begin{array}{lll}
1 & 0 & s \\
& 1 & 0 \\
& & 1
\end{array}\right] .
$$

Note that $\left\{\mathbf{v}_{s}\right\}$ is not contained in $\mathrm{SO}_{Q_{0}}(\mathbb{R})$.
Proof of Thm.3.5 assuming Thm.3.6. Say, we have a $\left\{\mathbf{v}_{s}\right\}_{s \geq 0}$-orbit (the other case is similar) based at $\Lambda^{\prime}$ for some $\Lambda^{\prime} \in \overline{\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda}$. Write $\mathbf{x}=\left(x_{1}, x_{2}, x_{3}\right) \in \Lambda^{\prime}$. Then

$$
Q_{0}\left(\mathbf{v}_{s} \cdot \mathbf{x}\right)=Q_{0}\left(x_{1}+s x_{3}, x_{2}, x_{3}\right)=\left(2 x_{3}^{2}\right) s+\left(2 x_{1} x_{3}-x_{2}^{2}\right) .
$$

First we can find some $\mathbf{x} \in \Lambda^{\prime}$ such that $Q_{0}(\mathbf{x})<0$ and $x_{3} \neq 0$ (I leave it to you to convince yourself that this is possible). Then there is some $s$ (replace $x_{1}$ by $-x_{1}$ if necessary) with $Q_{0}\left(\mathbf{v}_{s}\right.$. $\mathbf{x})=0$. By Lem.3.1, this implies $\mathrm{SO}_{Q_{0}}(\mathbb{R}) \mathbf{v}_{s} \cdot \Lambda \subset \overline{\mathrm{SO}_{Q_{0}}(\mathbb{R}) \cdot \Lambda}$ is unbounded.

Proof of Thm.1.1 assuming Thm.3.6. To prove Thm.1.1, by Lem.3.1, if $\mathrm{SO}_{Q_{0}}(\mathbb{R}) g_{0} . \mathbb{Z}^{3}$ is unbounded in $X_{3}$ then we are done. Now we assume otherwise. If $\mathrm{SO}_{Q_{0}}(\mathbb{R}) g_{0} \cdot \mathbb{Z}^{3}$ is compact, or equivalently, $\mathrm{SO}_{Q}(\mathbb{R}) . \mathbb{Z}^{3}$ is compact, then by Lem.3.7 below, $Q$ is proportional to a rational quadratic form, contradiction. Thus we have a $\left\{\mathbf{v}_{s}\right\}_{s \geq 0}$ (the other case $s \leq 0$ is similar) orbit in the closure of $\mathrm{SO}_{Q_{0}}(\mathbb{R}) g_{0} \cdot \mathbb{Z}^{3}$. Repeat the argument above, we find $s \in \mathbb{R}$ such that $Q_{0}\left(\mathbf{v}_{s} \cdot x\right)=0$ for some $x \in g_{0} \mathbb{Z}^{3}$. But $\mathbf{v}_{s} . g_{0} \mathbb{Z}^{3}$ is in the closure of $\mathrm{SO}_{Q_{0}}(\mathbb{R}) g_{0} \cdot \mathbb{Z}^{3}$, implying that we can find $\left(v_{n}\right) \subset g_{0} \mathbb{Z}^{3},\left(g_{n}\right) \subset \mathrm{SO}_{Q_{0}}(\mathbb{R})$ such that $g_{n} . v_{n} \rightarrow \mathbf{v}_{s} . x$. Hence

$$
Q_{0}\left(v_{n}\right)=Q_{0}\left(g_{n} v_{n}\right) \rightarrow Q_{0}\left(\mathbf{v}_{s} \cdot x\right)=0 .
$$

Thus the closure of $Q\left(\mathbb{Z}^{3}\right)=Q_{0}\left(g_{0} \cdot \mathbb{Z}^{3}\right)$ contains 0 .
LEmma 3.7. For a non-degenerate quadratic form $Q$, if $\mathrm{SO}_{Q}(\mathbb{Z})$ is cocompact in $\mathrm{SO}_{Q}(\mathbb{R})$, then $Q$ is a multiple of a rational quadratic form.

Note that if $Q$ is NOT a multiple of a rational quadratic form, then for some non-zero coefficients $\alpha, \beta$ of $Q$, one has $\alpha / \beta \notin \mathbb{Q}$. Hence there exists $\sigma \in \operatorname{Aut}(\mathbb{R} / \mathbb{Q})$ such that $\sigma(\alpha / \beta) \neq$ $\alpha / \beta$, in particular, $\sigma Q$ is not proportional to $Q$.

So it suffices to complete
Step 1. for every $\sigma \in \operatorname{Aut}(\mathbb{R} / \mathbb{Q})$, show $\mathrm{SO}_{Q}(\mathbb{R})^{\circ}=\mathrm{SO}_{\sigma(Q)}(\mathbb{R})^{\circ}$;
Step 2. for every pair $Q_{1}, Q_{2}$ of non-degenerate quadratic forms of the same rank, show $\mathrm{SO}_{\mathrm{Q}_{1}}(\mathbb{R})^{\circ}=\mathrm{SO}_{\mathrm{Q}_{2}}(\mathbb{R})^{\circ} \Longrightarrow Q_{1}=\lambda Q_{2}$ for some $\lambda \in \mathbb{R}_{\neq 0}$.

Step 1. First note that

$$
\mathrm{SO}_{\sigma(Q)}(\mathbb{R})=\sigma\left(\mathrm{SO}_{Q}(\mathbb{R})\right) \supset \mathrm{SO}_{Q}(\mathbb{Z})
$$

Consider the linear representation

$$
\mathrm{SL}_{3}(\mathbb{R}) \curvearrowright \text { Sym }:=\{\mathbb{R}-\text { Symmetric matrices }\}, \quad g \cdot M:=g M g^{t r},
$$

and the map (call it $\phi$ ) $g \mapsto g \cdot \sigma(Q)$ from $\mathrm{SO}_{Q}(\mathbb{R})$ to Sym. Then $\phi$ factors through

$$
\mathrm{SO}_{Q}(\mathbb{R}) / \mathrm{SO}_{Q}(\mathbb{Z}) \rightarrow \mathrm{Sym}
$$

and hence has compact (and bounded) image. Now we need two facts

1. $\mathrm{SO}_{Q}(\mathbb{R})^{\circ}$ is generated (as closed subgroup, which follows by a Lie algebra calculation) by one-parameter unipotent flows $\left\{\mathbf{u}_{t}:=\exp \mathbf{u} t\right\}_{t \in \mathbb{R}}$ (u is some nilpotent matrix in $\mathfrak{s o}_{Q}(\mathbb{R})$;
2. For every unipotent flow $\left\{\mathbf{u}_{t}\right\}$ and $M \in \operatorname{Sym}$, either $\left\{\mathbf{u}_{t} \cdot M\right\}$ is unbounded or $M$ is fixed by $\left\{\mathbf{u}_{t}\right\}$. (if you do not believe this, do some explicit calculation with upper triangular unipotent flows)
But we already saw that $\mathrm{SO}_{Q}(\mathbb{R}) . \sigma(Q)$ is bounded, thus $\mathrm{SO}_{Q}(\mathbb{R})^{\circ}$ fixes $\sigma(Q)$. So $\mathrm{SO}_{Q}(\mathbb{R})^{\circ}$ is contained in $\mathrm{SO}_{\sigma(Q)}(\mathbb{R})$. But they are both Lie subgroups of $\mathrm{SL}_{3}(\mathbb{R})$ of the same dimension, so we must have

$$
\mathrm{SO}_{Q}(\mathbb{R})^{\circ}=\mathrm{SO}_{\sigma(Q)}(\mathbb{R})^{\circ} .
$$

STEP 2. By conjugation we assume $Q_{1}=Q_{0}=2 x_{1} x_{3}-2 x_{2}^{2}$. One can compute that $\mathfrak{s o}_{Q_{0}}(\mathbb{R})$ contains (and is generated by)

$$
\left[\begin{array}{lll}
1 & & \\
& 0 & \\
& & -1
\end{array}\right],\left[\begin{array}{lll}
0 & 1 & 0 \\
& 0 & 1 \\
& & 0
\end{array}\right],\left[\begin{array}{lll}
0 & & \\
1 & 0 & \\
0 & 1 & 0
\end{array}\right]
$$

(note that they do not form an $\mathfrak{s l}_{2}$-triple, you should multiply the first and the second, but not the third, by 2) and hence $\mathrm{SO}_{Q_{0}}(\mathbb{R})$ contains

$$
a_{t}:=\left[\begin{array}{lll}
e^{t} & &  \tag{9}\\
& 1 & \\
& & e^{-t}
\end{array}\right], \quad u_{s}:=\left[\begin{array}{ccc}
1 & s & s^{2} / 2 \\
& 1 & s \\
& & 1
\end{array}\right], \quad u_{s}^{-}:=\left[\begin{array}{ccc}
1 & & \\
s & 1 & \\
s^{2} / 2 & s & 1
\end{array}\right] .
$$

Then a direct computation (at the level of Lie algebra is perhaps easier) shows that in order for $\mathfrak{s o}_{Q_{2}}(\mathbb{R})$ to contain these elements, $Q_{2}$ must be a scalar multiple of $Q_{1}$ and we are done.

## 4. Exercises

4.1. Non-commensurable lattices in $\mathrm{SL}_{2}(\mathbb{R})$, I. We apply ideas in Lec. 4 to a different example. Our ultimate goal is to show that two cocompact lattices in $\mathrm{SL}_{2}(\mathbb{R})$ is either commensurable or their product is dense in $\mathrm{SL}_{2}(\mathbb{R})$, which will (hopefully) be achieved in the next set of exercises.

Notations:

- $\mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R}) \times \mathrm{SL}_{2}(\mathbb{R}), \mathrm{H}:=\Delta\left(\mathrm{SL}_{2}(\mathbb{R})\right)$ and $\Gamma$ is a cocompact lattice in G ;
- $\mathfrak{g}:=\operatorname{Lie}(\mathrm{G})$ and $\mathfrak{h}:=\operatorname{Lie}(\mathrm{H})$;
- $A:=\left\{\left(\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right],\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right]\right), t \in \mathbb{R}\right\}=\left\{\Delta \mathbf{a}_{t}, t \in \mathbb{R}\right\}$;
- $U:=\left\{\left(\left[\begin{array}{cc}1 & t \\ 0 & 1\end{array}\right],\left[\begin{array}{cc}1 & t \\ 0 & 1\end{array}\right]\right), t \in \mathbb{R}\right\}=\left\{\Delta \mathbf{u}_{t}, t \in \mathbb{R}\right\} ;$
- $V:=\left\{\left(\left[\begin{array}{ll}1 & t \\ 0 & 1\end{array}\right],\left[\begin{array}{cc}1 & -t \\ 0 & 1\end{array}\right]\right), t \in \mathbb{R}\right\}=\left\{\mathbf{v}_{t}, t \in \mathbb{R}\right\} ;$
- $V^{+}:=\left\{\mathbf{v}_{t}, t \geq 0\right\}, V^{-}:=\left\{\mathbf{v}_{t}, t \leq 0\right\}$;
- $W:=A U V, W^{+}:=A U V^{+}, W^{-}:=A U V^{-}$.

EXERCISE 4.1. Show that $W$ is a group and $W^{+}, W^{-}$are semigroups.
EXERCISE 4.2. Let

$$
\mathfrak{h}^{\perp}:=\left\{(X,-X) \mid X \in \mathfrak{s l}_{2}(\mathbb{R})\right\} \subset \mathfrak{g} .
$$

Show that $\mathfrak{g}=\mathfrak{h} \oplus \mathfrak{h}^{\perp}$ and this decomposition is preserved by $\operatorname{Ad}(\mathrm{H})$.
Now take $\Lambda_{0} \in G / \Gamma$ such that H. $\Lambda_{0}$ is not closed. Define $Y_{0}:=\overline{\mathrm{H} . \Lambda_{0}}$ and

$$
\mathscr{O}:=\left\{y \in \mathrm{Y}_{0} \mid \text { H. } y \text { is open in } \mathrm{Y}_{0}\right\} .
$$

ExErcise 4.3. Show that $\mathscr{O} \neq \mathrm{Y}_{0}$.
Let $\mathrm{Y}_{1}$ be a nonempty $U$-minimal set in $\mathrm{Y}_{0} \backslash \mathscr{O}$.
EXERCISE 4.4. Show that $\mathrm{Y}_{1}$ is not a closed $U$-orbit.
EXERCISE 4.5. Assume $\mathrm{Y}_{1}$ is not preserved by $A$. Show that $\mathrm{Y}_{0}$ contains $a W$-orbit.
(Hint: consider $\operatorname{Aut}\left(\mathrm{Y}_{1}\right)$.)
EXERCISE 4.6. Assume $\mathrm{Y}_{1}$ is preserved by $A$. Show that $\mathrm{Y}_{0}$ contains a $W^{+}$-orbit or a $W^{-}-$ orbit.
(Hint: consider $\operatorname{Map}\left(\mathrm{Y}_{0}, \mathrm{Y}_{1}\right)$.)
4.2. Totally geodesic hyperbolic planes in $\mathbb{H}^{3}$, I. We apply ideas in Lec. 4 to yet another example. Our ultimate goal (hopefully achieved in the next set of exercises) is to show that the image of a totally geodesic immersion of a hyperbolic plane in a closed hyperbolic three manifold is either closed or dense.

Notations:

- $\mathrm{G}:=\mathrm{SL}_{2}(\mathbb{C}), \mathrm{H}:=\mathrm{SL}_{2}(\mathbb{R})$ and $\Gamma$ is a cocompact lattice in G ;
- $\mathfrak{g}:=\operatorname{Lie}(\mathrm{G})$ and $\mathfrak{h}:=\operatorname{Lie}(\mathrm{H})$;
- $A:=\left\{\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right], t \in \mathbb{R}\right\}=\left\{\mathbf{a}_{t}, t \in \mathbb{R}\right\} ;$
- $U:=\left\{\left[\begin{array}{cc}1 & t \\ 0 & 1\end{array}\right], t \in \mathbb{R}\right\}=\left\{\mathbf{u}_{t}, t \in \mathbb{R}\right\} ;$
- $V:=\left\{\left[\begin{array}{cc}1 & \text { it } \\ 0 & 1\end{array}\right], t \in \mathbb{R}\right\}=\left\{\mathbf{v}_{t}, t \in \mathbb{R}\right\} ;$
- $V^{+}:=\left\{\mathbf{v}_{t}, t \geq 0\right\}, V^{-}:=\left\{\mathbf{v}_{t}, t \leq 0\right\} ;$
- $W:=A U V, W^{+}:=A U V^{+}, W^{-}:=A U V^{-}$;

EXERCISE 4.7. Let $\mathfrak{h}^{\perp}:=\left\{i \cdot X, X \in \mathfrak{s l}_{2}(\mathbb{R})\right\}$. Show that $\mathfrak{g}=\mathfrak{h} \oplus \mathfrak{h}^{\perp}$. Moreover, this decomposition is preserved by the $\operatorname{Ad}(\mathrm{H})$-action.

Exercise 4.8. Let $\mathrm{H} . \Lambda_{0}$ be a non-closed H -orbit in $\mathrm{G} / \Gamma$. Show that $\mathrm{Y}_{0}:=\overline{\mathrm{H} . \Lambda_{0}}$ contains a $W^{+}$or a $W^{-}$-orbit.

## CHAPTER 3

## Orbit closure of orthogonal groups in the space of lattices

Back to the Top.
Notations and assumptions.

- $Q_{0}\left(x_{1}, x_{2}, x_{3}\right)=2 x_{1} x_{3}-x_{2}^{2}$;
- $\mathrm{H}:=\mathrm{SO}_{Q_{0}}(\mathbb{R}) \leq \mathrm{G}:=\mathrm{SL}_{3}(\mathbb{R}), \mathrm{X}_{3} \cong \mathrm{SL}_{3}(\mathbb{R}) / \mathrm{SL}_{3}(\mathbb{Z})$;
- $\mathrm{A}:=\left\{\mathbf{a}_{t}: \left.=\left[\begin{array}{ccc}e^{t} & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\}, \mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ccc}1 & s & s^{2} \\ 0 & 1 & s \\ 0 & 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\} ;$
- $\mathrm{V}:=\left\{\mathbf{v}_{s}: \left.=\left[\begin{array}{ccc}1 & 0 & s \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\} ;$
- $u_{0}:=\left[\begin{array}{lll}0 & 1 & 0 \\ 0 & 0 & 1 \\ 0 & 0 & 0\end{array}\right], v_{0}:=\left[\begin{array}{ccc}0 & 0 & 1 \\ 0 & 0 & 0 \\ 0 & 0 & 0\end{array}\right]$;
- $\mathfrak{h}$ is the Lie algebra of $H$ and $\mathfrak{h}^{\perp}$ denotes its orthogonal complement, see Sec.3.1 for explicit calculations;
- Fix some $x_{0} \in \mathrm{X}_{3}$ with H. $x_{0}$ being bounded and non-closed. Write $\mathrm{Y}_{0}$ for the closure of H. $x_{0}$.
See last lecture for the precise definition of $\mathfrak{h}^{\perp}$.


## 1. Overview

In this lecture we prove Theorem 3.6 from Chapter 2.
Theorem 1.1. Let $\Lambda \in \mathrm{X}_{3}$ be such that $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ is bounded, then either $\mathrm{SO}_{Q_{0}}(\mathbb{R}) . \Lambda$ is closed and hence compact, or the closure ofSO $Q_{Q_{0}}(\mathbb{R}) . \Lambda$ contains $a\left\{\mathbf{v}_{s}\right\}_{s \geq 0}$-orbit or a $\left\{\mathbf{v}_{s}\right\}_{s \leq 0}$-orbit.

By comparison, the ultimate knowledge regarding this is:
Theorem 1.2. Every $\mathrm{SO}_{Q_{0}}(\mathbb{R})$-orbit in $\mathrm{X}_{3}$ is either closed or dense.
Outline of proof. Recall that from Chapter 1, we wish to obtain something nontrivial that preserve $\mathrm{Y}_{0}$ (or a minimal subset of $\mathrm{Y}_{0}$ ) in the direction of the normalizer. The normalizer of H is basically H itself. We restrict our attention to an one-parameter unipotent subgroup U of $H$, then we can start to apply the same argument as in Ch.1. Take a U-minimal subset $\mathrm{Y}_{1}$ of $\mathrm{Y}_{0}$. The possibility of having a closed $U$ orbit is excluded. Then by arguments as in Ch.1, we would have some additional elements preserving $\mathrm{Y}_{1}$ in the normalizer of $U$. Under the current situation, two possibilities for these additional invariants exist. They could be contained in A or V (or in between). So for our purpose we may assume that the U-minimal set is A-stable. And to treat this case, instead of considering those preserving $Y_{1}$, we consider the set of $g \in G$ mapping $\mathrm{Y}_{0}$ to $\mathrm{Y}_{1}$. The lack of group structure here would cause us some difficulty, taken care of by $A$.

## 2. The proof

In this section we prove Thm.1.1.
Consider the following

$$
\mathscr{O}:=\left\{y \in \mathrm{Y}_{0} \mid \mathrm{H} \cdot y \text { is open in } \mathrm{Y}_{0}\right\}
$$

Thus $\mathscr{O}$ is an H -invariant open (possibly empty) subset of $\mathrm{Y}_{0}$, in other words, $\mathrm{Y}_{0} \backslash \mathscr{O}$ is an H invariant compact set.

Lemma 2.1. $\mathscr{O} \neq \mathrm{Y}_{0}$ unless $\mathrm{Y}_{0}=\mathrm{H} \cdot x_{0}$.
Proof. Otherwise each H-orbit is open, and hence closed in $\mathrm{Y}_{0}$. In particular H. $x_{0}$ is closed. But $\mathrm{Y}_{0}$ is not closed, so here is a contradiction.

Eventually we would know $\mathscr{O}$ is empty, but this is what we can do at the moment.
Now take $\mathrm{Y}_{1}$ to be a nonempty U -minimal set in $\mathrm{Y} \backslash \mathscr{O}$. There are three cases to consider
Case 1. $\mathrm{Y}_{1}$ is a compact U -orbit;
Case 2. $\mathrm{Y}_{1}$ does not fall in case 1 and $\mathrm{Y}_{1}$ is A-invariant ;
Case 3. $\mathrm{Y}_{1}$ does not fall in case 1 or 2 .
Actually, case 1 is not an option since $Y_{1}$ is bounded and hence has a lower bound on injectivity radius. So we are left with case 2 and 3 .

Before we proceed, let us note that given $x \in \mathrm{X}_{3}$, for $y$ close enough to $x$, there exists unique small $h_{y} \in \mathfrak{h}$ and small $w_{y} \in \mathfrak{h}^{\perp}$ such that

$$
y=\exp \left(h_{y}\right) \exp \left(w_{y}\right) \cdot x
$$

2.1. Case 2. In this case, we are going to show that

$$
\mathbf{v}_{s \geq 0} \text { or } \mathbf{v}_{s \leq 0} \subset\left\{g \in \mathrm{H} \mid g \mathrm{Y}_{1} \subset \mathrm{Y}_{0}\right\},
$$

which implies the conclusion of Thm.1.1.
Since $\mathrm{Y}_{1} \subset \mathrm{Y}_{0} \backslash \mathscr{O}$ and by the definition of $\mathscr{O}$, for every $x \in \mathrm{Y}_{1}$, there exists $y_{n} \rightarrow x$ with $y_{n} \in \mathrm{Y}_{0}$ and

$$
y_{n}=\exp \left(h_{n}\right) \exp \left(w_{n}\right) x
$$

where $h_{n} \in \mathfrak{h}, w_{n} \in \mathfrak{h}^{\perp}$ both converging to 0 and $w_{n} \neq 0$. Replacing $y_{n}$ by $\exp \left(-h_{n}\right) y_{n}$ we assume $h_{n}=0$. The case when $w_{n}$ belongs to Lie $(\mathrm{V})$ for infinitely many $n$ 's is easier and we assume this is not the case.

Lemma 2.2. Assume $w_{n}$ does not belong to Lie(V) for $n$ large enough. For $\delta>0$ small enough and $n$ large enough, there exists $t_{n, \delta}$ such that

1. $\left\|\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right\| \in\left[\frac{\delta}{10^{10}}, 10^{10} \delta\right]$;
2. every limit point of $\left(\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right)$ lies in $\operatorname{Lie}(V)$.

See Sec.3.4 for the proof. Now assume the lemma and choose $t_{n, \delta}$ as above. Define

$$
x_{n, \delta}:=\mathbf{u}_{t_{n, \delta}} \cdot x_{0}, \quad y_{n, \delta}:=\mathbf{u}_{t_{n, \delta}} \cdot y_{n},
$$

then

$$
y_{n, \delta}=\exp \left(\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right) \cdot x_{n, \delta} .
$$

By passing to a subsequence depending on $\delta$, we assume $\lim x_{n, \delta}=x_{\infty, \delta} \in \mathrm{Y}_{1}, \lim y_{n, \delta}=y_{\infty, \delta} \in$ $\mathrm{Y}_{0}$ and $\lim \operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}=s_{\delta} v_{0}$ for some $\left|s_{\delta}\right| \in\left[\frac{\delta}{10^{10}}, 10^{10} \delta\right]$. Hence

$$
y_{\infty, \delta}=\exp \left(s_{\delta} v_{0}\right) \cdot x_{\infty, \delta}=\mathbf{v}_{s_{\delta}} \cdot x_{\infty, \delta}
$$

with $s_{\delta} \rightarrow 0$ as $\delta$ does so. Thus

$$
\begin{aligned}
& \mathrm{Y}_{0} \supset \mathrm{U} \mathbf{v}_{s_{\delta}} \cdot x_{\infty, \delta}=\mathbf{v}_{s_{\delta}} \mathrm{U} \cdot x_{\infty, \delta} \\
& \Longrightarrow \mathrm{Y}_{0} \supset \overline{\mathbf{v}_{s_{\delta}} \mathrm{U} \cdot x_{\infty, \delta}}=\mathbf{v}_{s_{\delta}} \mathrm{Y}_{1} .
\end{aligned}
$$

The closed set

$$
\left\{g \in \mathrm{H} \mid g \mathrm{Y}_{1} \subset \mathrm{Y}_{0}\right\}
$$

is not necessarily a group. Hence we can not conclude the existence of a $\mathbf{v}_{\mathbb{R}}$ (or half of it) orbit inside $\mathrm{Y}_{0}$ immediately. This is where the assumption that $\mathrm{Y}_{1}$ is A-invariant steps in. Indeed,

$$
\mathbf{v}_{e^{2 t} s_{\delta}} \mathrm{Y}_{1}=\mathbf{a}_{t} \mathbf{v}_{s_{\delta}} \mathbf{a}_{t}^{-1} \mathrm{Y}_{1}=\mathbf{a}_{t} \mathbf{v}_{s_{\delta}} \mathrm{Y}_{1} \subset \mathbf{a}_{t} \mathrm{Y}_{0}=\mathrm{Y}_{0}, \quad \forall t \in \mathbb{R},
$$

so depending on the sign of $s_{\delta},\left\{g \in \mathrm{H} \mid g \mathrm{Y}_{1} \subset \mathrm{Y}_{0}\right\}$ contains $\mathbf{v}_{s \geq 0}$ or $\mathbf{v}_{s \leq 0}$. We are done.
2.2. Case 3. In this case, we are going to show that

$$
\mathrm{V} \subset\left\{g \in \mathrm{H} \mid g \mathrm{Y}_{1}=\mathrm{Y}_{1}\right\},
$$

which implies the conclusion of Thm.1.1.
Take $x \in \mathrm{Y}_{1}$. Since U. $x$ is not closed, we can find $y_{n}=\exp \left(h_{n}\right) \exp \left(w_{n}\right) x \in \mathrm{Y}_{1}$ with $h_{n} \in$ $\mathfrak{h}, w_{n} \in \mathfrak{h}^{\perp}, h_{n}, w_{n} \rightarrow 0$ and $h_{n}+w_{n} \notin \operatorname{Lie}(\mathrm{U})$. We can no longer assume $h_{n}=0$.

LEmma 2.3. For $\delta>0$ small enough and $n$ large enough, there exist $t_{n, \delta}$ and $s_{n, \delta}$ such that

$$
\mathbf{u}_{s_{n, \delta}} \cdot \mathbf{u}_{t_{n, \delta}} \exp \left(h_{n}\right) \exp \left(w_{n}\right) \mathbf{u}_{t_{n, \delta}}^{-1}=\exp \left(h_{n, \delta}\right) \exp \left(w_{n, \delta}\right),
$$

for some $h_{n, \delta} \in \mathfrak{h}, w_{n, \delta} \in \mathfrak{h}^{\perp}$ with

$$
\max \left\{\left\|h_{n, \delta}\right\|,\left\|w_{n, \delta}\right\|\right\} \in\left[\frac{\delta}{10^{100}}, 10^{100} \delta\right]
$$

and every limit point of $\left(h_{n, \delta} \oplus w_{n, \delta}\right)$ lies in $\operatorname{Lie}(\mathrm{A}) \oplus \operatorname{Lie}(\mathrm{V})$.
See Sec.3.8 for the proof. By passing to a subsequence, let

$$
\begin{aligned}
& y_{\infty, \delta}=\lim y_{n}^{\prime}:=\lim \mathbf{u}_{s_{n, \delta}} \mathbf{u}_{t_{n, \delta}} \cdot y_{n} ; \\
& x_{\infty, \delta}=\lim x_{n}^{\prime}:=\lim \mathbf{u}_{t_{n, \delta}} \cdot x .
\end{aligned}
$$

Also let $h_{\infty, \delta} \oplus w_{\infty, \delta}$ be a limit of $\left(h_{n, \delta} \oplus w_{n, \delta}\right)$. Write $g_{\delta}:=\exp \left(h_{\infty, \delta}\right) \exp \left(w_{\infty, \delta}\right)$. Note that $g_{\delta}$ normalizes $U$.

As in Chapter 1, we arrive at

$$
y_{\infty, \delta}=g_{\delta} \cdot x_{\infty, \delta} \in \mathrm{Y}_{1}, \quad x_{\infty, \delta} \in \mathrm{Y}_{1} .
$$

Hence

$$
g_{\delta} \mathrm{Y}_{1}=\overline{g_{\delta} U \cdot x_{\infty, \delta}}=\overline{\mathrm{U} \cdot y_{\infty, \delta}}=\mathrm{Y}_{1} .
$$

As

$$
\left\{g \in \mathrm{G} \mid g \mathrm{Y}_{1}=\mathrm{Y}_{1}\right\}
$$

is a closed subgroup, if we write $g_{\delta}=\exp v_{\delta}$ with $v_{\delta} \rightarrow 0$ in Lie(AV), then there exists some $v_{\neq 0} \in \operatorname{Lie}(A V)$ such that

$$
\exp (s v) \mathrm{Y}_{1}=\mathrm{Y}_{1}, \quad \forall s \in \mathbb{R} .
$$

If $v$ has non-trivial Lie(V)-component then we are done. Otherwise we go back to case 2 . Hence the proof completes.

## 3. Proof of the two lemmas

The reader is encouraged to prove Lem. 2.2 and 2.3 on his/her own since the proof presented here has simple ideas but messy details.

Both $\mathfrak{h}$ and $\mathfrak{h}^{\perp}$ are invariant under the adjoint action of H , and hence can be considered separately. In matrix terms,

$$
\operatorname{Ad}(g) \cdot M=g M g^{-1}, \quad \operatorname{ad}(X) \cdot M=X M-M X, \quad \exp (\operatorname{ad}(X))=\operatorname{Ad}(\exp (X)) .
$$

### 3.1. Computation of the Lie algebra.

3.1.1. Lie algebra of the orthogonal group. By definition, writing $M_{0}=\left[\begin{array}{ccc}0 & 0 & 1 \\ 0 & -1 & 0 \\ 1 & 0 & 0\end{array}\right]$,

$$
\mathfrak{s o}_{Q_{0}}=\left\{X \in \mathfrak{s l}_{3} \mid M_{0} X+X^{t r} M_{0}=0\right\} .
$$

Write $X=\left(x_{i j}\right)$, then we are solving

$$
\begin{aligned}
& \quad\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & -1 & 0 \\
1 & 0 & 0
\end{array}\right]\left[\begin{array}{lll}
x_{11} & x_{12} & x_{13} \\
x_{21} & x_{22} & x_{23} \\
x_{31} & x_{32} & x_{33}
\end{array}\right]+\left[\begin{array}{lll}
x_{11} & x_{21} & x_{31} \\
x_{12} & x_{22} & x_{32} \\
x_{13} & x_{23} & x_{33}
\end{array}\right]\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & -1 & 0 \\
1 & 0 & 0
\end{array}\right]=0 \\
& \Longleftrightarrow\left[\begin{array}{ccc}
x_{31} & x_{32} & x_{33} \\
-x_{21} & -x_{22} & -x_{23} \\
x_{11} & x_{12} & x_{13}
\end{array}\right]+\left[\begin{array}{lll}
x_{31} & -x_{21} & x_{11} \\
x_{32} & -x_{22} & x_{12} \\
x_{33} & -x_{23} & x_{13}
\end{array}\right]=0 \\
& \Longleftrightarrow x_{31}=x_{22}=x_{13}=0, x_{32}=x_{21}, x_{33}+x_{11}=0, x_{23}=x_{12}
\end{aligned}
$$

That is to say

$$
\mathfrak{s o}_{Q_{0}}=\left\{\left[\begin{array}{ccc}
x_{11} & x_{12} & 0 \\
x_{21} & 0 & x_{12} \\
0 & x_{21} & -x_{11}
\end{array}\right]\right\} .
$$

3.1.2. Computation of its complement. The notation $\mathfrak{s o}_{Q_{0}}^{\perp}$ below is justified by the fact that it is indeed the orthogonal complement of $\mathfrak{s o}_{Q_{0}}$ in $\mathfrak{s l}_{3}$ with respect to the killing form (Exercise: check this).

$$
\mathfrak{s o}_{Q_{0}}^{\perp}=\left\{X \in \mathfrak{s l}_{3} \mid M_{0} X-X^{t r} M_{0}=0\right\} .
$$

Write $X=\left(x_{i j}\right)$, then we are solving

$$
\begin{aligned}
& \quad\left[\begin{array}{lll}
x_{11} & x_{12} & x_{13} \\
x_{21} & x_{22} & x_{23} \\
x_{31} & x_{32} & x_{33}
\end{array}\right]\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & -1 & 0 \\
1 & 0 & 0
\end{array}\right]-\left[\begin{array}{lll}
x_{11} & x_{21} & x_{31} \\
x_{12} & x_{22} & x_{32} \\
x_{13} & x_{23} & x_{33}
\end{array}\right]\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & -1 & 0 \\
1 & 0 & 0
\end{array}\right]=0 \\
& \Longleftrightarrow\left[\begin{array}{ccc}
x_{31} & x_{32} & x_{33} \\
-x_{21} & -x_{22} & -x_{23} \\
x_{11} & x_{12} & x_{13}
\end{array}\right]=\left[\begin{array}{ccc}
x_{31} & -x_{21} & x_{11} \\
x_{32} & -x_{22} & x_{12} \\
x_{33} & -x_{23} & x_{13}
\end{array}\right] \\
& \Longleftrightarrow x_{32}=-x_{21}, x_{11}=x_{33} \text { and } x_{23}=-x_{12} .
\end{aligned}
$$

That is to say

$$
\mathfrak{s o}_{Q_{0}}^{\perp}=\left\{\left[\begin{array}{ccc}
x_{11} & x_{12} & x_{13} \\
x_{21} & -2 x_{11} & -x_{12} \\
x_{31} & -x_{21} & x_{11}
\end{array}\right]\right\} .
$$

3.2. Computation, conjugacy by unipotents. Take $w=\left(w_{i j}\right) \in \mathfrak{h}^{\perp}$, note that
$\operatorname{Ad}\left(\mathbf{u}_{s}\right) \cdot w=\exp \left(s \operatorname{ad}\left(u_{0}\right)\right) \cdot w=w+s \cdot \operatorname{ad}\left(u_{0}\right) w+\frac{s^{2}}{2} \operatorname{ad}\left(u_{0}\right)^{2} w+\frac{s^{3}}{3!} \operatorname{ad}\left(u_{0}\right)^{3} w+\frac{s^{4}}{4!} \operatorname{ad}\left(u_{0}\right)^{4} w$ where the higher order terms vanish.

Write

$$
\begin{aligned}
w= & {\left[\begin{array}{ccc}
w_{11} & w_{12} & w_{13} \\
w_{21} & -2 w_{11} & -w_{12} \\
w_{31} & -w_{21} & w_{11}
\end{array}\right] } \\
= & w_{31} \cdot\left[\begin{array}{lll}
0 & 0 & 0 \\
0 & 0 & 0 \\
1 & 0 & 0
\end{array}\right]+w_{21}\left[\begin{array}{ccc}
0 & 0 & 0 \\
1 & 0 & 0 \\
0 & -1 & 0
\end{array}\right]+w_{11}\left[\begin{array}{ccc}
1 & 0 & 0 \\
0 & -2 & 0 \\
0 & 0 & 1
\end{array}\right] \\
& +\frac{-w_{12}}{3}\left[\begin{array}{ccc}
0 & -3 & 0 \\
0 & 0 & 3 \\
0 & 0 & 0
\end{array}\right]+\frac{w_{13}}{6}\left[\begin{array}{lll}
0 & 0 & 6 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right]
\end{aligned}
$$

The reason why we write it in this form is that

$$
\begin{aligned}
& {\left[\begin{array}{lll}
0 & 0 & 0 \\
0 & 0 & 0 \\
1 & 0 & 0
\end{array}\right] \xrightarrow{\operatorname{ad} u_{0}}\left[\begin{array}{ccc}
0 & 0 & 0 \\
1 & 0 & 0 \\
0 & -1 & 0
\end{array}\right] \xrightarrow{\operatorname{ad} u_{Q}}\left[\begin{array}{ccc}
1 & 0 & 0 \\
0 & -2 & 0 \\
0 & 0 & 1
\end{array}\right] \xrightarrow{\operatorname{ad} u_{0}}\left[\begin{array}{ccc}
0 & -3 & 0 \\
0 & 0 & 3 \\
0 & 0 & 0
\end{array}\right] } \\
& \downarrow \operatorname{ad} u_{0} \\
& {\left[\begin{array}{ccc}
0 & 0 & 6 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right] }
\end{aligned}
$$

Using this, one can compute that
$\operatorname{Ad}\left(\mathbf{u}_{t}\right) w=$

$$
\left.\begin{array}{ccc}
\frac{t^{2}}{2} w_{31}+t w_{21}+w_{11} & \frac{t^{3}}{3!} w_{31}+\frac{t^{2}}{2} w_{21}+t w_{11}+\frac{-w_{12}}{3} & \frac{t^{4}}{4!} w_{31}+\frac{t^{3}}{3!} w_{21}+\frac{t^{2}}{2} w_{11}+t \frac{-w_{12}}{3}+\frac{w_{13}}{6}  \tag{10}\\
t w_{31}+w_{21} & * & * \\
w_{31} & * & *
\end{array}\right]
$$

where the terms marked as $*$ are determined by the others, since the matrix is an element in $\mathfrak{h}^{\perp}$.
3.3. Linear independence of characters. Intuitively, one sees that the upper right corner of Equa.(10) should dominate the rest. To turn this intuition into a solid statement is not so direct due to the possible cancellations between different monomials. By modifying the value of $t$, though, we can avoid this. For simplicity let

$$
\begin{equation*}
\delta_{t}:=\max \left\{\left|\frac{t^{4}}{4!} w_{31}\right|,\left|\frac{t^{3}}{3!} w_{21}\right|,\left|\frac{t^{2}}{2} w_{11}\right|,\left|t \frac{-w_{12}}{3}\right|,\left|\frac{w_{13}}{6}\right|\right\} \tag{11}
\end{equation*}
$$

Let $\|\cdot\|_{\text {sup }}$ denote the maximal value of the absolute values of entries of a matrix, then

$$
\begin{equation*}
\left\|\operatorname{Ad}\left(\mathbf{u}_{t}\right) \cdot w\right\|_{\text {sup }} \leq 5 \delta \tag{12}
\end{equation*}
$$

For simplicity write

$$
p_{w}(t):=\frac{t^{4}}{4!} w_{31}+\frac{t^{3}}{3!} w_{21}+\frac{t^{2}}{2} w_{11}+t \frac{-w_{12}}{3}+\frac{w_{13}}{6} .
$$

Lemma 3.1. For $t \geq 1$, we have that

$$
\max \left\{\left|p_{w}(t)\right|,\left|p_{w}(2 t)\right|,\left|p_{w}(3 t)\right|,\left|p_{w}(4 t)\right|,\left|p_{w}(5 t)\right|\right\} \geq \frac{\delta_{t}}{10^{10}}
$$

To prove this lemma, consider the matrix

$$
M_{0}:=\left[\begin{array}{ccccc}
1 & 1 & 1 & 1 & 1 \\
2^{4} & 2^{3} & 2^{2} & 2 & 1 \\
3^{4} & 3^{3} & 3^{2} & 3 & 1 \\
4^{4} & 4^{3} & 4^{2} & 4 & 1 \\
5^{4} & 5^{3} & 5^{2} & 5 & 1
\end{array}\right]
$$

Lemma 3.2. $\operatorname{det}\left(M_{0}\right)=4!3!2!\neq 0$, and coefficients of $M_{0}^{-1}$ satisfy

$$
\left|\left(M_{0}^{-1}\right)_{i j}\right| \leq \frac{4!5^{4} 4^{3} 3^{2} 2}{4!3!2!} \leq 10^{9}
$$

for every $i, j$.
Proof. $M_{0}$ is a Vandermonde matrix. Details left as an exercise.
Proof of Lemma 3.1.

$$
\left[\begin{array}{c}
p_{w}(t) \\
p_{w}(2 t) \\
p_{w}(3 t) \\
p_{w}(4 t) \\
p_{w}(5 t)
\end{array}\right]=\left[\begin{array}{ccccc}
1 & 1 & 1 & 1 & 1 \\
2^{4} & 2^{3} & 2^{2} & 2 & 1 \\
3^{4} & 3^{3} & 3^{2} & 3 & 1 \\
4^{4} & 4^{3} & 4^{2} & 4 & 1 \\
5^{4} & 5^{3} & 5^{2} & 5 & 1
\end{array}\right] \cdot\left[\begin{array}{c}
\frac{t^{4}}{4^{3}} w_{31} \\
\frac{t^{3}}{3} w_{21} \\
\frac{t^{2}}{2} w_{11} \\
\frac{-w_{12}}{3} \\
\frac{\frac{11}{3}}{6}
\end{array}\right]
$$

And

$$
\delta=\left\|\left[\begin{array}{c}
\frac{t^{4}}{4} w_{31} \\
\frac{t^{3}}{3!} w_{21} \\
\frac{t^{2}}{2} w_{11} \\
\frac{-w_{12}}{w_{13}} \\
\frac{w_{13}}{6}
\end{array}\right]\right\|_{\text {sup }}=\left\|M_{0}^{-1} \cdot\left[\begin{array}{c}
p_{w}(t) \\
p_{w}(2 t) \\
p_{w}(3 t) \\
p_{w}(4 t) \\
p_{w}(5 t)
\end{array}\right]\right\|_{\text {sup }} \leq 5\left\|M_{0}^{-1}\right\|_{\text {sup }} \cdot\left\|\left[\begin{array}{c}
p_{w}(t) \\
p_{w}(2 t) \\
p_{w}(3 t) \\
p_{w}(4 t) \\
p_{w}(5 t)
\end{array}\right]\right\|_{\text {sup }}
$$

Hence

$$
\left\|\left[\begin{array}{c}
p_{w}(t) \\
p_{w}(2 t) \\
p_{w}(3 t) \\
p_{w}(4 t) \\
p_{w}(5 t)
\end{array}\right]\right\|_{\text {sup }} \geq \frac{\delta}{5\left\|M_{0}^{-1}\right\|_{\text {sup }}} \geq \frac{\delta}{10^{10}} .
$$

3.4. Proof of Lemma 2.2. Let $w_{i j}(n)$ be the $(i, j)$-th coefficient of $w_{n}$. Let $\delta>0$, for $n$ large, we can find $t \in \mathbb{R}$ such that

$$
\begin{equation*}
\delta:=\max \left\{\left|\frac{t^{4}}{4!} w_{31}(n)\right|,\left|\frac{t^{3}}{3!} w_{21}(n)\right|,\left|\frac{t^{2}}{2} w_{11}(n)\right|,\left|t \frac{-w_{12}(n)}{3}\right|,\left|\frac{w_{13}(n)}{6}\right|\right\}, \tag{13}
\end{equation*}
$$

namely, Equa.(11) holds with $\delta_{t}=\delta$ and $w_{i j}=w_{i j}(n)$. Let $t_{n, \delta} \in\{t, 2 t, \ldots, 5 t\}$ such that the maximum in Lem.3.1 is attained. By Lem.3.1,

$$
\left\|\operatorname{Ad} \mathbf{u}_{t_{n, \delta}} \cdot w_{n}\right\|_{\sup } \geq \frac{\delta}{10^{10}}
$$

Also note that as $n \rightarrow \infty, t_{n, \delta}$ necessarily goes to $+\infty$. Equa.(12) says that

$$
\left\|\operatorname{Ad} \mathbf{u}_{t_{n, \delta}} \cdot w_{n}\right\|_{\text {sup }} \leq 5 \delta
$$

From Equa.(10), one sees that for $(i, j) \neq(1,3)$,

$$
\left|\left(\operatorname{Ad} \mathbf{u}_{t_{n, \delta}} \cdot w_{n}\right)_{i, j}\right| \leq \frac{4!\delta}{t_{n, \delta}},
$$

which shows that as $n$ goes to the infinity, only $\left(\operatorname{Ad} \mathbf{u}_{t_{n, \delta} .} \cdot w_{n}\right)_{1,3}$ survives. Now the proof is complete.
3.5. From SL2 to $\mathbf{S O}(\mathbf{Q})$. In this subsection, we provide an explicit morphism from $\mathrm{SL}_{2}(\mathbb{R})$ to $\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R})$.
3.5.1. $s l_{2}(\mathbb{R})$ as a quadratic space. Note that $\mathrm{SL}_{2}(\mathbb{R})$ acts on

$$
\mathfrak{s l}_{2}(\mathbb{R})=\{2 \times 2 \text { trace zero matrices }\}
$$

via the adjoint representation. And this action preserves the symmetric bilinear form

$$
-\operatorname{Tr}:(X, Y) \mapsto-\operatorname{Tr}(X \cdot Y) .
$$

To identify $\mathfrak{s l}_{2}(\mathbb{R})$ with $\mathbb{R}^{3}$, consider the basis

$$
E_{1}=\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right], \quad E_{2}=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right], \quad E_{3}=\left[\begin{array}{cc}
0 & 0 \\
-1 & 0
\end{array}\right] .
$$

And we fix an isomorphism $\mathbb{R}^{3} \cong \mathfrak{s l}_{2}(\mathbb{R})$ by sending $e_{i}$ to $E_{i}$ where $\left(e_{1}, e_{2}, e_{3}\right)$ is the standard basis of $\mathbb{R}^{3}$.

Then one can check that

$$
\left(-\operatorname{Tr}\left(E_{i} \cdot E_{j}\right)\right)_{i, j}=\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & -1 & 0 \\
1 & 0 & 0
\end{array}\right]
$$

which means that - Tr is identified with $Q_{0}$ under the fixed isomorphism.
3.5.2. adjoint action of sl2 in basis. Denote by $\rho: \mathrm{SL}_{2}(\mathbb{R}) \rightarrow \mathrm{SO}_{Q_{0}}(\mathbb{R})$ the morphism obtained by the above identification of $\mathfrak{s l}_{2}(R) \cong \mathbb{R}^{3}$. Let us compute $\mathrm{d} \rho: \mathfrak{s l}_{2}(R) \rightarrow \mathfrak{s o}_{Q_{0}}(\mathbb{R})$.

Let

$$
X=\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right] \in \mathfrak{s l}_{2}(\mathbb{R}) .
$$

Then

$$
\begin{aligned}
\operatorname{ad}(X) E_{1} & =\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right]\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]-\left[\begin{array}{cc}
0 & 1 \\
0 & 0
\end{array}\right]\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right] \\
& =\left[\begin{array}{ll}
0 & a \\
0 & c
\end{array}\right]-\left[\begin{array}{cc}
c & -a \\
0 & 0
\end{array}\right]=\left[\begin{array}{cc}
-c & 2 a \\
0 & c
\end{array}\right]=2 a E_{1}+(-\sqrt{2} c) E_{2}+0 E_{3},
\end{aligned}
$$

$\operatorname{ad}(X) E_{2}=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}a & b \\ c & -a\end{array}\right]\left[\begin{array}{cc}1 & 0 \\ 0 & -1\end{array}\right]-\frac{1}{\sqrt{2}}\left[\begin{array}{cc}1 & 0 \\ 0 & -1\end{array}\right]\left[\begin{array}{cc}a & b \\ c & -a\end{array}\right]$

$$
=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
a & -b \\
c & a
\end{array}\right]-\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
a & b \\
-c & a
\end{array}\right]=\left[\begin{array}{cc}
0 & -\sqrt{2} b \\
\sqrt{2} c & 0
\end{array}\right]=(-\sqrt{2} b) E_{1}+0 E_{2}+(-\sqrt{2} c) E_{3}
$$

and

$$
\begin{aligned}
\operatorname{ad}(X) E_{3} & =\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right]\left[\begin{array}{cc}
0 & 0 \\
-1 & 0
\end{array}\right]-\left[\begin{array}{cc}
0 & 0 \\
-1 & 0
\end{array}\right]\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right] \\
& =\left[\begin{array}{cc}
-b & 0 \\
a & 0
\end{array}\right]-\left[\begin{array}{cc}
0 & 0 \\
-a & -b
\end{array}\right]=\left[\begin{array}{cc}
-b & 0 \\
2 a & b
\end{array}\right]=0 E_{1}+(-\sqrt{2} b) E_{2}+(-2 a) E_{3}
\end{aligned}
$$

Hence we have that

$$
\mathrm{d} \rho:\left[\begin{array}{cc}
a & b \\
c & -a
\end{array}\right] \mapsto\left[\begin{array}{ccc}
2 a & -\sqrt{2} b & 0 \\
-\sqrt{2} c & 0 & -\sqrt{2} b \\
0 & -\sqrt{2} c & -2 a
\end{array}\right] .
$$

Sanity check: RHS is indeed a matrix in $\mathfrak{s o}_{Q_{0}}(\mathbb{R})$.
3.6. Image of a unipotent flow. Let

$$
\mathbf{u}_{s}^{\prime}=\left[\begin{array}{ll}
1 & s \\
0 & 1
\end{array}\right] .
$$

Then

$$
\rho\left(\mathbf{u}_{s}^{\prime}\right)=\exp \left(\mathrm{d} \rho\left[\begin{array}{ll}
0 & s \\
0 & 0
\end{array}\right]\right)=\exp \left(\left[\begin{array}{ccc}
0 & -\sqrt{2} s & 0 \\
0 & 0 & -\sqrt{2} s \\
0 & 0 & 0
\end{array}\right]\right)=\mathbf{u}_{-\sqrt{2} s}
$$

3.7. Exponential of a lower triangular matrix. Say we have

$$
\exp \left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]=\left[\begin{array}{cc}
(1+a)^{-1} & 0 \\
b & (1+a)
\end{array}\right]
$$

we would like to express $x, y$ in terms of $a, b$.
Indeed, by definition of exp,

$$
\text { LHS }=\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right]+\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]+\frac{1}{2}\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]^{2}+\frac{1}{3!}\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]^{3}+\ldots
$$

So we should compute the powers of this matrix first.

$$
\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]^{2}=\left[\begin{array}{cc}
x^{2} & 0 \\
0 & (-x)^{2}
\end{array}\right] \Longrightarrow\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]^{2 n}=\left[\begin{array}{cc}
x^{2 n} & 0 \\
0 & (-x)^{2 n}
\end{array}\right]
$$

And odd powers are

$$
\left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]^{2 n+1}=\left[\begin{array}{cc}
x^{2 n+1} & 0 \\
y x^{2 n} & (-x)^{2 n+1}
\end{array}\right]
$$

Thus

$$
\exp \left[\begin{array}{cc}
x & 0 \\
y & -x
\end{array}\right]=\left[\begin{array}{cc}
e^{x} & 0 \\
y\left(\frac{e^{x}-e^{-x}}{2 x}\right) & e^{-x}
\end{array}\right]=\left[\begin{array}{cc}
(1+a)^{-1} & 0 \\
b & (1+a)
\end{array}\right] .
$$

And thus

$$
\begin{equation*}
x=\ln (1+a), y=b\left(\frac{2 \ln (1+a)}{(1+a)-(1+a)^{-1}}\right) . \tag{14}
\end{equation*}
$$

The equality for $y$ is not needed. Also note that for $|a|<1$

$$
\begin{equation*}
|\ln (1+a)-a| \leq 2|a|^{2} \tag{15}
\end{equation*}
$$

and that if ( $a_{n}$ ) is a sequence contained in some fixed compact sub-interval of $(-1,+\infty)$ and $\left(b_{n}\right)$ is a sequence converging to 0 then the corresponding $\left(y_{n}\right)$ should converge to 0 .
3.8. Proof of Lemma 2.3. For convenience, let us repeat Lem.2.3.

Lemma 3.3. For $\delta>0$ small enough and $n$ large enough, there exist $t_{n, \delta}$ and $s_{n, \delta}$ such that

$$
\begin{equation*}
\mathbf{u}_{s_{n, \delta}} \cdot \mathbf{u}_{t_{n, \delta}} \exp \left(h_{n}\right) \exp \left(w_{n}\right) \mathbf{u}_{t_{n, \delta}}^{-1}=\exp \left(h_{n, \delta}\right) \exp \left(w_{n, \delta}\right), \tag{16}
\end{equation*}
$$

for some $h_{n, \delta} \in \mathfrak{h}, w_{n, \delta} \in \mathfrak{h}^{\perp}$ with

$$
\begin{equation*}
\max \left\{\left\|h_{n, \delta}\right\|,\left\|w_{n, \delta}\right\|\right\} \in\left[\frac{\delta}{10^{100}}, 10^{100} \delta\right] \tag{17}
\end{equation*}
$$

and every limit point of $\left(h_{n, \delta} \oplus w_{n, \delta}\right)$ lies in $\operatorname{Lie}(\mathrm{A}) \oplus \operatorname{Lie}(\mathrm{V})$.
Proof. Define $h_{n}^{\prime}:=\mathrm{d} \rho^{-1}\left(h_{n}\right)$. Write

$$
\exp \left(h_{n}^{\prime}\right)=\left[\begin{array}{cc}
1+a_{n} & b_{n} \\
c_{n} & 1+d_{n}
\end{array}\right] .
$$

For $s, t \in \mathbb{R}$, write $s^{\prime}:=s /(-\sqrt{2}), t^{\prime}=t /(-\sqrt{2})$. Hence $\rho\left(\mathbf{u}_{s^{\prime}}^{\prime}\right)=\mathbf{u}_{s}$ and $\rho\left(\mathbf{u}_{t^{\prime}}^{\prime}\right)=\mathbf{u}_{t}$.
Choose $s_{n, \delta}$ depending on $t_{n, \delta}$ (to be determined later) such that

$$
\mathbf{u}_{s_{n, \delta}^{\prime}}^{\prime} \mathbf{u}_{t_{n, \delta}^{\prime}}^{\prime} \exp \left(h_{n}^{\prime}\right)\left(\mathbf{u}_{t_{n, \delta}^{\prime}}^{\prime}\right)^{-1}=\left[\begin{array}{cc}
\left(1+d_{n}-t_{n, \delta}^{\prime} c_{n}\right)^{-1} & 0  \tag{18}\\
c_{n} & 1+d_{n}-t_{n, \delta}^{\prime} c_{n}
\end{array}\right] .
$$

See Chapter 1 for details. Define $h_{n, \delta}^{\prime}$ by

$$
\exp \left(h_{n, \delta}^{\prime}\right)=\left[\begin{array}{cc}
\left(1+d_{n}-t_{n, \delta}^{\prime} c_{n}\right)^{-1} & 0  \tag{19}\\
c_{n} & 1+d_{n}-t_{n, \delta}^{\prime} c_{n}
\end{array}\right]
$$

Write $w_{n}=\left(w_{i j}(n)\right)$. Choose $t$ such that

$$
\delta=\max \left\{\left|d_{n}-t^{\prime} c_{n}\right|,\left|\frac{t^{4}}{4!} w_{31}(n)\right|,\left|\frac{t^{3}}{3!} w_{21}(n)\right|,\left|\frac{t^{2}}{2} w_{11}(n)\right|,\left|t \frac{-w_{12}(n)}{3}\right|,\left|\frac{w_{13}(n)}{6}\right|\right\} .
$$

Also let

$$
\delta^{\prime}:=\max \left\{\left|\frac{t^{4}}{4!} w_{31}(n)\right|,\left|\frac{t^{3}}{3!} w_{21}(n)\right|,\left|\frac{t^{2}}{2} w_{11}(n)\right|,\left|t \frac{-w_{12}(n)}{3}\right|,\left|\frac{w_{13}(n)}{6}\right|\right\} .
$$

We choose $t_{n, \delta}$ from $t, 2 t, \ldots, 5 t$ such that the maximum in Lem.3.1 is attained (with $\delta_{t}$ replaced by $\delta^{\prime}$ ).

Define $h_{n, \delta}:=\mathrm{d} \rho\left(h_{n, \delta}^{\prime}\right)$ and

$$
w_{n, \delta}:=\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n} .
$$

Now everything is defined and it remains to check the conclusion of Lem.2.3.
First one can verify Equa.(16). By Equa.(18) and (19),

$$
\begin{aligned}
& \mathbf{u}_{s_{n, \delta}^{\prime}}^{\prime} \mathbf{u}_{t_{n, \delta}^{\prime}}^{\prime} \exp \left(h_{n}^{\prime}\right)\left(\mathbf{u}_{t_{n, \delta}^{\prime}}^{\prime}\right)^{-1}=\exp \left(h_{n, \delta}^{\prime}\right) \\
(\text { apply } \rho) \Longrightarrow & \mathbf{u}_{s_{n, \delta}} \mathbf{u}_{t_{n, \delta}} \exp \left(h_{n}\right) \mathbf{u}_{t_{n, \delta}}^{-1}=\exp \left(h_{n, \delta}\right) \\
\Longrightarrow & \mathbf{u}_{s_{n, \delta}} \mathbf{u}_{t_{n, \delta}} \exp \left(h_{n}\right) \exp \left(w_{n}\right) \mathbf{u}_{t_{n, \delta}}^{-1} \\
& =\exp \left(h_{n, \delta}\right) \exp \left(\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right)=\exp \left(h_{n, \delta}\right) \exp \left(w_{n, \delta}\right) .
\end{aligned}
$$

That $h_{n, \delta} \in \mathfrak{h}, w_{n, \delta} \in \mathfrak{h}^{\perp}$ follows from their definition. It remains to verify Equa.(17) and that every limit of ( $h_{n, \delta}$ ) is in $\operatorname{Lie}(\mathrm{A})$, every limit of $\left(w_{n, \delta}\right)$ is in $\operatorname{Lie}(\mathrm{V})$.

By the discussion below Equa.(15) and that $c_{n} \rightarrow 0$, we find

$$
\left|\left(h_{n, \delta}^{\prime}\right)_{2,1}\right| \rightarrow 0
$$

Thus every limit of $\left(h_{n, \delta}\right)$ is in $\operatorname{Lie}(\mathrm{A})$. That every limit of $\left(w_{n, \delta}\right)$ is in $\operatorname{Lie}(\mathrm{V})$ follows from the proof of Lem.2.2.

For $n$ sufficiently large such that $\left|d_{n}\right| \leq \delta$,

$$
\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right| \leq\left|d_{n}\right|+5\left|t^{\prime} c_{n}\right| \leq 6\left|d_{n}\right|+5\left|d_{n}-t^{\prime} c_{n}\right| \leq 11 \delta
$$

Combined with Equa.(15) we see that (assume $\delta \leq 1$ )

$$
\left|\left(h_{n, \delta}^{\prime}\right)_{2,2}\right| \leq\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right|+2\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right|^{2}=11 \delta+2 \cdot 11^{2} \delta^{2} \leq 300 \delta
$$

This shows that

$$
\left\|h_{n, \delta}\right\| \leq 10^{100} \delta
$$

If $\delta=\delta^{\prime}$, then the remaining conclusions follow from the proof of Lem.2.2.
If $\delta>\delta^{\prime}$, then $\delta=\left|d_{n}-t^{\prime} c_{n}\right|$. For $n$ large enough such that $\left|d_{n}\right| \leq 0.1 \delta$,

$$
\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right| \geq\left|t_{n, \delta}^{\prime} c_{n}\right|-\left|d_{n}\right| \geq\left|t^{\prime} c_{n}\right|-\left|d_{n}\right| \geq\left|d_{n}-t^{\prime} c_{n}\right|-2\left|d_{n}\right| \geq 0.8 \delta
$$

Now by Equa.(15),

$$
\left|\left(h_{n, \delta}^{\prime}\right)_{2,2}\right| \geq\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right|-2\left|d_{n}-t_{n, \delta}^{\prime} c_{n}\right|^{2} \geq 0.8 \delta-2 \cdot(11)^{2} \delta^{2}
$$

If $\delta$ is sufficiently small such that $2 \cdot(11)^{2} \delta \leq 0.1$, then

$$
\left|\left(h_{n, \delta}^{\prime}\right)_{2,2}\right| \geq 0.5 \delta \Longrightarrow\left\|h_{n, \delta}\right\| \geq 0.5 \delta
$$

Again, the rest of the claim follows by arguments in Lem.2.2.
Now we are done.

## 4. Exercises

4.1. orbits of diagonal groups. We say that a matrix $g \in \mathrm{SL}_{2}(\mathbb{R})$ is $\mathbb{R}$-diagonalizable iff there exists $h \in \mathrm{SL}_{2}(\mathbb{R})$ such that $h g h^{-1}$ is a diagonal matrix. Note that for a matrix $X_{\neq \pm \mathrm{id}} \in$ $\mathrm{SL}_{2}(\mathbb{R})$, being $\mathbb{R}$-diagonalizable is equivalent to being hyperbolic in the sense that $\operatorname{Tr}(X)>2$. Fix a discrete subgroup $\Gamma$ of $\mathrm{SL}_{2}(\mathbb{R})$, an $\mathbb{R}$-diagonalizable matrix $\gamma \in \Gamma$ is said to be primitive iff it can not be written as $\left(\gamma^{\prime}\right)^{n}$ for some $n \in \mathbb{Z}, n \neq \pm 1$ and some other $\gamma^{\prime} \in \Gamma$ that is $\mathbb{R}$ diagonalizable. By definition $\pm \mathrm{id}$ is never primitive. Let

$$
\operatorname{Prim}(\Gamma):=\{\gamma \text { is } \mathbb{R} \text {-diagonalizable and primitive }\} .
$$

EXERCISE 4.1. Assume $\Gamma \leq \mathrm{SL}_{2}(\mathbb{R})$ is a discrete subgroup containing $\{ \pm \mathrm{id}\}$. Find a bijection between

$$
\left\{\text { compact }\left\{a_{t}\right\} \text {-orbits }\right\} \cong \operatorname{Prim}(\Gamma) / \sim_{\Gamma}
$$

where $\sim_{\Gamma}$ is the equivalence relation defined by $g \sim_{\Gamma} h$ iff $g=\gamma h \gamma^{-1}$ for some $\gamma \in \Gamma$.
EXERCISE 4.2. Classify all compact $\left\{a_{t}\right\}_{t \in \mathbb{R}}$-orbits on $\mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$.
EXERCISE 4.3. Classify all divergent $\left\{a_{t}\right\}_{t \in \mathbb{R}}$-orbits on $\mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$.
Recall that an orbit $\left\{a_{t} \cdot x\right\}$ is said to be divergent iff for every compact set in $C \subset \mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$ there exists $t_{0}>0$ such that for all $|t|>t_{0}$, we have $a_{t} . x \notin C$.

## CHAPTER 4

## Nondivergence of unipotent flows on $X_{2}$

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The main reference for this chapter is Kleinbock's Clay notes [Kle10].
Notation:

- $\mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}, \mathrm{A}:=\left\{\mathbf{a}_{t}: \left.=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\} ;$
- $\mathrm{X}_{2}=\left\{\right.$ unimodular lattices in $\left.\mathbb{R}^{2}\right\}$


## 1. Summary

Definition 1.1. For $\varepsilon>0$, define

$$
\mathscr{C}_{\varepsilon}:=\left\{\Lambda \in \mathrm{X}_{2} \mid \operatorname{sys}(\Lambda) \geq \varepsilon\right\} .
$$

By Lem.2.9 from Ch. 2 (Mahler's criterion), $\mathscr{C}_{\varepsilon}$ is a compact set and every compact set in $\mathrm{X}_{2}$ is contained in $\mathscr{C}_{\varepsilon}$ for some $\varepsilon>0$.

Theorem 1.2. [Uniform non-divergence of unipotent flows for $\mathrm{X}_{2}$ ] For every compact set $K \subset \mathrm{X}_{2}$ and $\varepsilon \in(0,1)$, there exists $\delta=\delta(K, \varepsilon)>0$ such that the following holds. For every interval $(a, b)$ with $a<b$ in $\mathbb{R}$ and $\Lambda \in X_{2}$ satisfying $\mathbf{u}_{s_{0}} . \Lambda \in K$ for some $s_{0} \in(a, b)$, we have that

$$
\frac{1}{b-a} \operatorname{Leb}\left\{s \in(a, b) \mid \mathbf{u}_{s . \Lambda} \notin \mathscr{C}_{\delta}\right\} \leq \varepsilon
$$

Actually the choice of $\delta$ is also independent of the unipotent flow we use - you may replace $\mathbf{u}_{s}$ everywhere by its conjugates.

Theorem 1.3. If $\varepsilon \leq 1$ and $\Lambda \in \mathrm{X}_{2}$ are such that $\mathbf{u}_{s} . \Lambda \notin \mathscr{C}_{\varepsilon}$ for everys in some interval of infinite length (i.e., something like $(a,+\infty),(-\infty, b),(-\infty,+\infty)$ ), then $\Lambda$ contains a horizontal vector of length less than $\varepsilon$. That is to say, $\left(\nu_{1}, 0\right) \in \Lambda$ for some $0<\left|\nu_{1}\right|<\varepsilon$.

Corollary 1.4. For every two $x_{1}$, $x_{2}$ with compact U -orbits, there exists $u \in \mathrm{U}$ and $a \in \mathrm{~A}$ such that $x_{2}=a u . x_{1}$.

The reader might have noticed that the converse also holds since $U$-action fixes the horizontal direction. Also note that such $U$-orbits are closed and compact. In this case, one may think of $U$-action on $\Lambda$ as "Dehn-twist" along the closed geodesic represented by $\left(\nu_{1}, 0\right) \in \Lambda \cong$ $\pi_{1}\left(\mathbb{R}^{2} / \Lambda\right)$.

Here is a compact $U$-orbit $U . \mathbb{Z}^{2}$ :


## 2. The proof

Lemma 2.1. There exist $C_{1}>0$ and $\alpha_{1}>0$ such that for every interval $(a, b)$ in $\mathbb{R}$, every $\mathbf{v}=\left(v_{1}, v_{2}\right) \in \mathbb{R}^{2}$ and every $\rho \in(0,1)$, we have

$$
\frac{1}{b-a} \operatorname{Leb}\left\{s \in(a, b) \mid\left\|\mathbf{u}_{s} . v\right\|<\rho M_{0}\right\} \leq C_{1} \rho^{\alpha_{1}} .
$$

where $M_{0}:=\sup _{s \in(a, b)}\left\|\mathbf{u}_{s . v} v\right\|$.
Proof. Take $C_{1}=2 \sqrt{2}$ and $\alpha_{1}=1$.
Note $\mathbf{u}_{s} .\left(v_{1}, v_{2}\right)=\left(\nu_{1}+s v_{2}, v_{2}\right)$.
If $\left|v_{2}\right| \geq \frac{1}{\sqrt{2}} M_{0}$ then for every $s \in(a, b),\left\|\mathbf{u}_{s} . v\right\| \geq\left|v_{2}\right| \geq \frac{1}{\sqrt{2}} M_{0}$. So if $\rho \leq \frac{1}{\sqrt{2}}$, then we are already done. Otherwise, $C_{1} \rho^{\alpha_{1}} \geq 1$. Also ok.

So now we are left with the case when $\left|v_{1}+s_{0} \nu_{2}\right| \geq \frac{1}{\sqrt{2}} M_{0}$ for some $s_{0} \in(a, b)$. Refer to the picture below, we see that

$$
\frac{1}{b-a} \operatorname{Leb}\left\{s \in(a, b)\left|\left|v_{1}+s v_{2}\right|<\rho M_{0}\right\} \leq 2 \frac{\rho}{1 / \sqrt{2}}=C_{1} \rho .\right.
$$

It remains only to note that $\left|\nu_{1}+s v_{2}\right|<\left\|\mathbf{u}_{s} .\left(\nu_{1}, \nu_{2}\right)\right\|$.



To go from this Lemma to Thm.1.2 or 1.3, we need
Key observation. A rank 2 unimodular lattice $\Lambda \in \mathrm{X}_{2}$ is not allowed to contain two linearly independent vector of length strictly smaller than 1 . For otherwise, if $v, w$ is such a pair,

$$
\|\Lambda\| \leq\|\mathbb{Z} v \oplus \mathbb{Z} w\| \leq\|v\|\|w\|<1,
$$

contradicting against the assumption that $\Lambda$ is unimodular.


## No matter what the unimodular lattice is, you at most see a single line in a small neighborhood about the origin

Let

$$
\operatorname{Prim}(\Lambda):=\left\{v_{\neq 0} \in \Lambda \mid \mathbb{R} . \nu \cap \Lambda=\mathbb{Z} . v\right\}
$$

be the set of primitive vectors.
Proof of Thm.1.2. Find $\delta_{0} \in(0,1)$ such that $K \subset \mathscr{C}_{\delta_{0}}$. We shall determine $\delta$ later, depending on $\delta_{0}$ and $\varepsilon$.

Take $s_{0} \in(a, b)$ such that $\mathbf{u}_{s_{0}} . \Lambda_{0} \in K \subset \mathscr{C}_{\delta_{0}}$. Let

$$
I\left(\Lambda_{0}, \delta_{0}\right):=\left\{s \in(a, b) \mid \operatorname{sys}\left(\mathbf{u}_{s} . \Lambda_{0}\right)<\delta_{0}\right\}
$$

which decomposes into a disjoint union of open intervals

$$
I\left(\Lambda_{0}, \varepsilon_{0}\right)=\bigsqcup_{\alpha \in \mathscr{A}} I_{\alpha}
$$

with certain index set $\mathscr{A}$.


Take one $I_{\alpha}=\left(x_{\alpha}, y_{\alpha}\right)$. By the remark right before the proof, for $s \in I_{\alpha}$, there exists a unique $v_{s}$ (up to $\pm 1$ ) in $\operatorname{Prim}\left(\Lambda_{0}\right)$ with

$$
\left\|\mathbf{u}_{s} \cdot v_{s}\right\|<\delta_{0}
$$

By connectedness, this $v_{s}$ has to be independent of $s \in I_{\alpha}$. For this reason denote it by $v_{\alpha}$. By Lem.3.1 and the assumption that $\mathbf{u}_{s_{0}} . \Lambda_{0} \in \mathscr{C}_{\delta_{0}}$,

$$
\frac{1}{\left|I_{\alpha}\right|} \operatorname{Leb}\left\{s \in I_{\alpha} \mid\left\|\mathbf{u}_{s} \cdot v_{\alpha}\right\|<\rho \delta_{0}\right\}<C_{1} \rho^{\alpha_{1}} .
$$

We take $\rho=\rho(\varepsilon)$ such that $C_{1} \rho^{\alpha_{1}}<\varepsilon$. Let $\delta:=\rho \delta_{0}$.

$$
\left\{s \in(a, b) \mid\left\|\mathbf{u}_{s} \cdot v_{\alpha}\right\|<\delta\right\}=\bigsqcup_{\alpha \in \mathscr{A}}\left\{s \in I_{\alpha} \mid\left\|\mathbf{u}_{s} \cdot v_{\alpha}\right\|<\rho \delta_{0}\right\}
$$

implying

$$
\operatorname{Leb}\left\{s \in(a, b) \mid\left\|\mathbf{u}_{s} \cdot v_{\alpha}\right\|<\delta\right\}=\sum_{\alpha \in \mathscr{A}} \operatorname{Leb}\left\{s \in I_{\alpha} \mid\left\|\mathbf{u}_{s} \cdot v_{\alpha}\right\|<\rho \delta_{0}\right\}<\sum_{\alpha \in \mathscr{A}}\left|I_{\alpha}\right| \cdot \varepsilon \leq(b-a) \varepsilon .
$$

Proof of Lem.1.3. Let $I$ be this infinite interval. Since for each $s \in I$ there exists a unique (up to $\pm 1$ ) $v_{s}$ in $\operatorname{Prim}(\Lambda)$ with $\left\|\mathbf{u}_{s} . v_{s}\right\|<1$. By connectedness argument, this $v=v_{s}$ is independent of $s \in I$. Thus $\left\|\mathbf{u}_{s} . v\right\|<1$ for all $s \in I$. But coordinates of $\mathbf{u}_{s} . v$ are polynomials in $s$, being bounded then implies that $\mathbf{u}_{s} . v$ is constant. Therefore U fixes $v$ and we are done.

## 3. Exercises

3.1. Nondivergence in rank 1, a number field example. In these set of exercises, it is more convenient to write $\mathbb{R}^{4}$ as $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$.

EXERCISE 3.1. Show that $\mathbb{Z}[\sqrt{2}]$ is a principal ideal domain.
Thus every torsion free (finitely generated) $\mathbb{Z}[\sqrt{2}]$-module is free.
Fix an embedding of $\mathbb{Q}(\sqrt{2})$ in $\mathbb{R}$. Let $\sigma$ be the other embedding of $\mathbb{Q}(\sqrt{2})$ in $\mathbb{R}$. Consider the action of $\mathbb{Q}(\sqrt{2})$ on $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$ given by

$$
x .(v, w):=(x . v, \sigma(x) \cdot w) .
$$

Exercise 3.2. This is a linear action. Write down the matrix representation of this action. Namely, for every $x=a+b \sqrt{2} \in \mathbb{Q}(\sqrt{2})$, write down a 4 -by-4 matrix representing the action of $x$ on $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$ with respect to the standard basis.

Let $\Delta$ be a rank- $1 \mathbb{Z}[\sqrt{2}]$-submodule in $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$. We may write $\Delta=\mathbb{Z}[\sqrt{2}] .(\nu, w)$. Let $\|\Delta\|:=$ $\|\nu\| \cdot\|w\|$.

Exercise 3.3. Show that $\|\Delta\|$ is independent of the choice of generator for the $\mathbb{Z}[\sqrt{2}]$ module $\Delta$.

Define

$$
\mathrm{X}_{4}^{\prime}(\mathbb{Z}[\sqrt{2}]):=\left\{\Lambda \leq \mathbb{R}^{2} \oplus \mathbb{R}^{2} \text { lattice , } \Lambda \text { is preserved by } \mathbb{Z}[\sqrt{2}]\right\} .
$$

EXERCISE 3.4. Show that such a lattice is a rank-2 $\mathbb{Z}[\sqrt{2}]$-module.
Thus for $\Lambda \in \mathrm{X}_{4}^{\prime}(\mathbb{Z}[\sqrt{2}])$, we can find a $\mathbb{Z}[\sqrt{2}]$-basis $\left(v_{1}, w_{1}\right)$ and $\left(v_{2}, w_{2}\right)$ in $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$. Define $\|\Lambda\|:=\left\|\nu_{1} \wedge v_{2}\right\| \cdot\left\|w_{1} \wedge w_{2}\right\|$. Define $\operatorname{det}(\Lambda):=\left(\nu_{1} \wedge v_{2}, w_{1} \wedge w_{2}\right) \in(\mathbb{R} \oplus \mathbb{R}) / \mathbb{Z}[\sqrt{2}]^{\times}$. Here $\mathbb{Z}[\sqrt{2}]^{\times}$ denotes the invertible elements in this ring $\mathbb{Z}[\sqrt{2}]$.

Exercise 3.5. Show that indeed, the value of $\operatorname{det}(\Lambda)$ in $(\mathbb{R} \oplus \mathbb{R}) / \mathbb{Z}[\sqrt{2}]^{\times}$is independent of the choice of bases. Thus $\|\Lambda\|$ is also independent of the choice of bases.

EXERCISE 3.6. Find the relation between this newly defined $\|\Lambda\|$ and the old $\|\Lambda\|_{\text {Old }}$ defined as the volume of $\mathbb{R}^{4} / \Lambda$.

Define

$$
\mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}]):=\left\{\Lambda \in \mathrm{X}_{4}^{\prime}(\mathbb{Z}[\sqrt{2}]) \mid \operatorname{det} \Lambda=1\right\} .
$$

Here " 1 " is the image of $(1,1)$ in $(\mathbb{R} \oplus \mathbb{R}) / \mathbb{Z}[\sqrt{2}]^{\times}$. Equip $X_{4}(\mathbb{Z}[\sqrt{2}])$ with the Chabauty topology, viewing it as a collection of closed subgroups of $\mathbb{R}^{2} \oplus \mathbb{R}^{2}$.

EXERCISE 3.7. Show that the free $\mathbb{Z}[\sqrt{2}]$-module with basis $\left\{\left(e_{1}, e_{1}\right),\left(e_{2}, e_{2}\right)\right\}$ (denote this module as $\Lambda_{0}$ ) belongs to $\mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}])$ and that $g \mapsto g . \Lambda_{0}$ induces a homeomorphism

$$
\mathrm{SL}_{2}(\mathbb{R}) \times \mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z}[\sqrt{2}]) \cong \mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}]) .
$$

For $\Lambda \in X_{4}(\mathbb{Z}[\sqrt{2}])$, define

$$
\operatorname{sys}_{\mathbb{Z}[\sqrt{2}]}(\Lambda):=\inf _{\Delta \leq \Lambda}\|\Delta\|
$$

where $\Delta$ varies over all rank-1 $\mathbb{Z}[\sqrt{2}]$-submodule of $\Lambda$. For every $\varepsilon>0$, let

$$
\mathscr{C}_{\varepsilon}:=\left\{\Lambda \in \mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}]) \mid \operatorname{sys}_{\mathbb{Z}[\sqrt{2}]}(\Lambda) \geq \varepsilon\right\} .
$$

EXERCISE 3.8. For every $\varepsilon>0, \mathscr{C}_{\varepsilon}$ is a compact subset of $\mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}])$.
EXERCISE 3.9. Conversely, every compact subset of $\mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}])$ is contained in $\mathscr{C}_{\varepsilon}$ for some $\varepsilon>0$.

EXERCISE 3.10. For $\varepsilon>0$ small enough, for every $\Lambda \in X_{4}(\mathbb{Z}[\sqrt{2}])$, the set

$$
\{(\nu, w) \in \Lambda \mid\|v\|\|w\|<\varepsilon\}
$$

is either $\{0\}$ or generates a rank-1 $\mathbb{Z}[\sqrt{2}]$-submodule of $\Lambda$.

$$
\text { Let } \mathbf{u}_{t}:=\left(\left[\begin{array}{cc}
1 & t \\
0 & 1
\end{array}\right],\left[\begin{array}{cc}
1 & t \\
0 & 1
\end{array}\right]\right) \text { and } U:=\left\{\mathbf{u}_{t}, t \in \mathbb{R}\right\} .
$$

EXercise 3.11. Prove the following. For every $\varepsilon>0$, there exists $\delta>0$ such that for every $\Lambda \in \mathrm{X}_{4}(\mathbb{Z}[\sqrt{2}])$,

- either $\Lambda$ contains $a \mathbb{Z}[\sqrt{2}]$-submodule preserved by $U$ with norm smaller than $\varepsilon$,
- or

$$
\limsup _{T \rightarrow+\infty} \frac{1}{T} \operatorname{Leb}\left\{t \in[0, T] \mid \mathbf{u}_{t} . \Lambda \notin \mathscr{C}_{\delta}\right\} \leq \varepsilon .
$$

## CHAPTER 5

## Nondivergence on $X_{3}$ and the strong form of Oppenheim conjecture

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Notations:

- $\mathrm{H}:=\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R})$ with $Q_{0}\left(x_{1}, x_{2}, x_{3}\right):=2 x_{1} x_{3}-x_{2}^{2}$;
- $\mathrm{X}_{3}:=\left\{\right.$ unimodular lattices in $\left.\mathbb{R}^{3}\right\}$;
- $\mathbf{u}_{s}:=\left[\begin{array}{llr}1 & s & \frac{s^{2}}{2} \\ 0 & 1 & s \\ 0 & 0 & 1\end{array}\right]=\exp \left(s \cdot\left[\begin{array}{lll}0 & 1 & 0 \\ 0 & 0 & 1 \\ 0 & 0 & 0\end{array}\right]\right), \mathrm{U}:=\left\{\mathbf{u}_{s}: s \in \mathbb{R}\right\} \subset \mathrm{H} ;$
- $\mathbf{v}_{s}:=\left[\begin{array}{lll}1 & 0 & s \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]=\exp \left(s \cdot\left[\begin{array}{lll}0 & 0 & 1 \\ 0 & 0 & 0 \\ 0 & 0 & 0\end{array}\right]\right), \mathrm{V}:=\left\{\mathbf{v}_{s}: s \in \mathbb{R}\right\} \nsubseteq \mathrm{H}$;
- $\mathbf{a}_{t}:=\left[\begin{array}{ccc}e^{t} & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & e^{-t}\end{array}\right]=\exp \left(t \cdot\left[\begin{array}{ccc}1 & 0 & 0 \\ 0 & 0 & 0 \\ 0 & 0 & -1\end{array}\right]\right), \mathrm{A}:=\left\{\mathbf{a}_{t}: t \in \mathbb{R}\right\} \subset \mathrm{H} ;$
- $\mathrm{B}:=\left\{\mathbf{a}_{t} \mathbf{u}_{s}: s, t \in \mathbb{R}\right\} \subset \mathrm{H}$.


## 1. Summary

Finally, in this section we prove the strong form of Oppenheim conjecture. The general case can be reduced to the case of three variables, which we now state

Theorem 1.1. Let $Q$ be a non-degenerate indefinite ternary real quadratic form that is not proportional to a rational quadratic form. Then $Q\left(\mathbb{Z}^{3}\right)$ is dense in $\mathbb{R}$. Actually $Q\left(\operatorname{Prim}\left(\mathbb{Z}^{3}\right)\right)$ is dense in $\mathbb{R}$.

Theorem 1.2. For every non-closed orbit of H on $\mathrm{X}_{3}$, its closure contains a $\left\{\mathbf{v}_{s}\right\}_{s \geq 0}$ or $\left\{\mathbf{v}_{s}\right\}_{s \leq 0}$ orbit.

A stronger statement will be proved later. See Ch.12, Thm.1.1.
By similar arguments presented in Chapter 2, Thm.1.1 would follow from Thm.1.2 and the following (whose proof is left as an exercise):

Theorem 1.3. If an H -orbit is closed, then the stabilizer in H is discrete and of finite covolume in H . Also the corresponding quadratic form is a scalar multiple of some rational quadratic form.

To promote the weak version to this one the following non-divergence theorem is needed.
Theorem 1.4. For every $\varepsilon>0$, there exists a compact subset $\mathscr{C}$ of $\mathrm{X}_{3}$ such that for every $\Lambda \in \mathrm{X}_{3}$, at least one of the followings is true

1. The portion of time for $\mathbf{u}_{s} . \Lambda$ to spend outside $\mathscr{C}$ is smaller than $\varepsilon$, i.e.,

$$
\limsup _{T \rightarrow+\infty} \frac{1}{T}\left|\left\{s \in[0, T] \mid \mathbf{u}_{s} . \Lambda \notin \mathscr{C}\right\}\right| \leq \varepsilon
$$

2. $\Lambda \cap\{(x, 0,0), x \in \mathbb{R}\}$ contains a non-zero vector of length smaller than $\varepsilon$;
3. $\Lambda \cap\{(x, y, 0), x, y \in \mathbb{R}\}$ contains a lattice (of $\mathbb{R} e_{1} \oplus \mathbb{R} e_{2}$ ) of covolume smaller than $\varepsilon$.

Corollary 1.5. Let $\varepsilon \in(0,1)$ and pick $\mathscr{C}$ as in the above theorem. Then every orbit of B intersects non-trivially with $\mathscr{C}$.

Finally let us make a convenient definition. Let $e_{1}:=(1,0,0)$ and $e_{2}=(0,1,0)$.
Definition 1.6. We say that $\mathbb{R} . e_{1}$ is $\Lambda$-rational iff $\Lambda \cap \mathbb{R} e_{1}$ is a lattice in $\mathbb{R} e_{1}$, and $\mathbb{R} e_{1} \oplus \mathbb{R} e_{2}$ is $\Lambda$-rational iff $\Lambda \cap \mathbb{R} e_{1} \oplus \mathbb{R} e_{2}$ is a lattice in $\mathbb{R} e_{1} \oplus \mathbb{R} e_{2}$. In either of these two cases, we say that the orbit $\mathrm{U} . \Lambda$ degenerates.

This notion is justified by the fact that in these cases the orbit is essentially contained in certain (embedded) $\mathrm{SL}_{2}(\mathbb{R}) \ltimes \mathbb{R}^{2} / \mathrm{SL}_{2}(\mathbb{Z}) \ltimes \mathbb{Z}^{2}$, which is interpreted as the space of lattices of lower rank with a fixed volume together with a marked point in the quotient torus.

## 2. Proof of the theorem

Now let us prove Thm.1.2. Start with $\Lambda_{0}$ with H. $\Lambda_{0}$ non-closed. Let $\mathrm{Y}_{0}:=\overline{\mathrm{H} . \Lambda_{0}}$. Define $\mathscr{O}$ as in Chapter 3, the union of all H -orbits in $\mathrm{Y}_{0}$ that is open in $\mathrm{Y}_{0}$. Note that $\mathscr{O} \neq \mathrm{Y}_{0}$.

The old argument takes care of the case when $\mathrm{Y}_{0} \backslash \mathscr{O}$ contains no degenerate $U$-orbits.
Indeed under this assumption every U -orbit in $\mathrm{Y}_{0} \backslash \mathscr{O}$, by Thm.1.4, intersects with some fixed compact set non-trivially. Hence we can find a nonempty $U$-minimal set $Y_{1}$ in $Y_{0} \backslash \mathscr{O}$. As in Ch.3, there are two cases:

1. $\mathrm{Y}_{1}$ is A-stable, we consider $\operatorname{Map}\left(\mathrm{Y}_{1}, \mathrm{Y}_{0}\right):=\left\{g \mathrm{Y}_{1} \subset \mathrm{Y}_{0}\right\}$;
2. $\mathrm{Y}_{1}$ is not A -stable, we consider $\operatorname{Aut}\left(\mathrm{Y}_{1}\right):=\left\{g \mathrm{Y}_{1}=\mathrm{Y}_{1}\right\}$.

The arguments in Ch. 3 should go quite smoothly here. In case 1, you may need to do a further perturbation to guarantee the sequence you get has a convergent subsequence.
2.1. New story, general assumption. However, it is unavoidable that $\mathrm{Y}_{0} \backslash \mathscr{O}$ may contain some degenerate $U$-orbit. Let us take a nonempty B minimal set $\mathrm{Y}_{1} \subset \mathrm{Y}_{0} \backslash \mathscr{O}$ whose existence is guaranteed by the nondivergence corollary Coro.1.5. Take some $\Lambda_{1} \in Y_{1}$ such that U. $\Lambda_{1}$ degenerates. We will assume $\mathbb{R} e_{1} \oplus \mathbb{R} e_{2}$ is $\Lambda_{1}$-rational and leave the other case when $\mathbb{R} e_{1}$ is $\Lambda_{1}$-rational to the reader.
2.2. Case 1, no closed U-orbits. Assume $Y_{1}$ contains no closed U-orbit.

As we assumed, $\mathrm{U} . \Lambda_{1}$ is stuck in the following closed set (for simplicity write $\mathbb{R} e_{1,2}:=\mathbb{R} e_{1} \oplus$ $\mathbb{R} e_{2}$ )

$$
\mathrm{X}_{3}\left(\mathbb{R} e_{1,2}, c_{1}\right):=\left\{\Lambda \in \mathrm{X}_{3} \mid \mathbb{R} e_{1,2} \text { is } \Lambda \text {-rational, }\left\|\Lambda \cap \mathbb{R} e_{1,2}\right\|=c_{1}\right\}
$$

where $c_{1}:=\left\|\Lambda_{1} \cap \mathbb{R} e_{1,2}\right\|$. Also let

$$
\mathrm{X}_{2}\left(c_{1}\right):=\left\{\text { lattices in } \mathbb{R}^{2} \text { of covolume } c_{1}\right\} .
$$

Then we have a natural continuous surjection $\pi: \mathrm{X}_{3}\left(\mathbb{R} e_{1,2}, c_{1}\right) \rightarrow \mathrm{X}_{2}\left(c_{1}\right)$ with compact fibres that is equivariant with respect to

$$
\begin{gathered}
\rho_{\pi}:\left\{g \in \mathrm{SL}_{3}(\mathbb{R}), g \text { preserves } \mathbb{R} e_{1,2}, \operatorname{det}\left(\left.g\right|_{\mathbb{R} e_{1,2}}\right)=1\right\} \rightarrow \mathrm{SL}_{2}(\mathbb{R}) \\
\left.g \mapsto g\right|_{\mathbb{R} e_{1,2}} .
\end{gathered}
$$

In particular we have


Now we wish to find a U-minimal set in $\mathrm{Y}_{1}$.
2.3. Case 1.1, some $\pi\left(U . \Lambda_{2}\right)$ is compact. Assume for some $\Lambda_{2} \in Y_{1}, \pi\left(U . \Lambda_{2}\right)$ is closed and hence compact.

Then $\overline{U . \Lambda_{2}}$ is compact (since $\pi$ is a proper map) and let $\overline{U . \Lambda_{3}}$ be a nonempty minimal U -set in $\mathrm{Y}_{2}:=\overline{\mathrm{U} . \Lambda_{2}}$.

Then we can find pairs $\left(x_{n}, y_{n}\right)$ in $\mathrm{Y}_{2}$ such that $y_{n}=\exp \left(w_{n}\right) x_{n}$ with

- $w_{n}=\left[\begin{array}{lll}0 & 0 & * \\ 0 & 0 & * \\ 0 & 0 & 0\end{array}\right], w_{n} \neq 0, w_{n} \rightarrow 0$.


First assume $w_{n} \in \operatorname{Lie}(\mathbb{V})$ for infinitely many $n$, then $\exp \left(w_{n}\right) \in \operatorname{Aut}\left(Y_{2}\right):=\left\{g \in \mathrm{SL}_{3}(\mathbb{R}), g Y_{2}=\right.$ $\left.\mathrm{Y}_{2}\right\}$ and since the latter is a closed subgroup, we have the full $\mathrm{V} \subset \operatorname{Aut}\left(\mathrm{Y}_{2}\right)$.

Otherwise $w_{n}$ is not fixed by $\operatorname{Ad}(\mathrm{U})$ and for any $\delta>0$ and for $n$ large enough we can find $t_{n, \delta}$ such that

- $\left\|\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right\|=\delta$ (i.e. for some constant $C>1$, LHS belongs to $\left(C^{-1} \delta, C \delta\right)$ );
- every limit of $\left(\operatorname{Ad}\left(\mathbf{u}_{t_{n, \delta}}\right) \cdot w_{n}\right)$ is in $\operatorname{Lie}(\mathrm{V})$.

And by taking a limit we find

- $x_{\infty, \delta}, y_{\infty, \delta} \in \mathrm{Y}_{2}$ and $w_{\infty, \delta} \in \operatorname{Lie}(\mathrm{V})$ such that $y_{\infty, \delta}=\exp \left(w_{\infty, \delta}\right) x_{\infty, \delta} ;$
- $w_{\infty, \delta} \neq 0, w_{\infty, \delta} \rightarrow 0$ as $\delta \rightarrow 0$.

Arguing as above, we have $\mathrm{V} \subset \operatorname{Aut}\left(\mathrm{Y}_{2}\right)$.
2.4. Case 1.2, $\pi(\mathrm{U} . \Lambda)$ is never compact. Assume for every $\Lambda \in \mathrm{Y}_{1}, \pi(\mathrm{U} . \Lambda)$ is not compact.

Then there is some compact set such that every $U . \Lambda$ intersects non-trivially for every $\Lambda \in$ $\mathrm{Y}_{1}$. Therefore there is a nonempty U-minimal set in $\mathrm{Y}_{1}$ and the rest of the proof is not so different from Sec.2.3.
2.5. Case 2, exists a closed U-orbit. Assume $Y_{1}$ contains a closed U-orbit U. $\Lambda_{2}$.
2.6. Case 2.1, recurrence in non-centralizer direction. Assume that there exists a sequence $\left(y_{n}\right) \subset \mathrm{Y}_{0}$ converging to $\Lambda_{2}$ such that

- $y_{n}=\exp \left(w_{n}\right) \Lambda_{2}$ with $w_{n} \in \mathfrak{h}^{\perp}, w_{n} \notin \operatorname{Lie}(V)$.
[Recall from Ch.3, $\mathfrak{h}^{\perp}$ is a complement of $\mathfrak{h}=\mathfrak{5 0}_{Q_{0}}(\mathbb{R})$ in $\mathfrak{s l}_{3}(\mathbb{R})$ that is invariant under the adjoint action of $\left.\mathrm{SO}_{Q_{0}}(\mathbb{R})\right]$

Without loss of generality assume $\left\|w_{n}\right\|<1$ for all $n$. Let

$$
t_{n}:=\inf \left\{t \geq 0 \mid\left\|\operatorname{Ad} \mathbf{u}_{t} \cdot w_{n}\right\|=1\right\} .
$$

Let $\varphi_{n}:\left[0, t_{n}\right] \rightarrow\left\{w \in \mathfrak{h}^{\perp} \mid\|w\| \leq 1\right\}$ defined by

$$
\varphi_{n}(t):=\operatorname{Ad}\left(\mathbf{u}_{t}\right) \cdot w_{n} .
$$

By passing to a subsequence, assume $\varphi_{n}\left(\left[0, t_{n}\right]\right)$ converges (in Chabauty topology, or equivalently, w.r.t. Hausdorff distance). Let $E_{\infty}$ denote the limit set. It is connected and closed.
[Side remark: We want something more than the perturbation method as in Ch. 3 could possibly provide.]
2.6.1. Lucky case. Assume there is some $\delta_{0}>0$ such that

$$
E_{\infty} \cap\left\{\|w\| \leq \delta_{0}\right\} \subset \operatorname{Lie}(\mathrm{V}) .
$$

For $n$ large enough s.t. $\left\|w_{n}\right\|<\delta_{0}$, we define

$$
t_{n}\left(\delta_{0}\right):=\inf \left\{t \geq 0 \mid\left\|\operatorname{Ad} \mathbf{u}_{t} \cdot w_{n}\right\|=\delta_{0}\right\}
$$

By passing to a further subsequence we assume $\varphi_{n}\left(\left[0, t_{n}\left(\delta_{0}\right)\right]\right)$ converges to $E_{\infty}\left(\delta_{0}\right)$. By our assumption, $E_{\infty}\left(\delta_{0}\right)$ is a connected subset of $\left\{w \in \operatorname{Lie}(\mathrm{~V}),\|w\| \leq \delta_{0}\right\}$. So it is an interval. Since U commutes with $V$, for every $w \in E_{\infty}, \exp (w) \mathrm{U} . \Lambda_{2} \subset \mathrm{Y}_{0}$. Therefore,

$$
\mathbf{v}_{\left[0, \delta_{0}\right]} \cdot \Lambda_{2} \subset \mathrm{Y}_{0}, \text { or } \mathbf{v}_{\left[-\delta_{0}, 0\right]} \cdot \Lambda_{2} \subset \mathrm{Y}_{0} .
$$

W.L.O.G, assume $\mathbf{v}_{\left[0, \delta_{0}\right]} \cdot \Lambda_{2} \subset \mathrm{Y}_{0}$. Hence for every $t, s \in \mathbb{R}$,

$$
\begin{equation*}
\mathbf{v}_{\left[0, e^{2 t} \delta_{0}\right]} \mathbf{a}_{t} \mathbf{u}_{s} \cdot \Lambda_{2}=\mathbf{a}_{t} \mathbf{u}_{s} \mathbf{v}_{\left[0, \delta_{0}\right]} \cdot \Lambda_{2} \subset \mathbf{a}_{t} \mathbf{u}_{s} \mathrm{Y}_{0}=\mathrm{Y}_{0} \tag{20}
\end{equation*}
$$

By Thm.1.4, there exists a compact set such that for every $t>0$, there exists $s_{t}>0$ such that $\mathbf{a}_{t} \mathbf{u}_{s_{t}} . \Lambda_{2}$ lives in this compact set. In particular we may select $t_{n} \rightarrow+\infty$ and $s_{n} \in \mathbb{R}$ such that $\lim a_{t_{n}} s_{n} . \Lambda_{2}$ exists and call it $\Lambda_{\infty}$. Then by Equa.(20) and a continuity argument, we have

$$
\mathbf{v}_{[0,+\infty)} \cdot \Lambda_{\infty} \subset Y_{0} .
$$

So we are done.
2.6.2. Unlucky, try again! If the assumption in Sec.2.6.1 does not hold, then we can repeat what is done above the Sec.2.6.1. So we get some $E_{\infty}^{(2)}$. If lucky, then we go back to Sec.2.6.1. If not, then we can repeat this process again to get $E_{\infty}^{(3)}$. It suffices to note that this process should stop.

Indeed recall the computation we made in Ch.3, Equa.(10)
$\operatorname{Ad}\left(\mathbf{u}_{t}\right) w=$

$$
\left[\begin{array}{ccc}
\frac{t^{2}}{2} w_{31}+t w_{21}+w_{11} & \frac{t^{3}}{3!} w_{31}+\frac{t^{2}}{2} w_{21}+t w_{11}+\frac{-w_{12}}{3} & \frac{t^{4}}{4!} w_{31}+\frac{t^{3}}{3!} w_{21}+\frac{t^{2}}{2} w_{11}+t \frac{-w_{12}}{3}+\frac{w_{13}}{6}  \tag{21}\\
t w_{31}+w_{21} & * & * \\
w_{31} & * & *
\end{array}\right] .
$$

If this process would continue, from the computation we sees right away that for $w_{\infty}^{(i)} \in E_{\infty}^{(i)}$, $w_{\infty} \in E_{\infty}$,

$$
\begin{aligned}
\left(w_{\infty}\right)_{3,1} & =0 ; \\
\left(w_{\infty}^{(2)}\right)_{3,1} & =\left(w_{\infty}^{(2)}\right)_{2,1}=0 ; \\
\left(w_{\infty}^{(3)}\right)_{3,1} & =\left(w_{\infty}^{(3)}\right)_{2,1}=\left(w_{\infty}^{(3)}\right)_{1,1}=0 ; \\
\left(w_{\infty}^{(4)}\right)_{3,1} & =\left(w_{\infty}^{(4)}\right)_{2,1}=\left(w_{\infty}^{(4)}\right)_{1,1}=\left(w_{\infty}^{(4)}\right)_{1,2}=0 \Longrightarrow w_{\infty}^{(4)} \in \operatorname{Lie}(\mathrm{V}) .
\end{aligned}
$$

Thus we are always lucky at some point.
2.7. Case 2.2, recurrence only in centralizer direction. Assume the assumption made in Sec.2.6 is wrong. This can be rephrased as saying that there exists some $\delta_{0}>0$, assumed to be much smaller than $\operatorname{Inj} \operatorname{Rad}\left(\Lambda_{2}\right)$, such that

$$
\operatorname{Obt}^{-1}\left(\operatorname{Map}\left(\Lambda_{2}, \mathrm{Y}_{0}\right) \cap \mathscr{N}_{\delta_{0}}(\mathrm{id})\right) \subset \mathfrak{h} \oplus \operatorname{Lie}(\mathrm{V})
$$

where Obt $: \mathfrak{h} \oplus \mathfrak{h}^{\perp} \rightarrow \mathrm{SL}_{3}(\mathbb{R})$ is a local diffeomorphism (around $(0,0)$ ) defined by

$$
\operatorname{Obt}(h, w):=\exp (h) \exp (w) .
$$

This is the last and the most annoying case. It is here that we are using the fact that $\mathrm{Y}_{1}$ is B-minimal. We are going to derive a contradiction and show that this case is not allowed. The argument below is a more-or-less reproduction of [BM00, Page 182].
2.7.1. Step 1. $\mathrm{Y}_{1}$ is not a closed B-orbit.

Indeed, otherwise, one sees that $\mathrm{Y}_{1}$ is even compact by Thm.1.4. But this is impossible by considering $\mathbf{a}_{t} . \Lambda_{2}$ as $t \rightarrow-\infty$.
2.7.2. Step 2. Step 1 together with minimality imply that there exists $b_{n}=a_{n} u_{n} \in B$ with $a_{n} \rightarrow \infty$ such that $b_{n} . \Lambda_{2} \rightarrow \Lambda_{2}$. Note that if we write $a_{n}=\mathbf{a}_{t_{n}}$ then $t_{n} \rightarrow+\infty$.
2.7.3. Step 3. Since $\mathrm{Y}_{1} \subset \mathrm{Y}_{0} \backslash \mathscr{O}$ and by our assumption made in this subsection, we find $\left(v_{n}\right) \subset \operatorname{Lie}(\mathrm{V})$ such that $v_{n} \neq 0, v_{n} \rightarrow 0$ and $\exp \left(v_{n}\right) \cdot \Lambda_{2} \in \mathrm{Y}_{0}$ for all $n$.
2.7.4. Step 4. This is the key step.

Since $b_{n} . \Lambda_{2} \rightarrow \Lambda_{2}$, we can find for every large $n$, a unique $\lambda_{n}$ close to id such that $b_{n} \cdot \Lambda_{2}=$ $\lambda_{n} . \Lambda_{2}$. By assumption one can write $\lambda_{n}=h_{n} \exp \left(\nu\left(\lambda_{n}\right)\right)$ for some $h_{n} \in \mathrm{H}$ and $\nu\left(\lambda_{n}\right) \in \operatorname{Lie}(\mathrm{V})$. We want to argue that $h_{n} \in \mathrm{~B}$.

Now fix some large $n$ and will take $l$ large compared to $n$. We have

$$
b_{n} \cdot\left(\exp \left(v_{l}\right) \cdot \Lambda_{2}\right)=\exp \left(v_{l}^{\prime}\right) \cdot b_{n} \cdot \Lambda_{2}=\exp \left(v_{l}^{\prime}\right) \cdot \lambda_{n} \cdot \Lambda_{2}
$$

where $v_{l}^{\prime}=\operatorname{Ad}\left(b_{n}\right) \cdot v_{l} \in \operatorname{Lie}(\mathrm{~V})$. When $l$ is large compared to $n, v_{l}^{\prime}$ is small.
By assumption for $n$ large and $l$ larger,

$$
\exp \left(v_{l}^{\prime}\right) \cdot \lambda_{n}=\exp \left(v_{l}^{\prime}\right) h_{n} \exp \left(\nu\left(\lambda_{n}\right)\right) \in \mathscr{N}_{\mathrm{id}}\left(\delta_{0}\right) \cap \mathrm{H} \cdot \mathrm{~V}
$$

Although the computation of $\log (\exp (X) \exp (Y))$ is usually hard, we still have (again, for $l$ large)

$$
\begin{aligned}
& \exp \left(v_{l}^{\prime}\right) h_{n} \exp \left(v\left(\lambda_{n}\right)\right) \in \mathscr{N}_{\mathrm{id}}\left(\delta_{0}\right) \cap \mathrm{H} \cdot \mathrm{~V} \Longrightarrow \exp \left(\operatorname{Ad}\left(h_{n}^{-1}\right) \cdot v_{l}^{\prime}\right) \in \mathscr{N}_{\mathrm{id}}\left(\delta_{0}\right) \cap \mathrm{H} \cdot \mathrm{~V} \\
& \quad \Longrightarrow \exp \left(\operatorname{Ad}\left(h_{n}^{-1}\right) \cdot v_{l}^{\prime}\right) \in \mathscr{N}_{\mathrm{id}}\left(\delta_{0}\right) \cap \mathrm{V} \Longrightarrow \operatorname{Ad}\left(h_{n}^{-1}\right) \cdot v_{l}^{\prime} \in \operatorname{Lie}(\mathrm{V})
\end{aligned}
$$

[some $\delta_{0}$ should be smaller than the others, we leave it to the reader to fill in the details]
As $l$ varies, $v_{l}^{\prime}$ spans Lie $(\mathrm{V})$. Thus $h_{n}$ preserves $\mathrm{Lie}(\mathrm{V})$ and is contained in $\pm \mathrm{B}$, the normalizer of V in G. Since $h_{n}$ is close to the identity, $h_{n}$ belongs to B. [Rmk: since $h_{n}$ is close to identity, this is a Lie algebraic calculation of $\mathfrak{n}_{\mathfrak{g}}(\operatorname{Lie}(\mathrm{V}))$, the normalizer of $\operatorname{Lie}(\mathrm{V})$ in $\mathfrak{g}:=\mathfrak{s l}_{3}(\mathbb{R})$. That is, it suffices to compute the connected component of $N_{G}(V)$.]

Here is a pictorial summary:

2.7.5. Step 5 . Step 4 says that

$$
b_{n} \cdot \Lambda_{2}=h_{n} \exp \left(v\left(\lambda_{n}\right)\right) \cdot \Lambda_{2}
$$

for some $h_{n} \in$ B close to the identity. This is impossible! Why? Note that $\Lambda_{2}$ is a periodic U-orbit and everything here normalizes U. Hence both sides are U-periodic. However, the centralizer of $U$ preserves the period but $\mathbf{a}_{t_{n}}$ (recall $b_{n}=\mathbf{a}_{t_{n}} u_{n}$ with $t_{n} \rightarrow+\infty$ ) makes the period much larger. This is a contradiction.

## 3. Proof of Theorem 1.4

From now on we discuss how Thm. 1.4 is proved. From the start assume item 2 and 3 do not happen and want to prove item 1 holds.

A direct computation shows that a vector $v \in \mathbb{R}^{3} \oplus \wedge^{2} \mathbb{R}^{3}$ is fixed by $U$ iff $v \in \mathbb{R} e_{1} \oplus \mathbb{R} e_{1} \wedge e_{2}$. For a primitive subgroup $\Delta$ of $\Lambda_{0} \in X_{3}$, we still denote by $\Delta$ the vector (well-defined up to $\pm 1$ ) representing $\Delta$. For instance if $\Delta=\mathbb{Z} v \oplus \mathbb{Z} w$, then $\Delta$ is viewed as a vector $\pm \nu \wedge w \in \wedge^{2} \mathbb{R}^{3}$. Now assume $U . \Lambda_{0}$ does not degenerate, then every nonzero subgroup $\Delta$ is not fixed by $U$ and by the feature of polynomials,

$$
\lim _{t \rightarrow-\infty}\left\|\mathbf{u}_{t} \cdot \Delta\right\|=\lim _{t \rightarrow+\infty}\left\|\mathbf{u}_{t} \cdot \Delta\right\|=+\infty
$$

We can ensure at least the trajectory under $U$ of each subgroup can not be small for a long time:

Lemma 3.1. There exist $C_{2}>0$ and $\alpha_{2}>0$ such that for every interval $[a, b]$ in $\mathbb{R}$, every $\mathbf{x} \in \mathbb{R}^{3} \oplus \wedge^{2} \mathbb{R}^{3}$ and every $\rho \in(0,1)$, if $M_{0}:=\sup _{s \in[a, b]}\left\|\mathbf{u}_{s} \mathbf{x}\right\|$, then

$$
\frac{1}{b-a} \operatorname{Leb}\left\{s \in[a, b] \mid\left\|\mathbf{u}_{s} \mathbf{x}\right\|<\rho M_{0}\right\} \leq C_{2} \rho^{\alpha_{2}} ;
$$

The proof is left as an exercise.
The key observation we made last time does not hold anymore. The following notion is aimed to save the situation, providing a sufficient condition for being contained in a compact set.

Definition 3.2. For $\delta, \rho \in(0,1), \Lambda \in X_{3}$ is said to be $(\delta, \rho)$-protected (by the flag $\{\{0\} \subset \mathbb{Z} v \subset$ $\Delta \subset \Lambda\}$ ) iff there exists $0 \subset \mathbb{Z} v \subset \Delta \subset \Lambda$ where $\mathbb{Z} v$ and $\Delta$ are primitive subgroups of rank 1 and 2 such that

$$
\|v\|,\|\Delta\| \in(\rho \delta, \delta)
$$

Lemma 3.3. Take $\delta, \rho \in(0,1)$. If $\Lambda \in \mathrm{X}_{3}$ is $(\delta, \rho)$-protected then $\Lambda \in \mathscr{C} \rho$.
Proof. It suffices to prove that every non-zero vector $w$ in $\Lambda$ has norm at least $\rho$. So we may assume that $\|w\|<1$.

Pick $v$ and $\Delta$ as in the definition. Because $\Lambda$ is of covolume one, $w$ has to be contained in $\Delta$ since $\|w\|<1$. Moreover

$$
\rho \delta \leq\|\Delta\| \leq\|v\| \cdot\|w\| \leq \delta\|w\| \Longrightarrow\|w\| \geq \rho .
$$

Key observation. Here we have already employed the special feature of $\mathrm{X}_{3}$ (not valid for $\mathrm{X}_{\geq 4}$ ): once we find $\mathbb{Z} v$ and $\Delta$ two primitive subgroups such that $\|\mathbb{Z} v\|,\|\Delta\|<1$, then it is automatic that $\mathbb{Z} v$ is contained in $\Delta$. Therefore, in searching for a flag that ( $\delta, \rho$ )-protects $\Lambda$ we may look for $\mathbb{Z} v$ and $\Delta$ in an independent way (the condition of $\mathbb{Z} v \subset \Delta$ automatically holds).

Thus Thm.1.4 follows from Lem.3.3, the key observation and the following:
Lemma 3.4. For every $\varepsilon>0$, there exist $\varepsilon^{\prime}, \rho, \delta \in(0,1)$ such that for every $\Lambda$ nondegenerate, there exists $T_{0}$ such that for all $T \geq T_{0}$,

$$
\frac{1}{T} \operatorname{Leb}\left\{t \in[0, T] \mid \nexists \mathbf{x} \in \operatorname{Prim}^{1}\left(\mathbf{u}_{t} \Lambda\right),\|\mathbf{x}\| \in(\rho \delta, \delta), \mathbf{u}_{t} . \Lambda \notin \mathscr{C}_{\varepsilon^{\prime}}\right\} \leq \varepsilon
$$

and

$$
\frac{1}{T} \operatorname{Leb}\left\{t \in[0, T] \mid \nexists \Delta \in \operatorname{Prim}^{2}\left(\mathbf{u}_{t} \Lambda\right),\|\Delta\| \in(\rho \delta, \delta), \mathbf{u}_{t} . \Lambda \notin \mathscr{C}_{\varepsilon^{\prime}}\right\} \leq \varepsilon
$$

If we fix a compact set in $\mathrm{X}_{3}$ from the beginning and allow $\varepsilon^{\prime}, \rho, \delta$ to depend on this compact set, then conclusion holds for $T_{0}=0$ and all $\Lambda$ contained in this compact set.

Proof of Lemma 3.4. Let $C_{2}, \alpha_{2}$ be as in Lem.3.1.
Fix some $\delta \in(0,1)$. Take $\varepsilon^{\prime}:=\delta / 2$. Choose $\rho \in(0,1)$ small enough such that $C_{2}(2 \rho)^{\alpha_{2}}<$ $0.5 \varepsilon$. Assume that $\Lambda$ contains no degenerate vectors. We are going to prove the first inequality and the second one can be proved similarly, which is left as an exercise.

By taking $T_{0}$ large enough, we assume that for every $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$, for some $t \in\left(0, T_{0}\right)$, $\left\|\mathbf{u}_{t} \cdot v\right\| \geq \delta$ (and we can forget about the non-degeneracy condition from now on).

Indeed, take $t=1$, there are only finitely many $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$ such that $\left\|\mathbf{u}_{t} \cdot v\right\|<\delta$. List them as $\left\{\mathbb{Z} v_{1}, \ldots, \mathbb{Z} v_{l}\right\}$. Since $\mathbf{u}_{t}$ does not fix $v_{i}$ for every $i$ by non-degeneracy condition, we have that $\left\|\mathbf{u}_{t} . v_{i}\right\| \rightarrow+\infty$ as $t \rightarrow+\infty$. So we can pick $T_{0}$ such that $\left\|\mathbf{u}_{T_{0}} . v_{i}\right\|>\delta$ for every $i$ and this would do the job.

Consider the set $\left\{t \in(0, T), \mathbf{u}_{t} . \Lambda \notin \mathscr{C}_{\delta / 2}\right\}$, which is open and hence can be written as a disjoint union of open intervals. Take one of them, say $(a, b)$. At the moment, we have not excluded the possibility of $(a, b)=(0, T)$ yet.

For every $t \in(a, b)$, by definition, there is some $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$ such that $\left\|\mathbf{u}_{t} \cdot v\right\|<\delta / 2$. For every such $\mathbb{Z} v$ and $t$, define $\mathscr{O}(\mathbb{Z} v, t)$ to be the maximal open interval in $\mathbb{R}$ containing $t$ such that

$$
s \in \mathscr{O}(\mathbb{Z} v, t) \Longrightarrow\left\|\mathbf{u}_{t} \cdot v\right\|<\delta .
$$

From the definition, it is possible that $\mathscr{O}(\mathbb{Z} v, t)$ is not contained in $(a, b)$, or even $(0, T)$. On the other hand, it is impossible for $(0, T)$ to be contained in $\mathscr{O}(\mathbb{Z} v, t)$ by the choice of $T_{0}$. Thus,

$$
\sup _{(0, T) \cap(\mathbb{Z} v, t)}\left\|\mathbf{u}_{t} \cdot v\right\| \geq \delta .
$$

If $(a, b)$ contains some end point of $\mathscr{O}(\mathbb{Z} v, t)$ then this also holds replacing $(0, T)$ by $(a, b)$. Otherwise, we must have for $t=a$ or $t=b, \operatorname{sys}\left(\mathbf{u}_{t} . \Lambda\right)=\delta / 2$. No matter what, the following always holds

$$
\begin{equation*}
\sup _{(a, b) \cap \tilde{O}(\mathbb{Z}, t)}\left\|\mathbf{u}_{t} \cdot v\right\| \geq \frac{\delta}{2} . \tag{22}
\end{equation*}
$$

As $\mathbb{Z} v, t$ varies, $\{\mathscr{O}(\mathbb{Z} v, t) \cap(a, b)\}$ covers $(a, b)$. Now we claim that it is possible to select a subcovering with multiplicity at most 2 (the number 2 is not important, but it should be an absolute constant). The multiplicity of a covering refers to the maximal number of possible overlaps. Here is one possible way of proving the claim, you may wish to find your own.

Since each of $a, b$ belongs to some $\mathscr{O}(\mathbb{Z} v, t)$, we can find a finite collection of $\{\mathscr{O}(\mathbb{Z} v, t)\}$ that covers $(a, b)$. By passing to a further sub-covering if necessary, we assume it is minimal and is given by $\left\{\mathscr{O}\left(\mathbb{Z} v_{i}, t_{i}\right)=\left(a_{i}, b_{i}\right)\right\}$ with $a_{i}<a_{i+1}$. Then we must have

$$
a_{1}<a<a_{2}<b_{1}<a_{3}<b_{2}<a_{4}<\ldots .<a_{l}<b_{l-1}<b<b_{l},
$$

and the claim follows.
Let $I_{i}:=\left(a_{i}, b_{i}\right) \cap(a, b)$. By Equa.(22), $\sup _{s \in I_{i}}\left\|\mathbf{u}_{s} . v_{i}\right\| \geq \frac{\delta}{2}$. Then by Lem.3.1,

$$
\frac{1}{\left|I_{i}\right|}\left|\left\{s \in I_{i} \left\lvert\,\left\|\mathbf{u}_{s} \cdot v_{i}\right\| \leq(2 \rho) \cdot \frac{\delta}{2}\right.\right\}\right| \leq C_{2}(2 \rho)^{\alpha_{2}} \leq 0.5 \varepsilon
$$

Adding them together completes the proof.

## 4. Exercises

4.1. ( $C, \alpha$ )-good property of polynomials of bounded degree. Let $C, \alpha>0$ and $J$ be an interval in $\mathbb{R}$, recall a function $f: J \rightarrow \mathbb{R}$ is said to be $(C, \alpha)$-good on $J$ iff for every interval $I \subset J$ of finite length and every $\rho \in(0,1)$,

$$
\begin{equation*}
\frac{1}{|I|} \operatorname{Leb}\left\{t \in I\left||f(t)| \leq \rho M_{I}\right\} \leq C \rho^{\alpha} .\right. \tag{23}
\end{equation*}
$$

where $M_{I}:=\sup _{t \in I}|f(t)|$.
In this set of exercises we show that there are constants ( $C, \alpha$ ) such that every polynomial of degree at most three is $(C, \alpha)$-good on $\mathbb{R}$. The general case would follow from the same proof with some constant depending only on the degree.

Given four distinct points $\mathbf{v}=\left(\nu_{0}, \nu_{1}, \nu_{2}, \nu_{3}\right)$ in $\mathbb{R}$, for $k=0,1,2,3$, define

$$
L_{\mathbf{v}}^{k}(x):=\prod_{i \neq k} \frac{x-v_{i}}{v_{k}-v_{i}} .
$$

EXercise 4.1. Fix such av as above. Prove that for any choice offour real numbers ( $w_{0}, w_{1}, w_{2}, w_{3}$ ), there exists at most one polynomial $p$ of degree at most 3 such that $p\left(v_{i}\right)=w_{i}$.

EXERCISE 4.2. Same assumption as in last exercise. Show that $p(x):=\sum_{k=0}^{3} w_{k} \cdot L_{\mathbf{v}}^{k}(x)$ satisfies $p\left(v_{i}\right)=w_{i}$ for every $i=0,1,2,3$.

Exercise 4.3. Same assumption as in last exercise. Let $\varepsilon, \delta>0$ be two positive real numbers. Assume further that $\left|v_{i}-v_{j}\right| \geq \delta$ for every pair $(i, j)$ with $i \neq j$. Also assume $\left|w_{i}\right| \leq \varepsilon$ for all $i$. Show that for every $x \in[0,1],|p(x)| \leq 4 \varepsilon \delta^{-3}$ where $p$ is as in the last exercise.

Exercise 4.4. Let $I \subset[0,1]$ be a measurable subset with $\operatorname{Leb}(I)=9 \delta>0$. Show that there exists four points ( $\nu_{0}, v_{1}, v_{2}, v_{3}$ ) in I such that $\left|\nu_{i}-v_{j}\right| \geq \delta$ for every pair $(i, j)$ with $i \neq j$.

Exercise 4.5. Find $C, \alpha>0$ such that for every polynomial of degree at most three and $\rho \in(0,1)$, Equa.(26) holds when $I=[0,1]$.

Exercise 4.6. Show that every polynomial of degree at most three is ( $C, \alpha$ )-good on $\mathbb{R}$ with $C, \alpha$ same as in the last exercise.

## CHAPTER 6

## Nondivergence of unipotent flows on $X_{N}$

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Notations

- $\mathrm{X}_{N}:=\left\{\right.$ unimodular lattices in $\left.\mathbb{R}^{N}\right\} \cong \mathrm{SL}_{N}(\mathbb{R}) / \mathrm{SL}_{N}(\mathbb{Z})$;
- for a discrete subgroup $\Delta$ in $\mathbb{R}^{N}$, let $\|\Delta\|:=\operatorname{Vol}\left(\Delta_{\mathbb{R}} / \Delta\right)$ where $\Delta_{\mathbb{R}}$ denotes the $\mathbb{R}$-linear span of $\Delta$ in $\mathbb{R}^{N}$;
- for $\Lambda \leq \mathbb{R}^{N}$, sys $(\Lambda):=\inf _{v_{\neq 0} \in \Lambda}\|\nu\|$, and for $\delta>0, \mathscr{C}_{\delta}:=\left\{\Lambda \in \mathrm{X}_{N}: \operatorname{sys}(\Lambda) \geq \delta\right\}$;
- $\operatorname{Prim}^{k}(\Lambda):=\{$ primitive subgroups of $\Lambda$ of rank $k\}, \operatorname{Prim}(\Lambda):=\cup_{k=0}^{\operatorname{rank}(\Lambda)} \operatorname{Prim}^{k}(\Lambda)$.


## 1. Summary and definitions

We would like to illustrate the main ideas behind [Kle10, Section 3] using $\mathrm{X}_{4}$ as an example. The discussion can be generalized to $\mathrm{X}_{N}$ and even to $\mathbf{G}(\mathbb{R}) / \mathbf{G}(\mathbb{Z})$ for other semisimple algebraic groups $\mathbf{G}$. Warning: our presentation and sometimes definitions differ from [Kle10, Section 3] and is "less careful" in many ways.

The discussion is useful beyond unipotent flows on $\mathrm{X}_{N}$. We would like to mention [EMS97, MW02] here.

Definition 1.1. Fix $(C, \alpha)$ two positive constants. A map $\phi: I \rightarrow \mathrm{SL}_{N}(\mathbb{R})$ is said to be $(C, \alpha)$ good at $\Lambda \in \mathrm{X}_{N}$ iffor every primitive subgroup $\Delta$ of $\Lambda$, every interval $J \subset I$, every $\rho \in(0,1)$ (the case $\rho \geq 1$ is rather trivial), define $M(J, \Delta):=\sup _{s \in J}\left\|\phi_{s} . \Delta\right\|$, then we have

$$
\frac{1}{|J|}\left|\left\{s \in J \mid\left\|\phi_{s} \cdot \Delta\right\| \leq \rho \cdot M(J, \Delta)\right\}\right| \leq C \cdot \rho^{\alpha} .
$$

The main examples for us are unipotent flows.
Lemma 1.2. There are constants $C_{N}, \alpha_{N}>0$, depending only on $N$ such that for every nilpotent matrix $u$ in $\mathfrak{s l}_{N}(\mathbb{R})$ and for every (finite or infinite) interval I in $\mathbb{R}, \phi(t):=\exp (t . u)$ is $\left(C_{N}, \alpha_{N}\right)$-good at every $\Lambda \in \mathrm{X}_{N}$.

Proof. Exercise or see [Kle10].
Theorem 1.3. Fix $C, \alpha, \varepsilon, \delta$ positive constants. There exists a constant $\kappa=\kappa(C, \alpha, \varepsilon, \delta)>0$ such that the following holds. Let $\Lambda \in \mathrm{X}_{N}$ and $\phi: I \rightarrow \mathrm{SL}_{N}(\mathbb{R})$. Assume

- $\phi$ is $(C, \alpha)$-good at $\Lambda$;
- $\sup _{t \in I}\left\|\phi_{t} . \Delta\right\| \geq \delta$ for every $\Delta \in \operatorname{Prim}(\Lambda)$,
then

$$
\frac{1}{|I|} \operatorname{Leb}\left\{s \in I \mid \phi_{s} . \Lambda \notin \mathscr{C}_{\kappa}\right\} \leq \varepsilon .
$$

In the case of unipotent flows and an interval $I$ of infinite length, if the condition fails, then $\Lambda$ contains a primitive subgroup fixed by the unipotent flow with small norm.

## 2. Nondivergence and flags

The key notion is being $(\delta, \rho)$-protected, which provides a sufficient condition to guarantee non-divergence.

## Definition 2.1.

A subset $\mathscr{F}$ of $\operatorname{Prim}(\Lambda)$ is said to be a flag iffor every two elements $\Delta_{1}$ and $\Delta_{2}$ in $\mathscr{F}$, either $\Delta_{1} \subset \Delta_{2}$ or $\Delta_{1} \supset \Delta_{2}$. The length of a flag $\mathscr{F}$ is simply the cardinality of $\mathscr{F}$.

Definition 2.2.
Let $\delta, \rho \in(0,1)$. Let $\Lambda \in \mathrm{X}_{N}$ and $\mathscr{F}=\left\{\Delta_{1} \subsetneq \Delta_{2} \subsetneq \ldots \subsetneq \Delta_{l}\right\}$ be a flag in $\operatorname{Prim}(\Lambda)$. We say that $\Lambda$ is weakly $(\delta, \rho)$-protected by $\mathscr{F}$ iff

1. $\rho \cdot \delta \leq\left\|\Delta_{i}\right\| \leq \delta$ for every $i=1, . ., l$;
2. $\|\Delta\| \geq 0.5 \delta$ for every $\Delta \notin \mathscr{F}$ comparable with $\mathscr{F}$, i.e. $\mathscr{F} \cup\{\Delta\}$ is still a flag.

Now given a map $\phi: I \rightarrow \mathrm{SL}_{N}(\mathbb{R})$. We say that $s \in I$ is weakly $(\delta, \rho)$-protected by $\mathscr{F}$ iff

1. $\rho \cdot \delta \leq\left\|\phi_{s} . \Delta_{i}\right\| \leq \delta$ for every $i=1, . ., l$;
2. $\left\|\phi_{s} . \Delta\right\| \geq 0.5 \delta$ for every $\Delta \notin \mathscr{F}$ comparable with $\mathscr{F}$.

That is to say, $\phi_{s} . \Lambda$ is weakly $(\delta, \rho)$-protected by $\phi_{s} . \mathscr{F}$.
I shall drop the word "weakly" later. But keep in mind our definition is different from [Kle10] where $0.5 \delta$ is replaced by $\delta$.

From the definition, such a flag is not allowed to contain $\{0\}$ or $\Lambda$. Thus the maximal possible length is $N-1$.

One may wish to compare with the definition of Siegel sets.
Lemma 2.3 (Criterion of non-divergence in terms of flags). Fix $\delta, \rho \in(0,1)$. Assume for some reason that $\rho<0.5$. Then there exists a constant $\theta=\theta(\delta, \rho)>0$ (from the proof, can take $\theta=\rho^{N} \delta$ ) such that if $\Lambda \in \mathrm{X}_{N}$ is $(\delta, \rho)$-protected by some flag $\mathscr{F}$ of $\operatorname{Prim}(\Lambda)$, then $\|\Delta\| \geq \theta$ for every primitive subgroup $\Delta \leq \Lambda$. In particular $\operatorname{sys}(\Lambda) \geq \theta$.


Proof of a special case. Say $\mathscr{F}=\left\{\Delta_{1} \leq \Delta_{2}\right\}$, which gives a filtration of $\Lambda$. For $v \in \Lambda$, there are three cases. We will show $\|\nu\| \geq \rho \delta$.

Case l. $v \in \Lambda \backslash \Delta_{2}$.
Then $\Delta_{2}+\mathbb{Z} . v$ is compatible with $\mathscr{F}$, though it may not be primitive. $\left(\left(\Delta_{2}\right)_{\mathbb{R}}+\mathbb{R} . v\right) \cap \Lambda$ is a primitive subgroup compatible with $\mathscr{F}$ and contains $\Delta_{2}+\mathbb{Z} . v$. Thus

$$
\left\|\Delta_{2}+\mathbb{Z} \cdot v\right\| \geq\left\|\left(\left(\Delta_{2}\right)_{\mathbb{R}}+\mathbb{R} \cdot v\right) \cap \Lambda\right\| \geq 0.5 \delta .
$$

On the other hand

$$
\left\|\Delta_{2}+\mathbb{Z} \cdot v\right\| \leq\left\|\Delta_{2}\right\| \cdot\|v\| \leq \delta\|v\|
$$

Combined together gives $\|v\| \geq 0.5$.

Case 2. $v \in \Delta_{2} \backslash \Delta_{1}$.
Either $\Delta_{1}+\mathbb{Z} . v$ has the same rank as $\Delta_{2}$ or not. Anyway, we always have,

$$
\left\|\Delta_{1}+\mathbb{Z} . v\right\| \geq \min \{\rho \delta, 0.5 \delta\}=\rho \delta .
$$

On the other hand

$$
\left\|\Delta_{1}+\mathbb{Z} \cdot v\right\| \leq\left\|\Delta_{1}\right\| \cdot\|v\| \leq \delta\|v\| .
$$

Combined together gives $\|v\| \geq \rho$.
Case 3. $v \in \Delta_{1}$.
Then either $\mathbb{Z} . v$ has the same rank as $\Delta_{1}$, in which case $\|\mathbb{Z} . v\| \geq\left\|\Delta_{1}\right\| \geq \rho \delta$, or $\mathbb{Z} . v$ has smaller rank than $\Delta_{1}$, in which case $\|\mathbb{Z} . v\| \geq 0.5 \delta \geq \rho \delta$.

Proof in General. [Read this only if you feel necessary!] Let $\mathscr{F}=\left\{\Delta_{1} \subsetneq \Delta_{2} \subsetneq \ldots \subsetneq \Delta_{l}\right\}$ be the flag and $\Delta$ is a primitive subgroup of $\Lambda$. Let $V_{k}:=\mathbb{R}^{N} /\left(\Delta_{k}\right)_{\mathbb{R}}$ and $\pi_{k}$ be the natural quotient $\operatorname{map} \mathbb{R}^{N} \rightarrow V_{k}$.

Note that if $\Delta^{\prime} \leq \Lambda$ is contained in $\Delta_{k}$ for some $k \in\{1, \ldots, l\}$, then

$$
\begin{equation*}
\left\|\pi_{k-1}\left(\Delta^{\prime}\right)\right\|_{V_{k-1}}=\left\|\pi_{k-1}\left(\Delta^{\prime}+\Delta_{k-1}\right)\right\|_{V_{k-1}}=\frac{\left\|\Delta^{\prime}+\Delta_{k-1}\right\|}{\left\|\Delta_{k-1}\right\|} \geq \frac{\min \{\rho \delta, 0.5 \delta\}}{\delta} \geq \rho . \tag{24}
\end{equation*}
$$

Let $a$ be the largest index such that $\Delta_{a}$ is contained in $\Delta$. By default, $\Delta_{0}:=\{0\}$ if $\Delta_{1} \neq\{0\}$. If $a=l$, then we are done with $\theta=\rho \delta$. Assume otherwise.

$$
\begin{aligned}
\|\Delta\|= & \left\|\pi_{a+1}(\Delta)\right\|_{V_{a+1}} \cdot\left\|\Delta \cap \Delta_{a+1}\right\|=\left\|\pi_{a+1}\left(\Delta+\Delta_{a+1}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\| \\
= & \left\|\pi_{a+2}(\Delta)\right\|_{V_{a+2}} \cdot\left\|\pi_{a+1}\left(\Delta+\Delta_{a+1}\right) \cap \pi_{a+1}\left(\Delta_{a+2}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\| \\
= & \left\|\pi_{a+2}(\Delta)\right\|_{V_{a+2}} \cdot\left\|\pi_{a+1}\left(\left(\Delta+\Delta_{a+1}\right) \cap \Delta_{a+2}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\| \\
= & \left\|\pi_{a+2}(\Delta)\right\|_{V_{a+2}} \cdot\left\|\pi_{a+1}\left(\Delta \cap \Delta_{a+2}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\| \\
& \cdots \ldots . \\
= & \left\|\pi_{a+k-1}\left(\Delta \cap \Delta_{a+k}\right)\right\|_{V_{a+k-1}} \cdot \ldots \cdot\left\|\pi_{a+1}\left(\Delta \cap \Delta_{a+2}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\| \\
= & \left\|\pi_{a+k-1}\left(\Delta \cap \Delta_{a+k}\right)\right\|_{V_{a+k-1}} \cdot \ldots \cdot\left\|\pi_{a+1}\left(\Delta \cap \Delta_{a+2}\right)\right\|_{V_{a+1}} \cdot\left\|\pi_{a}\left(\Delta \cap \Delta_{a+1}\right)\right\|_{V_{a}} \cdot\left\|\Delta_{a}\right\|
\end{aligned}
$$

where $k$ is the smallest positive integer such that $\Delta$ is contained in $\Delta_{a+k}$. By invoking Equa.(24),

$$
\|\Delta\| \geq \rho^{k} \delta
$$

So we are done by taking $\theta:=\rho^{N} \delta$.

## 3. The proof

Instead of proving by induction, we have decided to unfold this process. This makes the proof much longer but hopefully less mysterious. Here is a guide for Step 1-3.


Step 1. By assumption for every $\Delta \in \operatorname{Prim}(\Lambda)$,

$$
\sup _{s \in I}\left\|\phi_{s} . \Delta\right\| \geq \delta
$$

Consider the open subset

$$
I^{\prime}:=\left\{s \in I \mid \exists \Delta \in \operatorname{Prim}(\Lambda),\left\|\phi_{s} . \Delta\right\|<0.9 \delta\right\}
$$

Write it as a disjoint union of open intervals

$$
I^{\prime}=\bigsqcup_{a \in \mathscr{I}_{0}} I_{a} .
$$

Thus for every $\Delta \in \operatorname{Prim}(\Lambda)$,

$$
\sup _{t \in I^{\prime}}\left\|\phi_{s} . \Delta\right\| \geq 0.9 \delta
$$

For $a \in \mathscr{I}_{0}$, consider (the 0.9 here is just to get a finite cover later, but it is not necessary to do so)

$$
\mathscr{A}_{a}:=\left\{(x, \Delta) \in I_{a} \times \operatorname{Prim}(\Lambda) \mid\left\|\phi_{x} . \Delta\right\|<0.9 \delta\right\} .
$$

For each $(x, \Delta) \in \mathscr{A}_{a}$, define

$$
I(x, \Delta):=\text { the connected component of }\left\{s \in I_{a} \mid\left\|\phi_{s} . \Delta\right\|<\delta\right\} \text { containing } x .
$$

For every $x \in I_{a}$, pick some $\Delta_{x}$ such that $I_{x}:=I\left(x, \Delta_{x}\right)$ is maximal among (the finitely many) $I(x, \Delta)$ as $(x, \Delta)$ varies in $\mathscr{A}_{a}$. By this choice, $I_{x}$ and $\Delta_{x}$ satisfy

1. for every $\Delta \in \operatorname{Prim}(\Lambda), \sup _{s \in I_{x}}\left\|\phi_{s} . \Delta\right\| \geq 0.9 \delta$;
2. $\sup _{s \in I_{x}}\left\|\phi_{s} . \Delta_{x}\right\| \leq \delta$.
$I_{a}$ admits a finite sub-covering by $I_{x}$ 's and by passing to a further sub-covering, we assume

$$
I_{a}=\bigcup_{x \in \mathscr{I}_{a}} I_{x} \quad \text { with multiplicity } \leq 2
$$

where $\mathscr{I}_{a}$ is certain finite subset of $I_{a}$ (finiteness is not important, multi $\leq 2$ is). Also define

$$
\mathscr{P}_{x}:=\left\{\Delta \in \operatorname{Prim}(\Lambda) \mid \Delta \text { is comparable to } \Delta_{x}\right\} .
$$

Step 2. Consider the open subset of $I_{x}$ :

$$
I_{x}^{\prime}:=\left\{s \in I_{x} \mid \exists \Delta \in \mathscr{P}_{x},\left\|\phi_{s} . \Delta\right\|<0.8 \delta\right\} .
$$

Write it as a disjoint union of open intervals

$$
I_{x}^{\prime}=\bigsqcup_{b \in \mathscr{I}_{x}} I_{b} .
$$

For $b \in \mathscr{I}_{x}$, consider

$$
\mathscr{A}_{b}:=\left\{(y, \Delta) \in I_{b} \times \mathscr{P}_{x} \mid\left\|\phi_{y} \cdot \Delta\right\|<0.8 \delta\right\} .
$$

For each $(y, \Delta) \in \mathscr{A}_{b}$, define

$$
I(y, \Delta):=\text { the connected component of }\left\{s \in I_{b} \mid\left\|\phi_{s} . \Delta\right\|<0.9 \delta\right\} \text { containing } y .
$$

For every $y \in I_{b}$, pick some $\Delta_{y}$ such that $I_{x, y}:=I\left(y, \Delta_{y}\right)$ is maximal among (the finitely many) $I(y, \Delta)$ as $(y, \Delta)$ varies in $\mathscr{A}_{b}$. By this choice, $I_{x, y}$ and $\Delta_{y}$ satisfy

1. for every $\Delta \in \mathscr{P}_{x}, \sup _{s \in I_{x, y}}\left\|\phi_{s} . \Delta\right\| \geq 0.8 \delta$;
2. $\sup _{s \in I_{x, y}}\left\|\phi_{s .} \Delta_{y}\right\| \leq 0.9 \delta$.

Similarly,

$$
I_{b}=\bigcup_{y \in \mathscr{F}_{b}} I_{x, y} \text { with multiplicity } \leq 2
$$

where $\mathscr{I}_{b}$ is some finite subset of $I_{b}$. Also define

$$
\mathscr{P}_{x, y}:=\left\{\Delta \in \operatorname{Prim}(\Lambda) \mid \Delta \text { is comparable to }\left\{\Delta_{x}, \Delta_{y}\right\}\right\} .
$$

Step 3. Consider the open subset of $I_{x, y}$ :

$$
I_{x, y}^{\prime}:=\left\{s \in I_{x, y} \mid \exists \Delta \in \mathscr{P}_{x, y},\left\|\phi_{s} . \Delta\right\|<0.7 \delta\right\} .
$$

Write it as a disjoint union of open intervals

$$
I_{x, y}^{\prime}=\bigsqcup_{c \in \mathscr{A}_{x, y}} I_{c} .
$$

For $c \in \mathscr{I}_{x, y}$, consider

$$
\mathscr{A}_{c}:=\left\{(z, \Delta) \in I_{c} \times \mathscr{P}_{x, y} \mid\left\|\phi_{z} . \Delta\right\|<0.7 \delta\right\} .
$$

For each $(z, \Delta) \in \mathscr{A}_{c}$, define

$$
I(z, \Delta):=\text { the connected component of }\left\{s \in I_{c} \mid\left\|\phi_{s} . \Delta\right\|<0.8 \delta\right\} \text { containing } z .
$$

For every $z \in I_{c}$, pick some $\Delta_{z}$ such that $I_{x, y, z}:=I\left(z, \Delta_{z}\right)$ is maximal among (the finitely many) $I(z, \Delta)$ as $(z, \Delta)$ varies in $\mathscr{A}_{c}$. By this choice, $I_{x, y, z}$ and $\Delta_{z}$ satisfy

1. for every $\Delta \in \mathscr{P}_{x, y}, \sup _{s \in I_{x, y, z}}\left\|\phi_{s} . \Delta\right\| \geq 0.7 \delta$;
2. $\sup _{s \in I_{x, y, z}}\left\|\phi_{s .} \Delta_{z}\right\| \leq 0.8 \delta$.

Similarly,

$$
I_{c}=\bigcup_{z \in \mathscr{I}_{c}} I_{x, y, z} \quad \text { with multiplicity } \leq 2
$$

where $\mathscr{I}_{c}$ is certain finite subset of $I_{c}$. Now $\left\{\Delta_{x}, \Delta_{y}, \Delta_{z}\right\}$ is already a complete flag modulo $\{0\}$ and $\Lambda$.

Good and bad points 1. For $x, a, y, b, z$, let

$$
I_{x, y, z}(\mathrm{Good}):=\left\{s \in I_{x, y, z} \mid\left\|\phi_{s .} \Delta_{z}\right\| \geq \rho \delta\right\}, \quad I_{x, y, z}(\mathrm{Bad}):=I_{x, y, z} \backslash I_{x, y, z} \text { (Good). }
$$

By ( $C, \alpha$ )-goodness, we choose $\rho \in(0,1)$ such that

$$
\left|I_{x, y, z}(\mathrm{Bad})\right| \leq(0.01 \varepsilon)\left|I_{x, y, z}\right| .
$$

Thus

$$
\begin{aligned}
\left|I_{x, y}^{\prime}(\mathrm{Bad})\right| & :=\left|\bigsqcup_{c \in \mathscr{I}_{x, y}} \bigcup_{z \in \mathscr{I}_{c}} I_{x, y, z}(\mathrm{Bad})\right| \leq \sum_{c} \sum_{z}\left|I_{x, y, z}(\mathrm{Bad})\right| \leq \sum_{c} \sum_{z}(0.01 \varepsilon) \cdot\left|I_{x, y, z}\right| \\
& \leq \sum_{c} 2(0.01 \varepsilon) \cdot\left|I_{c}\right|=(0.02 \varepsilon) \cdot\left|I_{x, y}^{\prime}\right| .
\end{aligned}
$$

Define $I_{x, y}^{\prime}(\mathrm{Good}):=I_{x, y}^{\prime} \backslash I_{x, y}^{\prime}(\mathrm{Bad})$, so $I_{x, y}^{\prime}=I_{x, y}^{\prime}(\mathrm{Good}) \sqcup I_{x, y}^{\prime}(\mathrm{Bad})$.
So far, we have the following regarding each $I_{x, y}$ :

1. $s \in I_{x, y} \backslash I_{x, y}^{\prime} \Longrightarrow\left\|\phi_{s} . \Delta\right\| \geq 0.7 \delta, \forall \Delta \in \mathscr{P}_{x, y}$;
2. $s \in I_{x, y}^{\prime}(\mathrm{Good}) \Longrightarrow \exists \Delta_{z} \in \mathscr{P}_{x, y}, \rho \delta \leq\left\|\phi_{s} . \Delta_{z}\right\| \leq 0.8 \delta$;
3. $\left|I_{x, y}^{\prime}(\mathrm{Bad})\right| \leq 2 \delta \cdot\left|I_{x, y}^{\prime}\right|$.

Good and bad points 2. Define

$$
I_{x, y}(\operatorname{Good}):=\left\{s \in I_{x, y} \mid\left\|\phi_{s} . \Delta_{y}\right\| \geq \rho \delta\right\}, \quad I_{x, y}(\mathrm{Bad}):=I_{x, y} \backslash I_{x, y}(\mathrm{Good}) .
$$

And $\rho$ is chosen such that

$$
\left|I_{x, y}(\mathrm{Bad})\right| \leq(0.01 \varepsilon)\left|I_{x, y}\right| .
$$

Thus,

$$
\begin{aligned}
\left|I_{x}^{\prime}(\mathrm{Bad})\right| & :=\left|\bigsqcup_{b \in \mathscr{I}_{x}} \bigcup_{y \in \mathscr{I}_{b}} I_{x, y}(\mathrm{Bad})\right| \leq \sum_{b} \sum_{y}\left|I_{x, y}(\mathrm{Bad})\right| \leq \sum_{b} \sum_{y}(0.01 \varepsilon) \cdot\left|I_{x, y}\right| \\
& \leq \sum_{b} 2(0.01 \varepsilon) \cdot\left|I_{b}\right|=(0.02 \varepsilon) \cdot\left|I_{x}^{\prime}\right| .
\end{aligned}
$$

Define $I_{x}^{\prime}(\mathrm{Good})$ by imposing $I_{x}^{\prime}=I_{x}^{\prime}(\mathrm{Good}) \sqcup I_{x}^{\prime}(\mathrm{Bad})$.
So far, regarding $I_{x}$ we have:

1. $s \in I_{x} \backslash I_{x}^{\prime} \Longrightarrow\left\|\phi_{s} . \Delta\right\| \geq 0.8 \delta, \forall \Delta \in \mathscr{P}_{x}$;
2. $s \in I_{x}^{\prime}(\operatorname{Good}) \cap I_{x, y} \Longrightarrow \rho \delta \leq\left\|\phi_{s} . \Delta_{y}\right\| \leq 0.9 \delta$;
3. $\left|I_{x}^{\prime}(\mathrm{Bad})\right| \leq 2 \delta \cdot\left|I_{x}^{\prime}\right|$.

Good and bad points 3. Finally, define

$$
I_{x}(\mathrm{Good}):=\left\{s \in I_{x} \mid\left\|\phi_{s} \cdot \Delta_{x}\right\| \geq \rho \delta\right\}, \quad I_{x}(\mathrm{Bad}):=I_{x} \backslash I_{x}(\mathrm{Good}) .
$$

And $\rho$ is chosen such that

$$
\left|I_{x}(\mathrm{Bad})\right| \leq 0.01 \varepsilon\left|I_{x}\right| .
$$

Thus,

$$
\begin{aligned}
\left|I^{\prime}(\mathrm{Bad})\right| & :=\left|\bigsqcup_{a \in \mathscr{I}_{0}} \bigcup_{x \in \mathscr{I}_{a}} I_{x}(\mathrm{Bad})\right| \leq \sum_{a} \sum_{x}\left|I_{x}(\mathrm{Bad})\right| \leq \sum_{a} \sum_{x} 0.01 \varepsilon \cdot\left|I_{x}\right| \\
& \leq \sum_{a} 2 \cdot 0.01 \varepsilon \cdot\left|I_{a}\right|=(0.02 \varepsilon) \cdot\left|I^{\prime}\right| .
\end{aligned}
$$

Define $I^{\prime}$ (Good) by imposing $I^{\prime}=I^{\prime}(\mathrm{Good}) \sqcup I^{\prime}(\mathrm{Bad})$. Here we have:

1. $s \in I \backslash I^{\prime} \Longrightarrow\left\|\phi_{s} . \Delta\right\| \geq 0.9 \delta, \forall \Delta \in \operatorname{Prim}(\Lambda)$;
2. $s \in I^{\prime}(\mathrm{Good}) \cap I_{x} \Longrightarrow \rho \delta \leq\left\|\phi_{s} . \Delta_{x}\right\| \leq \delta$;
3. $\left|I^{\prime}(\mathrm{Bad})\right| \leq 2 \delta \cdot\left|I^{\prime}\right|$.

Warp-up. Now we collect all the bad points together and let

$$
I(\mathrm{Bad}):=I^{\prime}(\mathrm{Bad}) \cup\left(\bigcup_{a \in \mathscr{\mathscr { I }}_{0}, x \in \mathscr{I}_{a}}^{\bigcup} I_{x}^{\prime}(\mathrm{Bad})\right) \cup\left(\bigcup_{a \in \mathscr{\mathscr { I }}_{0}, x \in \mathscr{I}_{a}}^{\bigcup} \bigcup_{b \in \mathscr{I}_{x}, y \in \mathscr{I}_{b}} I_{x, y}^{\prime}(\mathrm{Bad})\right)
$$

We have

$$
\begin{aligned}
\left|\bigcup_{a, x, b, y} I_{x, y}^{\prime}(\mathrm{Bad})\right| & \leq \sum_{a, x, b \in \mathscr{I}_{x}} \sum_{y \in \mathscr{\mathscr { F }}_{b}}\left|I_{x, y}^{\prime}(\mathrm{Bad})\right| \leq(0.02 \varepsilon) \cdot \sum_{a, x, b} \sum_{y}\left|I_{x, y}\right| \\
& \leq(0.04 \varepsilon) \cdot \sum_{a, x, b \in \mathscr{I}_{x}}\left|I_{b}\right| \leq(0.04 \varepsilon) \cdot \sum_{a, x}\left|I_{x}\right| \\
& \leq(0.08 \varepsilon) \cdot|I|
\end{aligned}
$$

and

$$
\begin{aligned}
\left|\bigcup_{a \in \mathscr{I}_{0}, x \in \mathscr{\mathscr { F }}_{a}} I_{x}^{\prime}(\mathrm{Bad})\right| & \leq \sum_{a \in \mathscr{I}_{0}} \sum_{x \in \mathscr{I}_{a}}\left|I_{x}^{\prime}(\mathrm{Bad})\right| \leq(0.02 \varepsilon) \cdot \sum_{a \in \mathscr{\mathscr { I }}_{0}} \sum_{x \in \mathscr{I}_{a}}\left|I_{x}\right| \\
& \leq(0.04 \varepsilon) \cdot \sum_{a \in \mathscr{I}_{0}}\left|I_{a}\right| \leq(0.04 \varepsilon) \cdot|I| .
\end{aligned}
$$

Hence

$$
\begin{equation*}
|I(\mathrm{Bad})| \leq(0.14 \varepsilon) \cdot|I|<\varepsilon|I| . \tag{25}
\end{equation*}
$$

Let $s \in I \backslash I($ Bad $)$.
Case 1. $s \in I \backslash I^{\prime}$, then $\left\|\phi_{s} . \Delta\right\| \geq 0.9 \delta>0.5 \delta, \forall \Delta \in \operatorname{Prim}(\Lambda)$ so it is $(\delta, \rho)$-protected by the trivial flag.

Case 2. $s \in I^{\prime} \backslash I(\mathrm{Bad})=\left(\sqcup I_{a}\right) \backslash I(\mathrm{Bad})=\left(\sqcup_{a} \cup_{x} I_{x}\right) \backslash I(\mathrm{Bad})$. Say $s \in I_{x} \backslash I(\mathrm{Bad})$. Then

$$
\rho \delta \leq\left\|\phi_{s} \cdot \Delta_{x}\right\| \leq \delta .
$$

Case 2.1. $s \in I_{x} \backslash I_{x}^{\prime}$. Then $\left\|\phi_{s} . \Delta\right\| \geq 0.8 \delta>0.5 \delta$ for all $\Delta \in \mathscr{P}_{x}$. This means that $s$ is $(\rho, \delta)$ protected by $\left\{\Delta_{x}\right\}$.

Case 2.2. $s \in I_{x}^{\prime} \backslash I(\mathrm{Bad})=\left(\sqcup I_{b}\right) \backslash I(\mathrm{Bad})=\left(\sqcup_{b} \cup_{y} I_{x, y}\right) \backslash I(\mathrm{Bad})$. Say $s \in I_{x, y} \backslash I(\mathrm{Bad})$. Then

$$
\rho \delta \leq\left\|\phi_{s} \cdot \Delta_{y}\right\| \leq \delta
$$

Case 2.2.1. $s \in I_{x, y} \backslash I_{x, y}^{\prime}$. Then $\left\|\phi_{s} . \Delta\right\| \geq 0.7 \delta>0.5 \delta$ for all $\Delta \in \mathscr{P}_{x, y}$. This means that $s$ is $(\rho, \delta)$-protected by $\left\{\Delta_{x}, \Delta_{y}\right\}$.

Case 2.2.2. $s \in I_{x, y}^{\prime} \backslash I(\mathrm{Bad})=\sqcup_{c} I_{c} \backslash I(\mathrm{Bad})=\sqcup_{c} \cup_{z} I_{x, y, z} \backslash I(\mathrm{Bad})$. Say $x \in I_{x, y, z} \backslash I(\mathrm{Bad})$, then

$$
\rho \delta \leq\left\|\phi_{s} \cdot \Delta_{z}\right\| \leq \delta .
$$

Thus $s$ is $(\delta, \rho)$-protected by $\left\{\Delta_{x}, \Delta_{y}, \Delta_{z}\right\}$.
Now every $s \in I \backslash I(\mathrm{Bad})$ falls into one of the cases $1,2.1,2.2 .1$ and 2.2 .2 , so it is $(\delta, \rho)$ protected. Hence Lem.2.3 implies if $s \in I \backslash I(\mathrm{Bad})$ then $\phi_{s} . \Lambda \in \mathscr{C}_{\theta}$ with $\theta=\theta(\delta, \rho)$. Now we take $\kappa:=\theta$. Combining with Equa.(25), we are done.

## 4. Exercises

4.1. More examples of $(C, \alpha)$-good functions. Let $C, \alpha>0$ and $J$ be an interval in $\mathbb{R}$, recall a function $f: J \rightarrow \mathbb{R}$ is said to be $(C, \alpha)$-good on $J$ iff for every interval $I \subset J$ of finite length and every $\rho \in(0,1)$,

$$
\begin{equation*}
\frac{1}{|I|} \operatorname{Leb}\left\{t \in I\left||f(t)| \leq \rho M_{I}\right\} \leq C \rho^{\alpha} .\right. \tag{26}
\end{equation*}
$$

where $M_{I}:=\sup _{t \in I}|f(t)|$.
Let $J$ be an interval of finite length. Let

$$
\mathscr{A}:=\left\{f=a e^{x}+b e^{-x}, a, b \in \mathbb{R}\right\} .
$$

Exercise 4.1. Show that there exist $C, \alpha>0$ (depending on $J$ and $\mathscr{A}$ ) such that for every function $f \in \mathscr{A}$ is $(C, \alpha)$-good on $J$.

EXercise 4.2. If $f_{1}, f_{2}$ are $(C, \alpha)$-good on $J$, then $x \mapsto \max \left\{\left|f_{1}(x)\right|,\left|f_{2}(x)\right|\right\}$ is also $(C, \alpha)$ good on $J$.

## CHAPTER 7

## Ergodicity and Mixing

Back to the Top.

## 1. Basic constructions

For details the reader may consult [EW11], especially chapter 8 and appendices therein.
Let $G$ be a "nice" ( $\sigma$-compact locally compact metrizable) topological group and $X$ a "nice" ( $\sigma$-compact locally compact metrizable) topological space. Assume $G$ acts on $X$ continuously, i.e. we have a continuous map $G \times X \rightarrow X$ satisfying some compatibility conditions.

Let $\mathscr{B}_{X}$ be the $\sigma$-algebra on $X$ generated by open sets in $X$. This is termed the Borel $\sigma$ algebra. Then the $G$-action is also measurable with respect to $\mathscr{B}_{X}$. Thus $G$ naturally acts on measures on $\left(X, \mathscr{B}_{X}\right)$.

Definition 1.1. A measure $\mu$ on $\mathscr{B}_{X}$ is called a Borel measure. It is called a probability measure iff $\mu(X)=1$. The collection of all probability measures is denoted as $\operatorname{Prob}(X)$. We view $\operatorname{Prob}(X)$ as a topological space equipped with the weak-* topology.

More precisely, we embed $\operatorname{Prob}(X)$ with the weakest topology such that

$$
\mu \mapsto \int f(x) \mu(x)
$$

is continuous for every

$$
f \in C_{c}(X):=\{\text { compactly supported real-valued continuous functions on } X\} .
$$

Being real-valued or complex-valued is not important.
Let

$$
\operatorname{Meas}(X)^{\leq 1}:=\{\text { finite measures } \mu \text { on } X, \mu(X) \leq 1\},
$$

also equipped with weak-* topology. We also let

$$
\operatorname{LFM}(X):=\{\text { locally finite measures on } X\}
$$

be equipped with weak-* topology. Note that $C_{c}(X)$ admits a countable dense subset.
Lemma 1.2. With weak-* topology, $\operatorname{Meas}(X)^{\leq 1}$ is a compact metrizable space. If $X$ is compact, then so is $\operatorname{Prob}(X)$.

REmARK 1.3. If we forget about the topological structure on $X$, and take some probability measure $\mu$, then up to completion, $\left(X, \mathscr{B}_{X}, \mu\right)$ is "isomorphic" to a convex combination of the natural measure on $[0,1]$ interval and atomic measures supported on single points (see [Wal82, Theorem 2.1]). Thus the study of ( $X, \mathscr{B}_{X}, \mu$ ) is rather boring without a group action, unlike the topological space $X$, when the classification of $X$ is already a huge problem.

We naturally has an action of $G$ on $\operatorname{Prob}(X), \operatorname{Meas}^{\leq 1}(X)$ and $\operatorname{LFM}(X)$ defined by

$$
g_{*} \mu(E):=\mu\left(g^{-1} E\right)
$$

for every measurable set $E$ and measure $\mu$.

LEMMA 1.4. The induced map $G \times \operatorname{LFM}(X) \rightarrow \operatorname{LFM}(X)$ is continuous.
A measure $\mu$ is said to be $G$-invariant iff $g_{*} \mu=\mu$ for all $g \in G$. The collection of $G$-invariant probability measures is denoted as $\operatorname{Prob}(X)^{G}$. Similarly define Meas ${ }^{\leq 1}(X)^{G}$ and $\operatorname{LFM}(X)^{G}$.

To distinguish different p.m.p(= probability measure preserving) actions of $G$, a convenient functor is given by taking the associated unitary representation.

Take a $\mu \in \operatorname{LFM}(X)^{G}$. Then the associated unitary representation is given by

$$
\begin{aligned}
G \times L^{2}(X, \mu) & \rightarrow L^{2}(X, \mu) \\
(g, \phi) & \mapsto g \cdot \phi(x):=\phi\left(g^{-1} x\right)
\end{aligned}
$$

LEMMA 1.5. This is indeed a unitary representation:

1. for each $g \in G$, the action on $L^{2}(X, \mu)$ is a unitary;
2. the representation is continuous
where $\mathscr{U}\left(L^{2}(X, \mu)\right)$, the set of unitary operators on $L^{2}(X, \mu)$, is equipped with the strong operator topology.

In more concrete terms, using the following lemma, the continuity claim just asserts that if $g_{n} \rightarrow g$ in $G$ and $\phi_{n} \rightarrow \phi$ in $L^{2}(X, \mu)$, then $g_{n} \cdot \phi_{n} \rightarrow g \cdot \phi$ in $L^{2}(X, \mu)$.

LEMMA 1.6. $L^{2}(X, \mu)$ admits a countable dense subset.
For two p.m.p. G-actions to be isomorphic, it is necessary for the associated unitary representations to be isomorphic. Properties of p.m.p. $G$-actions defined via the associated unitary representation are sometimes called "spectral properties".

## 2. Ergodicity and mixing

We assume $G$ and $X$ to be nice in this section.
DEFINITION 2.1. A p.m.p. G-action on $\left(X, \mathscr{B}_{X}, \mu\right)$ is said to be ergodic iffevery $G$-invariant measurable subset $E$ of $X$ is either $\mu$-null $(\mu(E)=0)$ or $\mu$-conull $(\mu(X \backslash E)=0)$.

So ergodicity is something like irreducibility.
LEMMA 2.2. If a p.m.p. G-action on $\left(X, \mathscr{B}_{X}, \mu\right)$ is ergodic, then every $\mu$-almost invariant measurable subset of $X$ is either $\mu$-null or $\mu$-conull.

A measurable subset $E \subset X$ is said to be $\mu$-almost invariant iff for every $g \in G$,

$$
\mu(g E \Delta E)=\mu((g E \backslash E) \cup(E \backslash g E))=0
$$

Since our group could be uncountable, this lemma is not obvious. Using this lemma, one can show that

LEMMA 2.3. A p.m.p. G-action on $\left(X, \mathscr{B}_{X}, \mu\right)$ is ergodic iff the associated unitary representation has no fixed vector orthogonal to constants.

Hint: Starting from a set $E$, one has the characteristic function $1_{E}$. Starting from a function $f$, one considers its level sets.

By this lemma, being ergodic is a spectral property. Another spectral property we need is mixing.

DEFINITION 2.4. A p.m.p. G-action on $\left(X, \mathscr{B}_{X}, \mu\right)$ is said to be mixing iff for every two measurable subsets $E, F \subset X$ and every divergent sequence $\left(g_{n}\right)$ in $G$, we have

$$
\lim _{n \rightarrow \infty} \mu\left(g_{n}^{-1} E \cap F\right)=\mu(E) \mu(F)
$$

This notion is useless for compact groups.

Lemma 2.5. A p.m.p. G-action on $\left(X, \mathscr{B}_{X}, \mu\right)$ is mixing iff for every two $\phi, \psi \in L^{2}(X, \mu)$ orthogonal to constants and every divergent sequence $\left(g_{n}\right)$ in $G$, we have

$$
\lim _{n \rightarrow \infty}\left\langle g_{n} \cdot \phi, \psi\right\rangle=0
$$

Here $\langle\phi, \psi\rangle:=\int \phi(x) \overline{\psi(x)} \mu(x)$.

## 3. Unitary representations of $\mathrm{SL}_{2}(\mathbb{R})$ are mixing

Notations

- $\mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R})$ and $\Gamma$ is a discrete subgroup of G ;
- $\mathrm{A}:=\left\{\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right], t \in \mathbb{R}\right\}=\left\{\mathbf{a}_{t}, t \in \mathbb{R}\right\} ;$
- $\mathrm{U}:=\left\{\left[\begin{array}{cc}1 & s \\ 0 & 1\end{array}\right], s \in \mathbb{R}\right\}=\left\{\mathbf{u}_{s}, s \in \mathbb{R}\right\} ;$
- $\mathrm{B}:=\mathrm{A} \cdot \mathrm{U}$.

For convenience let us make the following definition
Definition 3.1. A unitary representation $\pi: G \rightarrow \mathscr{U}(\mathscr{H})$ is mixing iff for every $v, w \in \mathscr{H}$ and every divergent sequence $\left(g_{n}\right)$ in $G, \lim _{n \rightarrow \infty}\left\langle g_{n} \cdot v, w\right\rangle=0$.

Unitary representations, if containing no non-zero vectors fixed by G , of $\mathrm{G}=\mathrm{SL}_{2}(\mathbb{R})$ are always mixing.

Theorem 3.2. Let $\pi$ be a unitary representation of G on a separable Hilbert space $\mathscr{H}$. Assume there is no non-zero $G$-fixed vectors. Then $\pi$ is mixing.

Proof. By " $K A K$-decomposition" (see the comment after the proof for an explanation), it suffices to show that $\left.\pi\right|_{\mathrm{A}}$ is mixing. So take $\left(a_{n}\right)$ to be a divergent sequence in A . By a diagonal argument, we find an infinite subsequence $\left(a_{n_{k}}\right)$ such that for every $\phi, \psi \in \mathscr{H}$,

$$
\lim _{k \rightarrow \infty}\left\langle a_{n_{k}} \phi, \psi\right\rangle \text { exists. }
$$

This defines a linear map $E: \mathscr{H} \rightarrow \mathscr{H}$ such that the above limit is equal to $\langle E \phi, \psi\rangle$. One can check $\|E\|_{\mathrm{op}} \leq 1$ where $\|\cdot\|_{\mathrm{op}}$ stands for the operator norm. It suffices to show that $E=0$, which is going to be achieved by showing that every vector contained in the image of $E$ is fixed by G.

By passing to a further subsequence we assume either $\left(\log \left(a_{n_{k}}\right)\right)_{1,1} \rightarrow+\infty$ or $-\infty$.
Define

$$
U^{-}:=\left\{x \in \mathrm{G} \mid \lim _{k \rightarrow+\infty} a_{n_{k}} x a_{n_{k}}^{-1}=1\right\}, U^{+}:=\left\{x \in \mathrm{G} \mid \lim _{k \rightarrow+\infty} a_{n_{k}}^{-1} x a_{n_{k}}=1\right\}
$$

There are two things we firstly note. Let $E^{*}$ be the adjoint of $E$.

1. $E \circ u=E$ for every $u \in U^{-}$. Indeed, for every pair $\phi, \psi$ in $\mathscr{H}$,

$$
\begin{aligned}
\langle E u \phi, \psi\rangle & =\lim \left\langle a_{n_{k}} u \phi, \psi\right\rangle=\lim \left\langle a_{n_{k}} u a_{n_{k}}^{-1} a_{n_{k}} \phi, \psi\right\rangle \\
& =\lim \left\langle a_{n_{k}} \phi, a_{n_{k}} u^{-1} a_{n_{k}}^{-1} \psi\right\rangle=\langle E \phi, \psi\rangle .
\end{aligned}
$$

The last step is because $\left(a_{n_{k}} u^{-1} a_{n_{k}}^{-1} \psi\right)$ converges to $\psi$ in norm. Hence $E \circ u=E$. By taking the adjoint, we get $u^{-1} \circ E^{*}=E^{*}$. Thus the image of $E^{*}$ is fixed by $U^{-}$.
2. $u \circ E=E$ for every $u \in U^{+}$. For every pair $\phi, \psi$,

$$
\begin{aligned}
\langle u E \phi, \psi\rangle & =\left\langle E \phi, u^{-1} \psi\right\rangle=\lim \left\langle a_{n_{k}} \phi, u^{-1} \psi\right\rangle=\lim \left\langle a_{n_{k}} a_{n_{k}}^{-1} u a_{n_{k}} \phi, \psi\right\rangle \\
& =\lim \left\langle a_{n_{k}} \phi, \psi\right\rangle=\langle E \phi, \psi\rangle .
\end{aligned}
$$

Hence $u \circ E=E$.
Next is the trick. As the $*$ operation is continuous with respect to W.O.T., $\left(a_{n_{k}}^{-1}\right)=\left(a_{n_{k}}^{*}\right)$ converges in W.O.T. to $E^{*}$.
3. $\operatorname{ker} E=\operatorname{ker} E^{*}$. Indeed,

$$
\langle E \phi, E \phi\rangle=\lim _{l} \lim _{k}\left\langle a_{n_{k}} \phi, a_{n_{l}} \phi\right\rangle=\lim _{l} \lim _{k}\left\langle a_{n_{l}}^{-1} \phi, a_{n_{k}}^{-1} \phi\right\rangle=\left\langle E^{*} \phi, E^{*} \phi\right\rangle
$$

(Exercise: show that in general $\operatorname{ker} E \neq \operatorname{ker} E^{*}$ for a bounded linear operator on a Hilbert space.)

Now we can finish the proof. 1. says that $E(1-u)=0, \forall u \in U^{-}$. Combined with 3 ., we get $E^{*}(1-u)=0, \forall u \in U^{-}$. Taking $*$ of this, we get $E=u^{-1} E$. Thus the image of $E$ is fixed by $U^{-}$. 2. asserts that the image of $E$ is fixed by $U^{+}$. Since $U^{-}$and $U^{+}$generates G, we are done.

Let us quickly explain, using linear algebra, why you can write a matrix $g \in \mathrm{SL}_{2}(\mathbb{R})$ as $k_{1} a k_{2}$ with $k_{i}$ in $\mathrm{SO}_{2}(\mathbb{R})$ and $a$ being diagonal. This fact was used to reducing the mixing in general to mixing of A. First we claim that we can write $g=k_{1}|g|$ where $k_{1}$ is orthogonal and $|g|$ is symmetric. Assuming the claim, since $|g|$ can be diagonalized under an orthogonal basis, we are done. Now let us prove the claim. The matrix $g g^{t r}$ is symmetric and hence diagonalizable. Moreover it has positive eigenvalues. Hence makes sense to take $|g|:=\sqrt{g g^{t r}}$. Then one defines $k_{1}:=g|g|^{-1}$ and it is direct to check that $\left\langle k_{1} v, k_{1} v\right\rangle=\langle v, v\rangle$ for every vector $v$. And we are done.

## 4. Examples

EXAMPLE 4.1. Let $G:=\mathbb{Z}$ generated by $1:=R_{\alpha}$ act on $\mathbb{R} / \mathbb{Z}$ by $R_{\alpha} \cdot x:=x+\alpha$ for some real number $\alpha$. Then this action preserves the natural Lebesgue measure $m$ on $\mathbb{R} / \mathbb{Z}$. It is ergodic iff $\alpha \in \mathbb{R} \backslash \mathbb{Q}$. Moreover, this action is not mixing.

Sketch of proof. There are two proofs. Assume $\alpha \notin \mathbb{Q}$.
Either you can argue that $R_{\alpha}$ generates a dense subgroup of $\mathbb{R} / \mathbb{Z}$ and then by continuity, $m$ has to be invariant under the full $\mathbb{R} / \mathbb{Z}$. Then argue that $m$ is the unique $\mathbb{R} / \mathbb{Z}$-invariant probability measure.

Or you can argue that there are no invariant $L^{2}$ functions by expanding them under the basis $\left\{x \mapsto e^{2 \pi i n x}\right\}_{n \in \mathbb{Z}}$.

I leave it to you to show that $R_{\alpha}$ is not mixing.
EXAMPLE 4.2. Let $G:=\mathbb{Z}$ act on $\mathbb{R}^{2} / \mathbb{Z}^{2}$ where the generator 1 acts by $(x, y) \mapsto(x+y, x+2 y)$. Then $G$ preserves the natural Lebesgue measure on $\mathbb{R}^{2} / \mathbb{Z}^{2}$ and the action is ergodic and mixing.

SKETCH OF PROOF. Two ways: 1. Fourier analysis; 2. use the idea presented in last section (you need something contracted by the $G$ action to make the argument work, what is this?).

Let $M:=\left[\begin{array}{ll}1 & 1 \\ 1 & 2\end{array}\right]$. For $t \in \mathbb{R}$, let $M^{t}:=\exp (t \cdot \log M)$. The above example is about the induced action of $M^{\mathbb{Z}}$ on $\mathbb{R}^{2} / \mathbb{Z}^{2}$. The reason why you have such an induced action is of course $\mathbb{Z}^{2}$ is preserved by $M^{\mathbb{Z}}$. For other $t$, this is not true. Nevertheless, each $M^{t}$ defines a homeomorphism

$$
\mathbb{R}^{2} / \mathbb{Z}^{2} \mapsto \mathbb{R}^{2} / M^{t} \cdot \mathbb{Z}^{2}
$$

Let

$$
X=\left\{(x, t) \mid t \in \mathbb{R} / \mathbb{Z}, x \in \mathbb{R}^{2} / M^{t} \cdot \mathbb{Z}^{2}\right\}
$$

EXAMPLE 4.3. Show that $X$ has a natural measure $m$. Moreover, the action of $M^{\mathbb{R}}$ is ergodic but not mixing.

This example tells you that in general an ergodic B-action (that is not extendable to an $\mathrm{SL}_{2}(\mathbb{R})$ p.m.p. action) may not be mixing. However, this B-action is not "totally ergodic" in the sense that some infinite subgroup does not act ergodically. I do not know an example of totally ergodic B-action that is not mixing. Note that by argument from the last section, it must be A-mixing.

## 5. Exercises

5.1. Non-commensurable lattices in $\mathrm{SL}_{2}(\mathbb{R})$, II. This is a continuation of Exercise 2.1-2.6 from Exercise Sheet 2. Notations are inherited and here are a few more:

- Let $X:=\mathrm{G} / \Gamma$ and $\widehat{\mathrm{m}}_{X}$ the unique G -invariant probability measure on $X$;
- Let $\Omega$ be a nonempty open bounded subset of $U V^{+}$(or $U V^{-}$);
- Let $\widetilde{\mu}_{0}$ be the restriction of the Haar measure on $U V$ to $\Omega$. Fix $x_{0} \in X$, let $\mu_{0}$ be the push-forward of $\widetilde{\mu}_{0}$ under the map $g \mapsto g . x_{0}$. By multiplying by a scalar, we normalize $\mu_{0}$ to be a probability measure $\widehat{\mu}_{0}$.

Exercise 5.1. Show that $\widehat{\mathrm{m}}_{X}$ is $A$-mixing.
EXERCISE 5.2. Using mixing to show that $\lim _{t \rightarrow+\infty}\left(\mathbf{a}_{t}\right)_{*} \widehat{\mu}_{0}=\widehat{\mathbf{m}}_{X}$.
Exercise 5.3. Let $\mathrm{Y}_{0}$ be as in Exer 2.3 from Exer. Sheet 2. Show that $\mathrm{Y}_{0}=X$.
Thus we have shown that H -orbits on X are either closed or dense.
Now let $\Gamma_{1}, \Gamma_{2}$ be two discrete subgroups in $\mathrm{SL}_{2}(\mathbb{R})$ (later we will assume them to be cocompact).

EXERCISE 5.4. The following two are equivalent

1. $\Gamma_{1} \cdot \Gamma_{2}$ is closed in $\mathrm{SL}_{2}(\mathbb{R})$;
2. $\mathrm{H} \cdot\left(\Gamma_{1} \times \Gamma_{2}\right)$ is closed in G .

## Exercise 5.5. The following two are equivalent

1. $\Gamma_{1} \cdot \Gamma_{2}$ is dense in $\mathrm{SL}_{2}(\mathbb{R})$;
2. $\mathrm{H} \cdot\left(\Gamma_{1} \times \Gamma_{2}\right)$ is dense in G .

From now on we assume $\Gamma_{1}, \Gamma_{2}$ are both cocompact in $\mathrm{SL}_{2}(\mathbb{R})$.

## Exercise 5.6. The following two are equivalent

1. $\Gamma_{1} \cdot \Gamma_{2}$ is closed in $\mathrm{SL}_{2}(\mathbb{R})$;
2. $\Gamma_{1}$ is commensurable with $\Gamma_{2}$ (namely, $\Gamma_{1} \cap \Gamma_{2}$ is of finite-index in both $\Gamma_{1}$ and $\Gamma_{2}$ ).
[It seems unclear to me how to prove this only assuming $\Gamma_{i}$ 's are lattices. There is an approach using random walk by Eskin-Margulis.]

Exercise 5.7. The followings are equivalent

1. $\Gamma_{1}$ is commensurable with $\Gamma_{2}$;
2. $\Gamma_{1} \cdot[\mathrm{id}]_{\Gamma_{2}}$ is a finite subset of $\mathrm{SL}_{2}(\mathbb{R}) / \Gamma_{2}$;
3. $\Gamma_{1} \cdot \Gamma_{2}$ is not dense in $\mathrm{SL}_{2}(\mathbb{R})$.
5.2. Totally geodesic hyperbolic planes in H3, II. Notations and assumptions are inherited from Sec. 3 from Exercise Sheet 2.

EXERCISE 5.8. Show that H -orbits on $\mathrm{G} / \Gamma$ are either closed or dense.
5.3. Mixing fails for non-semisimple groups. Notations

- $B=A \cdot U$ where $A:=\left\{\mathbf{a}_{t}=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right], t \in \mathbb{R}\right\}$ and $U=\left\{\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right], s \in \mathbb{R}\right\}$;
- $\mathscr{H}$ is a separable Hilbert space and $\Phi: B \rightarrow \mathscr{U}(\mathscr{H})$ is a unitary representation of $B$.

Exercise 5.9. Show that if $\mathscr{H}$ has no non-zero $\Phi(U)$-fixed vector ("U-ergodic"), then for every $\phi, \psi \in \mathscr{H}$ and $t_{n} \rightarrow+\infty, \lim _{n}\left\langle\Phi\left(\mathbf{a}_{t_{n}}\right) . \phi, \psi\right\rangle=0$ (" $A^{+}$-mixing").

Exercise 5.10. Same notations and assumptions as in last exercise. Show that for every $\phi, \psi \in \mathscr{H}$ and $t_{n}^{\prime} \rightarrow-\infty, \lim _{n}\left\langle\Phi\left(\mathbf{a}_{t_{n}^{\prime}}\right) \cdot \phi, \psi\right\rangle=0$ (" $A^{-}$-mixing").

Below is an example showing that " $U$-mixing" may not be true under the hypothesis made in last two exercises.

Let $\mathscr{H}_{0}:=L^{2}\left(\mathbb{R}_{>0}\right.$, Leb). Define, for $t, s \in \mathbb{R}$ and $\phi \in \mathscr{H}_{0}$,

$$
\left(\mathbf{a}_{t} \cdot \phi\right)(x):=e^{t} \phi\left(e^{2 t} x\right), \quad\left(\mathbf{u}_{s} \cdot \phi\right)(x):=e^{2 \pi i s x} \cdot \phi(x) .
$$

EXercise 5.11. Show that the above defined action of $A$ and $U$ extends to a group homomorphism $\Phi_{0}: B \rightarrow \operatorname{Hom}\left(\mathscr{H}_{0}, \mathscr{H}_{0}\right)$.

Here $\operatorname{Hom}\left(\mathscr{H}_{0}, \mathscr{H}_{0}\right)$ stands for linear maps from $\mathscr{H}_{0}$ to $\mathscr{H}_{0}$.
EXERCISE 5.12. Show that image of $\Phi_{0}$ consists of unitary operators.
EXERCISE 5.13. Show that $\Phi_{0}$ defines a unitary representation of B (namely, one should check continuity w.r.t. strong operator topology).

EXERCISE 5.14. Show directly that $\Phi_{0}$ is A-mixing. Namely, for a divergent sequence $\left(a_{n}\right) \subset$ $A$ and $\phi, \psi \in \mathscr{H}_{0}, \lim _{n}\left\langle\Phi_{0}\left(a_{n}\right) . \phi, \psi\right\rangle=0$.

Exercise 5.15. Show that there is no non-zero $\Phi_{0}(U)$-fixed vector. Yet $\Phi_{0}$ is not $U$-mixing.
5.4. Another example of Mautner phenomenon. Notations

- $N:=\left\{\left.\left[\begin{array}{ccc}1 & s & r \\ 0 & 1 & t \\ 0 & 0 & 1\end{array}\right] \right\rvert\, s, t, r \in \mathbb{R}\right\}, Z:=\left\{\mathbf{z}_{r}: \left.=\left[\begin{array}{ccc}1 & 0 & r \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right] \right\rvert\, r \in \mathbb{R}\right\} ;$
- $W:=\left\{\mathbf{w}_{t}: \left.=\left[\begin{array}{ccc}1 & 0 & 0 \\ 0 & 1 & t \\ 0 & 0 & 1\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\}, U:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ccc}1 & s & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\} ;$
- $\mathscr{H}$ is a separable Hilbert space and $\Phi: N \rightarrow \mathscr{U}(\mathscr{H})$ is a unitary representation of $N$.

Exercise 5.16. Verify the following

$$
\mathbf{w}_{t} \mathbf{u}_{s} \mathbf{w}_{-t}=\mathbf{u}_{s} \mathbf{z}_{-s t}, \forall s, t \in \mathbb{R} .
$$

Exercise 5.17. Show that a $\Phi(W)$-fixed vector is $\Phi(Z)$-fixed.
[Since $W \cdot Z$ is a normal subgroup of $N$ with quotient group $\mathbb{R}$, there exists a unitary representation $(\Phi, \mathscr{H})$ of $N$ and $v \in \mathscr{H}$ such that its stabilizer in $N$ is exactly $W$. Z.]

Now let $\Gamma$ be a lattice in $N$.
EXercise 5.18. Show that $\Gamma$ is not commutative, and hence, not virtually commutative (namely, every finite-index subgroup of $\Gamma$ is not commutative).

Exercise 5.19. Show that $\Gamma \cap Z$ is a lattice in $Z$.
Let $p: N \rightarrow N / Z$ ( $Z$ is normal in $N$ ) be the natural quotient map.
Exercise 5.20. Show that $p(\Gamma)$ is a lattice of $N / Z$.

Let $\widehat{\mathrm{m}}_{X}$ be the $N$-invariant probability measure on $N / \Gamma$ and let $\widehat{\mathrm{m}}_{\bar{X}}$ be the $N / Z$-invariant probability measure on $(N / Z) / p(\Gamma)$.

Exercise 5.21. Show that $\widehat{\mathrm{m}}_{X}$ is $W$-ergodic iff $\widehat{\mathrm{m}}_{\bar{X}}$ is $W$-ergodic.
EXERCISE 5.22. Fix $\Gamma$, show that there exists some one-parameter unipotent subgroup $\left\{\mathbf{v}_{s}\right\}$ of $N$ that acts ergodically on $\widehat{\mathrm{m}}_{X}$.

One more example.
Let $G:=\left\{\left[\begin{array}{lll}a & b & x \\ c & d & y \\ 0 & 0 & 1\end{array}\right] \left\lvert\,\left[\begin{array}{ll}a & b \\ c & d\end{array}\right] \in \mathrm{SL}_{2}(\mathbb{R})\right., x, y \in \mathbb{R}\right\}$.
$\Gamma:=\left\{\left[\begin{array}{lll}a & b & x \\ c & d & y \\ 0 & 0 & 1\end{array}\right] \left\lvert\,\left[\begin{array}{ll}a & b \\ c & d\end{array}\right] \in \mathrm{SL}_{2}(\mathbb{Z})\right., x, y \in \mathbb{Z}\right\}$.
EXERCISE 5.23. Use mixing and non-divergence of unipotent flow to show that $\mathrm{SL}_{2}(\mathbb{Z})$ is a lattice in $\mathrm{SL}_{2}(\mathbb{R})$.

EXercise 5.24. Show that $\Gamma$ is a lattice in $G$.
Let $\widehat{m}_{G / \Gamma}$ be the unique $G$-invariant probability measure on $G / \Gamma$.
Exercise 5.25. Show that $\widehat{\mathrm{m}}_{G / \Gamma}$ is $\mathrm{SL}_{2}(\mathbb{R})$-ergodic.
Here we embed $\mathrm{SL}_{2}(\mathbb{R})$ in the left upper corner of $G$. By what has been proved in the class, this implies that $\widehat{\mathrm{m}}_{G / \Gamma}$ is $\mathrm{SL}_{2}(\mathbb{R})$-mixing.

## CHAPTER 8

## Unipotent invariant finite measures on quotients of $\mathrm{SL}_{2}(\mathbb{R})$, I

Back to the Top.
In this chapter we introduce pointwise ergodic theorem and start our discussion of classification of unipotent invariant probability measures on quotients of $\mathrm{SL}_{2}(\mathbb{R})$.

## 1. Ergodicity and extremality

Unless otherwise specified, we assume $G$ and $X$ are nice. So $G$ is a locally compact and $\sigma$-compact metrizable group and $X$ is a $\sigma$-compact locally compact metrizable space. The space of probability measures $\operatorname{Prob}(X)$ with the weak* topology is not necessarily compact (unless $X$ is compact) but the $\operatorname{Meas}(X)^{\leq 1}$ is. And $\operatorname{Meas}(X)^{\leq 1}, \operatorname{Prob}(X)$ and $\operatorname{LFM}(X)$ are also nice (decompose $X$ into countable union of compact pieces and consider probability or finite measures supported on finite unions of them).

LEMMA 1.1. A $G$-invariant probability measure $\mu$ is ergodic iff it is extremal in the space of $G$-invariant probability measures. Or more succinctly, $\operatorname{Prob}(X)^{G, \operatorname{Erg}}=\operatorname{Extre}\left(\operatorname{Prob}(X)^{G}\right)$.

Being extremal means that $\mu$ can not be written as convex combination of different invariant probability measures. That is to say, if $\mu=a v_{1}+(1-a) v_{2}$ for some $a \in(0,1)$ and $v_{i} \in \operatorname{Prob}(X)^{G}$, then $v_{1}=v_{2}=\mu$. In particular, two different ergodic $\mu_{1}, \mu_{2}$ must be singular w.r.t. each other. Namely, we may partition $X=A \sqcup B$ into two measurable parts such that $\mu_{1}(B)=0$ and $\mu_{2}(A)=0$.

Sketch of Proof. If $\mu$ is not ergodic, then we can pick two complementary invariant measurable sets. Then $\mu$ is the sum of the restriction of $\mu$ to these two sets and is not extremal. Conversely, if $\mu=a v_{1}+(1-a) v_{2}$ then $v_{1}$ and $v_{2}$ are absolutely continuous w.r.t. $\mu$. So we find two $G$-invariant $L^{1}(\mu)$-functions representing " $\frac{\mathrm{d} v_{i} \text { " }}{\mathrm{d} \mu}$ which are forced to be constants unless $\mu$ is not ergodic.

By general facts from functional analysis (Hahn-Banach theorem), the convex combinations of $\operatorname{Extre}\left(\operatorname{Prob}(X)^{G}\right)$ are dense in $\operatorname{Prob}(X)^{G}$ (pretend $X$ to be compact first and then do the general case). A theorem of Choquet says that more precisely (See Thm.4.8 and 8.20 of the book of Einsiedler-Ward [EW11]),

Theorem 1.2 (Ergodic decomposition). For every $\mu \in \operatorname{Prob}(X)^{G}$ there exists a unique Borel probability measure $\lambda \in \operatorname{Prob}\left(\operatorname{Prob}(X)^{G}\right)$ such that

- $\lambda\left(\operatorname{Prob}(X)^{G, \operatorname{Erg}}\right)=1$;
- $\mu=\int_{v \in \operatorname{Prob}(X)}{ }^{G, \operatorname{Erg}} v \lambda(v)$.

Let me add that $\operatorname{Prob}(X)^{G, E r g}$ is not closed in general (Exercise: find such an example) but in the world of unipotent flows, this is closed due to a theorem of Mozes-Shah.

In virtue of this theorem, to classify invariant probability measures, we often start with ergodic ones.

## 2. Pointwise ergodic theorem for a flow

We can construct a new invariant probability measure from known ones by convex combination. But how to get one to start with? Well, in general such a measure may not exist (say, the $\mathrm{SL}_{2}(\mathbb{R})$-action on the space of lines of $\mathbb{R}^{2}$ ). But for a flow, namely a continuous $\mathbb{R}$-action (denote the action $\mathbb{R} \times X \rightarrow X$ by $\left.(t, x) \mapsto T_{t} . x\right)$ on a nice $X$, we can consider

$$
\frac{1}{T} \int_{0}^{T}\left(T_{t}\right)_{*} \delta_{x} \mathrm{dt}=\frac{1}{T} \int_{0}^{T} \delta_{T_{t} \cdot x} \mathrm{dt}
$$

as $T \rightarrow+\infty$. Here $\delta_{x}$ denotes the measure defined by $\delta_{x}(E)=1$ iff $x \in E$ and is zero otherwise. You can replace the $\delta$-measure supported on $\{x\}$ by any other probability measure. Using this construction, one shows that

Lemma 2.1. Let $\left(T_{t}\right)_{t \in \mathbb{R}}$ be a flow on $X$. If further assume $X$ is compact, then there exists a $\left(T_{t}\right)$-invariant probability measures.

Conversely, every ergodic flow-invariant probability measure may be constructed this way from a delta measure. Actually, more is true. This is the pointwise ergodic theorem.

Theorem 2.2. Let $T_{t}$ denote the action of $\mathbb{R}$ on a nice space $X$. Let $\mu$ be an ergodic Borel probability $\left(T_{t}\right)$-invariant measure on $X$. Then for every $f \in L^{1}\left(X, \mathscr{B}_{X}, \mu\right)$ there exists a measurable set $E_{f}$ of full measure $\left(\mu\left(E_{f}\right)=1\right)$ such that for every $x \in E_{f}$ we have

$$
\begin{equation*}
\lim _{T \rightarrow+\infty} \frac{1}{T} \int_{0}^{T} f\left(T_{t} \cdot x\right) \mathrm{dt}=\int f(x) \mu(x) \tag{27}
\end{equation*}
$$

Using the fact that $C_{c}(X)$ admits a countable dense subset for a nice $X$, a diagonal argument shows that

Corollary 2.3. Assumption as in the above theorem. There exists a full measure set $E$ such that for every $x \in E$,

$$
\begin{equation*}
\lim _{T \rightarrow+\infty} \frac{1}{T} \int_{0}^{T}\left(T_{t}\right)_{*} \delta_{x} \mathrm{dt}=\mu \tag{28}
\end{equation*}
$$

where the limit is taken with respect to the weak* topology.
There is no such general ergodic theorem beyond the world of amenable groups.
A point $x$ satisfying Equa.(27) (or (28)) is sometimes called f-generic (or generic). To emphasize both the group action and the invariant measure, one may also call $x$ a $\left(T_{t}, \mu\right)$-generic point. In general, it may be very difficult to describe the set of generic points. One beauty of unipotent flows is that you do have an explicit description of generic points in this case.

## 3. Ergodic measures for unipotent flows

Let us start with the easiest case.
Definition 3.1. Given a continuous action of $G$ on $X$. We say that the action is uniquely ergodic iff the action admits a unique invariant probability measure.

Theorem 3.2. Let $\Gamma$ be a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{R})$, then there exists a unique up-to-ascalar $\mathrm{SL}_{2}(\mathbb{R})$-invariant locally finite measure $\mathrm{m}_{\mathrm{X}}$ on $\mathrm{X}:=\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$.

A reference is Raghunathan's book [Rag72]. For existence and uniqueness of invariant (Haar) measures on a (nice) topological group, one may consult [DE14].

Thus when such a measure is finite, we get an example of uniquely ergodic action.
Lemma 3.3. Assume a flow $T_{t}$ on a compact space $X$ is uniquely ergodic with the unique invariant probability measure denoted by $\mu$, then for every $x \in \mathrm{X}$, Equa.(28) holds.

Proof is left as an exercise. You can not drop the compactness assumption.
Now we go to the world of unipotent flows. Some notations:

- $\mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R}), \Gamma$ is a discrete subgroup of G and $\mathrm{X}:=\mathrm{G} / \Gamma$.
- $\mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}, \mathrm{A}:=\left\{\mathbf{a}_{t}: \left.=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\} ;$
- B :=A•U.

Theorem 3.4. Assume $\Gamma$ is cocompact in G . Then the U -action on X is uniquely ergodic.
The existence is guaranteed. One needs to prove the uniqueness. This result is due to Furstenberg [Fur73].

A more general result is
Theorem 3.5. Let $\Gamma$ be a discrete subgroup of G . Then every $\mu \in \operatorname{Prob}(\mathrm{X})^{\mathrm{U}, \mathrm{Erg}}$ is one of the following:

1. supported on a closed (necc. compact) U-orbit;
2. $\mathrm{m}_{\mathrm{X}} /\left|\mathrm{m}_{\mathrm{X}}\right|$ with $\left|\mathrm{m}_{\mathrm{X}}\right|<\infty$.

In particular, if X has no compact U -orbit and $\left|\mathrm{m}_{\mathrm{X}}\right|=\mathrm{m}_{\mathrm{X}}(\mathrm{X})$ is not finite, then there is no finite U-invariant measure. Though this does not prevent the existence of dynamically interesting infinite U -invariant measures.

A reference for the material presented here is Ratner's paper [Ra92].

## 4. Outline of the proof and step 1

The proof of Thm. 3.5 to be presented here consists of two parts
Step 1. Upgrade from U-invariance to B-invariance if the measure is not supported on a compact U-orbit.
Step 2. Show that the action of $B$ is uniquely ergodic unless $m_{X}$ is infinite.
The first step is essentially achieved by a combination of ideas from Ch. 1 and pointwise ergodic theorem. It might be possible to do the second step by a duality argument in the style of Ch.1. We will do something different.
4.1. Step 1. Compared to Ch. 1 we will do the following adjustment

> compact topological spaces $\longrightarrow$ probability invariant measures
> minimal sets $\longrightarrow$ generic points

We shall actually use compact subsets of generic points so that we can take limits.
Lemma 4.1. Let $\mu$ be an ergodic U -invariant probability measure on X , then

1. either $\mu$ is supported on a closed U -orbit;
2. or $\mu$ is B -invariant.

Before the proof we make the following observation
Lemma 4.2. If $x, y$ are both $(\mathrm{U}, \mu)$ generic points and $y=g . x$ with $g \in \mathrm{G}$ normalizing U , then $g_{*} \mu=\mu$.

Proof. Since $g \in G$ normalizes $U$, we find some constant $c_{g}>0$ such that $g \mathbf{u}_{t} g^{-1}=\mathbf{u}_{c_{g} t}$. By definition of genericity we have

$$
\begin{aligned}
g_{*} \mu & =g_{*} \lim _{T \rightarrow+\infty} \frac{1}{T} \int_{t=0}^{T}\left(\mathbf{u}_{t}\right)_{*} \delta_{x} \mathrm{dt}=\lim _{T \rightarrow+\infty} \frac{1}{T} \int_{t=0}^{T} g_{*}\left(\mathbf{u}_{t}\right)_{*} \delta_{x} \mathrm{dt} \\
& =\lim _{T \rightarrow+\infty} \frac{1}{T} \int_{t=0}^{T}\left(\mathbf{u}_{c_{g}} \cdot t\right)_{*} g_{*} \delta_{x} \mathrm{dt}=\lim _{T \rightarrow+\infty} \frac{1}{T} \int_{t=0}^{T}\left(\mathbf{u}_{g_{g}} \cdot t\right)_{*} \delta_{y} \mathrm{dt}=\mu .
\end{aligned}
$$

Proof of Lemma 4.1. Without loss of generality assume $\mu$ is not supported on a closed U-orbit. In light of Lem.4. 2 above, we hope to find a pair $x, y$ that are both $(\mathrm{U}, \mu)$-generic and $y=\mathbf{a}_{t} . x$ with $t \neq 0$; and by varying the pair, we want $t$ to be arbitrarily close to 0 .

Recall that the argument from Ch. 1 basically goes like:
Step 1. find two sequences $\left(x_{n}\right)$ and $\left(y_{n}\right)$ with $d\left(x_{n}, y_{n}\right) \rightarrow 0$ and for each $n, x_{n}$ and $y_{n}$ are not on the same local U-orbit;
Step 2. if for infinitely many $n, x_{n}$ and $y_{n}$ are on the same local B-orbit, then we are done;
Step 3. otherwise, depending on $\delta>0$, we find $s_{n}, t_{n}$ such that every limit pair ( $x_{\infty}, y_{\infty}$ ) of $x_{n}^{\prime}:=\mathbf{u}_{t_{n}} x_{n}$ and $y_{n}^{\prime}:=\mathbf{u}_{s_{n}} y_{n}$ are differed by some $\mathbf{a}_{t}$ with $t \in\left[C^{-1} \delta, C \delta\right]$ for some constant $C>1$;
Step 4. as a complement to Step 3, it should be noted that the choice of $s_{n}$ is determined by $t_{n}$ and the choice of $t_{n}$ has the freedom of multiplying by a (multiplicatively) bounded number. This havs the effect of changing the $C$ in step 3 by another $C^{\prime}$;
Step 5. so far we have demonstrated $\mathbf{a}_{t}$ with $|t| \rightarrow 0, t \neq 0$ with $\mathbf{a}_{t} \in \mathrm{G}_{\mu}$, the stabilizer of $\mu$ in G. Since $\mathrm{G}_{\mu}$ is a closed subgroup, $A \subset \mathrm{G}_{\mu}$.

Below is a detailed account of carrying out the above strategy in the measure theoretic setting. You may try to figure out how by yourself.

We need to guarantee the limits ( $x_{\infty}, y_{\infty}$ ) to be generic. Since the set of generic points is usually not closed, we define

$$
E_{\mathrm{U}, \mu}:=\{(\mathrm{U}, \mu) \text {-generic points }\}
$$

and take $E$ to be a compact subset of $E_{\mathrm{U}, \mu}$ such that $\mu(E)>0.9$.
Take $T_{0}$ large enough such that the following set

$$
\begin{equation*}
F:=\left\{x \in \mathrm{X} \left\lvert\, \frac{1}{T} \operatorname{Leb}\left\{t \in[0, T], \mathbf{u}_{t} \cdot x \in E\right\} \geq 0.9\right., \forall T \geq T_{0}\right\} \tag{29}
\end{equation*}
$$

has $\mu(F)>0.9$. (how? First by ptws ergodic theorem applied to the indicator function of $E$, we see that

$$
\left\{x \in \mathrm{X} \left\lvert\, \lim _{T \rightarrow+\infty} \frac{1}{T} \operatorname{Leb}\left\{t \in[0, T], \mathbf{u}_{t} \cdot x \in E\right\}=\mu(E)>0.9\right.\right\}
$$

has full measure 1 . Thus as $S$ varies over positive integers, the increasing union of the following sets

$$
F_{S}:=\left\{x \in \mathrm{X} \left\lvert\, \frac{1}{T} \operatorname{Leb}\left\{t \in[0, T], \mathbf{u}_{t} \cdot x \in E\right\}>0.9\right., \forall T \geq S\right\}
$$

has measure 1. Thus we can find some $T_{0}$ such that $F_{T_{0}}$ has measure at least 0.9. )
We claim that there exist pairs $(x, y)$ in $F$ arbitrarily close to each other and yet not on the same local $U$-orbit (unless $\mu$ is supported on a compact $U$-orbit, which by assumption does not happen).

To be precise, two points $x, y$ are said to be on the same local U -orbit if $x=\mathbf{u}_{s} . y$ for some $s \in(-1,1)$.

If the claim were not true, then there exists $\varepsilon>0$ such that if $x, y \in F$ and $d(x, y)<\varepsilon$ then $x=\mathbf{u}_{s} y$ for some $|s|<1$. Cover $F$ by countably many measurable sets $\left\{B_{i}\right\}$ of diameter smaller than $\varepsilon$. Then $B_{i} \cap F \subset \mathbf{u}_{(-1,1)} \cdot x_{i}$ for some $x_{i}$. So $F \subset \bigcup_{i} \mathbf{u}_{(-1,1)} \cdot x_{i}$. Thus for some $x_{i}$,

$$
\mu\left(\left\{\mathbf{u}_{s} . x_{i}| | s \mid<1\right\}\right)>0 .
$$

By ergodicity this implies that U. $x$ has to close up (you can invoke ptws ergodic theorem to prove this but you do not have to) and that $\mu$ is the $U$-invariant measure supported on this orbit. Contradiction.

Recall calculation from Ch.1, with $s+t$ replaced by $t$,

$$
u_{t} A_{n} u_{s}^{-1}=\left[\begin{array}{cc}
1+a_{n}+t c_{n} & b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n}+(t-s)\left(1+d_{n}-s c_{n}\right)  \tag{30}\\
c_{n} & 1+d_{n}-s c_{n}
\end{array}\right]
$$

for

$$
A_{n}=\left[\begin{array}{cc}
1+a_{n} & b_{n} \\
c_{n} & 1+d_{n}
\end{array}\right] \quad \text { with } a_{n}, b_{n}, c_{n}, d_{n} \rightarrow 0 .
$$

Compared to Ch.1, let us make a little adjustment on the choice of $s_{n, \delta}$ and particularly $t_{n, \delta}$ to simplify matters. Assume $c_{n} \neq 0$. For a small number $\delta>0$, choose $s_{n, \delta}$ as before, namely,

$$
s_{n, \delta}:=\frac{d_{n}+\delta}{c_{n}}
$$

(choosing $s=\left(d_{n}-\delta\right) / c_{n}$ is also ok). We also need an additional parameter $\lambda=\lambda_{n, \delta} \in(0.1,1)$ to be determined in a moment. Let $s_{n, \delta}^{\prime}:=\lambda_{n, \delta} s_{n, \delta}$. Choose $t_{n, \delta}^{\prime}:=\phi_{n, \delta}\left(\lambda_{n, \delta}\right) \cdot s_{n, \delta}^{\prime}$ where

$$
\begin{equation*}
\phi_{n, \delta}\left(\lambda_{n, \delta}\right):=\frac{a_{n}-\left(1-\lambda_{n, \delta}\right) d_{n}+\lambda_{n, \delta} \delta}{1+\left(1-\lambda_{n, \delta}\right) d_{n}-\lambda_{n, \delta} \delta}+1=\frac{\left(a_{n}-d_{n}\right)+\left(d_{n}+\delta\right) \lambda_{n, \delta}}{\left(1+d_{n}\right)-\left(d_{n}+\delta\right) \lambda_{n, \delta}}+1 . \tag{31}
\end{equation*}
$$

This choice is such that the upper right corner of Equa.(30) converges to 0 asymptotically. Indeed with $\lambda=\lambda_{n, \delta}, s=s_{n, \delta}^{\prime}$ and $t=t_{n, \delta}^{\prime}$ we have

$$
\begin{gathered}
\quad b_{n}+s\left(d_{n}-a_{n}\right)-s^{2} c_{n}+(t-s)\left(1+d_{n}-s c_{n}\right) \\
\left(s c_{n}=d_{n}+\delta\right)=b_{n}+s\left(d_{n}-a_{n}\right)-s \lambda\left(d_{n}+\delta\right)+(t-s)\left(1+d_{n}\right)-(t-s) \lambda\left(d_{n}+\delta\right) \\
(\text { Equa.(31) })=b_{n}+s\left(-a_{n}+d_{n}-\lambda d_{n}-\lambda \delta\right)+(t-s)\left(1+d_{n}-\lambda d_{n}-\lambda \delta\right)=b_{n} \rightarrow 0 .
\end{gathered}
$$

Let us firstly cheat by assuming $\phi_{n, \delta}\left(\lambda_{n, \delta}\right) \equiv 1$ in Equa.(31). See Sec.4.2 below to see the true proof. We have

$$
s_{n, \delta}^{\prime}=t_{n, \delta}^{\prime}=\lambda_{n, \delta} s_{n, \delta}
$$

Take $\delta>0$ and $n$ large enough such that $s_{n, \delta}>T_{0}$. Then by the definition of $F$ (see Equa.(29)),

$$
\begin{align*}
& \operatorname{Leb}\left(\left\{\lambda \in(0.1,1) \mid \mathbf{u}_{\lambda_{n, \delta} .} . x_{n} \in E\right\}\right)>0.9-0.1=0.8 ;  \tag{32}\\
& \operatorname{Leb}\left(\left\{\lambda \in(0.1,1) \mid \mathbf{u}_{\lambda_{n, \delta} .} . y_{n} \in E\right\}\right)>0.9-0.1=0.8
\end{align*}
$$

In particular, their intersection is nonempty and we take some element $\lambda_{n, \delta}$. Define $x_{n, \delta}^{\prime \prime}:=$ $\mathbf{u}_{s_{n, \delta}^{\prime}} \cdot x_{n}$ and $y_{n, \delta}^{\prime \prime}:=\mathbf{u}_{s_{n, \delta}^{\prime}} \cdot y_{n}$, then $x_{n, \delta}^{\prime \prime} \cdot y_{n, \delta}^{\prime \prime} \in E$. By letting $n \rightarrow+\infty$ (pass to a subsequence if necessary) and by Equa.(30) above, we get

$$
y_{\infty, \delta}=\left[\begin{array}{cc}
\left(1-\lambda_{\infty, \delta} \delta\right)^{-1} & 0 \\
0 & 1-\lambda_{\infty, \delta} \delta
\end{array}\right] x_{\infty, \delta}
$$

where $x_{\infty, \delta}:=\lim x_{n, \delta}^{\prime \prime} \in E, y_{\infty, \delta}:=\lim y_{n, \delta}^{\prime \prime} \in E$ and $\lambda_{\infty, \delta}:=\lim \lambda_{n, \delta} \in[0.1,1]$. So we get a sequence of non-identity elements in A converging to id that maps some generic point ( $x_{\infty, \delta}$ ) to another one ( $y_{\infty, \delta}$ ). By Lem.4.2, they are contained in $\mathrm{G}_{\mu}$, which is a closed subgroup. Thus A is contained in $\mathrm{G}_{\mu}$ and the proof completes.

Finally, here is a summary-by-picture:

4.2. To avoid cheating... Here is the honest proof.

As functions on $[0,2]$ indexed by $n, \delta$, we can check that as $n \rightarrow \infty$ and $\delta \rightarrow 0$, the functions $\phi_{n, \delta}$ (resp., their derivatives) converge to the constant 1 (resp., 0 ) uniformly. Thus for $n$ sufficiently large and $\delta$ sufficiently small, we may and do assume that

$$
\phi_{n, \delta}(\lambda) \in[0.99,1.01], \quad \phi_{n, \delta}^{\prime}(\lambda) \in[-0.01,0.01], \quad \forall \lambda \in[0,2] .
$$

Let $\psi_{n, \delta}(\lambda):=\phi_{n, \delta}(\lambda) \cdot \lambda$. For $n$ large and $\delta$ small,

$$
\begin{equation*}
\psi_{n, \delta}^{\prime}(\lambda)=\phi_{n, \delta}^{\prime}(\lambda) \cdot \lambda+\phi_{n, \delta}(\lambda) \in[0.97,1.03], \quad \forall \lambda \in[0,2] \tag{33}
\end{equation*}
$$

So $\psi=\psi_{n, \delta}$ defines a diffeomorphism from $[0.1,1] \rightarrow \psi([0.1,1])$. Note that

$$
\begin{equation*}
[0.15,0.95] \subset \psi([0.1,1]) \subset[0.05,1.05] \tag{34}
\end{equation*}
$$

Let (abbr. $\psi:=\psi_{n, \delta}$ and $s:=s_{n, \delta}$ )

$$
\begin{align*}
& A:=\left\{\lambda \in[0.1,1] \mid \mathbf{u}_{\psi(\lambda) s} \cdot y_{n} \in E\right\} \\
& \Longrightarrow \psi(A)=\left\{\lambda \in \psi([0.1,1]) \mid \mathbf{u}_{\lambda s} \cdot y_{n} \in E\right\} ;  \tag{35}\\
& B:=\left\{\lambda \in[0.1,1] \mid \mathbf{u}_{\lambda s} \cdot x_{n} \in E\right\} .
\end{align*}
$$

By Equa.(32) and (34), we have $\operatorname{Leb}(\psi(A)), \operatorname{Leb}(B) \geq 0.8-0.1=0.7$. Also, by Equa.(33), for every $z \in A,\left|\psi^{\prime}(z)\right|^{-1} \geq 1.03^{-1}$. Therefore,

$$
\begin{aligned}
\operatorname{Leb}(A) & =\int 1_{A}(x) \mathrm{dx}=\int 1_{A}\left(\psi^{-1} y\right)\left|\left(\psi^{-1}\right)^{\prime}(y)\right| \mathrm{dy} \\
& =\int 1_{\psi(A)}(y)\left|\psi^{\prime}\left(\psi^{-1} y\right)\right|^{-1} \mathrm{dy} \geq 0.7 \cdot 1.03^{-1} \geq 0.6 .
\end{aligned}
$$

Consequently, $A \cap B \neq \varnothing$ and we choose some $\lambda_{n, \delta} \in A \cap B$. As above, define $s_{n, \delta}^{\prime}:=\lambda_{n, \delta} \cdot s_{n, \delta}$ and $t_{n, \delta}^{\prime}=\phi_{n, \delta}\left(\lambda_{n, \delta}\right) \cdot s_{n, \delta}^{\prime}=\psi_{n, \delta}\left(\lambda_{n, \delta}\right) \cdot s_{n, \delta}$. By Equa.(35), $x_{n, \delta}^{\prime \prime}:=\mathbf{u}_{s_{n, \delta}^{\prime}} \cdot x_{n}$ and $y_{n, \delta}^{\prime \prime}:=\mathbf{u}_{s_{n, \delta}^{\prime}} \cdot y_{n}$ belongs to $E$. The rest of the proof is the same as those below Equa.(32).

Now we have completed Step 1.

## 5. Exercises

### 5.1. Lattices and closedness of orbits.

- $G$ is a connected Lie group and $\Gamma$ is a discrete subgroup of $G$;
- $H \leq G$ is a closed subgroup.

EXERCISE 5.1. Assume $H \cap \Gamma$ is a lattice in $H$. Show that for a divergent sequence $\left(x_{n}\right)$ in $H / H \cap \Gamma, \operatorname{InjRad}\left(x_{n}\right) \rightarrow 0$.

EXERCISE 5.2. Assume $\Gamma$ satisfies the conclusion of the last exercise. Show that $Н Г / Г ~ i s ~$ closed in $G / \Gamma$.

- $U=\left\{\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right], s \in \mathbb{R}\right\}, \Gamma$ is a discrete subgroup of $\mathrm{SL}_{2}(\mathbb{R})$.

EXERCISE 5.3. Assume $U \cap \Gamma$ is cocompact in $U$, by duality we know that $\Gamma U / U$ is closed in $\mathrm{SL}_{2}(\mathbb{R}) / U$. The latter is homeomorphic to $\mathbb{R}^{2}-(0,0)$ under $g \mapsto g . e_{1}$. Thus $\Gamma . e_{1}$ is closed in $\mathbb{R}^{2}-(0,0)$. Show that, in fact, $\Gamma . e_{1}$ is closed in $\mathbb{R}^{2}$.

EXERCISE 5.4. Show that the conclusion might fail if we replace " $U \cap \Gamma$ is cocompact in $U$ " by "UГ is closed in $\mathrm{SL}_{2}(\mathbb{R})$ ".

EXERCISE 5.5. Show that $B=A \cdot U$ with $A:=\left\{\mathbf{a}_{t}=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right], t \in \mathbb{R}\right\}$ has no lattice.

### 5.2. More exercises.

ExERCISE 5.6. Let $\Gamma$ be a lattice in $\mathrm{SL}_{2}(\mathbb{R})$, and assume $\Gamma$ is not cocompact in $\mathrm{SL}_{2}(\mathbb{R})$. Let $X:=\mathrm{SL}_{2}(\mathbb{R}) / \Gamma$. Let d be a right invariant Riemannian metric on $\mathrm{SL}_{2}(\mathbb{R})$, which induces a quotient Riemannian metric $d_{X}$ on $X$, from which we can define a (volume) measure on $X$. Accept the fact that such a measure is necessarily the $\mathrm{SL}_{2}(\mathbb{R})$-invariant finite measure on $X$. Show that a sequence $\left(x_{n}\right) \subset X$ goes to $\infty$ iff $\operatorname{InjRad}\left(x_{n}\right) \rightarrow 0$ as $n \rightarrow \infty$.

ExErcise 5.7. Assume the notations and the conclusion of the exercise above. Show that $\left(g_{n} \Gamma / \Gamma\right) \subset X$ goes to $\infty$ iff there exists $\gamma_{n} \in \Gamma$ such that $\operatorname{dist}\left(\mathrm{id}, g_{n} \gamma_{n} g_{n}^{-1}\right) \rightarrow 0$.

EXERCISE 5.8. For a matrix $X=\left(x_{i, j}\right)$, let $\|X\|_{\text {sup }}:=\sup _{i, j}\left|x_{i, j}\right|$. By a direct computation, show that there exists a constant $C>0$, such that for every $\varepsilon>0$ and $X, Y \in \mathrm{SL}_{2}(\mathbb{R})$ with $\|\mathrm{id}-X\| \leq \varepsilon$ and $\|\mathrm{id}-Y\| \leq \varepsilon$, we have that

$$
\left\|\mathrm{id}-X Y X^{-1} Y^{-1}\right\| \leq C \cdot \varepsilon^{2}
$$

EXERCISE 5.9. Notations as in the exercise above. Show that there exists a neighborhood $\mathscr{N}$ of id in $\mathrm{SL}_{2}(\mathbb{R})$ such that for every discrete subgroup $\Gamma \leq \mathrm{SL}_{2}(\mathbb{R}), \Gamma \cap \mathscr{N}$ generates an abelian group.

ExERCISE 5.10. Notations as in the exercise above. Show that there exists a neighborhood $\mathscr{N}^{\prime}$ of id in $\mathrm{SL}_{2}(\mathbb{R})$ such that for every discrete subgroup $\Gamma \leq \mathrm{SL}_{2}(\mathbb{R})$, there exists $g \in \mathrm{SL}_{2}(\mathbb{R})$ such that $g \Gamma g^{-1} \cap \mathscr{N}^{\prime}=\{\mathrm{id}\}$.

ExERCISE 5.11. Let $\Gamma$ in $\mathrm{SL}_{2}(\mathbb{R})$ be a lattice. Use previous exercises to show that $\Gamma$ is not cocompact iff it contains non-identity unipotent matrices.

REMARK 5.1. The "if" direction is proved in the class. This is a special instance of KazhdanMargulis theorem.

EXERCISE 5.12. Let $a_{t}:=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right]$ and $u_{s}:=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right]$. In the class we have seen that for a discrete subgroup $\Gamma \leq \mathrm{SL}_{2}(\mathbb{R})$, if $x \in \mathrm{SL}_{2}(\mathbb{R}) / \Gamma$ belongs to a compact $u_{s}$-orbit, then $a_{t} \cdot x$ diverges as $t$ goes to $-\infty$. Now assume $\Gamma$ is a lattice. Show that the converse holds. Namely, if $a_{t} . x$ diverges as $t$ goes to $-\infty$, then $\left\{u_{s} \cdot x\right\}_{s \in \mathbb{R}}$ is compact.

## CHAPTER 9

## Unipotent invariant finite measures on quotients of $\mathrm{SL}_{2}(\mathbb{R})$, II

Back to the Top.
In this chapter we complete "step 2" from the last chapter. To do this we need the notion of conditional measures.

## 1. Conditional measures

As a reference, see [EW11, Ch.5] and [Cou16, Part IV and Ch.17].
Let $X$ be a nice space and $\mathscr{B}_{X}$ be its Borel $\sigma$-algebra. Let $\mu \in \operatorname{Prob}(X)$. Let $\mathscr{A}$ be a countably generated sub- $\sigma$-algebra of $\mathscr{B}_{X}$. Being countably generated means that, by definition, there exists a countable collection of measurable subsets $\operatorname{Gene}(\mathscr{A})=\left\{A_{i}\right\}$ of $X$ such that $\mathscr{A}$ is the smallest sub- $\sigma$-algebra containing them. Assume the complement of every $A_{i}$ is also contained in $\operatorname{Gene}(\mathscr{A})$. For $x \in X$, let the atom containing $x$ be $[x]^{\mathscr{A}}:=\bigcap_{A_{i} \ni x} A_{i}$.

Lemma 1.1. Actually $[x]^{\mathscr{A}}=\bigcap_{A \ni x, A \in \mathscr{A}} A$, hence $[x]^{\mathscr{A}}$ is independent of the choice of a (symmetric) countable generator Gene( $\mathscr{A}$ ).

Proof is left as an exercise.
Theorem 1.2. (Conditional measures) Let $\left(X, \mathscr{B}_{X}, \mu\right)$ and $\mathscr{A}$ be as above .

1. Existence of conditional measures.

There exists $X^{\prime} \in \mathscr{A}$ of full measure and a measurable map $X^{\prime} \rightarrow \operatorname{Prob}(X)$ denoted as $x \mapsto$ $\mu_{x}^{\mathscr{A}}$ such that $\mu_{x}^{\mathscr{}( }\left([x]^{\mathscr{A}}\right)=1$ and

$$
\begin{equation*}
\int_{A} \int f(y) \mu_{x}^{\mathscr{A}}(y) \mu(x)=\int_{A} f(x) \mu(x) \tag{36}
\end{equation*}
$$

for every $A \in \mathscr{A}$ and $f \in L^{1}\left(X, \mathscr{B}_{X}, \mu\right)$. Implicitly we have claimed that $x \mapsto \int f(y) \mu_{x}^{\mathscr{A}}(y)$ is integral on $A$.
2. Uniqueness of conditional measures.

If $x \mapsto v_{x}^{\mathscr{A}}$ is another measurable map from a possibly different full measure subset $X^{\prime \prime}$ to $\operatorname{Prob}(X)$ satisfying Equa.(36) for every compactly supp. cont. function $f \in C_{c}(X)$ and $A=X^{\prime \prime}$, then for some full measure set $X^{\prime \prime \prime} \subset X^{\prime} \cap X^{\prime \prime}$ we have $\mu_{x}^{\mathscr{A}}=v_{x}^{\mathscr{A}}$ for $x \in X^{\prime \prime \prime}$.

Example 1.3. Let $\mathscr{A}=\mathscr{B}_{X}$. Then $[x]^{\mathscr{A}}=\{x\}$ and $\mu_{x}^{\mathscr{A}}=\delta_{x}$ for every $x \in X$.
EXAMPLE 1.4. Let $\mathscr{A}$ be the sigma algebra generated by a finite measurable partition $\left\{P_{1}, \ldots, P_{l}\right\} \subset$ $\mathscr{B}_{X}$ of $X$, then $[x]^{\mathscr{A}}=P_{i}$ iff $x \in P_{i}$ and $\mu_{x}^{\mathscr{A}}=\frac{\mu \mid P_{i}}{\mu\left(P_{i}\right)}$.

Example 1.5. Let $X=[0,1] \times[0,1]$ and $\mu=$ Leb be the standard Lebesgue measure defined by $\mid \mathrm{dx} \wedge$ dy|. Let $\mathscr{A}:=\left\{A \times[0,1] \mid A \in \mathscr{B}_{[0,1]}\right\}$. Then for every $(x, y) \in X,[(x, y)]_{(x, y)}^{\mathcal{A}}=\{x\} \times[0,1]$ and $\mu_{(x, y)}^{\mathscr{d}}$ is induced by $|\mathrm{dy}|$.

This example can be generalized to foliations on manifolds where $X$ is a small open set with a local foliation chart, which provides $\mathscr{A}$.

Example 1.6. Everything same as in the last example except that we let $\mu$ be the standard Lebesgue measure supported on $\Delta:=\{(x, x), x \in[0,1]\}$. Then $[(x, y)]_{(x, y)}^{\mathscr{A}}=\{x\} \times[0,1]$ and $\mu_{(x, y)}^{\mathscr{L}}=$ $\delta_{y}$.

EXAMPLE 1.7. If you have a probability measure preserving map $\pi:\left(X, \mathscr{B}_{X}, \mu\right) \rightarrow\left(Y, \mathscr{B}_{Y}, v\right)$ with $X, Y$ nice. Let $\mathscr{A}:=\pi^{-1} \mathscr{B}_{Y}$. In this case, Equa.(36) can be viewed as a "fibre integration formula" (you can replace the $\mu$ on the LHS by $v$ ). Here atoms are fibres of $\pi$. In some sense, all countably generated sub $\sigma$-algebra $\mathscr{A}$ arises from such $a \pi$.

EXAMPLE 1.8. Let $G, X$ both be nice and assume $G$ preserves $\mu$. Let

$$
\mathscr{A}:=\left\{A \in \mathscr{B}_{X} \mid A \text { is almost } G \text {-invariant }\right\}
$$

A measurable subset $A$ is almost $G$-invariant if $\mu(g . A \Delta A)=0$ for all $g \in G$. Then Equa.(36) provides an explicit form of ergodic decomposition.

## 2. Step 2 of the measure classification

Notations:

- $\mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R}), \Gamma$ is a discrete subgroup of G and $\mathrm{X}:=\mathrm{G} / \Gamma$.
- let $m_{X}$ be a G-invariant locally finite measure on $X$ and let $\widehat{m}_{X}:=\frac{m_{X}}{m_{X}(X)}$ if $m_{X}(X)<+\infty$;
- $\mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}, \mathrm{A}:=\left\{\mathbf{a}_{t}: \left.=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\} ;$
- B:=A•U;
- $\mathrm{V}:=\left\{\mathbf{v}_{s}: \left.=\left[\begin{array}{ll}1 & 0 \\ s & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$.

Recall that we are left to prove the following.
Theorem 2.1. If there exists a B -invariant, U -ergodic probability measure $\mu$ on X , then $\Gamma$ is a lattice and $\mu$ is equal to $\widehat{\mathrm{m}}_{\mathrm{X}}$.

By the discussion from Chapter 7, we have the following
Lemma 2.2. Same assumption. The measure $\mu$ is ergodic (actually mixing) with respect to $a^{\mathbb{Z}}$-action for every $a_{\neq \mathrm{id}} \in \mathrm{A}$.

Let $\mu$ be a B-invariant, $a$-ergodic probability measure. Here $a$ is a fixed element of A such that $a^{n} v a^{-n} \rightarrow \mathrm{id}$ as $n \rightarrow+\infty$ for every $v \in \mathrm{~V}$. We need to show that $\mu$ coincides with the $\mathrm{m}_{\mathrm{X}}$ (up to a scalar) and in particular, $\mathrm{m}_{\mathrm{X}}$ is finite.

Fix some $o$ in the support of $\mu$. Choose (symmetric) neighborhoods of identity $\mathscr{N}_{\varepsilon}^{\mathrm{B}}$ (resp., $\mathscr{N}_{\varepsilon}{ }^{\mathrm{V}}$ ) in B (resp., V ) that are very small compared to the injectivity radius at $o$. We say two points $x, y$ are on the same local B (resp., V) orbit iff $x \in \mathscr{N}_{\varepsilon}^{\mathrm{B}} \cdot y$ (resp., $x \in \mathscr{N}_{\varepsilon}^{\mathrm{V}} \cdot y$ ). Choose $\delta>0$ even smaller compared to $\varepsilon$.

Let

$$
\operatorname{Gene}(f, \mu):=\left\{x \in \mathrm{X} \left\lvert\, \lim _{N \rightarrow+\infty} \frac{1}{N} \sum_{n=1}^{N} f\left(a^{n} x\right)=\int f(y) \mu(y)\right.\right\} .
$$

Note that this set is $\mathrm{V} \cdot \mathrm{A}$-invariant. Let $E_{f}$ be its intersection with $\mathscr{N}_{\delta}(o)$.
We define a sub- $\sigma$-algebra $\mathscr{A}$ on $\mathscr{N}_{\delta}(o)$ by specifying its atoms: $x$ and $y$ belong to the same atom iff $x$ and $y$ are on the same local B-orbit. Let $E_{f}^{\prime} \subset E_{f}$ be those $x$ such that the conditional measure $\mu_{x}^{\mathscr{A}}$ is the restriction of some (left-)B-invariant measure when we identify $[x]^{\mathscr{A}}$ as a subset of $\mathscr{N}_{\varepsilon}^{\mathrm{B}} \subset \mathrm{B}$ via the orbit map. Then $\mu$ being B-invariant, $E_{f}^{\prime}$ is a conull set in $E_{f}$. Let $\widetilde{E}_{f}$ consist of elements in $\mathscr{N}_{\delta}(o)$ that are on the local V-orbit of some element in $E_{f}^{\prime}$. Thus $\widetilde{E}_{f}$ is conull in $\mathscr{N}_{\delta}(o)$ with respect to $\mu$ and $\mathrm{m}_{\mathrm{X}}$.

As an exercise, the reader is invited to fill in the various missing details here. Consult Sec. 3 if it helps. Here is a picture.


Now we are ready to conclude the proof.
First assume $\mathrm{m}_{\mathrm{X}}<\infty$. Every point $x \in \widetilde{E}_{f}$ is $a$-generic for $\mu$. But since the $a^{\mathbb{Z}}$-action on $\mathrm{m}_{\mathrm{X}}$ is also ergodic and $\mathrm{m}_{\mathrm{X}}\left(\widetilde{E}_{f}\right)>0$, we can find a point $x \in \widetilde{E}_{f}$ generic for $\mathrm{m}_{\mathrm{X}}$. Thus $\int f(x) \mu(x)=$ $\int f(x) \mathrm{m}_{\mathrm{X}}(x)$ by pointwise ergodic theorem. Since $f$ is arbitrary we are done.

Now assume $\mathrm{m}_{\mathrm{X}}=\infty$. Then the associated unitary representation is absence of constants. Thus by mixing, for every $\phi, \psi \in L^{2}\left(\mathrm{X}, \mathrm{m}_{\mathrm{X}}\right)$, we have

$$
\lim _{n \rightarrow \infty} \int \phi\left(a^{n} \cdot x\right) \psi(x) \mathrm{m}_{\mathrm{X}}(x)=0
$$

Take $\phi=f$ and $\psi=1_{\widetilde{E}_{f}}$, then

$$
\lim _{n \rightarrow \infty} \int_{\widetilde{E}_{f}} f\left(a^{n} \cdot x\right) \mathrm{m}_{\mathrm{X}}(x)=\lim _{n \rightarrow \infty} \int f\left(a^{n} \cdot x\right) 1_{\widetilde{E}_{f}}(x) \mathrm{m}_{\mathrm{X}}(x)=0
$$

Let us compute, for $f \in C_{c}(\mathrm{X})$,

$$
\begin{aligned}
\mathrm{m}_{\mathrm{X}}\left(\widetilde{E}_{f}\right) \int f(x) \mu(x) & =\int_{\tilde{E}_{f}}\left(\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} f\left(a^{n} \cdot x\right)\right) \mathrm{m}_{\mathrm{X}}(x) \\
\text { (bounded convergence thm) } & =\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1}\left(\int_{\widetilde{E}_{f}} f\left(a^{n} \cdot x\right) \mathrm{m}_{\mathrm{X}}(x)\right)=0,
\end{aligned}
$$

which is impossible if $f>0$ at some point in $\operatorname{Supp}(\mu)$. Hence $\mathrm{m}_{\mathrm{X}}=\infty$ leads to a contradiction. See [Ra92, Page 27,28] for an alternative way of concluding the proof.

## 3. More details on locally invariant measures

Let $G$ be a Lie group and $\Omega \subset G$ be a nonempty open subset. A measure $\mu$ on $\Omega$ is said to be locally left invariant under $G$ iff for every measurable subset $A \subset \Omega$ and $g \in G$ such that $g A \subset \Omega$, we have $\mu(g A)=\mu(A)$.

Lemma 3.1. A locally left invariant locally finite measure $\mu$ is the restriction of some left $G$-invariant measure on $G$.

Proof. Fix a countable set $\left(g_{i}\right)_{i \in \mathbb{Z}}=0$ in $G$ such that $G=\bigcup g_{i} . \Omega$. Assume $g_{0}=\mathrm{id}$.

$$
A_{0}:=\Omega, A_{1}:=g_{1} . \Omega \backslash \Omega, A_{2}:=g_{2} . \Omega \backslash\left(\Omega \cup g_{1} \Omega\right) \ldots
$$

Then $G=\bigsqcup_{i \in \mathbb{Z}_{\geq 0}} A_{i}$.

Define a measure $\mu^{\prime}$ on $G$ by

$$
\begin{equation*}
\mu^{\prime}(E):=\sum_{i \geq 0} \mu\left(g_{i}^{-1}\left(E \cap A_{i}\right)\right) \tag{37}
\end{equation*}
$$

for every measurable subset $E$. Then one can prove that $\mu^{\prime}$ is left $G$-invariant. Here are more details.

Take $g \in G$. For simplicity let $E_{i}:=E \cap A_{i}$, then

$$
\mu^{\prime}(g \cdot E)=\sum_{j} \mu^{\prime}\left(g \cdot E_{j}\right)=\sum_{i, j} \mu\left(g_{i}^{-1}\left(g \cdot E_{j} \cap A_{i}\right)\right) .
$$

Note that $g_{i}^{-1}\left(g . E_{j} \cap A_{i}\right) \subset \Omega$ and $\left(g \circ g_{j}\right)^{-1}\left(g . E_{j} \cap A_{i}\right) \subset \Omega$. By local left-invariance we get

$$
\mu\left(g_{i}^{-1}\left(g \cdot E_{j} \cap A_{i}\right)\right)=\mu\left(g_{j}^{-1} g^{-1}\left(g \cdot E_{j} \cap A_{i}\right)\right)
$$

Note that

$$
\bigsqcup_{i} g_{j}^{-1} g^{-1}\left(g \cdot E_{j} \cap A_{i}\right)=g_{j}^{-1} g^{-1}\left(g \cdot E_{j}\right)=g_{j}^{-1} \cdot E_{j} .
$$

Thus from Equa.(37) we get

$$
\mu^{\prime}(g \cdot E)=\sum_{i, j} \mu\left(g_{j}^{-1} g^{-1}\left(g \cdot E_{j} \cap A_{i}\right)\right)=\sum_{j} \mu\left(g_{j}^{-1} \cdot E_{j}\right)=\mu^{\prime}(E) .
$$

So we are done.

To check local-invariance, the following is helpful.
Lemma 3.2. Let $\delta>0$. Let $\Omega$ be a connected open subset of $G$ and $\mu$ be a measure on $\Omega$. Assume $\mu$ is locally invariant with respect to $g \in \mathscr{N}_{\delta}(\mathrm{id}) \subset G$. Then $\mu$ is locally invariant.

Note that being open connected and being open path-connected is equivalent for a subset of a manifold.

Proof. Fix a countable dense subset $\left\{\epsilon_{i}\right\}$ of $\mathcal{N}_{\delta}$ (id). For every $g \in G$, the collection of finite-length words in $\left\{\epsilon_{i}\right\}$ representing $g$ is a countable set, we may index it by $\mathbb{Z}_{\geq 1}$. More precisely,

$$
\mathscr{W}(g):=\left\{w_{k}:\{1,2, \ldots, l\} \rightarrow\left\{\epsilon_{i}\right\} \mid l, k \in \mathbb{Z}_{\geq 1}, w_{k}(l) \cdot \ldots \cdot w_{k}(1)=g\right\}
$$

Write $l_{k}$ for the length of the word $w_{k}$. Let $E \subset \Omega$ with $g . E \subset \Omega$. We need to show $\mu(E)=\mu(g . E)$. Now for every $w \in \mathscr{W}(g)$, consider

$$
\begin{aligned}
& E_{1}:=\left\{x \in E \mid w_{1}(k) \cdot \ldots \cdot w_{1}(1) \cdot x \in \Omega, \forall k=1, \ldots, l_{1}\right\} \\
& E_{2}:=\left\{x \in E \backslash E_{1} \mid w_{2}(k) \cdot \ldots \cdot w_{2}(1) \cdot x \in \Omega, \forall k=1, \ldots, l_{2}\right\}
\end{aligned}
$$

Then $E_{i}$ 's are disjoint from each other. Moreover, since $\Omega$ is path-connected and $\left\{\epsilon_{i}\right\}$ is dense in $\mathscr{N}_{\delta}(\mathrm{id})$, we have

$$
E=\bigsqcup_{i \in \mathbb{Z} \geq 1} E_{i} .
$$

Moreover for each $i$,

$$
\mu\left(E_{i}\right)=\mu\left(w_{i}(1) \cdot E_{i}\right)=\mu\left(w_{i}(2) w_{i}(1) \cdot E_{i}\right)=\ldots=\mu\left(g \cdot E_{i}\right) .
$$

Therefore,

$$
\mu(E)=\sum \mu\left(E_{i}\right)=\sum \mu\left(g \cdot E_{i}\right)=\mu(g \cdot E) .
$$

## 4. Exercises

## CHAPTER 10

## Equidistribution of unipotent flows on quotients of $\mathrm{SL}_{2}(\mathbb{R})$

## Back to the Top.

Notations

- $\mathrm{X}_{2}:=\left\{\right.$ unimodular lattices in $\left.\mathbb{R}^{2}\right\} \cong \mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z}), \mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R})$;
- $\mathrm{U}:=\left\{\mathbf{u}_{s}: \left.=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\} ; \mathrm{A}:=\left\{\left.\mathbf{a}_{t}=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right] \right\rvert\, t \in \mathbb{R}\right\} ;$
- $\widehat{\mathrm{m}}_{\mathrm{X}_{2}}$ is the $\mathrm{SL}_{2}(\mathbb{R})$-invariant probability measure on $\mathrm{X}_{2}$;
- $\operatorname{Prim}(\Lambda)$ is the set of non-zero primitive vectors in $\Lambda$ for $\Lambda \leq \mathbb{R}^{2}$ discrete;
- $\operatorname{Prim}^{1}(\Lambda)$ is the set of rank-1 primitive subgroups of $\Lambda$.


## 1. Equidistribution on the modular surface

In this section, we illustrate the idea of [DS84] in the case $\mathrm{X}_{2}$. The general case will be discussed in Sec.3.

Theorem 1.1. Let $\Lambda_{0} \in \mathrm{X}_{2}$ be such that $\mathrm{U} . \Lambda_{0}$ is not compact. Then

$$
\lim _{S \rightarrow+\infty} \mu_{S}:=\lim _{S \rightarrow+\infty} \int_{0}^{S}\left(\mathbf{u}_{s}\right)_{*} \delta_{\Lambda_{0}} \mathrm{ds}=\widehat{\mathrm{m}}_{\mathrm{X}_{2}} .
$$

Consider

$$
\mathscr{T}:=\left\{\Lambda \in X_{2} \mid \mathrm{U} . \Lambda \text { is compact }\right\} .
$$

Lemma 1.2. The set of compact U -orbits is a tube: $\mathscr{T}=\left\{\mathbf{a}_{t} \mathbf{u}_{s} \cdot \mathbb{Z}^{2}, t \in \mathbb{R}, s \in \mathbb{R} / \mathbb{Z}\right\}$. And $\mathrm{U} . \Lambda$ is compact iff $\Lambda$ contains a non-zero horizontal vector.

This has been proved in Chapter 4.
Our proof of Thm.1.1 decomposes as:
Step 1. Passing to a subsequence, assume the limit of $\left(\mu_{S}\right)_{S}$ exists and call it $\mu$. Thanks to the non-divergence theorem (see Ch.4), we also know $\mu$ is a probability measure.
Step 2. Also $\mu$ is readily seen to be U -invariant since it comes from an averaging process.
Step 3. Show $\mu(\mathscr{T})=0$.
Step 4. Use the ergodic decomposition to conclude.
Details of Step 1 and 2 are left as an exercise. Let us take up Step 3.
Proof of Step 3. Fix $t_{1}<t_{2}$, let

$$
\mathscr{T}_{\left[t_{1}, t_{2}\right]}:=\left\{\mathbf{a}_{t} \mathbf{u}_{s} \cdot \mathbb{Z}^{2}, t \in\left[t_{1}, t_{2}\right], s \in \mathbb{R} / \mathbb{Z}\right\} .
$$

Thus it suffices to show that $\mu\left(\mathscr{T}_{\left[t_{1}, t_{2}\right]}\right)=0$ for all $-\infty<t_{1}<t_{2}<+\infty$. By the definition of weak* convergence, it suffices to find an open neighborhood $\mathscr{N}_{\varepsilon}$, for every $\varepsilon>0$, of $\mathscr{T}_{\left[t_{1}, t_{2}\right]}$ such that $\limsup \mu_{S}\left(\mathscr{N}_{\varepsilon}\right) \leq \varepsilon$. Letting $\varepsilon \rightarrow 0$ the finishes the proof.

This will be achieved by Thm.1.5 below.

Note that $\mathbf{u}_{s} \Lambda_{0}$ being close to $\mathscr{T}_{\left[t_{1}, t_{2}\right]}$ means that, for certain $v \in \operatorname{Prim}\left(\Lambda_{0}\right)$, we have $\mathbf{u}_{s} \cdot v$ is close to

$$
A_{\left[t_{1}, t_{2}\right]}:=\left\{\mathbf{a}_{t} \mathbf{u}_{s} . e_{1} \mid t \in\left[t_{1}, t_{2}\right], t \in \mathbb{R}\right\}=\left[e^{t_{1}}, e^{t_{2}}\right] \times\{0\} .
$$

For $C, \delta>0$, consider the box

$$
\operatorname{Box}_{C, \delta}:=[-C, C] \times[-\delta, \delta] .
$$

Define

$$
I(C, \delta):=\left\{s \geq 0 \mid \operatorname{Prim}\left(\mathbf{u}_{s} . \Lambda_{0}\right) \cap \operatorname{Box}_{C, \delta} \neq \varnothing\right\} .
$$

For $\mathbb{Z} v \in \operatorname{Prim}^{1}\left(\Lambda_{0}\right)$, consider

$$
I(C, \delta, \mathbb{Z} v):=\left\{s \geq 0 \mid \mathbf{u}_{s} . v \in \operatorname{Box}_{C, \delta}\right\} .
$$

Since $-\operatorname{Box}_{C, \delta}=\operatorname{Box}_{C, \delta}$, this is independent of the choice of the generator of $\mathbb{Z} \nu$. From the definition,

$$
\begin{equation*}
I(C, \delta)=\bigcup_{\mathbb{Z} \nu \in \operatorname{Prim}^{1}\left(\Lambda_{0}\right)} I(C, \delta, \mathbb{Z} \nu) . \tag{38}
\end{equation*}
$$

The key observation is that
Lemma 1.3. Assume $\delta \cdot C \leq 0.1$. Then for two $\mathbb{Z} v \neq \mathbb{Z} w \in \operatorname{Prim}^{1}\left(\Lambda_{0}\right), I(C, \delta, \mathbb{Z} v) \cap I(C, \delta, \mathbb{Z} w)=$ $\varnothing$. In other words, Equa.(38) above is a disjoint union when $\delta \cdot C \leq 0.1$.

Proof. Otherwise the lattice $\mathbf{u}_{s} . \Lambda_{0}$ would contain two linearly independent vectors $v, w$ in $[-C, C] \times[-\delta, \delta]$. Thus the triangle spanned by $v, w$ is also contained in $[-C, C] \times[-\delta, \delta]$, implying $\|\nu \wedge w\| \leq 2(4 C \delta)<1$. This contradicts against the assumption $\Lambda_{0}$ is unimodular.

For $\varepsilon>0$, define

$$
C_{1}(\varepsilon):=\varepsilon^{-1}, \quad \delta_{1}(\varepsilon):=0.1 \varepsilon
$$

For every $\mathbb{Z} \nu \in \operatorname{Prim}^{1}\left(\Lambda_{0}\right)$, there are three cases
Case 1. $I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\varnothing$;
Case 2. $I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right) \neq \varnothing$ and $\mathbb{Z} \nu \nsubseteq \mathbb{R} e_{1}$; in this case $I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\mathbb{R} \geq 0$;
Case 3. $I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right) \neq \varnothing$ and $\mathbb{Z} \nu \nsubseteq \mathbb{R} e_{1} ;$ in this case $I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)$ is a closed interval of the form $\left[a_{v}, b_{v}\right]$.
Case 2 is excluded since $\Lambda_{0}$ contains no non-zero horizontal vector by assumption (see Lem.1.2).
Now take $S>0$, there are sub-cases for case 3:
3.1 $S<a_{\nu}$ or $b_{\nu}<0$; in this case $[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\varnothing$;
$3.2 a_{\nu} \leq 0 \leq b_{\nu} \leq S$; in this case $[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\left[0, b_{\nu}\right]$;
$3.30<a_{\nu} \leq b_{\nu}<S$; in this case $[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\left[a_{\nu}, b_{\nu}\right]$;
$3.40 \leq a_{\nu} \leq S \leq b_{\nu}$; in this case $[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\left[a_{\nu}, S\right]$;
$3.5[0, S] \subset\left[a_{\nu}, b_{\nu}\right]$.


Proposition 1.4. Take $C_{2}$ satisfying $1<C_{2}<0.5 C_{1}(\varepsilon)=0.5 \varepsilon^{-1}$. Then

$$
\limsup _{S \rightarrow+\infty} \frac{1}{S} \operatorname{Leb}\left(I\left(C_{2}, \delta_{1}(\varepsilon)\right) \cap[0, S]\right) \leq 4 C_{2} \varepsilon
$$

From the proof it will be clear that the inequality holds for $S$ large enough without taking the limit.

Only case $3.2,3.3$ and 3.4 above will contribute, for which we have three lemmas Lem.2.2, 2.1, and 2.3 below.

Proof. If every $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$ falls in case 1 or case 3.1 (for every $S>0$ ), then LHS in Prop.1.4 is zero and the inequality trivially holds. Otherwise, find $S>0$ large enough such that for some $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$, we are in case $3.2-3.5$. By choosing $S$ larger, case 3.5 can be excluded.

$$
\begin{aligned}
& \operatorname{Leb}\left(I\left(C_{2}, \delta_{1}(\varepsilon)\right) \cap[0, S]\right)=\left|\bigsqcup_{\nu \in \text { case3 }}[0, S] \cap I\left(C_{2}, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \\
& \quad(\text { Lem.2.2, 2.1, 2.3 }) \leq 4 C_{2} \varepsilon \cdot \sum\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq 4 C_{2} \varepsilon \cdot S
\end{aligned}
$$

Now we can prove:
Theorem 1.5. For every $\varepsilon>0$, there is a neighborhood $\mathscr{N}_{\varepsilon}$ of $\mathscr{T}_{t_{1}, t_{2}}$ such that

$$
\limsup _{S \rightarrow+\infty} \mu_{S}\left(\mathscr{N}_{\varepsilon}\right) \leq \varepsilon
$$

Consequently for every limit point $\mu$ of $\left(\mu_{S}\right), \mu\left(\mathscr{T}_{t_{1}, t_{2}}\right)=0$.
Proof. Take $C_{2}>1$, depending on $t_{1}, t_{2}$, such that $\operatorname{Box}_{C_{2}, \delta_{1}(\varepsilon)}$ contains $\left[e^{t_{1}}, e^{t_{2}}\right] \times\{0\}$. Let $\varepsilon^{\prime}:=\frac{\varepsilon}{4 C_{2}}$. When $\varepsilon>0$ is small enough, $C_{2}<0.5\left(\varepsilon^{\prime}\right)^{-1}$. Define $\mathscr{N}_{\varepsilon}$ to be those lattices whose primitive vectors intersect non-trivially with $\mathrm{Box}_{\mathrm{C}_{2}, \delta_{1}\left(\varepsilon^{\prime}\right)}$. Then Prop.1.4 concludes the proof.

Thus we have completed Step 3.
Proof of Step 4. Say we have a U-invariant probability measure $\mu$ with $\mu(\mathscr{T})=0$. By classification of ergodic U-invariant probability measures $v$ on X (see Ch. 4 Thm.3.5), either $v$ is supported on $\mathscr{T}$ or $v=\widehat{\mathrm{m}}_{\mathrm{X}_{2}}$. Let

$$
\mu=\int_{\operatorname{Prob}\left(X_{2}\right) \mathrm{U}, \operatorname{Erg}} v \lambda(v)
$$

be the ergodic decomposition of $\mu$, then

$$
0=\mu(\mathscr{T})=\int v(\mathscr{T}) \lambda(v) .
$$

Thus $\lambda$-almost every $v, v(\mathscr{T})=0 \Longrightarrow v=\widehat{\mathrm{m}}_{\mathrm{X}_{2}}$. So $\mu=\widehat{\mathrm{m}}_{\mathrm{X}_{2}}$.

## 2. Supplementary lemmas

Lemma 2.1. (For case 3.3 above) Assume $\Lambda_{0} \cap \mathbb{R} e_{1}=\{0\}$, then for $C>0$,

$$
\left|I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq C \varepsilon \cdot\left|I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right|
$$

Proof. If the LHS is 0 , then nothing needs to be done. Otherwise, assume w.l.o.g that $v=\left(\nu_{1}, v_{2}\right)$ with $\nu_{2}>0$. Then

$$
\left[\begin{array}{ll}
1 & s \\
0 & 1
\end{array}\right]\binom{v_{1}}{v_{2}}=\binom{v_{1}+s v_{2}}{v_{2}}
$$

and

$$
\begin{equation*}
I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} v\right)=\frac{1}{v_{2}}\left[-v_{1}-C,-v_{1}+C\right] . \tag{39}
\end{equation*}
$$

Thus

$$
\left|I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} v\right)\right|=\frac{2 C}{v_{2}}=C \varepsilon \cdot \frac{2 \varepsilon^{-1}}{v_{2}}=C \varepsilon \cdot\left|I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} v\right)\right| .
$$

Lemma 2.2. (For Case 3.2 above) Assume $\Lambda_{0} \cap \mathbb{R} e_{1}=\{0\}$. Take $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$ and $S>0$ satisfying case 3.2 above. Also let $1<C \leq 0.5 \varepsilon^{-1}$. Then

$$
\left|[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq 4 C \varepsilon \cdot\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| .
$$

Proof. W.l.o.g, keep assuming that $v=\left(\nu_{1}, v_{2}\right)$ with $v_{2}>0$. In case 3.2,

$$
[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} v\right)=\left[0, b_{v}\right]
$$

If $[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)$ is empty nothing needs to be done. Otherwise, by Equa.(39),

$$
0<-v_{1}+C \Longrightarrow v_{1}<C .
$$

Then

$$
\begin{aligned}
\left|[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq \frac{2 C}{v_{2}} & =\frac{2 C}{-v_{1}+\varepsilon^{-1}} \cdot \frac{-v_{1}+\varepsilon^{-1}}{v_{2}} \\
& =\frac{2 C}{-v_{1}+\varepsilon^{-1}} \cdot\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right|
\end{aligned}
$$

It remains to observe

$$
\frac{2 C}{-v_{1}+\varepsilon^{-1}} \leq \frac{2 C}{-C_{2}+\varepsilon^{-1}} \leq \frac{2 C}{-0.5 \varepsilon^{-1}+\varepsilon^{-1}}=4 C \varepsilon .
$$

Lemma 2.3. [For Case 3.4 above] Assume $\Lambda_{0} \cap \mathbb{R} e_{1}=\{0\}$. Take $\mathbb{Z} v \in \operatorname{Prim}^{1}(\Lambda)$ and $S>0$ satisfying case 3.4 above. Also let $1<C \leq 0.5 \varepsilon^{-1}$. Then

$$
\left|[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq 4 C \varepsilon \cdot\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| .
$$

Proof. W.l.o.g, keep assuming that $v=\left(\nu_{1}, v_{2}\right)$ with $v_{2}>0$. In case 3.4

$$
[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)=\left[a_{v}, S\right]
$$

If $[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)$ is empty, nothing needs to be done. Otherwise, by Equa.(39),

$$
\frac{-v_{1}-C}{v_{2}} \leq S \Longrightarrow v_{1}+v_{2} S \geq C
$$

Under this condition we have

$$
\left|[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq \frac{2 C}{v_{2}}
$$

and

$$
\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right|=S-\frac{\left(-v_{1}-\varepsilon^{-1}\right)}{v_{2}}=\frac{\varepsilon^{-1}+v_{2} S+v_{1}}{v_{2}} \geq \frac{\varepsilon^{-1}-C}{v_{2}} \geq \frac{0.5 \varepsilon^{-1}}{v_{2}} .
$$

Thus,

$$
\left|[0, S] \cap I\left(C, \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| \leq \frac{2 C}{0.5 \varepsilon^{-1}}\left|[0, S] \cap I\left(C_{1}(\varepsilon), \delta_{1}(\varepsilon), \mathbb{Z} \nu\right)\right| .
$$

Note that $\frac{2 C}{0.5 \varepsilon^{-1}}=4 C \varepsilon$.

## 3. [Not readable at the moment]Other non-cocompact lattices

[Needs further revision!]
Let $\Gamma \leq \mathrm{G}:=\mathrm{SL}_{2}(\mathbb{R})$ be a lattice. Let $\mathrm{X}:=\mathrm{G} / \Gamma$. We are going to assume some light hyperbolic geometry. Readers who are less familiar with hyperbolic geometry are welcome to take $\Gamma=$ $\mathrm{SL}_{2}(\mathbb{Z})$. Main ideas are preserved in this case.

The discussion here is more "geometric" compared to the last section.
First we have the non-divergence theorem (compare Chapter 4).
THEOREM 3.1. For every $\varepsilon>0$, there exists a compact subset of $\mathscr{C} \subset X$ such that for every $x \in X$, either

$$
\limsup _{S \rightarrow+\infty} \frac{1}{S} \operatorname{Leb}\left\{s \in[0, S], \mathbf{u}_{s} . x \notin \mathscr{C}\right\} \leq \varepsilon
$$

or $\mathrm{U} . \mathrm{x}$ is compact.
Proof is left as an exercise.
Let

$$
\mathscr{T}:=\{x \in \mathrm{X} \mid \mathrm{U} . x \text { is compact }\} .
$$

One can show that
THEOREM 3.2. There exist finitely many points $y_{1}, \ldots, y_{l}$ in X with compact U -orbits such that if $\mathscr{T}_{i}:=\mathrm{AU} . y_{i}$ then

$$
\mathscr{T}=\bigsqcup_{i=1, . ., l} \mathscr{T}_{i}
$$

Fix $x_{0} \notin \mathscr{T}$, let

$$
\mu_{S}:=\frac{1}{S} \int_{0}^{S}\left(\mathbf{u}_{s}\right)_{*} \delta_{x_{0}} \mathrm{ds}
$$

and take $\mu$ to be a weak* limit. Let us explain why $\mu(\mathscr{T})=0$, which follows if $\mu\left(\mathscr{T}_{i}\right)=0$ for every $i=1, \ldots, l$. From now on we focus on a single index $i_{0}$. W.l.o.g., assume $y_{i_{0}}=[\mathrm{id}]_{\Gamma}$, where $[\bullet]_{\Gamma}$ stands for the image of $\bullet$ in the quotient by $\Gamma$.
3.1. Lifts of tubes. Define, for $-\infty \leq t_{1}<t_{2} \leq+\infty$,

$$
\begin{aligned}
& \mathscr{T}_{t_{1}, t_{2}, i_{0}}:=\left\{\mathbf{a}_{t} U \cdot y_{i_{0}} \mid t_{1}<t<t_{2}, s \in \mathbb{R}\right\} \\
& \widetilde{\mathscr{T}}_{t_{1}, t_{2}, i_{0}}:=\left\{\mathbf{a}_{t} U . \widetilde{y}_{i_{0}} \mid t_{1}<t<t_{2}, s \in \mathbb{R}\right\}, \widetilde{\mathscr{T}}_{i_{0}}:=\text { AU. } \widetilde{y}_{i_{0}} .
\end{aligned}
$$

where $\widetilde{y}_{i_{0}}=[\mathrm{id}]_{\Gamma \cap \pm \mathrm{U}} \in \mathrm{G} / \Gamma \cap \pm \mathrm{U}$. In general, one should lift $y_{i}=:\left[g_{i}\right]_{\Gamma}$ to $\widetilde{y}_{i}=\left[g_{i}\right]_{\Gamma \cap \pm g_{i} U g_{i}^{-1}}$.


Theorem 3.3. Fix some $-\infty<t_{1}<t_{2}<+\infty$. For every $\varepsilon>0$, there exists a neighborhood $\mathscr{N}_{\varepsilon}$ of $\mathscr{T}_{t_{1}, t_{2}}$ such that

$$
\limsup _{S \rightarrow+\infty} \frac{1}{S} \operatorname{Leb}\left\{s \in[0, S] \mid \mathbf{u}_{s} . x_{0} \in \mathscr{N}_{\varepsilon}\right\} \leq \varepsilon .
$$

In light of the case of $\mathrm{X}_{2}$, we are going to find two neighborhoods $\mathscr{N}_{\varepsilon} \subset \mathscr{N}_{\varepsilon}^{\prime}$ such that the time a noncompact $U$-orbit spends in $\mathscr{N}_{\varepsilon}$ is much shorter than that in $\mathscr{N}_{\varepsilon}^{\prime}$.

Consider the natural projection $p: \mathrm{G} / \pm \mathrm{U} \cap \Gamma \rightarrow \mathrm{G} / \Gamma$. It is an injection restricted to $\widetilde{\mathscr{T}_{i}}$ and is a closed embedding when restricted to the closure of $\widetilde{\mathscr{T}}_{s, t}$ for every pair $s<t$. For $t_{1}^{\prime}<t_{1}$ and $t_{2}^{\prime}=t_{2}^{\prime}(\varepsilon)$ to be determined, there exists an open neighborhood $\widetilde{\Omega}_{\varepsilon}$ of $\widetilde{\mathscr{T}}_{t_{1}^{\prime}, t_{2}^{\prime}}$ such that

$$
\begin{equation*}
\left.p\right|_{\tilde{\Omega}_{\varepsilon}}: \widetilde{\Omega}_{\varepsilon} \rightarrow p\left(\widetilde{\Omega}_{\varepsilon}\right) \tag{40}
\end{equation*}
$$

is a homeomorphism.
Definition 3.4. For $T \in \mathbb{R}$,

$$
\begin{aligned}
\operatorname{cusp}_{i_{0}}(T) & :=\left\{k \mathbf{a}_{t} U \cdot y_{i_{0}} \mid k \in \mathrm{SO}_{2}(\mathbb{R}), t<T\right\}=\mathrm{SO}_{2}(\mathbb{R}) \mathscr{T}_{-\infty, T, i_{0}} ; \\
\widetilde{\operatorname{cusp}}_{i_{0}}(T) & :=\left\{k \mathbf{a}_{t} U \cdot \widetilde{y}_{i_{0}} \mid k \in \mathrm{SO}_{2}(\mathbb{R}), t<T\right\}=\mathrm{SO}_{2}(\mathbb{R}) \mathscr{T}_{-\infty, T, i_{0}},
\end{aligned}
$$

Lemma 3.5. There exists $T_{0} \in \mathbb{R}$ such that the following holds. Under $p, \widetilde{\operatorname{cusp}}_{i_{0}}\left(T_{0}\right)$ is mapped homeomorphically onto $\operatorname{cusp}_{i_{0}}\left(T_{0}\right)$. For $s \in \mathbb{R}$, there exists $T(s)<T_{0}$ such that $\mathscr{T}_{T_{0}, s, i_{0}}$ does not intersect cusp $_{i_{0}}(T(s)$ ).

Anticipating the proof, $t_{2}^{\prime}$ will be chosen depend on $\varepsilon$ and linear algebra. Then we choose $t_{1}^{\prime}$ to be $T_{1}^{\prime}+1$ where $T_{1}^{\prime}:=T\left(t_{2}^{\prime}\right)$ comes from Lem.3.5. Then $\widetilde{\Omega}_{\varepsilon}$ is chosen by Equa.(40).

LEMMA 3.6. The natural projection $p$ restricted to $\operatorname{cusp}_{i_{0}}\left(T_{1}^{\prime}\right) \cup \widetilde{\Omega}_{\varepsilon}$ is a homeomorphism onto its image.
3.2. Linearization. Let us define $q$ to be the natural quotient $G / \Gamma \cap \pm U \rightarrow G / \pm U$ and $\phi: \mathrm{G} / \pm \mathrm{U} \rightarrow \mathbb{R}^{2} / \pm 1$ by $\phi(g):=g . e_{1} / \pm 1$. For notational convenience, we will be working with $\mathbb{R}^{2}$ rather than $\mathbb{R}^{2} / \pm 1$. Here is a diagram.


The $\widetilde{\operatorname{cusp}}_{i_{0}}$ is already $q$-saturated: $q^{-1} q\left(\widetilde{\operatorname{cusp}}_{i_{0}}\right)=\widetilde{\operatorname{cusp}}_{i_{0}}$. More concretely,

$$
\phi \circ q\left(\widetilde{\operatorname{cusp}}_{i_{0}}\right)=\left\{v_{\neq 0} \in \mathbb{R}^{2} \mid\|v\|<e^{T_{i}}\right\} / \pm 1
$$



$\widetilde{\Omega}_{\varepsilon}$ may not be $q$-saturated. However, its image under $\phi \circ q$ is an open neighborhood of

$$
\phi \circ q\left(\widetilde{\mathscr{T}}_{t_{1}^{\prime}, t_{2}^{\prime}, i_{0}}\right)=\left(e^{t_{1}^{\prime}}, e^{t_{2}^{\prime}}\right) \times\{0\} / \pm 1 .
$$

Then one can show that there exists a smaller open nbhd $\Omega^{\prime}$ of $q\left(\widetilde{\mathscr{T}}_{t_{1}^{\prime}, t_{2}^{\prime}, i_{0}}\right)$ such that its preimage under $q$ is contained in $\widetilde{\Omega}_{\varepsilon}$. Thus we can choose $\delta=\delta(\varepsilon)>0$ small enough such that

$$
\widetilde{\Omega}_{\varepsilon}^{\prime}:=(\phi \circ q)^{-1}\left(\left(e^{t_{1}^{\prime}}, e^{t_{2}^{\prime}}\right) \times(-\delta, \delta)\right) / \pm 1 .
$$

is contained in $\widetilde{\Omega}_{\varepsilon}$. To combine $\widetilde{\operatorname{cusp}}_{i_{0}}$ with $\widetilde{\Omega}_{\varepsilon}^{\prime}$, choose a even smaller $\delta$ such that

$$
\widetilde{\mathcal{N}_{\varepsilon}^{\prime}}:=(\phi \circ q)^{-1}\left(\operatorname{Box}\left(e^{t_{2}^{\prime}}, \delta\right)\right) / \pm 1
$$

is contained in $\widetilde{\operatorname{cusp}}_{i_{0}} \cup \widetilde{\Omega}_{\varepsilon}^{\prime}$ where $\operatorname{Box}\left(e^{t_{2}^{\prime}}, \delta\right)=\left[-e^{t_{2}^{\prime}}, e^{t_{2}^{\prime}}\right] \times[-\delta, \delta]$. So $p$ restricted to $\widetilde{\mathcal{N}_{\varepsilon}^{\prime}}$ is injective.


Also let

$$
\widetilde{\mathscr{N}_{\varepsilon}}:=(\phi \circ q)^{-1}\left(\operatorname{Box}\left(e^{t_{2}+1}, \delta\right)\right) / \pm 1 .
$$

Let $\mathscr{N}_{\varepsilon}^{\prime}:=p\left(\widetilde{\mathscr{N}_{\varepsilon}^{\prime}}\right)$ and $\mathscr{N}_{\varepsilon}:=p\left(\widetilde{\mathscr{N}_{\varepsilon}}\right)$. They are open neighborhoods of $\mathscr{T}_{t_{1}, t_{2}, i_{0}}$.
3.3. Some linear algebra. At this point, one can adapt the strategy of previous sections to prove Thm.3.3 and hence analogues of Thm.1.1 for other lattices. The constant $t_{2}^{\prime}=t_{2}^{\prime}(\varepsilon)$ is determined in this process.

## 4. Exercises

### 4.1. Equidistribution via mixing.

- $G=\mathrm{SL}_{2}(\mathbb{R}), U=\left\{\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right], s \in \mathbb{R}\right\}, A=\left\{\mathbf{a}_{t}=\left[\begin{array}{cc}e^{t} & 0 \\ 0 & e^{-t}\end{array}\right], t \in \mathbb{R}\right\} ;$
- $V=\left\{\mathbf{v}_{r}=\left[\begin{array}{cc}1 & s \\ 0 & 1\end{array}\right], r \in \mathbb{R}\right\}$;
- $\Gamma$ is a lattice in $G$, let $X:=G / \Gamma$ and $\widehat{\mathrm{m}}_{X}$ be the unique $G$-invariant probability measure on X;
- Fix a right invariant Riemannian metric on $G$. Use this metric to induce a distance function $d(\cdot, \cdot)$ on $G$, let $d_{X}\left(\left[g_{1}\right]_{\Gamma},\left[g_{2}\right]_{\Gamma}\right):=\inf _{\gamma_{1}, \gamma_{2} \in \Gamma} d\left(g_{1} \gamma_{1}, g_{2} \gamma_{2}\right)$;
- for every $\delta, s_{0}>0$, let

$$
\operatorname{Box}\left(\delta, s_{0}\right):=(-\delta, \delta) \times(-\delta, \delta) \times\left(0, s_{0}\right) ;
$$

- let $\operatorname{Leb}_{\delta, s_{0}}$ be the restriction of standard Lebesgue measure restricted to $\operatorname{Box}\left(\delta, s_{0}\right)$;
- by abuse of notation we also denote by $\operatorname{Leb}_{\delta, s_{0}}$ for its push-forward under the map $(r, t, s) \mapsto \mathbf{v}_{r} \cdot \mathbf{a}_{t} \cdot \mathbf{u}_{s} ;$
- for $x \in X$, let $\mathrm{Obt}_{x}$ denote the map $G \rightarrow X$ defined by $g \mapsto g$.x.

Exercise 4.1. Fix $x \in X, \delta, s_{0}>0$. Show that there exists a non-negative function $f \in$ $L^{\infty}\left(X, \mathrm{~m}_{X}\right)$ such that $\left(\mathrm{Obt}_{x}\right)_{*} \operatorname{Leb}_{\delta, s_{0}}=f \cdot \widehat{\mathrm{~m}}_{X}$.

Exercise 4.2. Show that for every $\varepsilon>0$, there exists $\delta>0$ such that for every $s_{0}>0, t>0$, $(r, u, s) \in \operatorname{Box}\left(\delta, s_{0}\right)$ and $x \in X$ we have

$$
d_{X}\left(\mathbf{a}_{t} \cdot\left(\mathbf{v}_{r} \mathbf{a}_{u}\right) \cdot \mathbf{u}_{s} . x, \mathbf{a}_{t} \mathbf{u}_{s} . x\right)<\varepsilon
$$

Recall that mixing implies that for $\phi, \psi \in L^{2}\left(X, \widehat{\mathrm{~m}}_{X}\right)$,

$$
\lim _{t \rightarrow \pm \infty} \int \phi\left(\mathbf{a}_{t} \cdot x\right) \psi(x) \widehat{\mathrm{m}}_{X}(x)=\int \phi(x) \widehat{\mathrm{m}}_{X}(x) \cdot \int \psi(x) \widehat{\mathrm{m}}_{X}(x)
$$

Exercise 4.3. For every $s_{0}>0, x_{0} \in X$ and $f \in C_{c}(X)$, we have

$$
\lim _{t \rightarrow+\infty} \frac{1}{s_{0}} \int_{0}^{s_{0}} f\left(\mathbf{a}_{t} \mathbf{u}_{s} \cdot x_{0}\right) \mathrm{ds}=\int f(x) \widehat{\mathrm{m}}_{X}(x)
$$

Exercise 4.4. Show that if $\left(U . x_{n}\right)$ is a sequence of compact $U$-orbits of periods $S_{n} \rightarrow+\infty$, then for every compactly supported continuous function $f$,

$$
\lim _{n \rightarrow+\infty} \frac{1}{S_{n}} \int_{0}^{S_{n}} f\left(\mathbf{u}_{s} \cdot x_{n}\right) \mathrm{ds}=\int f(x) \widehat{\mathrm{m}}_{X}(x) .
$$

EXERCISE 4.5. Show that the above convergence (in Exer.4.3) is "uniform" in the following sense. For every $f \in C_{c}(X), \varepsilon, s_{0}>0$ and $x_{0} \in X$, there exists $\delta>0$ such that for every $y \in X$ with $d_{X}\left(x_{0}, y\right)<\delta$, we have for all $t>0$,

$$
\left|\frac{1}{s_{0}} \int_{0}^{s_{0}} f\left(\mathbf{a}_{t} \mathbf{u}_{s} \cdot x_{0}\right) \mathrm{ds}-\frac{1}{s_{0}} \int_{0}^{s_{0}} f\left(\mathbf{a}_{t} \mathbf{u}_{s} . y\right) \mathrm{ds}\right|<\varepsilon .
$$

EXERCISE 4.6. Use the above exercise to give another proof of the equidistribution of horocycle flows. Show that if $U . x_{0}$ is not compact in $X$, then for every $f \in C_{c}(X)$,

$$
\lim _{S \rightarrow+\infty} \frac{1}{S} \int_{0}^{S} f\left(\mathbf{u}_{s} \cdot x_{0}\right) \mathrm{ds}=\int f(x) \widehat{\mathrm{m}}_{X}(x)
$$

## CHAPTER 11

## Ergodic decomposition of unipotent invariant measures

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The main reference of this chapter is [Sha91b] and [MS95, Section 2].
Notations

- $\mathrm{X}:=\mathrm{G} / \Gamma$ with $\mathrm{G}:=\mathrm{SL}_{n}(\mathbb{R})$ and $\Gamma:=\mathrm{SL}_{n}(\mathbb{Z})$;
- $U=\left\{\mathbf{u}_{s}, s \in \mathbb{R}\right\}$ is a one-parameter unipotent subgroup of $G$.

Definition 0.1. A subgroup $U$ of $\mathrm{SL}_{n}(\mathbb{R})$ is said to be a one-parameter unipotent subgroup iff there exists a nilpotent matrix $u \in \mathfrak{s l}_{n}(\mathbb{R})$ such that $U=\{\exp (t . u), t \in \mathbb{R}\}$.

In this and the following few chapters, we are going to assume the measure classification theorem of unipotent flows (Thm.1.1 below) and demonstrate how it is applied. Further discussion of its proof is delayed to a later chapter.

## 1. Ergodic U-invariant measures

The following is the description of ergodic U-invariant probability measures due to Ratner [Ra91a].

Theorem 1.1. Let $\mu$ be an ergodic U -invariant probability measure on X , then there exists $x \in \mathrm{X}$ and a closed connected subgroup $H \leq \mathrm{G}$ containing U such that

1. H.x is closed and supports an H-invariant probability measure $\widehat{\mathrm{m}}_{H . x}$;
2. $\mu=\widehat{\mathrm{m}}_{H . x}$.

In short, one says that ergodic U-invariant probability measures are homogeneous (the word "algebraic" is also used). By writing $x=[g]_{\Gamma}$ and replacing $H$ by $g^{-1} H g$, the theorem may be rephrased as

Theorem 1.2. Let $\mu$ be an ergodic U -invariant probability measure on X , then there exists $g \in \mathrm{G}$ and a closed connected subgroup $H \leq \mathrm{G}$ containing $g^{-1} \mathrm{U} g$ such that

1. $[H]_{\Gamma}:=H \Gamma / \Gamma$ is closed and supports an $H$-invariant probability measure $\widehat{\mathrm{m}}_{[H]_{\Gamma}}$;
2. $\mu=g_{*} \widehat{\mathrm{~m}}_{[H]_{\Gamma}}$.

In particular, $\operatorname{supp}(\mu)=g[H]_{\Gamma}$.
EXAMPLE 1.3. If $\mathrm{G}=\mathrm{SL}_{2}(\mathbb{R})$ and $\mathrm{U}:=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$, then candidates of $H$ are $\{\mathrm{G}$, conjugates of U$\}$. If further assume $\Gamma$ is cocompact, then G is the only candidate. For non cocompact lattices, there are finitely many candidates up to $\Gamma$-conjugacy.

We have discussed this example in depth in Ch.8,9,10.
EXAMPLE 1.4. If $\mathrm{G}=\mathrm{SL}_{2}(\mathbb{C})\left(\Gamma=\mathrm{SL}_{2}(\mathbb{Z}[i])\right)$ and $\mathrm{U}:=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$, then, up to conjugacy, candidates of $H$ are $\left\{\mathrm{U}, \mathrm{V}, \mathrm{SL}_{2}(\mathbb{R}), \mathrm{G}\right\}$, where $\mathrm{V}=\left\{\left.\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{C}\right\}$.

Where does the pair $(g, H)$ live?
Definition 1.5. For two subgroups $A, B$ of G , define

$$
N(A, B):=\left\{g \in \mathrm{G} \mid g A g^{-1} \supset B\right\} .
$$

In this terminology, $g$ as above belongs to $N(H, \mathrm{U})$. When $H=U, N(U, U)$ is just the normalizer of U in G . When $H=\mathrm{G}, N(\mathrm{G}, U)=\mathrm{G}$. How about $H$ ? Note that we do not want U to appear in the definition of this space.

Definition 1.6. Let $\mathscr{H}$ be the collection of subgroups $L$ of $G$ satisfying

1. L is a connected and closed subgroup;
2. $[L]_{\Gamma}$ is closed and supports an L-invariant probability measure $\widehat{\mathrm{m}}_{[L]_{\Gamma}}$;
3. some one-parameter unipotent subgroup of L acts ergodically on $\widehat{\mathrm{m}}_{[L]_{\Gamma}}$.

Thus $H$ as above belongs to $\mathscr{H}$.
From $\mu \in \operatorname{Prob}(\mathrm{X})^{U, E r g}$ we get a pair $(g, H)$. However, the pair $(g, H)$ is not unique, for

$$
\left.g_{*} \mathrm{~m}_{[H]_{\Gamma}}=(g h)_{*} \mathrm{~m}_{[H]_{\Gamma}}=(g h \gamma)_{*} \mathrm{~m}_{[\gamma-1}{ }^{-1} H \gamma\right]_{\Gamma}
$$

for every $h \in H$ and $\gamma \in \Gamma$. Thus ( $g h \gamma, \gamma^{-1} H \gamma$ ) and ( $g, H$ ) correspond to the same $\mu$. The lemma below says that is all.

Lemma 1.7. Let $H_{1}, H_{2}$ be two connected closed subgroups of G such that $H_{i} \Gamma(i=1,2)$ are both closed. Let $g_{1}, g_{2} \in G$. Then $g_{1} H_{1} \Gamma=g_{2} H_{2} \Gamma$ iff there exist $h_{2} \in H_{2}$ and $\gamma_{2} \in \Gamma$ such that

$$
g_{2} h_{2} \gamma_{2}=g_{1}, \gamma_{2}^{-1} H_{2} \gamma_{2}=H_{1} .
$$

Proof. It only suffices to prove the " $\Longrightarrow$ " direction. The other direction follows directly.
So assume $g_{1} H_{1} \Gamma=g_{2} H_{2} \Gamma$. Then

$$
H_{1} \Gamma=g_{3} H_{2} \Gamma, g_{3}:=g_{1}^{-1} g_{2}
$$

Thus id $\epsilon g_{3} H_{2} \Gamma$ and

$$
1=g_{3} h_{2} \gamma_{2}, \exists h_{2} \in H_{2}, \gamma_{2} \in \Gamma_{2} .
$$

This already implies that

$$
g_{2} h_{2} \gamma_{2}=g_{2} g_{3}^{-1}=g_{1} .
$$

Now we have

$$
H_{1} \Gamma=\gamma_{2}^{-1} h_{2}^{-1} H_{2} \Gamma=\gamma_{2}^{-1} H_{2} \gamma_{2} \Gamma .
$$

By inspecting a small neighborhood of $[\mathrm{id}]_{\Gamma}$ and use the fact that $\left[H_{1}\right]_{\Gamma}$ and $\left[\gamma_{2}^{-1} H_{2} \gamma_{2}\right]_{\Gamma}$ are both embedded submanifolds, we see that $H_{1}=\gamma_{2}^{-1} H_{2} \gamma_{2}$.

## 2. Candidates of homogeneous orbit closure

Take some $x_{0} \in \mathrm{X}$. Eventually, we would know that the closure of U. $x_{0}$ is homogeneous and the homogeneous measure is finite and U-ergodic. But this does not follow immediately from the Thm.1.1. Nevertheless, we can say something even without appealing to Thm.1.1.

Definition 2.1. Let $\mathscr{A}$ (depending on $x_{0} \in \mathrm{X}$ and U ) be the collection of subgroups $L$ of G satisfying

1. L is a connected closed subgroup of G containing U ;
2. L. $x_{0}$ is closed.

Lemma 2.2. The collection $\mathscr{A}$ has a smallest element. Indeed, if $L_{1}, L_{2} \in \mathscr{A}$ then $\left(L_{1} \cap L_{2}\right)^{\circ} \in$ $\mathscr{A}$.

Proof. First we remark that for a closed subgroup $L \leq \mathrm{G}$, if $L \Gamma$ is closed then $L^{\circ} \Gamma$ is also closed. Indeed, one shows that every orbit of $L^{\circ}$ on $L / L \cap \Gamma$, which is homeomorphic to $L \Gamma / \Gamma$, is open and hence closed. So it suffices to show that if $L_{3}:=L_{1} \cap L_{2}$, then $L_{3} \Gamma$ is closed. This follows from a similar reasoning. Indeed, every orbit of $L_{3}$ on $\left[L_{1}\right]_{\Gamma} \cap\left[L_{2}\right]_{\Gamma}$ is open and hence closed. To see why every orbit is open, one may take a local neighborhood.

As an exercise, fill in the missing details in the proof.
Take $g_{0} \in$ G such that $x_{0}=\left[g_{0}\right]_{\Gamma}$.
Theorem 2.3. Let $H:=H_{\mathscr{A}}$ be the smallest element of $\mathscr{A}$, then

1. the closed set $H . x_{0}$ supports a finite $H$-invariant measure $\mathrm{m}_{H . x_{0}}$;
2. the measure $\mathrm{m}_{H . x_{0}}$ is U -ergodic;
3. there exists a $\mathbb{Q}$-algebraic subgroup $\mathbf{H}^{\prime}$ of $\mathrm{SL}_{n}$ such that $g_{0}^{-1} H g_{0}=\mathbf{H}^{\prime}(\mathbb{R})^{\circ}$. Actually, $\mathbf{H}^{\prime}$ is the smallest $\mathbb{Q}$-algebraic subgroup containing $g_{0}^{-1} U g_{0}$. In particular, $H$ is algebraic.

The last statement can be skipped if you are allergic to algebraic groups.
Before the proof, note that there is a locally finite measure $\mathrm{m}_{H . x_{0}}$ that is only "quasiinvariant under $H^{\prime \prime}$ (for instance, the one induced from a right invariant Riemannian metric). A priori, it is not clear why it is H -invariant. But one can still talk about ergodicity and the associated unitary representation (with suitably twisted action). You may ignore this minor issue by pretending $\mathrm{m}_{H . x_{0}}$ to be $H$-invariant from the start.

Here is a sketch of proof.
Step 1. By Mautner's phenomenon (see [Moo80, Theorem 1.1] and some supplementary arguments in [Sha91b, Proposition 2.7]), there exists a closed normal subgroup $F \triangleleft H$ containing $U$ such that for every unitary representation of $H$, every $U$-fixed vector is $F$-fixed. We already "know" this if $H$ is semisimple by arguments in Ch.7. See exercises attached to Ch. 7 for an example beyong the semisimple case. Thus to show U-ergodicity, suffices to show $F$ ergodicity.

Step 2. Let $\Gamma_{H}$ be the stabilizer of $x_{0}$ in $H$. Explicitly, $\Gamma_{H}=H \cap g_{0} \Gamma g_{0}^{-1}$. Define

$$
F^{\prime}:=\overline{F \cdot \Gamma_{H}}
$$

Since $F$ is normal, $F^{\prime}$ is a closed subgroup of $H$. Since $F^{\prime}$ is right invariant under $\Gamma_{H}, F^{\prime} \Gamma_{H} / \Gamma_{H}$ is closed in $H / \Gamma_{H}$. Thus $F^{\prime} . x_{0}$ is closed. And $F^{\prime}$ contains $U$. By minimality of $H, F^{\prime}=H$.

Step 3. Now we show $F$-ergodicity of $\mathrm{m}_{H . x_{0}}$. Let $\Omega$ be an $F$-invariant measurable set of $H / \Gamma_{H}$. Assume $\mathrm{m}_{H . x_{0}}(\Omega)>0$, we need to show that its complement has zero measure. Since $F$ is normal, we see that the preimage $\widetilde{\Omega}$ of $\Omega$ in $H$ right invariant under the group $F \cdot \Gamma_{H}$. Let $\mathrm{m}_{H}$ be a right $H$-invariant locally finite measure on $H$. Then $\mu:=1_{\Omega} \cdot \mathrm{m}_{H}$ is right $F \cdot \Gamma_{H^{-}}$ invariant. Since $\mu$ is a locally finite measure, by continuity, the stabilizer of $\mu$ in $H$ (w.r.t. the action from the right) is a closed subgroup. Thus $\mu$ is right $F^{\prime}$-invariant, hence $H$-invariant. By the uniqueness of invariant measures, $\mu=\mathrm{m}_{H}$ (up to a scalar, which has to be 1 ). In particular, the complement of $\widetilde{\Omega}$ has zero measure. This implies that the complement of $\Omega$ also has zero measure.

Step 4. It remains to show that $\mathrm{m}_{H . x_{0}}$ is a finite measure. In fact every U-ergodic locally finite measure $v$ is finite. Let us see why. By pointwise ergodic theorem (see [Wal82, Theorem 1.14, Section 1.6]), for every $f \in L^{1}(v)$, for $v$-almost every X,

$$
f^{*}(x):=\lim _{S \rightarrow+\infty} \frac{1}{S} \int_{0}^{S} f\left(\mathbf{u}_{s} \cdot x\right) \text { ds exists. }
$$

Moreover $f^{*} \in L^{1}(v)$ and is U -invariant. By ergodicity, $f^{*}$ is a constant, which has to be 0 if $v$ is an infinite measure.

On the other hand, by non-divergence of unipotent flow, there exists a compact set $C$ such that if $f$ is the indicator function of $C$, then $f^{*} \neq 0$.

Thus $v$ has to be finite. This finishes the proof of 1 and 2 of Thm.2.3.
Step 5. To save notation, we assume $g_{0}=\mathrm{id}$ here.
Let $\mathbf{L}$ be the smallest $\mathbb{Q}$-algebraic subgroup of $\mathrm{SL}_{n}$ containing $U$. Let $\pi_{1}: \mathbf{L} \rightarrow \mathbf{T}$ be the maximal quotient (algebraic) torus of $\mathbf{L}$. $\pi_{1}$ is defined over $\mathbb{Q}$. Since $U$ is unipotent and $\pi_{1}$ preserves this property, the image of $\pi_{1}(U)$ consists of unipotent elements. But torus $\mathbf{T}$ only contains semisimple elements. Thus $U$ is contained in the kernel of $\pi_{1}$, which is in the form of a semisimple (algebraic) group semidirect product with a unipotent (algebraic) group. In particular, $\mathbf{L}$ admits no nontrivial characters (: =algebraic group morphisms to $\mathbb{C}^{\times}=\mathrm{GL}_{1}$ ). By a theorem of Borel-Harish-Chandra (see for instance [Bor19, Corollary 13.2]), $\mathrm{L} \cap \Gamma\left(\mathrm{L}:=\mathbf{L}(\mathbb{R})^{\circ}\right)$ is a lattice in $L$ and in particular, $\mathrm{L} / \Gamma$ is closed. See exercises attached to Chapter 8 . By minimality of $H, H \subset \mathrm{~L}$. Our goal is to show $H=\mathrm{L}$ (this is what we mean by saying $H$ is "algebraic"). Note that the Zariski closure of $H$ is equal to $\mathbf{L}$.

Step 6. Let $\mathfrak{h}$ be the Lie algebra of $H$. By Levi's decomposition (reference? probably Bourbaki's book?), there exists a semisimple sub Lie algebra $\mathfrak{m}$ and a solvable ideal $\mathfrak{r}$ of $\mathfrak{h}$ such that $\mathfrak{h}=\mathfrak{m} \ltimes \mathfrak{r}$. By [Bor91, ChII, Corollary 7.9], since $\mathfrak{m}=[\mathfrak{m}, \mathfrak{m}], \mathfrak{m}$ is already algebraic (i.e., the corresponding Lie subgroup $M$ is algebraic). Let $\mathbf{M}$ be the corresponding $\mathbb{R}$-algebraic subgroup. We seek to show that $\mathfrak{r}$ consists of nilpotent matrices and hence is algebraic. By [Bor91, ChII, Corollary 7.7], this shows that $H$ is algebraic. Since $H$ normalizes $\mathfrak{r}$ and "normalizing $\mathfrak{r}$ " is an algebraic condition, we have that $\mathbf{L}$ normalizes $\mathfrak{r}$. Thus $\mathfrak{r}$ is an ideal of $\mathfrak{l}$. Let $\pi_{2}: \mathfrak{l} \rightarrow \mathfrak{l} / \mathfrak{u} \cong \mathfrak{m}$ (here $\mathfrak{u}$ is the Lie algebra of the unipotent radical of $\mathbf{L}$ ), then $\pi_{2}(\mathfrak{r})$ is an ideal of $\mathfrak{l} / \mathfrak{u}$. But every non-zero ideal of a semisimple Lie algebra is semisimple and can not be solvable. Thus $\pi_{2}(\mathfrak{r})=0$, or $\mathfrak{r} \subset \mathfrak{u}$, which consists of nilpotent matrices. The rest of the claim in 3. of Thm.2.3 now follows from Borel density theorem.

## 3. Tubes and ergodic components

Assume U acts on $g_{*} \mathrm{~m}_{[H]}$ ergodically, it is still possible for some $h \in H, \mathrm{U}[g h]_{\Gamma}$ is trapped in a closed homogeneous set of smaller dimension.

Definition 3.1. For $H \in \mathscr{H}$, define

$$
\begin{aligned}
& \operatorname{Sing}(H, \mathrm{U}):=\bigcup_{L \in \mathscr{H}, L \nsupseteq H} N(L, \mathrm{U}) ; \\
& \mathrm{NS}(H, \mathrm{U}):=N(H, \mathrm{U}) \backslash \operatorname{Sing}(H, \mathrm{U}) ; \\
& T(H, \mathrm{U}):=\mathrm{NS}(H, \mathrm{U}) \Gamma / \Gamma .
\end{aligned}
$$

Lemma 3.2. Let $H_{1}, H_{2} \in \mathscr{H}$. If $\mathrm{NS}\left(H_{1}, \mathrm{U}\right) \Gamma \cap \mathrm{NS}\left(H_{2}, \mathrm{U}\right) \Gamma \neq \varnothing$, then $H_{1}$ is $\Gamma$-conjugate to $H_{2}$ and $\mathrm{NS}\left(H_{1}, \mathrm{U}\right) \Gamma=\mathrm{NS}\left(H_{2}, \mathrm{U}\right) \Gamma$.

Proof. So assume $\operatorname{NS}\left(H_{1}, \mathrm{U}\right) \Gamma \cap \mathrm{NS}\left(H_{2}, \mathrm{U}\right) \Gamma \neq \varnothing$, which means that there exist $g_{1} \in \mathrm{NS}\left(H_{1}, \mathrm{U}\right)$ and $\gamma_{1} \in \Gamma$ such that $g_{1} \gamma_{1} \in \operatorname{NS}\left(H_{2}, \mathrm{U}\right)$. By definition, we have

$$
g_{1}^{-1} U g_{1} \subset H_{1} \cap \gamma_{1} H_{2} \gamma_{1}^{-1}
$$

We know (the connected component of) $H^{\prime}:=H_{1} \cap \gamma_{1} H_{2} \gamma_{2}^{-1}$ has a closed orbit based at [id] ${ }_{\Gamma}$. But we do not know whether it supports a finite $H^{\prime}$-invariant measure. This is where we apply Thm.2.3 (to the unipotent group $g_{1}^{-1} U g_{1}$ and $x_{0}=[i d]_{\Gamma}$ ) to conclude that there exists $L \subset H^{\prime}$, $L \in \mathscr{H}$ such that $g_{1}^{-1} U g_{1} \subset L$. So $g_{1} \in N(L, U)$.

Therefore $H_{1}=\gamma_{1} H_{2} \gamma_{1}^{-1}$ for otherwise $L$ will be strictly contained in at least one of $H_{1}$ or $\gamma_{1} H_{2} \gamma_{1}^{-1}$ and this would imply $g_{1} \notin \mathrm{NS}\left(H_{1}, \mathrm{U}\right)$ or $g_{1} \gamma_{1} \notin \mathrm{NS}\left(H_{2}, \mathrm{U}\right)$, contradicting against our assumption. $\mathrm{NS}\left(H_{1}, \mathrm{U}\right) \Gamma=\mathrm{NS}\left(H_{2}, \mathrm{U}\right) \Gamma$ follows immediately.

Since U acts ergodically on $\mathrm{m}_{\mathrm{X}}$ (it is even mixing!), we have $\mathrm{G} \in \mathscr{H}$ and

$$
\mathrm{X}=\bigsqcup_{[H] \in \mathscr{H} / \sim \mathrm{T}} T(H, \mathrm{U})
$$

thanks to the Lem.3.2.
Definition 3.3. For $[H] \in \mathscr{H} / \sim_{\Gamma}$, let

$$
\mu^{[H]}:=\left.\mu\right|_{T(H, U)} .
$$

EXAMPLE 3.4. If $\mathrm{X}=\mathrm{SL}_{2}(\mathbb{R}) / \mathrm{SL}_{2}(\mathbb{Z})$ and $\mathrm{U}=\left\{\left.\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$, then $\mathscr{H} / \sim_{\Gamma}=\left\{\mathrm{U}, \mathrm{SL}_{2}(\mathbb{R})\right\}$ (if you pass to a smaller subgroup of $\mathrm{SL}_{2}(\mathbb{Z})$ then this set has other $\mathrm{SL}_{2}(\mathbb{Q})$-conjugates of U that are not conjugate over $\Gamma)$. Here $T(\mathrm{U}, \mathrm{U})$ consists of compact orbits of $\mathrm{U}, T(\mathrm{G}, \mathrm{U})$ is the complement of $T(\mathrm{U}, \mathrm{U})$.

See Sec. 4 for more examples.
Assuming Lem.3.6, we have proved
Theorem 3.5. For a U-invariant probability measure $\mu$,

$$
\mu=\sum_{[H] \in \mathscr{H} / \sim \Gamma} \mu^{[H]}
$$

and each $\mu^{[H]}$ is U -invariant.
Lemma 3.6. $\mathscr{H}$ is countable.
Proof. For every $H \in \mathscr{H}, H \cap \Gamma$ is a lattice in $H$. Thus $H \cap \Gamma$ is finitely generated. Note that this seems not obvious unless $H \cap \Gamma$ is cocompact. In the case at hand, $H$ is algebraic by Thm.2.3. Thus $H \cap \Gamma$ is an arithmetic lattice of $H$ and finite generation follows from the theory of Siegel sets, see [Bor19]. See also [Gel14, Lecture 3, Section 5] for another possibly more geometric proof. Hence the set $\{H \cap \Gamma, H \in \mathscr{H}\}$ is countable.

Since $H$ can be recovered from $H \cap \Gamma$ by

$$
H=\left(\overline{\overline{H \cap \Gamma}} \cap \mathrm{SL}_{n}(\mathbb{R})\right)^{\circ},
$$

we are done. Here $\overline{\overline{H \cap \Gamma}}$ means the closure of $H \cap \Gamma$ in $\mathrm{SL}_{n}(\mathbb{C})$ with respect to the topology defined by polynomials.

We have not used Thm.1.1 yet. For a finite positive measure $\mu$ on X , let $\widehat{\mu}:=\mu / \mu(X)$ be the unique probability measure proportional to $\mu$.

Theorem 3.7. Assume $\mu^{[H]} \neq 0$. For almost every U-ergodic component $v$ of $\widehat{\mu^{[H]}}$, there exists $g_{v} \in N(H, \mathrm{U})$ such that $v=\left(g_{v}\right)_{*} \mathrm{~m}_{[H]_{\Gamma}}$.

Proof. First we have the (abstract) ergodic decomposition

$$
\widehat{\mu^{[H]}}=\int_{\operatorname{Prob}(X)^{\mathrm{U}, \mathrm{Erg}}} v \lambda(v) .
$$

Thus for almost every $v, v(T(H, \mathrm{U}))=1$. Take such a $v$, by Thm.1.2, there exists $H_{1} \in \mathscr{H}$ and $g_{1} \in N\left(H_{1}, \mathrm{U}\right)$ such that $v=\left(g_{1}\right)_{*} \widehat{\mathrm{~m}}_{\left[H_{1}\right] \Gamma}$. By pointwise ergodic theorem, we can find a full measure set of $h_{1} \in H_{1}$ such that

$$
\lim _{S \rightarrow+\infty} \frac{1}{S} \int_{0}^{S}\left(\mathbf{u}_{s}\right)_{*} \delta_{\left[g_{1} h_{1}\right] \Gamma} \mathrm{ds}=\left(g_{1}\right)_{*} \widehat{\mathrm{~m}}_{\left[H_{1}\right] \Gamma} .
$$

In particular, $\overline{\mathrm{U} .\left[g_{1} h_{1}\right]_{\Gamma}}=g_{1}\left[H_{1}\right]_{\Gamma}$. One sees that $g_{1} h_{1} \in N\left(H_{1}, \mathrm{U}\right)$ and we claim that $g_{1} h_{1} \in$ $\mathrm{NS}\left(H_{1}, \mathrm{U}\right)$. Otherwise, there exists $L \supsetneqq H_{1}$ with $L \in \mathscr{H}$ such that $g_{1} h_{1} \in N(L, \mathrm{U})$. This implies that $\overline{\mathrm{U} .\left[g_{1} h_{1}\right]_{\Gamma}} \subset g_{1} h_{1}[L]_{\Gamma}$. Since $\operatorname{dim} L$ is strictly smaller than $\operatorname{dim} H_{1}$, we have a contradiction.

So now $\left[g_{1} h_{1}\right]_{\Gamma} \in T\left(H_{1}, \mathrm{U}\right)$, moreover, all such $h_{1}$ 's are of full measure in $H_{1}$ and consequently $v\left(T\left(H_{1}, \mathrm{U}\right)\right)=\left(g_{1}\right)_{*} \widehat{\mathrm{~m}}_{\left[H_{1}\right]_{\Gamma}}\left(T\left(H_{1}, \mathrm{U}\right)\right)=1$. But $v(T(H, \mathrm{U}))=1$. Thus $T\left(H_{1}, \mathrm{U}\right)$ and $T(H, \mathrm{U})$ have nontrivial intersection. By Lem.3.2, for some $\gamma_{1} \in \Gamma, H_{1}=\gamma_{1} H \gamma_{1}^{-1}$ and $T\left(H_{1}, \mathrm{U}\right)=$ $T(H, \mathrm{U})$. Hence $\left[g_{1} H_{1}\right]_{\Gamma}=g_{1} \gamma_{1}[H]_{\Gamma}$. Let $g_{v}:=g_{1} \gamma_{1}$. One can check that $g_{v} \in N(H, \mathrm{U})$ and $v=\left(g_{v}\right)_{*} \widehat{\mathrm{~m}}_{[H]_{\mathrm{\Gamma}}}$

## 4. Two examples

Here we include two examples, a little bit beyond $\mathrm{SL}_{2}(\mathbb{R})$, to illustrate what kind of objects we are dealing with. You are welcome to test the general theory using these (still rather special) examples!

In both examples, set

- $G=\mathrm{SL}_{2}(\mathbb{C}), \Gamma=\mathrm{SL}_{2}(\mathbb{Z}[i]), U=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\} ;$
- $\mathbf{U}(\mathbb{C}):=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{C}\right\}$
- for $t \in \mathbb{C}^{\times}$, let $\mathbf{a}_{t}:=\left[\begin{array}{cc}t & 0 \\ 0 & t^{-1}\end{array}\right]$.

One can show that $\Gamma$ is a lattice in $G$ (using non-divergence of unipotent flows, for instance). And $G / \Gamma$ can be embedded in $\mathrm{SL}_{4}(\mathbb{R}) / \mathrm{SL}_{4}(\mathbb{Z})$ (so this example does not escape away from the setting in this chapter).

### 4.1. Example 1.

- $H:=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{C}\right\}, \mathfrak{h}$ is the Lie algebra of $H$;
- $K_{H}:=\left\{\mathbf{a}_{z}, z \in \mathbb{C},|z|=1\right\}$.

Lemma 4.1. $N(H, U)=N_{G}(H)=\left\{\mathbf{a}_{t} \cdot \mathbf{u}_{s}, t \in \mathbb{C}^{\times}, s \in \mathbb{C}\right\}=: B$.
Proof. Let $g \in G$. Indeed, $g$ belongs to $N(H, U)$ iff $\operatorname{Ad}(g) \cdot \mathfrak{h}$ contains $\mathfrak{u}$. By Bruhat decomposition (ref??),

$$
G=B w B \sqcup B
$$

where $w=\left[\begin{array}{cc}0 & 1 \\ -1 & 0\end{array}\right]$. If $g \in B$, then $\operatorname{Ad}(g) \cdot \mathfrak{h}=\mathfrak{h} \supset \mathfrak{u}$. On the other hand, if $g=b_{1} w b_{2}$ for $b_{i} \in B$ then

$$
\operatorname{Ad}(g) \cdot \mathfrak{h}=\operatorname{Ad}\left(b_{1}\right) \operatorname{Ad}(w) \cdot \mathfrak{h}=\operatorname{Ad}\left(b_{1}\right) \cdot\left[\begin{array}{cc}
0 & 0 \\
* & 0
\end{array}\right] \Longrightarrow \operatorname{Ad}(g) \cdot \mathfrak{h} \cap \mathfrak{h}=\{0\} .
$$

So we are done.
The orbits of $B$ on $G / \Gamma$ are all dense, and hence not easy to draw. Since $N(H, U)(=B$ here) is stable under right translation by $N_{G}(H)$ and therefore $N_{G}(H) \cap \Gamma$ (call it $\Gamma_{N}$ for simplicity). Thus $N(H, U)$ being closed implies that $N(H, U) / \Gamma_{N} \subset G / \Gamma_{N}$ is closed. We will draw pictures for $N(H, U) / \Gamma_{N}$. (warning! pictures are just for illustration, they may be wrong in many aspects!)

By the way, a quick computations show that

$$
\Gamma_{N}=\left\{\left[\begin{array}{cc}
1 & \mathbb{Z}[i] \\
0 & 1
\end{array}\right],\left[\begin{array}{cc}
-1 & \mathbb{Z}[i] \\
0 & -1
\end{array}\right],\left[\begin{array}{cc}
i & \mathbb{Z}[i] \\
0 & -i
\end{array}\right]\right\} .
$$

Here is a picture for $N(H, U) / \Gamma_{N}$ with $U \Gamma_{N} / \Gamma_{N}$ contained in here:


What about $\operatorname{Sing}(H, U)$ ? The possible $L \in \mathscr{H}$ and $L \subsetneq H$ are given as follows. For $z \in \mathbb{C}$, let $U^{z}:=\left\{\left.\mathbf{u}_{s \cdot z}=\left[\begin{array}{cc}1 & s z \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$. Then every proper nontrivial connected subgroup of $H$ is of this form. And $U^{z} \in \mathscr{H}$ iff $U^{z} \cap \Gamma$ is a lattice in $U^{z}$ ff $\mathbb{R} . z \cap \mathbb{Z}[i] \leq \mathbb{R} . z$ is a lattice.

Note that

$$
N\left(U^{z}, U\right)=\left[\begin{array}{cc}
\sqrt{z}^{-1} & 0 \\
0 & \sqrt{z}
\end{array}\right] \cdot N_{G}\left(U^{z}\right)=\left[\begin{array}{cc}
\sqrt{z}^{-1} & 0 \\
0 & \sqrt{z}
\end{array}\right] \cdot\left\{\mathbf{a}_{t}, t \in \mathbb{R}^{\times}\right\} \cdot \mathbf{U}(\mathbb{C})
$$

And $\operatorname{Sing}(H, U)$ is the union of these $N\left(U^{z}, U\right)$ as $z$ varies over $\mathbb{Z}[i]$.


### 4.2. Example 2.

- $H:=\mathrm{SL}_{2}(\mathbb{R}), U^{i}:=\left\{\mathbf{u}_{i s}, s \in \mathbb{R}\right\}$.

LEMMA 4.2. $N(H, U)=U^{i} \cdot \mathrm{SL}_{2}(\mathbb{R}) \sqcup U^{i} \cdot \mathrm{SL}_{2}(\mathbb{R}) \cdot\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]$.
Note that $U^{i} \cdot \mathrm{SL}_{2}(\mathbb{R}) \cdot\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]=\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right] \cdot U^{i} \cdot \mathrm{SL}_{2}(\mathbb{R})=\mathbf{U}(\mathbb{C}) \cdot \mathrm{SL}_{2}(\mathbb{R}) \cdot\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]$.

Proof. As before, $N(H, U)=\{g \in G, \operatorname{Ad}(g) \cdot \mathfrak{h} \supset \mathfrak{u}\}$. It is direct to observe that RHS is a subset of LHS. It remains to do the converse.

Recall Bruhat decomposition again: $G=B w B \sqcup B$. If $g \in B$, then we are done since $B$ is contained in the right hand side.

Now assume $g \in B w B$. Every element $b$ of $B$ can be written as $\mathbf{a}_{t} \mathbf{u}_{s}$ with $t \in \mathbb{C}^{\times}, s \in \mathbb{C}$. Since $w$ normalizes $\left\{\mathbf{a}_{t}, t \in \mathbb{C}^{\times}\right\}$, we can write

$$
g^{-1}=u_{2} \mathbf{a}_{t_{1}} w u_{1}, \quad \exists u_{1}, u_{2} \in \mathbf{U}(\mathbb{C}), t_{1} \in \mathbb{C}^{\times}
$$

Thus (to save notation we omit Ad in the following)

$$
\begin{aligned}
g^{-1} \cdot \mathfrak{u} & =\left(u_{2} \mathbf{a}_{t_{1}} w u_{1}\right) \cdot \mathfrak{u}=\left(u_{2} \mathbf{a}_{t_{1}}\right) \cdot\left[\begin{array}{cc}
0 & 0 \\
\mathbb{R} & 0
\end{array}\right] \\
& =u_{2} \cdot\left[\begin{array}{cc}
0 & 0 \\
t_{1}^{-2} \mathbb{R} & 0
\end{array}\right]=\left[\begin{array}{cc}
* & * \\
t_{1}^{-2} \mathbb{R} & *
\end{array}\right] \subset \mathfrak{h}=\mathfrak{s l}_{2}(\mathbb{R})
\end{aligned}
$$

Thus $t_{1}^{-2} \mathbb{R} \subset \mathbb{R} \Longrightarrow t_{1} \in \mathbb{R} \cup i \mathbb{R}$. In either case (write $u_{1}=\mathbf{u}_{z_{1}}$ for some $z_{1} \in \mathbb{C}$ ),

$$
g^{-1} \cdot \mathfrak{u}=u_{2} \cdot\left[\begin{array}{cc}
0 & 0 \\
\mathbb{R} & 0
\end{array}\right]=\left[\begin{array}{cc}
1 & z_{1} \\
0 & 1
\end{array}\right] \cdot\left[\begin{array}{cc}
0 & 0 \\
\mathbb{R} & 0
\end{array}\right] \cdot\left[\begin{array}{cc}
1 & -z_{1} \\
0 & 1
\end{array}\right]=\mathbb{R} \cdot\left[\begin{array}{cc}
z_{1} & -z_{1}^{2} \\
1 & -z_{1}
\end{array}\right] \subset \mathfrak{s l}_{2}(\mathbb{R})
$$

Thus $z_{1} \in \mathbb{R}$. And the proof completes.
The above proof also shows that
Lemma 4.3. $N_{G}(H)=\mathrm{SL}_{2}(\mathbb{R}) \sqcup \mathrm{SL}_{2}(\mathbb{R}) \cdot\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]$.
And hence one can check that
LEMMA 4.4. $\Gamma_{N}:=N_{G}(H) \cap \Gamma=\mathrm{SL}_{2}(\mathbb{Z}) \sqcup \mathrm{SL}_{2}(\mathbb{Z}) \cdot\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]$.
Below is a picture of $N(H, U) / \Gamma_{N}$ sitting inside $G / \Gamma_{N}$ as a closed subset. Note that its projection to $G / \Gamma$ is dense (you probably saw this in some exercise section).


Up to $\Gamma_{N}$-conjugacy, the only proper nontrivial connected subgroup of $H$ containing $U$ is just $U$ itself. Thus $\operatorname{Sing}(H, U)=N(U, U) \Gamma_{N}$.

Lemma 4.5. $N(U, U)=N_{G}(U)$ and is generated by $\left\{\mathbf{a}_{t} \cdot \mathbf{u}_{s}, t \in \mathbb{R}^{\times}, s \in \mathbb{R}\right\} \cup\left[\begin{array}{cc}i & 0 \\ 0 & -i\end{array}\right]$.
Thus the picture is not so new. Note that since $U^{i}$ commutes with $U, U^{i}$ translates of $U \Gamma_{N} / \Gamma_{N}$ does not "twist" the appearance of $U \Gamma_{N} / \Gamma_{N}$ (unlike $U^{i}$ translates of $H \Gamma_{N} / \Gamma_{N}$ ).


Remark/ if you let $s, t \rightarrow+\infty$ at the same time, it is unclear to me
Question : (asymptotically) how $U_{i s} a_{t} U T_{N} / I_{N}$ look like relative to $U_{i s} H T_{N} / T_{N}$

## 5. Exercises

## CHAPTER 12

## Linearization technique

Back to the Top.
In this chapter we are going to discuss the linearization method due to Dani-Margulis [DM93]. We will illustrate the method by proving an equidistribution statement, which on the one hand implies Oppenheim conjecture, on the other hand forms one ingredient of quantitative Oppenheim. Same method can be used to deduce equidistribution of unipotent flows, and hence to classify orbit closures of orbits of unipotent flows from the measure classification theorem. This will be left as a difficult exercise.

The main reference of this chapter is [MS95, Section 3]. Other related resources include Shah [Sha91b, Sha91a, Sha09], Ratner [Ra91a, Ra91b], Eskin-Mozes-Shah [EMS96], Eskin-Margulis-Mozes [EMM98, Section 4]. An effective treatment appears in [LMMS19].

Recall the notations when we discuss Oppenheim conjecture.

- $\mathrm{G}=\mathrm{SL}_{3}(\mathbb{R}), \Gamma=\mathrm{SL}_{3}(\mathbb{Z}), \mathrm{X}:=\mathrm{G} / \Gamma$;
- $\widehat{\mathrm{m}}_{\mathrm{X}}$ is the unique G -invariant probability measure on X ;
- $\mathrm{H}_{0}:=\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R})$ with $Q_{0}\left(x_{1}, x_{2}, x_{3}\right)=2 x_{1} x_{3}-x_{2}^{2}$;
- $\mathrm{U}=\left\{\mathbf{u}_{s}: \left.=\exp \left(s \cdot\left[\begin{array}{ccc}0 & 1 & 0 \\ 0 & 0 & 1 \\ 0 & 0 & 0\end{array}\right]\right) \right\rvert\, s \in \mathbb{R}\right\} \subset \mathrm{H}_{0}$;
- $\mathrm{A}=\left\{\mathbf{a}_{t}: \left.=\exp \left(t \cdot\left[\begin{array}{ccc}1 & 0 & 0 \\ 0 & 0 & 0 \\ 0 & 0 & -1\end{array}\right]\right) \right\rvert\, t \in \mathbb{R}\right\} \subset \mathrm{H}_{0}$;
- also we fix some $g_{0} \in G$ and $x_{0}:=\left[g_{0}\right]_{\Gamma}:=g_{0} \Gamma / \Gamma$.


## 1. Statement

As we explained, Oppenheim conjecture follows once we prove
Theorem 1.1. If $Q:=Q_{0} \circ g_{0}$ is irrational, then $\mathrm{H}_{0} g_{0} \Gamma / \Gamma$ is dense in X .
Actually something weaker is proved in Ch.5, Thm.1.2, which is sufficient. Now we would like to explain how to use Ratner's description of ergodic U-invariant probability measures to prove this stronger claim.

The idea is as follows. Take

$$
K_{0}:=\left(\mathrm{H}_{0} \cap \mathrm{SO}_{3}(\mathbb{R})\right)^{\circ},
$$

a maximal compact subgroup of $\mathrm{H}_{0}^{\circ}$. Then $\mathbf{a}_{t} K_{0} \cdot x_{0} \subset \mathrm{H}_{0} \cdot x_{0}$ and we seek to show that as $t \rightarrow+\infty$ ( $-\infty$ is also ok), $\mathbf{a}_{t} K_{0} \cdot x_{0}$ becomes dense in X. And this is achieved by the following equidistribution theorem

THEOREM 1.2. Let $\widehat{\mathrm{m}}_{K_{0} . x_{0}}$ be the unique $K_{0}$-invariant probability measure on $K_{0} \cdot x_{0}$. Then

$$
\lim _{t \rightarrow+\infty}\left(\mathbf{a}_{t}\right)_{*} \widehat{\mathrm{~m}}_{K_{0} \cdot x_{0}}=\widehat{\mathrm{m}}_{\mathrm{X}}
$$

in weak* topology.

REMARK 1.3. From the proof, you will see that $\widehat{\mathrm{m}}_{K_{0} \cdot x_{0}}$ can be replaced by any other probability measure that is absolutely continuous with respect to this one without affecting the conclusion.

REMARK 1.4. Instead of $K_{0}$, you can also use other subgroups of $\mathrm{H}_{0}$ and prove analogues of the theorem above. Actually it would be easier if we replace $K_{0}$ by a bounded open subset of $\mathrm{H}_{0}$. However, I prefer to do this in preparation for our later discussion on quantitative Oppenheim. This change only has an effect on Sec.11.

## 2. Step 1, nondivergence

Let $\mu$ be a limit of $\left(\mu_{t}\right):=\left(\left(\mathbf{a}_{t}\right)_{*} \widehat{\mathrm{~m}}_{K_{0} . x_{0}}\right)$ as $t \rightarrow+\infty$.
Lemma 2.1. $\mu \in \operatorname{Prob}(\mathrm{X})$.
In other words, there is no escape of mass. This is a consequence of $(C, \alpha)$-good property and a lemma in representation theory/linear algebra.

## 3. Step 2, unipotent invariance

Lemma 3.1. $\mu$ is U -invariant.
Proof. Since $\widehat{\mathrm{m}}_{K_{0} . x_{0}}$ is $K_{0}$-invariant, $\mu_{t}$ is $\mathbf{a}_{t} K_{0} \mathbf{a}_{t}^{-1}$-invariant. Hence $\mu$ is invariant under the limit group, which turns out to be U.

More details: Let $\mathfrak{k}_{0}$ be the Lie algebra of $K_{0}$. Take $v_{t} \in \operatorname{Ad}\left(\mathbf{a}_{t}\right) \cdot \mathfrak{k}_{0}$, if $\lim v_{t}=v$, then by continuity of the induced map $\mathrm{G} \times \operatorname{LFM}(X) \rightarrow \operatorname{LFM}(X), \mu$ is $\exp (\nu)$-invariant.

Recall that the Lie algebra of $\mathrm{H}_{0}$ is

$$
\mathfrak{s o}_{Q_{0}}=\left\{\left[\begin{array}{ccc}
x_{11} & x_{12} & 0 \\
x_{21} & 0 & x_{12} \\
0 & x_{21} & -x_{11}
\end{array}\right]\right\} .
$$

And the Lie algebra of $\mathrm{SO}_{3}(\mathbb{R})$ is given by anti-symmetric matrices. Thus by taking their intersection:

$$
\mathfrak{k}_{0}=\left\{\left[\begin{array}{ccc}
0 & x_{12} & 0 \\
-x_{12} & 0 & x_{12} \\
0 & -x_{12} & 0
\end{array}\right]\right\} .
$$

And

$$
\operatorname{Ad}\left(\mathbf{a}_{t}\right) \mathfrak{k}_{0}=\left\{\left[\begin{array}{ccc}
0 & e^{t} x_{12} & 0 \\
-e^{-t} x_{12} & 0 & e^{t} x_{12} \\
0 & -e^{-t} x_{12} & 0
\end{array}\right]\right\} .
$$

So depending on $s \in \mathbb{R}$, we take

$$
v_{t}:=\left[\begin{array}{ccc}
0 & s & 0 \\
-e^{-2 t} s & 0 & s \\
0 & -e^{-2 t} s & 0
\end{array}\right] \in \operatorname{Ad}\left(\mathbf{a}_{t}\right) \mathfrak{k}_{0} .
$$

Then as $s$ varies, $\lim v_{t}$ fills $\mathfrak{u}$, the Lie algebra of $U$.
Thus by the first two steps we get (by passing to a subsequence)

$$
\lim \mu_{t}=\mu \in \operatorname{Prob}(\mathrm{X})^{\mathrm{U}} .
$$

## 4. Step 3, ergodic components and tubes

By Thm. 3.5 and 3.7 from Chapter 11, to show $\mu=\widehat{\mathrm{m}}_{\mathrm{X}}$, it suffices to show that for every $H \in \mathscr{H}, H \neq \mathrm{G}, \mu(T(H, \mathrm{U}))=0$. Note that Ratner's theorem Ch.11, Thm.1.1 is only used to go from $\mu(T(H, \mathrm{U}))=0$ to $\mu=\widehat{\mathrm{m}}_{\mathrm{x}}$. To show $\mu(T(H, \mathrm{U}))=0$, we do not need it. The way to achieve this is via:

Lemma 4.1. For every compact subset $E$ of $T(H, U)$ and $\varepsilon>0$, there exists a neighborhood $\mathscr{N}_{\varepsilon}$ of E such that

$$
\limsup _{t \rightarrow+\infty} \mu_{t}\left(\mathscr{N}_{\varepsilon}\right) \leq \varepsilon .
$$

In view of Ch. 10 (cf. [DS84]), we hope to find a bigger $\mathscr{N}_{\varepsilon}^{\prime}$ such that

$$
\mu_{t}\left(\mathscr{N}_{\varepsilon}\right) \leq \varepsilon \mu_{t}\left(\mathscr{N}_{\varepsilon}^{\prime}\right) .
$$

Since $\mu_{t}$ is a probability measure, this finishes the proof.

## 5. Step 4, a lemma on linear representations

Though we do not know how to find $\mathscr{N}_{\varepsilon} \subset \mathscr{N}_{\varepsilon}^{\prime}$ at the moment, we do have something like this happening in a representation (rather than the complicated $\mathrm{G} / \Gamma$ ) due to the ( $C, \alpha$ )-good property. To give us more freedom (see below, the choice of $\Phi$ ) for things to come, we need a slightly more flexible statement.

Definition 5.1. Fix a non-empty connected bounded open set $\mathrm{D} \subset \mathfrak{k}_{0}$, let

$$
\psi_{t}: \mathrm{D} \rightarrow \mathrm{G}, \quad x \mapsto \psi_{t}(x):=\mathbf{a}_{t} \exp (x) .
$$

Lemma 5.2. Let $V$ be a representation of $G$. Let $W$ be a linear subspace of $V$. For every compact subset $E$ of $W$ and every $\varepsilon>0$, there exists another compact set $F \subset W$ such that the following is true. For every open neighborhood $\Phi$ of $F$, there exists an open neighborhood $\Psi$ of $E$ such that for every $t \in \mathbb{R}, v \in V$, every ball $\mathrm{B} \subset \mathrm{D}$, at least one of the following is true

1. $\overline{\psi_{t}(\mathrm{~B}) \cdot v} \subset \Phi$;
2. $\operatorname{Leb}\left\{x \in \mathrm{~B} \mid \psi_{t}(x) . v \in \Psi\right\} \leq \varepsilon \operatorname{Leb}\left\{x \in \mathrm{~B} \mid \psi_{t}(x) . v \in \Phi\right\}$.

Remark 5.3. The first possibility can often be excluded due to "algebraic" reasons (for instance, see Sec.11). And the second option is what we want.

Remark 5.4. You can replace Leb by any other measure equivalent to Leb, it is just that the choice of $\Psi$ may depend on this measure. Actually in application we have in mind, Leb should be replaced by some measure which maps to $\widehat{\mathrm{m}}_{K_{0} x_{0}}$ under the exponential and the orbit map. We are going to ignore this issue in the following.

## 6. Step 5, representation and dynamics, naive ideas

Let $\Gamma_{N}:=\Gamma \cap N_{\mathrm{G}}(H)$.
Definition 6.1. Let

$$
N_{\mathrm{G}}(H)^{(1)}:=\left\{g \in N_{\mathrm{G}}(H) \mid \operatorname{det}(\operatorname{Ad}(g), \mathfrak{h})= \pm 1\right\}
$$

Lemma 6.2. $\Gamma_{N}=\Gamma \cap N_{\mathrm{G}}(H)^{(1)}$.
Take a representation $V_{H}$ of G and a vector $\nu_{H} \in V_{H}$ such that the stabilizer of $\nu_{H}$ (or just $\pm v_{H}$ ) in G is equal to $N_{\mathrm{G}}(H)^{(1)}$. Moreover, we want $V_{H}$ to be equipped with a $\mathbb{Q}$-structure (i.e., fix a copy of $\mathbb{Q}^{\operatorname{dim} V_{H}}$ in $V_{H}$, call it $V_{H}(\mathbb{Q})$ ) and $v_{H} \in V_{H}(\mathbb{Q})$. A priori, $N_{\mathrm{G}}(H)^{(1)}$ is not known to be "observable" , the existence of such a pair ( $V_{H}, \nu_{H}$ ) is not obvious. But one can take $V_{H}:=\wedge^{\operatorname{dim} H_{s l}}{ }_{n}$ and $\nu_{H}:=\nu_{1} \wedge \ldots \wedge v_{\operatorname{dim} H}$ where $\left(v_{1}, \ldots, v_{\operatorname{dim} H}\right)$ is a basis of $\mathfrak{h}$. For this specific
choice of $\nu_{H}$, the stabilizer of $\pm v_{H}$ in G is equal to $N_{\mathrm{G}}(H)^{(1)}$. You may also have other choices. For instance when $\mathrm{G}=\mathrm{SL}_{2}(\mathbb{R})$ and $H$ is equal to the upper triangular unipotent group, then $V_{H}$ can be taken to be the standard representation $\mathbb{R}^{2}$ and $\nu_{H}=e_{1}$.

To go from the representation $V_{H}$ to $\mathrm{G} / \Gamma$, the following diagram is very natural.


Here $p$ and $q$ are natural projections and $\phi([g]):=g . \nu_{H}$. Strictly speaking $\phi$ may only be injective replacing $V_{H}$ by $V_{H} / \pm 1$, but we will ignore this minor issue.

Here is something naive one can do at this stage. Recall $E \subset T(H, \mathrm{U})$ is a compact set.

1. Take a compact subset $\widetilde{E} \subset N(H, \mathrm{U})$ such that $E=[\widetilde{E}]_{\Gamma}$, the image of $\widetilde{E}$ in $\mathrm{G} / \Gamma$;
2. Let $E^{\vee}:=\phi \circ q(\widetilde{E})=\widetilde{E} . v_{H}$;
3. Apply Lem.5.2 above to $E=E^{\vee}$ and $W$ to be determined (you may take $W=V$ and see why it does NOT work). Then we get $F$ (depending also on $\varepsilon$ ) by Lem.5.2, which asserts that for every open neighborhood $\Phi$ (we do not have a favorite $\Phi$ yet, so just fix some) there exists an open neighborhood $\Psi$ of $E^{\vee}$ such that something holds.
4. We simply take $\mathscr{N}_{\varepsilon}:=p\left((\phi \circ q)^{-1} \Psi\right)$ and $\mathscr{N}_{\varepsilon}^{\prime}:=p\left((\phi \circ q)^{-1} \Phi\right)$.

To simplify notations,
Definition 6.3. For $t \in \mathbb{R}$ and $[\gamma]_{\Gamma_{N}} \in \Gamma / \Gamma_{N}$,

$$
\begin{aligned}
\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right) & :=\left\{y \in \mathrm{D} \mid \psi_{t}(y) . x_{0} \in \mathscr{N}_{\varepsilon}\right\} \\
\mathrm{D}_{t}\left(\Psi,[\gamma]_{\Gamma_{N}}\right) & :=\left\{y \in \mathrm{D} \mid \psi_{t}(y) g_{0} \gamma \cdot \nu_{H} \in \Psi\right\} .
\end{aligned}
$$

And define $\mathrm{D}_{t}(\Psi):=\bigcup_{[\gamma] \in \Gamma / \Gamma_{N}} \mathrm{D}_{t}(\Psi,[\gamma])$. Similarly define $\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}^{\prime}\right), \mathrm{D}_{t}(\Phi,[\gamma])$ and $\mathrm{D}_{t}(\Phi)$.
Thus from the definition (the naive definition of $\mathscr{N}_{\varepsilon}$ above, we will work with a different $\mathscr{N}_{\varepsilon}$ later in Sec.8)

$$
\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right)=\bigcup \mathrm{D}_{t}\left(\Psi,[\gamma]_{\Gamma_{N}}\right) .
$$

## 7. Step 6, self-intersection

Now we seek to refine the rather crude strategy proposed in Step 5 so that it would actually work.

First of all in general, unlike the $\mathrm{SL}_{2}(\mathbb{R})$-case, the projection $N(H, \mathrm{U}) / \Gamma_{N} \rightarrow \mathrm{G} / \Gamma$ is not injective.

Lemma 7.1. If $g \in G$ is such that for two different $\left[\gamma_{1}\right]_{\Gamma_{N}} \neq\left[\gamma_{2}\right]_{\Gamma_{N}} \in \Gamma / \Gamma_{N}$ we have $g \gamma_{i} \in$ $N(H, \mathrm{U})$ for $i=1,2$, then $g \in \operatorname{Sing}(H, \mathrm{U}) \Gamma$.

So ideally we would like to avoid $\operatorname{Sing}(H, \mathrm{U})$ (or its projection to $\mathrm{G} / \Gamma_{N}$, or $\mathrm{G} / \Gamma$ ) from our discussion. But this is impossible! Since usually $\operatorname{Sing}(H, \mathbb{U})$ is dense in $N(H, \mathrm{U})$ modulo $\Gamma_{N}$, every non-empty open set intersects non-trivially with it. Lucky for us, each time we only work with certain compact set $F$ (to be found) in $V_{H}$ (and we have the freedom of choosing its neighborhood). And the subset of $\operatorname{Sing}(H, \mathrm{U})$ that is "relevant to $F$ " is indeed closed, see Lem.7.5.

To detect $N(H, \mathrm{U})$ inside $V_{H}$, it is convenient (though maybe not necessary) to have:
Definition 7.2. Let $W_{H}$ be the $\mathbb{R}$-linear subspace of $V_{H}$ spanned by $N(H, U) . v_{H}$.
This $W_{H}$ would be the $W$ when we apply Lem.5.2 above.

Lemma 7.3. We have

$$
(\phi \circ q)^{-1}\left(W_{H}\right)=N(H, \mathrm{U}) / \Gamma_{N} .
$$

The reader is reminded that being compact in $V_{H}$ is not the same as being compact in $\mathrm{G} / N_{\mathrm{G}}(H)^{(1)}$, as the $\mathrm{SL}_{2}(\mathbb{R})$-case already told us, unless $\mathrm{G} . \nu_{H}$ is closed in $V_{H}$, which is true if $H$ is reductive by [Kem78] or if $\Gamma$ is arithmetic and cocompact in G , for other reductive G's.

Here is the important observation
Definition 7.4. Let F be a compact subset of $W_{H}$, let

$$
\operatorname{Sing}(F):=\left\{g \in \mathrm{G} \mid g \cdot \nu_{H} \in F, g \gamma \cdot \nu_{H} \in F \exists \gamma \in \Gamma \backslash \Gamma_{N}\right\}
$$

Thus $\operatorname{Sing}(F) \subset \operatorname{Sing}(H, \mathrm{U})$. The fact we need is that
Lemma 7.5. $\operatorname{Sing}(F) \Gamma$ is closed.
Sketch of proof. First note that $\Gamma . v_{H}$ is discrete in $V_{H}$. This is rather straight-forward since $v_{H}$ is a rational vector and the image of $\Gamma$ in $\mathrm{SL}\left(V_{H}\right)$ is commensurable with $\mathrm{SL}_{N}(\mathbb{Z})$ for $N=\operatorname{dim} V_{H}$. For non-arithmetic lattices, see [DM93] for a proof.

Therefore if $\left(g_{n}\right)$ is bounded $\bmod \Gamma$ and $\left(g_{n} . v_{H}\right)$ is bounded in $V_{H}$, then $\left(g_{n}\right)$ is bounded modulo $\Gamma_{N}$. The conclusion follows quickly from here.

Note that

$$
\operatorname{Sing}(F) \Gamma:=\left\{g \in \mathrm{G} \mid g \gamma_{1} \cdot v_{H} \in F, g \gamma_{2} \cdot v_{H} \in F, \exists\left[\gamma_{1}\right] \neq\left[\gamma_{2}\right] \in \Gamma / \Gamma_{N}\right\}
$$

Consequently, by a continuity argument and the discreteness of $\Gamma . \nu_{H}$,
Lemma 7.6. Let $E^{\prime}$ be a compact set in $X \backslash[\operatorname{Sing}(F)]_{\Gamma}$. Then there exists an open neighbor$\operatorname{hood} \Phi$ of $F$ such that for every $[g]_{\Gamma} \in E^{\prime}$,

$$
\#\left\{[\gamma] \in \Gamma / \Gamma_{N} \mid g \gamma . \nu_{H} \in \Phi\right\} \leq 1 .
$$

## 8. Step 7, define the neighborhood

Let us explain how to find $\mathscr{N}_{\varepsilon}$. Fix $E \subset T(H, \mathrm{U})$ and $\varepsilon>0$.
Define $E^{\vee}$ as in Sec.6. By taking $W=W_{H}$ (see Def.7.2), Lem.5.2 offers some compact set $F$ of $W_{H}$. By Lem.7.5, $[\operatorname{Sing}(F)]_{\Gamma}$ is closed and is contained $[\operatorname{Sing}(H, \mathrm{U})]_{\Gamma}$ by Lem.7.1.

Now we take $E^{\prime}$ to be any compact set away from $[\operatorname{Sing}(H, \mathrm{U})]_{\Gamma}$ whose interior contains $E$. We find an open neighborhood $\Phi$ of $F$ such that the conclusion of Lem.7.6 holds. Then $\Psi$, an open neighborhood of $E$, is chosen according to Lem.5.2.

Just in case one gets confused, here is a diagram summarizing the logical dependence:


Now

$$
\mathscr{N}_{\varepsilon}:=\operatorname{Int}\left(E^{\prime}\right) \cap p\left((\phi \circ q)^{-1} \Psi\right)
$$

## 9. Step 8, a covering argument

The proof will be concluded with the help of a covering argument, something we encountered when discussing nondivergence of unipotent flow on $\mathrm{X}_{N}$. The argument here seems to differ from that of [EMS96].

Without loss of generality, assume D itself is a ball (the general case can be reduced to this one). The $\mathrm{D}^{(3)}(\bullet)$ is almost the same as $\mathrm{D}(\bullet)$ except that in Def.6.3, we replace D by the disk with the same center but whose radius is 3 times the radius of D (this is in order to apply Besicovitch's covering lemma, see Stein's book on real analysis, Chapter 3, Problem 3).

We further assume (this will be explained later in Sec.11)

$$
\begin{equation*}
\text { for } t \text { large enough, } \quad \psi_{t}(\mathrm{D}) g_{0} \gamma \cdot \nu_{H} \nsubseteq \Phi, \forall \gamma \in \Gamma \text {. } \tag{41}
\end{equation*}
$$

Recall the definition $\mathrm{D}_{t}(\Phi)$ and $\mathrm{D}_{t}(\Phi,[\gamma])$ from Def.6.3.
For each $[\gamma]$ such that $\mathrm{D}_{t}(\Phi,[\gamma])$ is non-empty. Find balls $\left\{B_{i}\right\}_{i \in \mathscr{F}_{t,[\gamma]}} \subset \mathrm{D}_{t}^{(3)}(\Phi,[\gamma])$ whose centers cover $\mathrm{D}_{t}(\Phi,[\gamma])$. Here $\mathscr{\mathscr { I }}_{t,[\gamma]}$ is some index set. We claim that we can find a covering such that for every $i \in \mathscr{I}_{t,[\gamma]}$, there exists $y \in \overline{B_{i}}$ such that

$$
\begin{equation*}
\psi_{t}(y) g_{0} \gamma \cdot v_{H} \notin \Phi . \tag{42}
\end{equation*}
$$

Indeed, for each $y \in \mathrm{D}_{t}(\Phi,[\gamma])$, take $B_{y}$ to be the largest open ball centered at y . Then this collection would satisfy Equa.(42) by Equa.(41).


Let $\mathscr{I}_{t}:=\bigsqcup_{[\gamma] \in \Gamma / \Gamma_{N}} \mathscr{I}_{t,[\gamma]}$. Then $\mathrm{D}_{t}(\Phi)$ is covered by the centers of $\left\{B_{i}\right\}_{i \in \mathscr{I}_{t}}$. By Besicovitch covering lemma, there exists a constant $C_{0}>0$, depending only on the dimension of $\mathfrak{k}_{0}$, and a subset $\mathscr{J}_{t} \subset \mathscr{I}_{t}$ such that $\left\{B_{i}\right\}_{i \in \mathscr{\mathscr { L }}}$ is a covering of $\mathrm{D}_{t}(\Phi)$ of multiplicity (i.e., the maximal number of possible overlaps among $B_{i}$ 's) bounded by $C_{0}$.

Let $\mathscr{F}_{t,[\gamma]}:=\mathscr{J}_{t} \cap \mathscr{I}_{t,[\gamma]}$.
Let me summarize the discussion in this subsection by the following lemma:
Lemma 9.1. Take $t$ such that Equa.(41) holds. There exists a covering of $\mathrm{D}_{t}(\Phi)$ by open balls $\left(B_{j}\right)_{j \in \mathscr{L}_{t}}$ together with a partition of the index set $\mathscr{J}_{t}=\sqcup_{[\gamma] \in \Gamma / \Gamma_{N}} \mathscr{J}_{t,[\gamma]}$ satisfying the following:

1. For $j \in \mathscr{J}_{t,[\gamma]}, B_{j} \subset \mathrm{D}^{(3)}(\Phi,[\gamma])$.
2. For $j \in \mathscr{J}_{t,[\gamma]}$, there exists $y \in \overline{B_{j}}$ such that

$$
\mathbf{a}_{t} \exp (y) g_{0} \gamma \cdot v_{H} \notin \Phi .
$$

3. The multiplicity of the covering is at most $C_{0}$ for a constant $C_{0}>0$ depending only on the dimension of $\mathfrak{k}_{0}$. Or more formally,

$$
\sum_{j \in \mathscr{I}_{t}} 1_{B_{j}} \leq C_{0}
$$

## 10. Step 9, finish the proof under some assumption

Lemma 10.1. There is a constant $C_{1}>0$ such that for tsatisfying Equa.(41)

$$
\operatorname{Leb}\left(\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right)\right) \leq C_{1} \varepsilon \operatorname{Leb}(\mathrm{D}) .
$$

Thus Lem.4.1 follows from this lemma provided Equa.(41) is verified.
Proof. Take $C_{1}:=3^{\text {dim }}{ }_{0} C_{0}$.
Take $y \in \mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right)$, then by Lem.7.6, there exists a unique $\left[\gamma_{y}\right] \in \Gamma / \Gamma_{N}$ such that

$$
\psi_{t}(y) g_{0} \gamma_{y} \cdot \nu_{H} \in \Phi
$$

On the other hand, since $y \in \mathrm{D}_{t}(\Psi) \subset \mathrm{D}_{t}(\Phi)$, there exists $[\gamma] \in \Gamma / \Gamma_{N}$ and $j \in \mathscr{J}_{t,[\gamma]}$ such that

$$
\psi_{t}(y) g_{0} \gamma \cdot v_{H} \in \Phi, \text { and } y \in B_{j} .
$$

By uniqueness, $[\gamma]=\left[\gamma_{y}\right]$. Let $\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon},[\gamma]\right):=\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right) \cap \mathrm{D}_{t}(\Psi,[\gamma])$.
We have proved that for every $[\gamma] \in \Gamma / \Gamma_{N}$,

$$
\begin{equation*}
\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon},[\gamma]\right)=\bigcup_{j \in \mathscr{\mathscr { A }}_{t,[\gamma]}} B_{j} \cap \mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon},[\gamma]\right) . \tag{43}
\end{equation*}
$$

By comparison, it may not be true that (even if you replace $\Phi$ by the smaller $\Psi$ )

$$
\mathrm{D}_{t}(\Phi,[\gamma])=\bigcup_{j \in \mathscr{\mathscr { F }},[r]} B_{j} \cap \mathrm{D}_{t}(\Phi,[\gamma])
$$

Now everything follows from this, the linear algebra lemma Lem.5.2 and the covering argument Lem.9.1. More details:

$$
\begin{aligned}
\operatorname{Leb}\left(\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon}\right)\right) & =\sum_{[\gamma] \in \Gamma / \Gamma_{N}} \operatorname{Leb}\left(\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon},[\gamma]\right)\right) \\
\text { (Equa.(43)) } & \leq \sum_{[\gamma] \in \Gamma / \Gamma_{N}} \sum_{j \in \mathcal{\mathscr { G }}_{t,[\gamma]}} \operatorname{Leb}\left(\mathrm{D}_{t}\left(\mathscr{N}_{\varepsilon},[\gamma]\right) \cap B_{j}\right) \\
& \leq \sum_{[\gamma] \in \Gamma / \Gamma_{N}} \sum_{j \in \mathscr{\mathscr { I }}_{t,[\gamma]}} \operatorname{Leb}\left(\mathrm{D}_{t}(\Psi,[\gamma]) \cap B_{j}\right) \\
\text { (Lem.5.2 and Equa.(42) ) } & \leq \sum_{[\gamma] \in \Gamma / \Gamma_{N}} \sum_{j \in \mathcal{\mathscr { G }}_{t,[\gamma]}} \varepsilon \operatorname{Leb}\left(B_{j}\right) \\
(\text { Lem.9.1 }) & \leq \varepsilon C_{0} \operatorname{Leb}\left(\mathrm{D}_{t}^{(3)}(\Phi)\right) \leq C_{0} \varepsilon \operatorname{Leb}\left(\mathrm{D}^{(3)}\right)=C_{1} \varepsilon \operatorname{Leb}(\mathrm{D}) .
\end{aligned}
$$

The promised $\mathscr{N}_{\varepsilon}^{\prime}$ did not show up explicitly. You may take it to be $p\left((\phi \circ q)^{-1} \Phi\right)$ in light of the discussion above.

## 11. Step 10, linear expansion

Note that the discussion so far only uses

- the limit measure $\mu$ is unipotent-invariant;
- ( $C, \alpha$ )-good properties.

In particular, as long as $\mu$ can be shown to be unipotent invariant, the discussion above applies equally well if you replace $\mathbf{a}_{t_{n}}$ by any other sequences $\left(g_{n}\right)$ in $G$ and $\exp (\mathrm{D})$ by any other bounded smooth curve/manifold in G equipped with a smooth measure.

Now we explain why Equa.(41) holds, for our particular choice of $\mathbf{a}_{t}$ and $\exp (\mathrm{D})$.
Recall that we may think of (the connected component of) $\mathrm{H}_{0}$ as the image of $\mathrm{SL}_{2}(\mathbb{R})$ under the Adjoint representation. And $K_{0}$ may be thought of as the image of $\mathrm{SO}_{2}(\mathbb{R}),\left\{\mathbf{a}_{t}\right\}$ the image of $\mathbf{b}_{t}:=\operatorname{diag}\left(e^{t}, e^{-t}\right)$.

Lemma 11.1. Let $V$ be an irreducible nontrivial representation of $\mathrm{SL}_{2}(\mathbb{R})$. Let $\Omega$ be a nonempty open subset of $\mathrm{SO}_{2}(\mathbb{R})$. Then for every constant $C>0$, there exists $T_{0}>0$ (depending on $C, \Omega$, the choice of metric on $V$ ) such that for every $t>T_{0}$, every $v_{\neq} 0 \in V$

$$
\sup _{\omega \in \Omega}\left\|\mathbf{b}_{t} \omega \cdot v\right\| \geq C\|v\| .
$$

REMARK 11.2. After the proof is given, it should be clear that $\mathrm{SL}_{2}(\mathbb{R})$ can be replaced by any other simple Lie group, $\mathrm{SO}_{2}(\mathbb{R})$ replaced by a maximal compact subgroup, $\mathbf{b}_{t}$ replaced by any one-parameter diagonalizable subgroup that is stable under Cartan involution associated with this maximal compact subgroup. Moreover, once $V$ is fixed, $C$ can be taken to be $\kappa_{1} e^{\kappa_{2}|t|}$ for some $\kappa_{1}, \kappa_{2}>0$ and the condition $t>T_{0}$ can be removed.

REmark 11.3. A weaker statement, with "for every $C>0$ " replaced by "there exists some $c>0$ " (and ignore the $t>T_{0}$ condition) holds in much greater generality, see [RS18]. And this condition is sufficient to conclude the limit measure supports on a unique tube (see [RZ16]).

Proof of Equa.(41) assuming Lem.11.1. Assume otherwise, find some $\gamma_{t} \in \Gamma$ such that

$$
\mathbf{a}_{t} \exp (\mathrm{D}) g_{0} \gamma_{t} \cdot v_{H} \subset \Phi
$$

for $t$ inside certain sequence tending to $+\infty$.

Decompose $V=V_{1} \oplus V_{2}$ in a $\mathrm{H}_{0}$-equivariant way such that $V_{1}=V^{\mathrm{H}_{0}}$, the vectors fixed by $\mathrm{H}_{0}$. Write $\pi_{i}$ for the projection $V \rightarrow V_{i}$ w.r.t. this decomposition. Without loss of generality we assume $V_{1} \perp V_{2}$ by changing the Euclidean metric. Thus for $t \in R, y \in \mathrm{D}$.

$$
\begin{aligned}
\mathbf{a}_{t} \exp (y)\left(g_{0} \gamma_{t} \cdot v_{H}\right) & =\mathbf{a}_{t} \exp (y)\left(\pi_{1}\left(g_{0} \gamma_{t} \cdot v_{H}\right)+\pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right) \\
& =\pi_{1}\left(g_{0} \gamma_{t} \cdot v_{H}\right)+\mathbf{a}_{t} \exp (y) \pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right) \\
\Rightarrow\left\|\mathbf{a}_{t} \exp (y) g_{0} \gamma_{t} \cdot v_{H}\right\| & =\left\|\pi_{1}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right\|+\left\|\mathbf{a}_{t} \exp (y) \pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right\| .
\end{aligned}
$$

For the 2nd term, the above Lem.11.1 implies that for $t$ large enough, for suitable choice of $y_{t}$,

$$
\left\|\mathbf{a}_{t} \exp \left(y_{t}\right) \pi_{2}\left(g_{0} \gamma_{t} \cdot \nu_{H}\right)\right\| \geq\left\|\pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right\| .
$$

So $\mathbf{a}_{t} \exp \left(y_{t}\right)$ action does not decrease the norm of $g_{0} \gamma_{t} \cdot v_{H}$. Since $\Phi$ is bounded, this implies that

$$
\left(g_{0} \gamma_{t} \cdot v_{H}\right) \text { is bounded. }
$$

But $\Gamma . v_{H}$, and hence $g_{0} \Gamma . v_{H}$ is discrete in $V_{H}$. A discrete, bounded set has no choice but being finite. After passing to a subsequence, we assume $\gamma_{t}=\gamma_{1}$ for all $t$ (in some infinite subsequence tending to $+\infty$ ).

Now if $g_{0} \gamma_{1} \cdot v_{H} \notin V_{1}$, then $\pi_{2}\left(g_{0} \gamma_{1} \cdot v_{H}\right) \neq 0$. Take $C_{2}>0$ such that every element in $\Phi$ has norm at most $C_{2}$. Apply Lem.11.1 to $C=1.1 C_{2}\left\|\pi_{2}\left(g_{0} \gamma_{1} \cdot v_{H}\right)\right\|^{-1}$, then we find $y_{t}^{\prime}$, for $t$ large enough, such that

$$
\left\|\mathbf{a}_{t} \exp \left(y_{t}^{\prime}\right) \pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right\| \geq 1.1 C_{2}\left\|\pi_{2}\left(g_{0} \gamma_{1} \cdot v_{H}\right)\right\|^{-1}\left\|\pi_{2}\left(g_{0} \gamma_{t} \cdot v_{H}\right)\right\|=1.1 C_{2}
$$

So $\mathbf{a}_{t} \exp \left(y_{t}^{\prime}\right) g_{0} \gamma_{1} \cdot v_{H}$ can not live in $\Phi$, a contradiction.
Thus $g_{0} \gamma_{1} \cdot v_{H} \in V_{1}$, or in other words, $g_{0} \gamma_{1} \cdot v_{H}$ is fixed by $\mathrm{H}_{0}$. Recall the stabilizer of $v_{H}$ in G is $N_{\mathrm{G}}(H)^{(1)}$, thus, $g_{0}^{-1} \mathrm{H}_{0} g_{0} \subset \gamma_{1} N_{\mathrm{G}}(H)^{(1)} \gamma_{1}^{-1} \subset \gamma_{1} N_{\mathrm{G}}(H) \gamma_{1}^{-1}$.

A Lie algebra computation shows that $\operatorname{Ad}\left(g_{0}\right)^{-1} \mathfrak{h}_{0}$ is a maximal proper Lie subalgebra. Actually, the only non-zero and non-full $\operatorname{Ad}\left(\mathrm{H}_{0}\right)$-stable Lie subalgebra of $\mathfrak{s l}_{3}$ is $\mathfrak{h}_{0}$. Thus $\operatorname{Ad}\left(g_{0}^{-1}\right) \mathfrak{h}_{0}=$ $\operatorname{Ad}\left(\gamma_{1}\right) \mathfrak{h}$ and $g_{0}^{-1} \mathrm{H}_{0}^{\circ} g_{0}=\gamma_{1} H \gamma_{1}^{-1}$. In particular $g_{0}^{-1} \mathrm{H}_{0}^{\circ} g_{0} \cap \Gamma$ is a lattice in $g_{0}^{-1} \mathrm{H}_{0}^{\circ} g_{0}$. This implies that $Q_{0} \circ g_{0}$ is proportional to a rational quadratic form, a contradiction.

Proof of Lem.11.1. Decompose $V$ w.r.t. the $\mathbf{b}_{t}$ action

$$
V=V^{-} \oplus V^{0} \oplus V^{+}
$$

into contracting/fixed/expanding subspaces. Namely, this decomposition is stable under $\mathbf{b}_{t}$ action. Moreover $V^{0}=V^{\left\{\mathbf{b}_{t}\right\}}$ and for some $c_{1}, \kappa_{1}>0$,

$$
\begin{aligned}
& \left\|\mathbf{b}_{t} \cdot v\right\| \geq c_{1} e^{\kappa_{1} t}\|v\|, \forall v \in V^{+} \\
& \left\|\mathbf{b}_{t} \cdot v\right\| \leq c_{1}^{-1} e^{-\kappa_{1} t}\|v\|, \forall v \in V^{-} .
\end{aligned}
$$

Let $\pi^{-}, \pi^{0}, \pi^{+}$be the corresponding projections. We claim that there exists $c_{2}>0$ such that

$$
\begin{equation*}
\sup _{\omega \in \Omega}\left\|\pi^{+}(\omega \cdot v)\right\| \geq c_{2}\|v\|, \forall v \in V . \tag{44}
\end{equation*}
$$

Once this is done, the proof completes. It suffices to verify Equa.(44) under the assumption $\|v\|=1$. If not true, then we can find a sequence of unit vectors ( $v_{n}$ ) such that

$$
\sup _{\omega \in \Omega}\left\|\pi^{+}\left(\omega . v_{n}\right)\right\| \rightarrow 0
$$

Let $v_{\infty}$ be any limit of $\left(v_{n}\right)$. Since $\Omega$ is bounded, we have

$$
\pi^{+}\left(\omega \cdot \nu_{\infty}\right)=0, \forall \omega \in \Omega
$$

In other words,

$$
\Omega . v_{\infty} \subset V^{-} \oplus V^{0} .
$$

Since this is a condition defined by vanishing of some polynomials and $\Omega$ is Zariski dense in $\mathrm{SO}_{2}(\mathbb{R})$, we have

$$
\mathrm{SO}_{2}(\mathbb{R}) \cdot v_{\infty} \subset V^{-} \oplus V^{0} .
$$

Since $w_{0}:=\left[\begin{array}{cc}0 & 1 \\ -1 & 0\end{array}\right] \in \mathrm{SO}_{2}(\mathbb{R})$ and $w_{0} \mathbf{b}_{t} w_{0}^{-1}=\mathbf{b}_{-t}$, we see that $w_{0} V^{-}=V^{+}, w_{0} V^{0}=V^{0}$ and $w_{0} V^{+}=V^{-}$. So

$$
\mathrm{SO}_{2}(\mathbb{R}) \cdot v_{\infty} \subset V^{0} .
$$

So every vector in $V^{0}$ is fixed by $\mathrm{SO}_{2}(\mathbb{R})$ and $\left\{\mathbf{b}_{t}\right\}$, which generate the full $\mathrm{SL}_{2}(\mathbb{R})$. This is a contradiction.

## 12. Exercises

### 12.1. An example of equidistribution of unipotent flows. Notations

- $G=\mathrm{SL}_{2}(\mathbb{C}), \Gamma=\mathrm{SL}_{2}(\mathbb{Z}[i])$ and $X:=G / \Gamma$;
- $U=\left\{\left.\mathbf{u}_{s}=\left[\begin{array}{ll}1 & s \\ 0 & 1\end{array}\right] \right\rvert\, s \in \mathbb{R}\right\}$ and $x_{0}=\left[g_{0}\right] \in G / \Gamma$.

Let $\left(S_{n}\right)$ be a sequence of positive real numbers tending to $+\infty$ such that the following limit exists:

$$
\mu:=\lim _{S_{n} \rightarrow+\infty} \frac{1}{S_{n}} \int_{0}^{S_{n}}\left(\mathbf{u}_{s}\right)_{*} \delta_{\left[g_{0}\right]} \mathrm{ds} .
$$

Assume the fact that such a $\mu$ belongs to $\operatorname{Prob}(X)^{U}$.
Recall the definitions of $\mathscr{H}, T(H, U), \ldots$ (see Lec.11, Def.1.6, Def.3.1). And $V_{H}, v_{H}$ same as in Lec. 12.

Exercise 12.1. Let $H \in \mathscr{H}, H \neq G$. Show that if $\mu(T(H, U))>0$, then there exists a bounded set $\Phi \subset V_{H}$ and a sequence $\left(\gamma_{n}\right) \subset \Gamma$ such that

$$
\mathbf{u}_{\left[0, S_{n}\right]} g_{0} \gamma_{n} . \nu_{H} \subset \Phi .
$$

EXERCISE 12.2. Same notations as the exercise above. Conclude that there exists $\gamma \in \Gamma$ such that

$$
\mathbf{u}_{[0,+\infty)} g_{0} \gamma \cdot v_{H} \subset \Phi .
$$

EXERCISE 12.3. Same notations as the exercise above. Conclude that $g_{0}^{-1} U g_{0} \subset N_{G}\left(\gamma H \gamma^{-1}\right)^{(1)}$.
EXERCISE 12.4. Use exercises above to show that if $x_{0}=\left[g_{0}\right] \notin[\operatorname{Sing}(G, U)]_{\Gamma}$, then

$$
\lim _{S_{n} \rightarrow+\infty} \frac{1}{S_{n}} \int_{0}^{S_{n}}\left(\mathbf{u}_{s}\right)_{*} \delta_{\left[g_{0}\right]} \mathrm{ds}=\widehat{\mathrm{m}}_{G / \Gamma}
$$

[Hint: use Lec.11, Thm.2.3 if it helps.]
Exercise 12.5. Conclude that if $x_{0}=\left[g_{0}\right] \notin[\operatorname{Sing}(G, U)]_{\Gamma}$, then $U . x_{0}$ is dense in $G / \Gamma$.
12.2. Homogeneous sets of bounded volume. Notations

- $G:=\mathrm{SL}_{N}(\mathbb{R})$ and $\Gamma:=\mathrm{SL}_{N}(\mathbb{Z})$.
- Fix a right $G$-invariant Riemannian metric on $G$, which induces Riemannian metrics on $G / \Gamma$ and also on immersed submanifolds. Volumes below are all induced from this.
For $C>0$, let

$$
\begin{aligned}
\mathscr{A} & :=\{H \leq G \mid H \text { is a closed connected subgroup of } G, \operatorname{Vol}(H / H \cap \Gamma)<\infty .\} \\
\mathscr{A}_{C} & :=\{H \leq G \mid H \text { is a closed connected subgroup of } G, \operatorname{Vol}(H / H \cap \Gamma)<C .\}
\end{aligned}
$$

DEFINITION 12.1. Given a sequence $\left(H_{n}\right)$ of closed subgroups of $G$, we say that $\left(H_{n}\right)$ converges iff for every (infinite) subsequence $\left(n_{k}\right)$ and $h_{n_{k}} \in H_{n_{k}}$ such that $\lim _{k} h_{n_{k}}$ exists, there exists $h_{n}^{\prime} \in H_{n}$ for each $n$, such that

$$
\lim _{k} h_{n_{k}}=\lim _{n} h_{n}^{\prime} .
$$

EXERCISE 12.6. Given a sequence $\left(H_{n}\right)$ of closed subgroups of $G$, there exists a subsequence that converges.

From now on we fix a convergent sequence $\left(H_{n}\right)$. And assume each $H_{n}$ is connected. Let

$$
L:=\left\{g \in G \mid g=\lim _{n} h_{n}, \exists h_{n} \in H_{n}\right\}
$$

EXERCISE 12.7. Show that $L$ is a closed subgroup.
EXERCISE 12.8. There exists a subsequence $n_{k}$ such that $\left(\mathfrak{h}_{n_{k}}\right)$ (the Lie algebra of $H_{n_{k}}$ ) converges.

From now on we assume $\left(\mathfrak{h}_{n}\right)$ converges to $\mathfrak{h}_{\infty}$.
EXERCISE 12.9. Find an example of $\left(H_{n}\right)$ such that $\mathfrak{h}_{\infty}$ is not the Lie algebra of $L$.
Now we further assume that $\left\{H_{n}\right\} \subset \mathscr{A}_{C_{0}}$ for some $C_{0}>0$.
EXERCISE 12.10. Show that under the assumption above, $\mathfrak{h}_{\infty}=\operatorname{Lie}(L)$.
EXERCISE 12.11. Show that $\left(H_{n} \cap \Gamma\right)$ converges and its limit is given by

$$
\Gamma_{\infty}:=\left\{\gamma \in \Gamma \mid \exists n_{0}, \forall n>n_{0}, \gamma \in H_{n} \cap \Gamma\right\} .
$$

EXERCISE 12.12. Show that $\mathrm{Vol}_{H_{n}}$ converges to $\mathrm{Vol}_{L}$ in the weak* topology.
EXERCISE 12.13. Show that $\Gamma_{\infty}$ is a lattice in L. Indeed show that

$$
\operatorname{Vol}\left(L / \Gamma_{\infty}\right) \leq \limsup \operatorname{Vol}\left(H_{n} / H_{n} \cap \Gamma\right)
$$

[Hint, consider compact parts of a fundamental domain]
It is a fact that once you know $\Gamma_{\infty}$ is a lattice in $L$, then it is finitely generated.
EXERCISE 12.14. Assume the fact above. Show that there exists $n_{0}$ such that for all $n>n_{0}$, $\Gamma \cap H_{n} \supset \Gamma_{\infty}$.

Continuing this way, using more inputs from the theory of algebraic groups, one can show that

Theorem 12.2 (Dani-Margulis). We have that

$$
\#\left\{H \cap \Gamma \mid H \in \mathscr{A}_{C_{0}}\right\}<\infty
$$

Exercise 12.15. For $H \in \mathscr{A}$ and $g \in G$, show that

$$
\operatorname{Vol}(g H \Gamma / \Gamma)=\frac{\left\|\operatorname{Ad}(g) \cdot v_{H}\right\|}{\left\|v_{H}\right\|} \operatorname{Vol}(H \Gamma / \Gamma)
$$

Here $v_{H}$ is a vector in $\wedge^{\operatorname{dim} H_{\mathfrak{s l}}^{n}}$ defined by $v_{1} \wedge \ldots \wedge v_{\operatorname{dim} H}$ where $\left(v_{1}, \ldots, v_{\operatorname{dim} H}\right)$ is a basis for $\mathfrak{h}$, the Lie algebra of $H$.

EXERCISE 12.16. Assume the theorem above, show that $\Gamma . v_{H}$ is a discrete subset of $\wedge \operatorname{dim}_{H_{\mathfrak{s l}}^{n}}$.
12.3. Orbit counting and equidistribution. Notations

- $G=\mathrm{SL}_{2}(\mathbb{R}), \Gamma=\mathrm{SL}_{2}(\mathbb{Z}), H=\left\{\left.\left[\begin{array}{cc}x & 2 y \\ y & x\end{array}\right] \right\rvert\, x^{2}-2 y^{2}=1\right\}$;
- $V:=\{2$-by-2 real matrices with trace 0$\}$;
- $V(\mathbb{Z}):=\{2$-by- 2 integer matrices with trace 0$\}$
- $M_{0}:=\left[\begin{array}{ll}0 & 2 \\ 1 & 0\end{array}\right]$ and $p_{0}(x):=x^{2}-2$;
- for a matrix $M$, its characteristic polynomial is denoted by $\operatorname{char}_{M}(x):=\operatorname{det}(x I-M)=$ $x^{2}-\operatorname{Tr}(M) x+\operatorname{det}(M)$;
- $X_{p_{0}}(\mathbb{R}):=\left\{M \in V, \operatorname{char}_{M}(x)=p_{0}(x)\right\}, X_{p_{0}}(\mathbb{Z}):=\left\{M \in V(\mathbb{Z}), \operatorname{char}_{M}(x)=p_{0}(x)\right\}$;
- for a 2-by-2 matrix $M=\left[\begin{array}{ll}a & b \\ c & d\end{array}\right]$, define $\operatorname{ht}(M):=\sqrt{a^{2}+b^{2}+c^{2}+d^{2}}$;
- $B_{R}:=\left\{M \in X_{p_{0}}(\mathbb{R}) \mid \operatorname{ht}(M) \leq R\right\}$.

EXERCISE 12.17. Show that every pair of matrices $M_{1}, M_{2} \in X_{p_{0}}(\mathbb{R})$, there exists $g \in G$ such that $g M_{1} g^{-1}=M_{2}$.

Let $G$ acts on $X_{p_{0}}(\mathbb{R})$ by $g . M:=g M g^{-1}$. The above exercise shows that this action is transitive.

EXERCISE 12.18. The stabilizer of $M_{0}$ in $G$ is equal to $H$.
EXERCISE 12.19. $H \cap \Gamma$ is a lattice in $H$.
EXERCISE 12.20. Show that the action of $\Gamma$ on $X_{p_{0}}(\mathbb{Z})$ is transitive.
[Hint: $\mathbb{Z}[\sqrt{2}]$ is a PID]
Further notations

- $\mathrm{m}_{G / H}$ is a $G$-invariant locally finite measure on $G / H$;
- similarly, $\mathrm{m}_{G}$ and $\mathrm{m}_{H}$ denote Haar measures on $G$ and $H$ respectively.

Note that $G$ and $H$ are unimodular: left Haar measures are the same as right Haar measures.
DEFINITION 12.3. We say that a triple $\left(\mathrm{m}_{G}, \mathrm{~m}_{H}, \mathrm{~m}_{G / H}\right)$ is compatible iff for every compactly supported function $f \in C_{c}(G)$, we have

$$
\begin{equation*}
\int_{G / H} \int_{H} f(g h) \mathrm{m}_{H}(h) \mathrm{m}_{G / H}([g])=\int_{G} f(g) \mathrm{m}_{G}([g]) \tag{45}
\end{equation*}
$$

EXERCISE 12.21. Show that for every triple of Haar measures $\left(\mathrm{m}_{G}, \mathrm{~m}_{H}, \mathrm{~m}_{G / H}\right)$, there exists a constant $c>0$ such that for every $f \in C_{c}(G)$,

$$
\int_{G / H} \int_{H} f(g h) \mathrm{m}_{H}(h) \mathrm{m}_{G / H}([g])=c \cdot \int_{G} f(g) \mathrm{m}_{G}([g]) .
$$

From now on we fix the unique triple $\left(\mathrm{m}_{G}, \mathrm{~m}_{H}, \mathrm{~m}_{G / H}\right)$ satisfying

1. ( $\left.\mathrm{m}_{G}, \delta_{\Gamma}, \widehat{\mathrm{m}}_{G / \Gamma}\right)$ and $\left(\mathrm{m}_{H}, \delta_{H \cap \Gamma}, \widehat{\mathrm{~m}}_{H / H \cap \Gamma}\right)$ are compatible. Here $\delta_{\Gamma}$ (resp. $\delta_{H \cap \Gamma}$ ) denotes the counting measure on $\Gamma$ (resp. $H \cap \Gamma$ ).
2. $\left(\mathrm{m}_{G}, \mathrm{~m}_{H}, \mathrm{~m}_{G / H}\right)$ is compatible.

Its existence is guaranteed by the Exer.12.21 above.
ExErcise 12.22. Find the asymptotics of

$$
\mathrm{m}_{G / H}\left(B_{R}\right):=\mathrm{m}_{G / H}\left(\left\{[g] \in G / H \mid \operatorname{ht}\left(g \cdot M_{0}\right) \leq R\right\}\right) .
$$

DEFINITION 12.4. Define $\varphi_{R}: G / \Gamma \rightarrow \mathbb{R}$ by

$$
\varphi_{R}([g]):=\#\left(g \Gamma \cdot M_{0} \cap B_{R}\right)
$$

We say that $\frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \varphi_{R}$ converges to 1 weakly iff for all $\psi \in C_{c}(G / \Gamma)$,

$$
\begin{equation*}
\lim _{R \rightarrow+\infty} \frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \int_{G / \Gamma} \varphi_{R}([g]) \psi([g]) \widehat{\mathrm{m}}_{G / \Gamma}([g])=\int \psi([g]) \widehat{\mathrm{m}}_{G / \Gamma}([g]) . \tag{46}
\end{equation*}
$$

EXERCISE 12.23. Show that if $\frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \varphi_{R}$ converges to 1 weakly then for every $[g] \in G / \Gamma$,

$$
\lim _{R \rightarrow+\infty} \frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \varphi_{R}([g])=1 .
$$

In particular, in light of Exer.12.20,

$$
\# X_{p_{0}}(\mathbb{Z}) \cap B_{R} \sim \mathrm{~m}_{G / H}\left(B_{R}\right)
$$

[Hint: use Exer.12.22].
Exercise 12.24. Show that the left hand side of Equa.(46) (excluding the limit) is equal to

$$
\frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \int_{\left\{g . M_{0} \in B_{R}\right\}}\left(\int \psi(x) g_{*} \widehat{\mathrm{~m}}_{H \Gamma / \Gamma}(x)\right) \mathrm{m}_{G / H}([g])
$$

EXERCISE 12.25. Use "linearization technique" to show that for every sequence $\left(g_{n}\right)$ such that $\left(\left[g_{n}\right]\right)$ diverges in $G / H$, we have

$$
\lim _{n \rightarrow+\infty}\left(g_{n}\right)_{*} \widehat{\mathrm{~m}}_{H \Gamma / \Gamma}=\widehat{\mathrm{m}}_{G / \Gamma}
$$

Exercise 12.26. Use Exer. 12.25 to conclude that $\frac{1}{\mathrm{~m}_{G / H}\left(B_{R}\right)} \varphi_{R}$ converges to 1 weakly.

## CHAPTER 13

## Quantitative Oppenheim I, reducing to dynamics

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Main reference: [EMM98, Section 3].
Notations

- Let $Q_{0}\left(x_{1}, x_{2}, x_{3}, x_{4}\right):=2 x_{1} x_{4}+x_{2}^{2}+x_{3}^{2}$, a real quadratic form of signature $(3,1)$ on $\mathbb{R}^{4}$.
- Let $\left(\mathbf{e}_{1}, \ldots, \mathbf{e}_{4}\right)$ be the standard basis of $\mathbb{R}^{4}$; and for a vector $\mathbf{v}$, define its coefficients by $v=\sum(\nu)_{i} \mathbf{e}_{i}$ and we also write $v=\left((\nu)_{1}, \ldots,(\nu)_{4}\right)$.
- Let $\left(\mathbf{f}_{1}, \ldots, \mathbf{f}_{4}\right)$ be another ONB(=orthogonal normal basis) defined by $\mathbf{f}_{2}=\mathbf{e}_{2}, \mathbf{f}_{3}=\mathbf{e}_{3}$ and $\mathbf{f}_{1}=\frac{\mathbf{e}_{1}+\mathbf{e}_{4}}{\sqrt{2}}, \mathbf{f}_{4}=\frac{\mathbf{e}_{1}-\mathbf{e}_{4}}{\sqrt{2}}$. If $v=\sum a_{i} \mathbf{f}_{i}$, we also write $v=\left(a_{1}, \ldots, a_{4}\right)_{\mathbf{f}}$.
- One can verify that $Q_{0}\left(\left(x_{1}, \ldots, x_{4}\right)_{\mathbf{f}}\right)=x_{1}^{2}+x_{2}^{2}+x_{3}^{2}-x_{4}^{2}$.
- $\mathrm{K}:=\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R}) \cap \mathrm{SO}_{4}(\mathbb{R})$.
- $\mathbf{a}_{t}:=\operatorname{diag}\left(e^{-t}, 1,1, e^{t}\right)$, contained in $\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R})$.


## 1. Detect points by probabilistic methods

Assume $Q_{0} \circ g_{0}$ is irrational. Define

$$
\begin{gathered}
V_{(a, b)}(\mathbb{Z}):=\left\{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4} \mid Q_{0}(\mathbf{v}) \in(a, b)\right\}, \\
N_{T}:=\# V_{a, b}(\mathbb{Z}, T), V_{a, b}(\mathbb{Z}, T):=\left\{\mathbf{v} \in V_{(a, b)}(\mathbb{Z}) \mid\|\mathbf{v}\| \leq T\right\} .
\end{gathered}
$$

Consider the function

$$
1_{\square}(x, y):=1_{(1,2\rfloor}(x) \cdot 1_{(a, b)}(y) .
$$

Hence

$$
N_{2 T}-N_{T}=\sum_{\mathbf{v} \in g_{0} . \mathbb{Z}^{4}} 1_{\square}\left(\frac{\|\mathbf{v}\|}{T}, Q_{0}(\mathbf{v})\right) .
$$

Find a compactly supported continuous function $h$ approximating $1_{\square}$ from above. Then one can find some (non-negative) $f \in C_{c}\left(\mathbb{R}_{>0} \times \mathbb{R}^{3}\right)$ such that

$$
\begin{equation*}
h(x, y)=\frac{1}{x^{2}} \int f\left(x, w_{2}, w_{3}, y^{\prime}\right)\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right| \tag{47}
\end{equation*}
$$

where $y^{\prime}:=\frac{y-w_{2}^{2}-w_{3}^{2}}{2 x}$.
1.1. A coarse upper bound. By abbreviating $V_{a, b}(\mathbb{Z}, 2 T-T):=V_{a, b}(\mathbb{Z}, 2 T) \backslash V_{a, b}(\mathbb{Z}, T)$, we have

$$
\begin{align*}
N_{2 T}-N_{T} & \leq \sum_{\mathbf{v} \in g_{0} . \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T-T)} h\left(\frac{\|\mathbf{v}\|}{T}, Q_{0}(\mathbf{v})\right) \\
& =\sum_{\mathbf{v} \in g_{0} . \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T-T)} \frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, \frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2\|\mathbf{v}\| T^{-1}}\right)\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right| \tag{48}
\end{align*}
$$

Each summand here is either 0 or $\geq 1$ since we are keeping the index $\mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T-T)$.
Now we need the following lemma, to be proved later (see Lem.2.10 where this is proved).

Lemma 1.1. Given $f \in C_{c}\left(\mathbb{R}_{>0} \times \mathbb{R}^{3}\right)$ and $\varepsilon \in(0,1)$, there exists $T_{0}=T_{0}(f, \varepsilon)>0$ such that for every $T>T_{0}$, for every $\mathbf{v} \in \mathbb{R}^{4}$ we have

$$
\left|\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{\ln T} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)-\frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}| |<\varepsilon
$$

where

$$
w_{4}:=\frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2\|\mathbf{v}\| T^{-1}}
$$

is a function in $\left(w_{2}, w_{3}\right)$, for every fixed $\mathbf{v}$ and $T$.
Apply Lem.1.1 with some $\varepsilon<0.5$, then for $T$ sufficiently large, each $\mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T-T)$, either

$$
\frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, \frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2\|\mathbf{v}\| T^{-1}}\right)=\frac{1}{2 C_{4}} T^{2} \int 2 f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)=0
$$

or $\geq 0.5$.
Therefore

$$
\begin{align*}
N_{2 T}-N_{T} & \leq \sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T-T)} 2 \cdot \frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{K}(k) \\
& \leq \sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}} 2 \frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{K}(k)=2 \frac{1}{2 C_{4}} T^{2} \int \tilde{f}\left(\mathbf{a}_{t} k g_{0} \cdot \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k) . \tag{49}
\end{align*}
$$

where

$$
\tilde{f}: \mathrm{X}_{4} \rightarrow \mathbb{R} \text { defined by } \tilde{f}(\Lambda):=\sum_{\mathbf{v} \in \Lambda} f(\mathbf{v}) .
$$

If $\widetilde{f}$ were a bounded function, then immediately we see that for some constant $C=C(f)>0$, $N_{2 T}-N_{T} \leq T^{2} C \Longrightarrow N_{2^{n}} T_{0} \leq T_{0}^{2} C\left(1+4^{1}+\ldots+4^{n-1}\right)+N_{T_{0}}=\frac{1-4^{n}}{1-4} T_{0}^{2} C+N_{T_{0}} \leq\left(2^{n} T_{0}\right)^{2} C+N_{T_{0}}$.
This shows that for $T$ large,

$$
N_{T} \leq 2 C T^{2} .
$$

Unfortunately our $\tilde{f}$ is not bounded. Nevertheless we still have
Theorem 1.2. There exists a constant $C=C(f)>0$ such that

$$
\int \tilde{f}\left(\mathbf{a}_{t} k g_{0} \cdot \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k) \leq C
$$

for all $t>0$.
By arguments outlined above and Thm.1.2 we get
Theorem 1.3. There exists a constant $C>0$ such that $N_{T} \leq C T^{2}$ for $T$ sufficiently large.
1.2. The exact upper/lower bound. Equipped with Thm.1.3, let us revisit Equa.(48):

$$
\begin{align*}
\frac{N_{2 T}-N_{T}}{T^{2}} & \leq T^{-2} \sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T)} h\left(\frac{\|\mathbf{v}\|}{T}, Q_{0}(\mathbf{v})\right) \\
& =T^{-2} \sum_{\mathbf{v} \in g_{0}, \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T)} \frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, Q_{0}(\mathbf{v})\right)\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right| . \tag{50}
\end{align*}
$$

Fix an $\varepsilon>0$, the range of $T$ such that Lem.1.1 is not applicable is bounded. Thus

$$
\begin{equation*}
\frac{N_{2 T}-N_{T}}{T^{2}} \leq T^{-2} \sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T)}\left(\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{K}(k)+O(\varepsilon)\right)+O_{\varepsilon}\left(T^{-2}\right) . \tag{51}
\end{equation*}
$$

By Thm.1.3, the number of indices is bounded by $C(2 T)^{2}$, hence

$$
\begin{align*}
\frac{N_{2 T}-N_{T}}{T^{2}} & \leq T^{-2}\left(\sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}, \mathbf{v} \in V_{a, b}(\mathbb{Z}, 2 T)} \frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{K}(k)\right)+O_{\varepsilon}\left(T^{-2}\right)+O(\varepsilon) \\
& \leq \frac{1}{2 C_{4}}\left(\sum_{\mathbf{v} \in g_{0} \cdot \mathbb{Z}^{4}} \int f\left(\mathbf{a}_{t} k \cdot \mathbf{v}\right) \widehat{\mathrm{m}}_{K}(k)\right)+O_{\varepsilon}\left(T^{-2}\right)+O(\varepsilon)  \tag{52}\\
& =\frac{1}{2 C_{4}} \int \tilde{f}\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)+O_{\varepsilon}\left(T^{-2}\right)+O(\varepsilon)
\end{align*}
$$

Hence (let $\varepsilon \rightarrow 0$ after taking the limit $\lim _{T}$ )

$$
\limsup _{T \rightarrow+\infty} \frac{N_{2 T}-N_{T}}{T^{2}} \leq \lim _{t \rightarrow+\infty} \int \frac{1}{2 C_{4}} \widetilde{f}\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k) .
$$

That the RHS is a true limit is justified below.
The exact lower bound is proved similarly.
Theorem 1.4. Assume $Q_{0} \circ g_{0}$ is not rational, then for every $f \in C_{C}\left(\mathbb{R}^{4}\right)$,

$$
\lim _{t \rightarrow+\infty} \int \widetilde{f}\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)=\int_{\mathrm{X}_{4}} \tilde{f}(x) \widehat{\mathrm{m}}_{\mathrm{X}_{4}}(x)=C_{6} \int_{\mathbb{R}^{4}} f(\mathbf{v}) \mathrm{dv}
$$

where $C_{6}>0$ depending only on the dimension.
Let us evaluate $\int_{\mathbb{R}^{4}} f(\mathbf{v})$ dv for our $f$. By change of variables $y^{\prime}=: \frac{y-w_{2}^{2}-w_{3}^{2}}{2 x}$,
$\int f(\mathbf{v}) \mathrm{dv}=\int f\left(x, w_{2}, w_{3}, y^{\prime}\right) \mathrm{dxdy}^{\prime} \mathrm{dw}_{2} \mathrm{dw}_{3}=\int \frac{1}{2 x} f\left(x, w_{2}, w_{3}, \frac{y-w_{2}^{2}-w_{3}^{2}}{2 x}\right) \mathrm{dxdydw}_{2} \mathrm{dw}_{3}$.
where we have used

$$
\mathrm{dy}^{\prime}=\frac{\mathrm{dy}-2 w_{2} \mathrm{dw}_{2}-2 w_{3} \mathrm{dw}_{3}}{2 x}-\frac{\mathrm{dx}}{2 x^{2}}\left(y-w_{2}^{2}-w_{3}^{2}\right)
$$

Recall Equa.(47), we have

$$
\int f(\mathbf{v}) \mathrm{dv}=\int \frac{x}{2} h(x, y) \mathrm{dx} \mathrm{dy} .
$$

As $h(x, y)$ approximates $1_{\square}$ we get

$$
\int \frac{x}{2} h(x, y) \mathrm{dxdy} \rightarrow \int_{y=a}^{b} \int_{x=1}^{2} \frac{x}{2} \mathrm{dxdy}=\frac{2^{2}-1}{4}(b-a) .
$$

Thus, by collecting the constants $C_{7}:=\frac{1}{2 C_{4}} C_{6} \frac{2^{2}-1}{4}$,

$$
\lim _{T \rightarrow+\infty} \frac{N_{2 T}-N_{T}}{T^{2}}=C_{7}(b-a) .
$$

Now a geometric series argument shows that
Corollary 1.5.

$$
\lim _{T \rightarrow+\infty} \frac{N_{T}}{T^{2}}=\frac{1}{3} C_{7}(b-a) .
$$

## 2. Proof of the Lemma

### 2.1. Nontrivial contribution to the integral.

Definition 2.1. For $(x, y, z) \in \mathbb{R}^{3}$ with $x \neq 0$ and $a \in \mathbb{R}$, we let

$$
\phi_{a}(x, y, z):=\frac{a-y^{2}-z^{2}}{2 x}
$$

in other words, $\phi_{a}(x, y, z)$ is the unique real number such that

$$
Q_{0}\left(x, y, z, \phi_{a}(x, y, z)\right)=a
$$

DEFINITION 2.2. Given $f \in C_{C}\left(\mathbb{R}_{>0} \times \mathbb{R} \times \mathbb{R} \times \mathbb{R}\right)$, we fix $C_{1}=C_{1}(f)>1$ such that

$$
\operatorname{Supp}(f) \subset\left(C_{1}^{-1}, C_{1}\right) \times\left(-C_{1}, C_{1}\right)^{3}
$$

Also fix $C_{2}>\left|a_{0}\right|,\left|b_{0}\right|$.
The following two directly follow from the definition.
LEMMA 2.3. Let $\mathbf{v}_{\neq 0} \in \mathbb{R}^{4}$ and $T>1$. Let $\left(w_{2}, w_{3}\right) \in \mathbb{R}^{2}$ be such that

$$
f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, \phi_{Q_{0}(\mathbf{v})}\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}\right)\right) \neq 0
$$

then

1. $C_{1}^{-1} T \leq\|\mathbf{v}\| \leq C_{1} T$ and $\left|w_{2}\right|,\left|w_{3}\right| \leq C_{1}$;
2. $\left.\left\lvert\, \phi_{Q_{0}(\mathbf{v})} \frac{\|\mathbf{v}\|}{T}\right., w_{2}, w_{3}\right) \mid \leq C_{1}$;
3. $\left|Q_{0}(\mathbf{v})\right| \leq 4 C_{1}^{2}$.

For a vector $\mathbf{w}, \mathbf{w}(i) \in \mathbb{R}$ is defined by $\mathbf{w}=\sum \mathbf{w}(i) \mathbf{e}_{i}$.
Lemma 2.4. Let $\mathbf{v}_{\neq 0} \in \mathbb{R}^{4}$ and $T>1$. Let $\mathbf{w} \in K . v$. If $f\left(\mathbf{a}_{\ln T} . \mathbf{w}\right) \neq 0$, then

1. $C_{1}^{-1} T \leq \mathbf{w}(1) \leq C_{1} T,|\mathbf{w}(2)|,|\mathbf{w}(3)| \leq C_{1}$ and $|\mathbf{w}(4)| \leq C_{1} T^{-1}$;
2. $\|\mathbf{v}\| \leq 2 C_{1} T$;
3. $\|\mathbf{v}\| \geq C_{1}^{-1} T$;
4. $\left|Q_{0}(\mathbf{v})\right| \leq 4 C_{1}^{2}$.

Proof. For item 3, $\|\mathbf{v}\|=\|\mathbf{w}\| \geq \mathbf{w}(1) \geq C_{1}^{-1} T$.
For item $4, Q_{0}(\mathbf{v})=Q_{0}(\mathbf{w})=\mathbf{w}(1) \mathbf{w}(4)+\mathbf{w}(2)^{2}+\mathbf{w}(3)^{2} \leq 2 C_{1}^{2}+C_{1}^{2}+C_{1}^{2}=4 C_{1}^{2}$.
2.2. Representative in a $K$-orbit. By working with the basis $\mathbf{f}$, one sees that for every $\mathbf{v} \in$ $\mathbb{R}^{4}$, there exists $k_{\mathbf{v}} \in K$ such that

$$
k_{\mathbf{v}} \cdot \mathbf{v}=\left(u_{1}, 0,0, u_{4}\right)_{\mathbf{f}} \text { for some } u_{1}, u_{4} \geq 0
$$

Indeed, if we set

$$
r_{1}(\mathbf{v}):=\frac{\|\mathbf{v}\|+Q_{0}(\mathbf{v})}{2}, r_{2}(\mathbf{v}):=\frac{\|\mathbf{v}\|-Q_{0}(\mathbf{v})}{2}
$$

or equivalently,

$$
r_{1}(\mathbf{v}):=\mathbf{v}_{\mathbf{f}}(1)^{2}+\mathbf{v}_{\mathbf{f}}(2)^{2}+\mathbf{v}_{\mathbf{f}}(3)^{2}, r_{2}(\mathbf{v}):=\mathbf{v}_{\mathbf{f}}(4)^{2}
$$

where we assume $\mathbf{v}=\left(\mathbf{v}_{\mathbf{f}}(1), \ldots, \mathbf{v}_{\mathbf{f}}(4)\right)_{\mathbf{f}}$. Then there exists $k \in K$ such that

$$
k . \mathbf{v}=\left(\sqrt{r_{1}}, 0,0, \sqrt{r_{2}}\right)_{\mathbf{f}}=: \mathbf{v}^{*}
$$

To summarize the discussion in the basis $\mathbf{e}$ :
LEMMA 2.5. For every $\mathbf{v} \in R^{4}$ there exists a unique $\mathbf{v}^{*} \in \mathbb{R}^{4}$ satisfying

1. $Q_{0}\left(\mathbf{v}^{*}\right)=Q_{0}(\mathbf{v})$;
2. $\left\|\mathbf{v}^{*}\right\|=\|\mathbf{v}\|$;
3. $\mathbf{v}^{*}(1) \geq\left|\mathbf{v}^{*}(4)\right|$ and $\mathbf{v}^{*}(2)=\mathbf{v}^{*}(3)=0$.

Also $\mathbf{v}^{*} \in K . \mathbf{v}$.
What we are going to need is the following slightly perturbed version.
Lemma 2.6. Let $\mathbf{v} \in \mathbb{R}^{4}$ and $\left(w_{2}, w_{3}\right) \in \mathbb{R}$ satisfying $\left|w_{2}\right|,\left|w_{3}\right| \leq C_{1}$. Assume $\|\mathbf{v}\|^{2} \geq Q_{0}(\mathbf{v})+$ $4 C_{1}^{2}$. Then there exists a unique $\mathbf{v}^{*}\left(w_{2}, w_{3}\right)=\mathbf{w} \in \mathbb{R}^{4}$ such that

1. $Q_{0}(\mathbf{w})=Q_{0}(\mathbf{v})$;
2. $\|\mathbf{w}\|=\|\mathbf{v}\|$;
3. $\mathbf{w}(1) \geq|\mathbf{w}(4)|$ and $\mathbf{w}(2)=w_{2}, \mathbf{w}(3)=w_{3}$.

Also $\mathbf{w} \in K . \mathbf{v}$.
SKetch of proof. Indeed under the assumption above

$$
\left|\frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2}\right| \leq \frac{1}{2}\left(Q_{0}(\mathbf{v})+2 C_{1}^{2}\right)
$$

and

$$
\|\mathbf{v}\|^{2}-w_{2}^{2}-w_{3}^{2} \geq Q_{0}(\mathbf{v})+4 C_{1}^{2}-C_{1}^{2}-C_{1}^{2}=Q_{0}(\mathbf{v})+2 C_{1}^{2} .
$$

Hence the equation

$$
\left\{\begin{array}{l}
x y=\frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2} \\
x^{2}+y^{2}=\|\mathbf{v}\|^{2}-w_{2}^{2}-w_{3}^{2}
\end{array}\right.
$$

admits a unique solution with $x \geq|y|$.
Here is a picture $\left(x=w_{1}, y=w_{4}\right)$


### 2.3. Approximates $I$, the points.

Lemma 2.7. Assumption as in Lem.2.3. Further assume $T \geq 8 C_{1}^{3}$ and $T^{2} \geq 16 C_{1}^{4}$. Define $\mathbf{w}=\mathbf{v}^{*}\left(w_{2}, w_{3}\right)$ as in Lem.2.6. Then for $C_{3}=46 C_{1}^{7}$,

$$
\operatorname{dist}_{\infty}\left(\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, \phi_{Q_{0}(\mathbf{v})}\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}\right)\right), \mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \leq C_{3} T^{-2} .
$$

Note that $T \geq 8 C_{1}^{3} \Longrightarrow\|\mathbf{v}\| \geq 4 C_{1}^{2}+4 C_{1}^{2} \geq Q_{0}(\mathbf{v})+4 C_{1}^{2}$ by Lem.2.3. Thus Lem.2.6 is applicable.

Proof. First we have

$$
|\mathbf{w}(4)|^{2} \leq|\mathbf{w}(1)||\mathbf{w}(4)|=\left|Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}\right| \leq 4 C_{1}^{2}+2 C_{1}^{2}=6 C_{1}^{2}
$$

Hence the difference of the first coordinate:

$$
\begin{aligned}
& \left|\|\mathbf{v}\|^{2}-\mathbf{w}(1)^{2}\right|=\mathbf{w}(2)^{2}+\mathbf{w}(3)^{2}+\mathbf{w}(4)^{2} \leq 8 C_{1}^{2} \\
\Rightarrow & \left|T^{-1}\|\mathbf{v}\|-T^{-1} \mathbf{w}(1)\right| \leq T^{-1} \frac{8 C_{1}^{2}}{\|\mathbf{v}\|+\mathbf{w}(1)} \leq T^{-1} \frac{8 C_{1}^{2}}{\|\mathbf{v}\|} \leq 8 C_{1}^{3} T^{-2} \leq C_{3} T^{-2} .
\end{aligned}
$$

From here we also see that

$$
|\mathbf{w}(1)| \geq\|\mathbf{v}\|-8 C_{1}^{3} T^{-1} \geq \frac{1}{2} C_{1}^{-1} T+\left(\frac{1}{2} C_{1}^{-1} T-8 C_{1}^{3} T^{-1}\right) \geq \frac{1}{2} C_{1}^{-1} T .
$$

Here we are using the assumption $T^{2} \geq 16 C_{1}^{4} \Longrightarrow \frac{1}{2} C_{1}^{-1} T-8 C_{1}^{3} T^{-1} \geq 0$.
Now the difference of the last coordinate (note that $w_{2}=\mathbf{w}(2)$ and $w_{3}=\mathbf{w}(3)$ from Lem.2.6)

$$
\begin{aligned}
& \left|\frac{Q_{0}(\mathbf{v})-\mathbf{w}(2)^{2}-\mathbf{w}(3)^{2}}{2\|\mathbf{v}\| T^{-1}}-\frac{Q_{0}(\mathbf{v})-\mathbf{w}(2)^{2}-\mathbf{w}(3)^{2}}{2 \mathbf{w}(1) T^{-1}}\right| \\
\leq & \frac{1}{2}\left(6 C_{1}^{2}\right) T\left|\frac{1}{\|\mathbf{v}\|}-\frac{1}{\mathbf{w}(1)}\right|=\frac{\left(6 C_{1}^{2}\right) T}{2} \frac{\|\mathbf{v}\|-\mathbf{w}(1) \mid}{\|\mathbf{v}\| \mathbf{w}(1)} \\
\leq & \frac{\left(6 C_{1}^{2}\right) T}{2} \frac{8 C_{1}^{3} T^{-1}}{1 / 2 C_{1}^{-2} T^{2}}=48 C_{1}^{7} T^{-2} \leq C_{3} T^{-2} .
\end{aligned}
$$

Lemma 2.8. Assumption as in Lem.2.4. Define $w_{2}:=\mathbf{w}(2)$ and $w_{3}:=\mathbf{w}(3)$. Then for $C_{3}=$ $48 C_{1}^{7}$,

$$
\operatorname{dist}_{\infty}\left(\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, \phi_{Q_{0}(\mathbf{v})}\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}\right)\right), \mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \leq C_{3} T^{-2} .
$$

Proof. The difference of the first coordinate:

$$
\begin{gathered}
\left|\|\mathbf{v}\|^{2}-\mathbf{w}(1)^{2}\right|=\mathbf{w}(2)^{2}+\mathbf{w}(3)^{2}+\mathbf{w}(4)^{2} \leq 3 C_{1}^{2} \\
\Rightarrow \\
\left|T^{-1}\|\mathbf{v}\|-T^{-1} \mathbf{w}(1)\right| \leq T^{-1} \frac{3 C_{1}^{2}}{\mathbf{w}(1)} \leq 3 C_{1}^{3} T^{-2} \leq C_{3} T^{-2} .
\end{gathered}
$$

And the difference of the last coordinate

$$
\begin{aligned}
& \left|\frac{Q_{0}(\mathbf{v})-\mathbf{w}(2)^{2}-\mathbf{w}(3)^{2}}{2\|\mathbf{v}\| T^{-1}}-\frac{Q_{0}(\mathbf{v})-\mathbf{w}(2)^{2}-\mathbf{w}(3)^{2}}{2 \mathbf{w}(1) T^{-1}}\right| \\
\leq & \frac{T}{2}\left(6 C_{1}^{2}\right)\left|\frac{1}{\|\mathbf{v}\|}-\frac{1}{\mathbf{w}(1)}\right|=\frac{\left(6 C_{1}^{2}\right) T}{2} \frac{|\|\mathbf{v}\|-\mathbf{w}(1)|}{\|\mathbf{v}\| \mathbf{w}(1)} \\
\leq & \frac{\left(6 C_{1}^{2}\right) T}{2} \frac{3 C_{1}^{3} T^{-1}}{C_{1}^{-2} T^{2}}=9 C_{1}^{7} T^{-2} \leq C_{3} T^{-2} .
\end{aligned}
$$

2.4. Approximates II, the measures. Let $S(r)$ be the sphere of radius $r$ in $\mathbb{R}^{3}$ centered at the origin. Let $\widehat{\mathrm{m}}_{S(r)}$ be the normalized (to be a probability measure) volume measure on $S(r)$.

Assume $r_{1}(\mathbf{v}) \geq 2 C_{1}^{2}$. For $\left(x_{2}, x_{3}\right) \in \mathbb{R}^{2}$ with $\left|x_{2}\right|,\left|x_{3}\right| \leq C_{1}$, there exists a unique $x_{1}>0$ such that

$$
x_{1}^{2}+x_{2}^{2}+x_{3}^{2}=r_{1}(\mathbf{v})
$$

Let $\mathrm{D}\left(C_{1}\right)$ be the image of $\left\{\left(x_{2}, x_{3}\right),\left|x_{2}\right|,\left|x_{3}\right| \leq C_{1}\right\}$ in $S\left(\sqrt{r_{1}}\right)$ thus defined. And identify $\left|\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}\right| \mid{\left|x x_{i}\right| \leq C_{1}}$ as a measure on $\mathrm{D}\left(C_{1}\right) \subset S\left(\sqrt{r_{1}}\right)$ by this. Equivalently one may first restrict the differential form $\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}$ to $\mathrm{D}\left(C_{1}\right)$ and then take the measure associated with it.

Lemma 2.9. Let $\mathbf{v} \in \mathbb{R}^{4}$ be satisfying $\frac{\|\mathbf{v}\|^{2}}{2} \geq\left|Q_{0}(\mathbf{v})\right|$ and $\|\mathbf{v}\|^{2} \geq 16 C_{1}^{2}$.

$$
\left\|\|\mathbf{v}\|^{2} \widehat{\mathrm{~m}}_{S\left(\sqrt{r_{1}}\right)}-2 C_{4}\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right|\right\|_{\mathrm{D}\left(C_{1}\right)} \leq \frac{1}{\|\mathbf{v}\|^{2}}\left(C_{5} Q_{0}(\mathbf{v})+C_{5}\right)
$$

where $C_{4}>0$ is a constant depending only on the dimension and $C_{5}>1$ depends on $C_{1}$. See Equa.(53), (54) below.


Note that our assumption implies that $r_{1}(\mathbf{v})=1 / 2\left(\|\mathbf{v}\|^{2}+Q_{0}(\mathbf{v})\right) \geq 4 C_{1}^{2}$. Thus the paragraph above the proposition makes sense.

Proof. First let us write $\widehat{\mathrm{m}}_{S\left(\sqrt{r_{1}}\right)}$ in terms of differential forms. By taking the differential

$$
x_{1}^{2}+x_{2}^{2}+x_{3}^{2}=r^{2} \Longrightarrow 2 x_{1} \mathrm{dx}_{1}+2 x_{2} \mathrm{dx}_{2}+2 x_{3} \mathrm{dx}_{3}=2 r \mathrm{dr} .
$$

Thus

$$
\mathrm{dx}_{1} \wedge \mathrm{dx}_{2} \wedge \mathrm{dx}_{3}=\frac{r \mathrm{dx}_{2} \wedge \mathrm{dx}_{3}}{x_{1}} \wedge \mathrm{dr}
$$

So up to constant (depending possibly on $r$ ), the spherical measure can be induced from $\frac{r \mathrm{dx}_{2} \wedge \mathrm{dx}_{3}}{x_{1}}$. To make it have total mass independent of $r$, we consider

$$
\mathrm{dx}_{1} \wedge \mathrm{dx}_{2} \wedge \mathrm{dx}_{3}=\frac{\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}}{r x_{1}} \wedge r^{2} \mathrm{dr}
$$

Since the volume of ball of radius $R$ is some constant multiple of $R^{3} / 3=\int_{0}^{R} r^{2} \mathrm{dr}$, there exists some constant $C_{4}>0$ depending only on the dimension such that

$$
\begin{equation*}
\widehat{\mathrm{m}}_{S\left(\sqrt{r_{1}}\right)}=C_{4} \frac{\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}}{\sqrt{r_{1}} x_{1}} . \tag{53}
\end{equation*}
$$

By assumption,

$$
2 r_{1}=\|\mathbf{v}\|^{2}+Q_{0}(\mathbf{v}) \geq\|\mathbf{v}\|^{2}-\left|Q_{0}(\mathbf{v})\right| \geq \frac{1}{2}\|\mathbf{v}\|^{2} \Longrightarrow r_{1} \geq 4 C_{1}^{2}
$$

Thus for $\left(x_{1}, x_{2}, x_{3}\right) \in S\left(\sqrt{r_{1}}\right)$,

$$
2 \sqrt{r_{1}} x_{1}=2 \sqrt{r_{1}} \sqrt{r_{1}-x_{2}^{2}-x_{3}^{2}} \geq\|\mathbf{v}\| \sqrt{r_{1}-2 C_{1}^{2}} \geq\|\mathbf{v}\| \sqrt{\frac{1}{2} r_{1}} \geq \frac{\|\mathbf{v}\|^{2}}{8} .
$$

On the other hand

$$
\left|\|\mathbf{v}\|^{2}-2 r_{1}\right|=\left|Q_{0}(\mathbf{v})\right|
$$

and

$$
\left|2 r_{1}-2 \sqrt{r_{1}} x_{1}\right|=2 \sqrt{r_{1}}\left|\frac{r_{1}-\left(r_{1}-x_{2}^{2}-x_{3}^{2}\right)}{\sqrt{r_{1}}+\sqrt{r_{1}-x_{2}^{2}-x_{3}^{2}}}\right| \leq 2\left|x_{2}^{2}+x_{3}^{2}\right| \leq 4 C_{1}^{2} .
$$

Therefore, when restricted to $\mathrm{D}\left(C_{1}\right)$, we have

$$
\begin{aligned}
\left|\|\mathbf{v}\|^{2} \widehat{\mathrm{~m}}_{S\left(\sqrt{r_{1}}\right)}-2 C_{4}\right| \mathrm{dx}_{2} \wedge \mathrm{dx}_{3}| | & =2 C_{4}\left|\frac{\|\mathbf{v}\|^{2}}{2 \sqrt{r_{1}} x_{1}}-1\right|\left|\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}\right| \\
& =2 C_{4}\left|\frac{\|\mathbf{v}\|^{2}-2 \sqrt{r_{1}} x_{1}}{2 \sqrt{r_{1}} x_{1}}\right|\left|\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}\right| \\
& \leq 2 C_{4}\left|\frac{\left|Q_{0}(\mathbf{v})\right|+4 C_{1}^{2}}{\frac{1}{4}\|\mathbf{v}\|^{2}}\right|\left|\mathrm{dx}_{2} \wedge \mathrm{dx}_{3}\right|
\end{aligned}
$$

Thus if integrating a function taking value in $[-M, M]$, the difference is at most

$$
2 C_{4}\left|\frac{Q_{0}(\mathbf{v})+4 C_{1}^{2}}{\frac{1}{8}\|\mathbf{v}\|^{2}}\right|\left(2 C_{1}\right)^{2} \cdot M=\|\mathbf{v}\|^{-2} \cdot\left|64 C_{4} C_{1}^{2}\left(\left|Q_{0}(\mathbf{v})\right|+4 C_{1}^{2}\right)\right| \cdot M
$$

Taking

$$
\begin{equation*}
C_{5}:=256 C_{4} C_{1}^{4} \tag{54}
\end{equation*}
$$

completes the proof.
2.5. Proof. Fix $v \in \mathbb{R}^{4}$, we identify $S(r)$ with a subset of $\mathbb{R}^{4}$ by embedding

$$
\left(x_{1}, x_{2}, x_{3}\right) \mapsto\left(x_{1}, x_{2}, x_{3}, \sqrt{r_{2}(\mathbf{v})}\right)_{\mathbf{f}}
$$

Let us state Lem.1.1 again:
LEMmA 2.10. Given $f \in C_{c}\left(\mathbb{R}_{>0} \times \mathbb{R}^{3}\right)$ and $\varepsilon \in(0,1)$, there exists $T_{0}=T_{0}(f, \varepsilon)>0$ such that for every $T>T_{0}$, for every $\mathbf{v} \in \mathbb{R}^{4}$ we have

$$
\left|\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{\ln T} k . \mathbf{v}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)-\frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}| |<\varepsilon
$$

where

$$
w_{4}:=\frac{Q_{0}(\mathbf{v})-w_{2}^{2}-w_{3}^{2}}{2\|\mathbf{v}\| T^{-1}}
$$

is a function in $\left(w_{2}, w_{3}\right)$, for every fixed $\mathbf{v}$ and $T$.
Proof. We are going to choose some $T_{0} \geq 10 C_{1}^{3}$.
Rewrite

$$
\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{\ln T} k . \mathbf{v}\right) \widehat{\mathrm{m}}_{\mathrm{K}}(k)=\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \widehat{\mathrm{m}}_{\mathrm{K} . \mathbf{v}}(\mathbf{w})
$$

By Lem.2.4, 2.6, if $T \geq T_{0}$, by change of variable $\mathbf{w} \mapsto\left(w_{2}, w_{3}\right):=(\mathbf{w}(2), \mathbf{w}(3))$ :

$$
\begin{aligned}
\frac{1}{2 C_{4}} T^{2} \int f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \widehat{\mathrm{m}}_{\mathrm{K} \cdot \mathbf{v}}(\mathbf{w}) & =\frac{1}{2 C_{4}} T^{2} \int_{f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \neq 0} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \widehat{\mathrm{m}}_{\mathrm{K} \cdot \mathbf{v}}(\mathbf{w}) \\
& =\frac{1}{2 C_{4}} T^{2} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{v}^{*}\left(w_{2}, w_{3}\right)\right) \widehat{\mathrm{m}}_{S\left(\sqrt{r_{1}}\right)}\left(w_{2}, w_{3}\right)
\end{aligned}
$$

Note that when $f\left(\mathbf{a}_{\ln T} k . \mathbf{v}\right) \neq 0$ for some $k \in K, T \geq 10 C_{1}^{3} \Longrightarrow\|\mathbf{v}\|^{2} \geq Q_{0}(\mathbf{v})+4 C_{1}^{2}$ by Lem.2.4. So Lem.2.6 is applicable to $\mathbf{v}$ and $\left(w_{2}, w_{3}\right):=(\mathbf{w}(2), \mathbf{w}(3))$. Moreover, Lem.2.6 implies that $\mathbf{w}=$ $\mathbf{v}^{*}(\mathbf{w}(2), \mathbf{w}(3))$.

By Lem.2.3, the RHS is equal to

$$
\frac{T^{2}}{\|\mathbf{v}\|^{2}} \int f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right|=\frac{T^{2}}{\|\mathbf{v}\|^{2}} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\left|\mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\right|
$$

Recall from Lem.2.3 and 2.4 that when $f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \neq 0$ or when $f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right) \neq 0$, we always have

$$
\frac{1}{C_{1}} T \leq\|\mathbf{v}\| \leq 2 C_{1} T
$$

and

$$
\begin{equation*}
\left|Q_{0}(\mathbf{v})\right| \leq 4 C_{1}^{2} \tag{55}
\end{equation*}
$$

Now it suffices to show that

$$
\left|\|\mathbf{v}\|^{2} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \widehat{\mathrm{m}}_{S\left(\sqrt{r_{1}}\right)}\left(w_{2}, w_{3}\right)-2 C_{4} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}| |<\varepsilon .
$$

By Lem.2.7 and 2.8, for $T$ large enough,

$$
\left|2 C_{4} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}\left|-2 C_{4} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\frac{\|\mathbf{v}\|}{T}, w_{2}, w_{3}, w_{4}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}| |<0.5 \varepsilon
$$

By Lem.2.9 and Equa.(55), for $T$ large enough,

$$
\left|\|\mathbf{v}\|^{2} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right) \widehat{\mathrm{m}}_{S\left(\sqrt{\left.r_{1}\right)}\right.}\left(w_{2}, w_{3}\right)-2 C_{4} \int_{\mathrm{D}\left(C_{1}\right)} f\left(\mathbf{a}_{\ln T} \cdot \mathbf{w}\right)\right| \mathrm{dw}_{2} \wedge \mathrm{dw}_{3}| |<0.5 \varepsilon
$$

Combining these two, we are done.

## 3. Exercises

## CHAPTER 14

## Quantitative Oppenheim II, height function and nondivergence

Back to the Top.
Main reference: [EMM98].
If you are new to this circle of ideas, a first example to keep in mind maybe: $\mathbf{a}_{t}:=\operatorname{diag}\left(e^{t}, e^{-t}\right)$, $K:=\mathrm{SO}_{2}(\mathbb{R}), X=\mathrm{X}_{2}$. Most arguments are trivialized here, yet you could see the main idea.

## Notations

- $Q_{0}\left(x_{1}, x_{2}, x_{3}, x_{4}\right):=2 x_{1} x_{4}+x_{2}^{2}+x_{3}^{2}$ real quadratic form of signature $(3,1)$ on $\mathbb{R}^{4}$.
- Let $\left(\mathbf{e}_{1}, \ldots, \mathbf{e}_{4}\right)$ be the standard basis of $\mathbb{R}^{4}$; and for a vector $v$, define its coefficients by $v=\sum(\nu)_{i} \mathbf{e}_{i}$ and we also write $v=\left((\nu)_{1}, \ldots,(\nu)_{4}\right)$.
- Let $\left(\mathbf{f}_{1}, \ldots, \mathbf{f}_{4}\right)$ be another ONB(=orthogonal normal basis) defined by $\mathbf{f}_{2}=\mathbf{e}_{2}, \mathbf{f}_{3}=\mathbf{e}_{3}$ and $\mathbf{f}_{1}=\frac{\mathbf{e}_{1}+\mathbf{e}_{4}}{\sqrt{2}}, \mathbf{f}_{4}=\frac{\mathbf{e}_{1}-\mathbf{e}_{4}}{\sqrt{2}}$. If $\nu=\sum a_{i} \mathbf{f}_{i}$, we also write $v=\left(a_{1}, \ldots, a_{4}\right)$.
- One can verify that $Q_{0}\left(\left(x_{1}, \ldots, x_{4}\right)_{\mathbf{f}}\right)=x_{1}^{2}+x_{2}^{2}+x_{3}^{2}-x_{4}^{2}$.
- $\mathrm{K}:=\mathrm{SO}_{\mathrm{Q}_{0}}(\mathbb{R}) \cap \mathrm{SO}_{4}(\mathbb{R})$.
- $\mathbf{a}_{t}:=\operatorname{diag}\left(e^{-t}, 1,1, e^{t}\right)$, contained in $\mathrm{SO}_{Q_{0}}(\mathbb{R})$.


## 1. Outline of the proof

Recall by last lecture, it remains to show the following
THEOREM 1.1. Let $f$ be a compactly supported continuous function on $\mathbb{R}^{4}$ and let $\tilde{f}: \mathrm{X}_{4} \rightarrow$ $\mathbb{R}$ be its Siegel transform. Let $g_{0} \in G$ be such that $Q_{0} \circ g_{0}$ is irrational. Then

$$
\lim _{t \rightarrow+\infty} \int_{K} \widetilde{f}\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k)=\int \widetilde{f}(x) \widehat{\mathrm{m}}_{\mathrm{X}_{4}}(x)
$$

As we explained, the difficulty here is that $\tilde{f}$ is usually an integrable but unbounded function. And it suffices to show that the contribution of the part outside a large compact set is small. The following observation reduces the general task to a rather special function.

Definition 1.2. For a lattice $\Lambda \leq \mathbb{R}^{4}$, let

$$
\operatorname{ht}_{\infty}(\Lambda):=\max _{i=1, \ldots, 3} \sup _{\Delta \in \operatorname{Prim}^{i}(\Lambda)} \frac{1}{\|\Delta\|}=\max _{i=1, \ldots, 3}\left(\operatorname{sys}^{(i)}(\Lambda)\right)^{-1}
$$

Lemma 1.3. Let $f$ be a bounded, non-negative function with compact support on $\mathbb{R}^{4}$. Then there exists a constant $C_{1}=C_{1}(f)>1$ such that

$$
\tilde{f}(\Lambda) \leq C_{1} \cdot \mathrm{ht}_{\infty}(\Lambda), \forall \Lambda \in \mathrm{X}_{4}
$$

Proof is left as an exercise.
Theorem 1.4. For every $\varepsilon>0$, there exists a compact set $\mathscr{C}_{\varepsilon}$ of $\mathrm{X}_{4}$ such that for all $t>0$,

$$
\int\left(\mathrm{ht}_{\infty} \cdot \mathrm{l}_{\mathrm{X}_{4} \backslash \mathscr{E}_{\varepsilon}}\right)\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k) \leq \varepsilon
$$

Proof of Thm.1.1 assuming Thm.1.4. Without loss of generality assume $f \geq 0$.
Fix $\varepsilon>0$, choose $\mathscr{C}_{\varepsilon} \subset \mathrm{X}_{4}$ as in Thm.1.4. Choose a compactly supported continuous function $1 \geq \varphi_{\varepsilon} \geq 1_{\mathscr{C}_{\varepsilon}}$. Thus by equidistribution theorem obtained in Ch.12, Thm.1.2.

$$
\lim _{t \rightarrow+\infty} \int\left(\tilde{f} \cdot \varphi_{\varepsilon}\right)\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k)=\int\left(\tilde{f} \cdot \varphi_{\varepsilon}\right)(x) \widehat{\mathrm{m}}_{\mathrm{X}_{4}}(x) .
$$

On the other hand by Thm.1.4 and Lem.1.3

$$
\begin{aligned}
\limsup _{t \rightarrow+\infty} \int\left(\tilde{f} \cdot\left(1-\varphi_{\varepsilon}\right)\right)\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k) & \leq \limsup _{t \rightarrow+\infty} \int\left(C_{1} \mathrm{ht}_{\infty} \cdot 1_{\mathrm{X}_{4} \backslash \mathscr{E}_{\varepsilon}}\right)\left(\mathbf{a}_{t} k g_{0} \mathbb{Z}^{4}\right) \widehat{\mathrm{m}}_{K}(k) \\
& \leq C_{1} \varepsilon .
\end{aligned}
$$

Combining both and letting $\varepsilon \rightarrow 0$ we are done.

In fact, something stronger than Thm. 1.4 will be proved.
Proposition 1.5. For $\delta \in(0,1)$ (we only need for some $\delta>0$ ) and $\Lambda_{0} \in \mathrm{X}_{4}$, there exists $C_{2}=C_{2}\left(\delta, \Lambda_{0}\right)>0$ such that for all $t>0$

$$
\int \mathrm{ht}_{\infty}^{1+\delta}\left(\mathbf{a}_{t} k \cdot \Lambda_{0}\right) \widehat{\mathrm{m}}_{K}(k) \leq C_{2}
$$

This will be deduced from the following two propositions.
Proposition 1.6. For every $\varepsilon>0$, there exist $C_{4}(\varepsilon)>1$ and $t_{0}(\varepsilon)>0$ such that for all $\Lambda \in X_{4}$ (this is important!'), we have

$$
\int \mathrm{ht}_{\delta}^{\mathrm{new}}\left(\mathbf{a}_{t_{0}(\varepsilon)} k . \Lambda\right) \widehat{\mathrm{m}}_{K}(k) \leq \varepsilon \mathrm{ht}_{\delta}^{\mathrm{new}}(\Lambda)+C_{4}(\varepsilon)
$$

where $\mathrm{h} \mathrm{h}_{\delta}^{\text {new }}: \mathrm{X}_{4} \rightarrow \mathbb{R}_{>0}$ is some function satisfying

$$
C_{5}^{-1} \mathrm{ht}_{\infty}^{1+\delta} \leq \mathrm{ht}_{\delta}^{\text {new }} \leq C_{5} \mathrm{ht}_{\infty}^{1+\delta} .
$$

Actually, we will find constants $c_{0}>0$ and $\kappa_{i}>0$ for $i=0,1,2,3,4$ such that

$$
\left.\mathrm{ht}_{\delta}^{\mathrm{new}}(\Lambda)=\sum_{i=1,2,3} c_{0}^{\kappa_{i}} \operatorname{sys}^{(i)}(\Lambda)\right)^{-1-\delta} .
$$

To yield the result by applying this operator repeatedly, we need the following:
Proposition 1.7. For every open neighborhood $V$ of identity in $H$, there exists a neighborhood $U$ of identity in $K$ such that for all $t, s \geq 0$

$$
\mathbf{a}_{t} U \mathbf{a}_{s} \subset K \cdot V \cdot \mathbf{a}_{t+s} \cdot K
$$

Proof of Prop.1.5. From the description of $h t_{\delta}^{\text {new }}$ as in Prop.1.6, we can find $V_{0}$, an open neighborhood of identity in $H$, such that

$$
\frac{1}{2} \mathrm{ht}_{\delta}^{\mathrm{new}}(\Lambda) \leq \mathrm{ht}_{\delta}^{\mathrm{new}}(v . \Lambda) \leq 2 \mathrm{ht}_{\delta}^{\mathrm{new}}(\Lambda), \forall v \in V_{0}, \Lambda \in \mathrm{X}_{4} .
$$

Find $U_{0}$ by Prop.1.7. Let $\varepsilon:=\frac{1}{4} \widehat{\mathrm{~m}}_{K}\left(U_{0}\right)$. Applying Prop.1.6 we get some $C_{4}, t_{0}$. Let $C_{6}:=\frac{C_{4}}{\widehat{m}_{K}\left(U_{0}\right)}$.
Fix $\Lambda_{0} \in X_{4}$, define a continuous function $\phi: G \rightarrow \mathbb{R}_{>0}$ by

$$
\phi(g):=\int \mathrm{ht}_{\delta}^{\mathrm{new}}\left(g k . \Lambda_{0}\right) \widehat{\mathrm{m}}_{K}(k) .
$$

Thus it suffices to show that $\phi\left(\mathbf{a}_{t}\right)$, as $t$ varies in $(0,+\infty)$, is bounded by Prop. 1.6.
The function $\phi$ enjoys the following properties

1. $\phi$ is bi- $K$-invariant;
2. for every $v \in V_{0}$ and $g \in G, \frac{1}{2} \phi(g) \leq \phi(\nu g) \leq 2 \phi(g)$.

Combined with Prop.1.7, we see that for all $t \geq t_{0}$,

$$
\phi\left(\mathbf{a}_{t_{0}} k \mathbf{a}_{t-t_{0}}\right) \geq \frac{1}{2} \phi\left(\mathbf{a}_{t}\right) .
$$

Also observe that

$$
\begin{aligned}
\frac{1}{\widehat{\mathrm{~m}}_{K}\left(U_{0}\right)} \int_{U_{0}} \phi\left(\mathbf{a}_{t_{0}} k g\right) \widehat{\mathrm{m}}_{K}(k) & \leq \frac{1}{\widehat{\mathrm{~m}}_{K}\left(U_{0}\right)} \int_{K} \phi\left(\mathbf{a}_{t_{0}} k g\right) \widehat{\mathrm{m}}_{K}(k) \\
& \leq \frac{1}{\widehat{\mathrm{~m}}_{K}\left(U_{0}\right)} \cdot\left(\frac{1}{4} \widehat{\mathrm{~m}}_{K}\left(U_{0}\right) \phi(g)+C_{4}\right) \\
& =\frac{1}{4} \phi(g)+C_{6} .
\end{aligned}
$$

Therefore, for $t>t_{0}$,

$$
\begin{aligned}
\phi\left(\mathbf{a}_{t}\right) & =\frac{1}{\widehat{\mathrm{~m}}_{K}\left(U_{0}\right)} \int_{U_{0}} \phi\left(\mathbf{a}_{t}\right) \widehat{\mathrm{m}}_{K}(k) \\
& \leq 2 \frac{1}{\widehat{\mathrm{~m}}_{K}\left(U_{0}\right)} \int_{U_{0}} \phi\left(\mathbf{a}_{t_{0}} k \mathbf{a}_{t-t_{0}}\right) \widehat{\mathrm{m}}_{K}(k) \\
& \leq \frac{1}{2} \phi\left(\mathbf{a}_{t-t_{0}}\right)+C_{6} .
\end{aligned}
$$

Now, for $t>0$, choose the unique $n_{t} \in \mathbb{Z}_{\geq 0}$ such that $t^{\prime}:=t-n_{t} t_{0} \in\left(0, t_{0}\right]$. By applying the above inequality $n_{t}$ times we get

$$
\phi\left(\mathbf{a}_{t}\right) \leq \frac{1}{2^{n_{t}}} \phi\left(\mathbf{a}_{t^{\prime}}\right)+C_{6}\left(1+\frac{1}{2}+\left(\frac{1}{2}\right)^{2}+\ldots\right)
$$

Hence $\phi\left(\mathbf{a}_{t}\right)$, as $t$ varies in $(0,+\infty)$, is bounded.

## 2. Wavefront lemma

We explain how Prop.1.7 is proved.
Proof. I am pretending $K=\mathrm{SO}_{4}(\mathbb{R})$ here. The justification of the arguments here without this false assumption is left to you.

Every matrix $g$ of determinant one can be written as

$$
g=k_{1} d k_{2}, k_{i} \in \mathrm{SO}_{n}(\mathbb{R}), d \text { is a diagonal matrix. }
$$

The order of the diagonal entries of $d$ can be permuted by changing $k_{1}, k_{2}$. The middle matrix is uniquely determined if we further assume

$$
d=\operatorname{diag}\left(d_{1}, \ldots, d_{n}\right), \text { with } d_{1} \geq d_{2} \geq \ldots \geq d_{n}>0
$$

We let $\alpha_{i}(g):=d_{1} \cdot \ldots \cdot d_{i}$. It suffices to show that, when $k \in K$ is close to identity, for every $i$, $\alpha_{i}\left(\mathbf{a}_{t_{1}} k \mathbf{a}_{t_{2}}\right)$ is closed to $\alpha_{i}\left(\mathbf{a}_{t_{1}+t_{2}}\right)$ multiplicatively.

To do this, note that

$$
\alpha_{i}(g)=\sup _{\mathbf{v} \in \wedge^{i} \mathbb{R}^{n},\|\mathbf{v}\|=1}\|g \cdot \mathbf{v}\|=\sup _{\mathbf{v}, \mathbf{w} \in \wedge^{i} \mathbb{R}^{n},\|\mathbf{v}\|=\|\mathbf{w}\|=1}|\langle g \cdot \mathbf{v}, \mathbf{w}\rangle| .
$$

For $\varepsilon \in(0,1)$, choose $U=U(\varepsilon) \subset K$ such that for all $i$,

$$
\left|\left\langle u . e_{1} \wedge \ldots \wedge e_{i}, e_{1} \wedge \ldots \wedge e_{i}\right\rangle\right| \geq \frac{1}{1+\varepsilon} .
$$

Now take $u \in U$. On the one hand,

$$
\begin{aligned}
\left|\left\langle\mathbf{a}_{t_{1}} u \mathbf{a}_{t_{2}} \cdot \mathbf{v}, \mathbf{w}\right\rangle\right| & =\left|\left\langle u \mathbf{a}_{t_{2}} \cdot \mathbf{v}, \mathbf{a}_{t_{1}} \cdot \mathbf{w}\right\rangle\right| \\
& \leq\left\|\mathbf{a}_{t_{2}} \cdot \mathbf{v}\right\| \cdot\left\|\mathbf{a}_{t_{1}} \cdot \mathbf{w}\right\| \leq \alpha_{i}\left(\mathbf{a}_{t_{1}+t_{2}}\right) .
\end{aligned}
$$

On the other hand,

$$
\begin{aligned}
& \left|\left\langle\mathbf{a}_{t_{1}} u \mathbf{a}_{t_{2}} . e_{1} \wedge \ldots \wedge e_{i}, e_{1} \wedge \ldots \wedge e_{i}\right\rangle\right| \\
= & \alpha_{i}\left(\mathbf{a}_{t_{1}+t_{2}}\right)\left|\left\langle u . e_{1} \wedge \ldots \wedge e_{i}, e_{1} \wedge \ldots \wedge e_{i}\right\rangle\right| \geq \frac{1}{1+\varepsilon} \alpha_{i}\left(\mathbf{a}_{t_{1}+t_{2}}\right) .
\end{aligned}
$$

So we are done.

## 3. The height function

Prop.1.6 relies on the following proposition on representations. It is here that we are avoiding the case of signature $(2,1)$ and $(2,2)$.

Proposition 3.1. For every $\varepsilon>0$ there exists $t_{1}=t_{1}(\varepsilon)>0$ such that for all $t \geq t_{1}, \delta \in(0,1)$ and for all pure wedges $\mathbf{v}_{\neq 0} \in \wedge^{i} \mathbb{R}^{n}(n=4$ here $)$, we have

$$
\int\left\|\mathbf{a}_{t} k \cdot \mathbf{v}\right\|^{-1-\delta} \widehat{\mathrm{m}}_{K}(k) \leq \varepsilon\|\mathbf{v}\|^{-1-\delta}
$$

Proof. Omitted for now.
A "pure wedge" (also called "decomposable vector") refers to a vector $\mathbf{v} \in \wedge^{i} \mathbb{R}^{n}$ that can be written as $v_{1} \wedge \ldots \wedge v_{k}$ for some $v_{i} \in \mathbb{R}^{n}$.
3.1. Preparations. Fix $\varepsilon \in(0,1)$, find $t_{1}(\varepsilon)$ as in Prop.3.1. Find $C_{7}=C_{7}(\varepsilon)>1$ such that

$$
C_{7}^{-1}\|\mathbf{v}\| \leq\left\|\mathbf{a}_{t_{1}} \cdot \mathbf{v}\right\| \leq C_{7}\|\mathbf{v}\|, \forall \mathbf{v} \in \sqcup \wedge^{i} \mathbb{R}^{4}
$$

Fix a strictly convex function $\kappa>0$ on $[0,4]$. And find $C_{8}>1$ such that

$$
\kappa_{j} \geq \frac{\kappa_{j-i}+\kappa_{j+i}}{2}+C_{8}^{-1} ; \quad \kappa_{0}=\kappa_{4}=1
$$

for all $j \in\{1,2,3\}$ and $j \pm i \in\{0,1,2,3,4\}$.
Choose $c_{0} \in(0,1)$ small enough, depending on $\varepsilon$,

$$
c_{0}^{2 C_{8}^{-1}} \leq C_{7}^{2} c_{0}^{2 C_{8}^{-1}} \leq\left(\varepsilon C_{7}^{-1}\right)^{100}
$$

Define

$$
\begin{equation*}
\operatorname{ht}_{\delta}^{\text {new }}(\Lambda)=\sum_{i=1,2,3} c_{0}^{\kappa_{i}}\left(\operatorname{sys}^{(i)}(\Lambda)\right)^{-1-\delta} \tag{56}
\end{equation*}
$$

3.2. The proof. For each $l=1,2,3$ find $\Delta_{1}^{(l)} \in \operatorname{Prim}^{l}(\Lambda)$ such that $\operatorname{sys}^{(l)}(\Lambda)=\left\|\Delta_{1}^{(l)}\right\|$.
3.2.1. Good indices. We define $\operatorname{Good}(\Lambda) \subset\{1,2,3\}$ by

$$
\begin{equation*}
l \in \operatorname{Good}(\Lambda) \Longleftrightarrow \forall \Delta \in \operatorname{Prim}^{l}(\Lambda) \backslash \Delta_{1}^{(l)}, C_{7}^{2}\|\Delta\|^{-1}<\operatorname{sys}^{(l)}(\Lambda)^{-1} \tag{57}
\end{equation*}
$$

Thus for $l \in \operatorname{Good}(\Lambda), \Delta \in \operatorname{Prim}^{l}(\Lambda) \backslash \Delta_{1}^{(l)}$ and $k \in K$,

$$
\begin{align*}
& \left\|\mathbf{a}_{t_{1}} k \cdot \Delta\right\|^{-1-\delta} \leq C_{7}^{1+\delta}\|\Delta\|^{-1-\delta}<C_{7}^{-1-\delta} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta}=C_{7}^{-1-\delta}\left\|\Delta_{1}^{(l)}\right\|^{-1-\delta} \leq\left\|\mathbf{a}_{t_{1}} k \cdot \Delta_{1}^{(l)}\right\|^{-1-\delta} \\
& \Longrightarrow \forall k \in K, \operatorname{sys}^{(l)}\left(\mathbf{a}_{t_{1}} k \cdot \Lambda\right)^{-1-\delta}=\left\|\mathbf{a}_{t_{1}} k \cdot \Delta_{1}^{(l)}\right\|^{-1-\delta} \tag{58}
\end{align*}
$$

This implies that

$$
\begin{align*}
\int c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}\left(\mathbf{a}_{t_{1}} k . \Lambda\right)^{-1-\delta} \widehat{\mathrm{m}}_{K}(k) & =\int c_{0}^{\kappa_{l}}\left\|\mathbf{a}_{t_{1}} k \cdot \Delta_{1}^{(l)}\right\|^{-1-\delta} \widehat{\mathrm{m}}_{K}(k) \\
& \leq \varepsilon c_{0}^{\kappa_{l}}\left\|\mathbf{a}_{t_{1}} k . \Delta_{1}^{(l)}\right\|^{-1-\delta}  \tag{59}\\
& =\varepsilon \cdot c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}\left(\mathbf{a}_{t_{1}} k \cdot \Lambda\right)^{-1-\delta}
\end{align*}
$$

3.2.2. $\operatorname{Bad}$ indices. $\operatorname{Bad}(\Lambda):=\{1,2,3\} \backslash \operatorname{Good}(\Lambda)$. In other words, we can find $\Delta_{2}^{(l)} \in \operatorname{Prim}^{l}(\Lambda) \backslash$ $\Delta_{1}^{(l)}$ such that

$$
C_{7}^{2}\left\|\Delta_{2}^{(l)}\right\|^{-1} \geq \operatorname{sys}^{(l)}\left(\mathbf{a}_{t_{1}} k . \Lambda\right)^{-1}
$$

Recall the following inequalities

$$
\left\|\Delta_{1}^{(l)}\right\| \cdot\left\|\Delta_{2}^{(l)}\right\| \geq\left\|\Delta_{1}^{(l)} \cap \Delta_{2}^{(l)}\right\| \cdot\left\|\Delta_{1}^{(l)}+\Delta_{2}^{(l)}\right\|
$$

from which we deduce that (let $\left.a:=\operatorname{rank} \Delta_{1}^{(l)}-\operatorname{rank} \Delta_{1}^{(l)} \cap \Delta_{2}^{(l)}\right)$

$$
c_{0}^{2 \kappa_{l}}\left\|\Delta_{2}^{(l)}\right\|^{-1-\delta}\left\|\Delta_{2}^{(l)}\right\|^{-1-\delta} \leq\left(c_{0}^{\kappa_{l-a}}\left\|\Delta_{1}^{(l)} \cap \Delta_{2}^{(l)}\right\|^{-1-\delta}\right) \cdot\left(c_{0}^{\kappa_{l+a}}\left\|\Delta_{1}^{(l)}+\Delta_{2}^{(l)}\right\|^{-1-\delta}\right) \cdot c_{0}^{2 \kappa_{l}-\kappa_{l-a}-\kappa_{l+a}}
$$

For the LHS we have

$$
C_{7}^{-4}\left(c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta}\right)^{2} \leq C_{7}^{-2(1+\delta)}\left(c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta}\right)^{2} \leq \mathrm{LHS}
$$

and for the RHS,

$$
\operatorname{RHS} \leq\left(c_{0}^{\kappa_{l-a}} \operatorname{sys}^{(l-a)}(\Lambda)^{-1-\delta}\right) \cdot\left(c_{0}^{\kappa_{l+a}} \text { sys }^{(l+a)}(\Lambda)^{-1-\delta}\right) \cdot c_{0}^{C_{8}^{-1}}
$$

Since $c_{0}^{C_{8}^{-1}} \leq \varepsilon^{50} C_{7}^{-50}$, by combining the above equations we get

$$
\left(c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta}\right)^{2} \leq \varepsilon^{50} C_{7}^{-46}\left(c_{0}^{\kappa_{l-a}} \operatorname{sys}^{(l-a)}(\Lambda)^{-1-\delta}\right) \cdot\left(c_{0}^{\kappa_{l+a}} \operatorname{sys}^{(l+a)}(\Lambda)^{-1-\delta}\right)
$$

Thus

$$
c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta} \leq \varepsilon^{20} C_{7}^{-23} \max _{l^{\prime}=0, \ldots, 4}\left\{c_{0}^{\kappa_{l^{\prime}}} \operatorname{sys}^{\left(l^{\prime}\right)}(\Lambda)^{-1-\delta}\right\}
$$

Now we choose $l_{1}=l_{1}(\Lambda, \delta)$ such that the maximum of RHS is achieved. Then $l_{1} \in \operatorname{Good}(\Lambda) \cup$ $\{0,4\}$. Also take $l_{0} \in \operatorname{Bad}(\Lambda)$. Then for every $k \in K$,

$$
\begin{aligned}
c_{0}^{\kappa_{l_{0}}} \operatorname{sys}^{\left(l_{0}\right)}\left(\mathbf{a}_{t_{1}} k . \Lambda\right)^{-1-\delta} & \leq C_{7}^{1+\delta} c_{0}^{\kappa_{l_{0}}} \text { sys }^{\left(l_{0}\right)}(\Lambda)^{-1-\delta} \leq \varepsilon^{20} C_{7}^{-20} c_{0}^{{ }_{l_{l}}} \text { sys }^{\left(l_{1}\right)}(\Lambda)^{-1-\delta} \\
& \leq \varepsilon^{20} C_{7}^{-18} c_{0}^{\kappa_{l_{1}}} \operatorname{sys}^{\left(l_{1}\right)}\left(\mathbf{a}_{t_{1}} k . \Lambda\right)^{-1-\delta}
\end{aligned}
$$

3.2.3. Wrap-up. To save notation define

$$
\begin{aligned}
\alpha_{l}(\Lambda) & :=c_{0}^{\kappa_{l}} \operatorname{sys}^{(l)}(\Lambda)^{-1-\delta} \\
\pi_{*}\left(\alpha_{l}\right)(\Lambda) & :=\int \alpha_{l}\left(\mathbf{a}_{t_{1}} k . \Lambda\right) \widehat{\mathrm{m}}_{K}(k) .
\end{aligned}
$$

So for $l \in \operatorname{Good}(\Lambda)$, we have

$$
\pi_{*}\left(\alpha_{l}\right)(\Lambda) \leq \varepsilon \alpha_{l}(\Lambda)
$$

For $l \in \operatorname{Bad}(\Lambda)$, we have $\left(l_{1}=l_{1}(\Lambda)\right.$ as above $)$

$$
\pi_{*}\left(\alpha_{l}\right)(\Lambda) \leq \varepsilon^{20} C_{7}^{-18} \pi_{*}\left(\alpha_{l}\right)(\Lambda)
$$

There are two cases.
Case I, $l_{1} \in\{0, n\}$. In this case, for all $l, \alpha_{l}(\Lambda) \leq \max \left\{c_{0}^{\kappa_{0}}, c_{0}^{\kappa_{n}}\right\}=c_{0}$. Thus ht ${ }_{\delta}^{\text {new }}(\Lambda) \leq 3 c_{0}$. And

$$
\pi_{*}\left(\mathrm{ht}_{\delta}^{\mathrm{new}}\right)(\Lambda) \leq 3 c_{0} C_{7}^{2}
$$

Case II, $l_{1} \in \operatorname{Good}(\Lambda)$.

$$
\begin{aligned}
\pi_{*}\left(\mathrm{ht}_{\delta}^{\text {new }}\right)(\Lambda) & =\sum \pi_{*}\left(\alpha_{l}\right)(\Lambda) \\
& \leq \varepsilon \sum_{l \in \operatorname{Good}(\Lambda)} \alpha_{l}(\Lambda)+\varepsilon^{20} C_{7}^{-18} \pi_{*}\left(\alpha_{l_{1}}\right)(\Lambda) \\
& \leq \varepsilon \sum_{l \in \operatorname{Good}(\Lambda)} \alpha_{l}(\Lambda)+\varepsilon^{21} C_{7}^{-18} \alpha_{l_{1}}(\Lambda) \\
& \leq 2 \varepsilon \sum_{l \in \operatorname{Good}(\Lambda)} \alpha_{l}(\Lambda) \leq 2 \varepsilon \operatorname{ht}_{\delta}^{\text {new }}(\Lambda)
\end{aligned}
$$

In either case, the following holds

$$
\begin{equation*}
\pi_{*}\left(\mathrm{ht}_{\delta}^{\text {new }}\right)(\Lambda) \leq 3 c_{0} C_{7}^{2}+2 \varepsilon \mathrm{ht}_{\delta}^{\text {new }}(\Lambda) \tag{60}
\end{equation*}
$$

for all $\Lambda \in \mathrm{X}_{4}$. Recall $c_{0}$ and $C_{7}$ are only dependent on $\varepsilon$.

## 4. Exercises

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