Sébastien Gouëzel

# Central limit theorem and stable laws for intermittent maps 

Received: 10 December 2002 / Revised version: 6 September 2003 /
Published online: 4 November 2003 - © Springer-Verlag 2003


#### Abstract

In the setting of abstract Markov maps, we prove results concerning the convergence of renormalized Birkhoff sums to normal laws or stable laws. They apply to onedimensional maps with a neutral fixed point at 0 of the form $x+x^{1+\alpha}$, for $\alpha \in(0,1)$. In particular, for $\alpha>1 / 2$, we show that the Birkhoff sums of a Hölder observable $f$ converge to a normal law or a stable law, depending on whether $f(0)=0$ or $f(0) \neq 0$. The proof uses spectral techniques introduced by Sarig, and Wiener's Lemma in non-commutative Banach algebras.


## 1. Introduction and statement of results

### 1.1. Introduction

Recently, general methods have been devised to prove the Central Limit Theorem for the Birkhoff averages of mixing dynamical systems. More precisely, if $T: X \rightarrow X$ is a map from a space to itself preserving a probability measure $\nu$, and $f: X \rightarrow \mathbb{R}$ is a function with zero average, we say that the Central Limit Theorem (CLT) holds for ( $T, v, f$ ) if

$$
\frac{S_{n} f}{\sqrt{n}}=\frac{1}{\sqrt{n}} \sum_{i=0}^{n-1} f \circ T^{i} \rightarrow \mathcal{N}\left(0, \sigma^{2}\right)
$$

the convergence being in distribution (with respect to the measure $\nu$ ).
To prove this kind of results, the first step is often to show that the correlations $\operatorname{Cor}\left(f, f \circ T^{n}\right)=\int f \cdot f \circ T^{n} \mathrm{~d} v$ are summable. Then one may use this to construct a reverse martingale, for which the CLT is known to be true. Clearly, when the correlations are not summable, this method does not work.

When the map $T$ is uniformly expanding, perturbative spectral methods can also be used, giving convergence to normal laws ([GH88]) or other limit laws, namely stable laws ([GLJ93] or [AD01]). In the particular case of the continued

[^0]fraction transformation, Lévy had already obtained this kind of convergence using elementary methods ([Lév52]).

In this paper, we will be interested in cases where $T$ has a neutral fixed point (or in the Markov analogues of this situation), where the classical methods to estimate the decay of correlations fail (however, see [Hu]), and the spectral methods do not work directly, since the expansion is not uniform. We will prove that in some cases, the CLT still holds. Moreover, in situations where it does not hold, we will show the convergence to stable laws, even when the function $f$ is arbitrarily smooth, for example $C^{\infty}$ or analytic. This kind of result has been announced by Fisher and Lopes (see [FL01] and references therein) for piecewise linear maps.

Our main results (Theorems 1.1 and 1.2) will be stated in the context of abstract Markov maps. Applications to one-dimensional maps are then given: if a map has a neutral fixed point at 0 of the form $x+x^{1+\alpha}$ with $\alpha \in(1 / 2,1)$ and is expanding elsewhere, then the Birkhoff sums of a Hölder observable $f$ converge to a normal law or a stable law depending on whether $f(0)=0$ or $f(0) \neq 0$. Applications are also given for other types of neutral fixed points, and for unbounded observables, for example $x^{-\beta}$, leading to new results even in the case $\alpha \in(0,1 / 2)$ (see also [Rau02] in this case, where he studies the decay of correlations for non Hölder observables).

The philosophy of the method is that, if it is possible to induce on a subset $Y$ of the space so that the induced map is uniformly expanding, then the behavior of the Birkhoff sums $\sum_{0}^{n-1} f\left(T^{i} x\right)$ is in fact dictated by the behavior on $Y$ of the Birkhoff sums $\sum_{0}^{p-1} f_{Y}\left(T_{Y}^{i} x\right)$, where $T_{Y}=T^{\varphi}$ is the induced map on $Y$ ( $\varphi$ is the first return time), and $f_{Y}(x)=\sum_{0}^{\varphi(x)-1} f\left(T^{i} x\right)$. As convergence to normal laws or stable laws is known in the uniformly expanding case (see [AD01] for example), we get the same kind of results in our nonuniform setting.

### 1.2. The result for Markov maps

### 1.2.1. Definition of Markov maps

For these definitions, see for example [ADU93], [Aar97] or [Sar02].
A Markov map is a non-singular transformation $T$ of a Lebesgue space ( $X, \mathcal{B}, m$ ) of finite mass, together with a measurable partition $\alpha$ of $X$ such that if $a \in \alpha$ then $m(a)>0, T a$ is a union $(\bmod m)$ of elements of $\alpha$, and $T: a \rightarrow T a$ is invertible. Moreover, it is assumed that the completion of $\bigvee_{0}^{\infty} T^{-i} \alpha$ is $\mathcal{B}$, i.e. the partition separates the points.

A Markov map is said to be topologically mixing if $\forall a, b \in \alpha, \exists N, \forall n \geqslant$ $N, b \subset T^{n} a$. This corresponds to topological mixing for the topology defined by the cylinders $\left[a_{0}, \ldots, a_{n-1}\right]=\bigcap_{i=0}^{n-1} T^{-i} a_{i}\left(\right.$ where $\left.a_{0}, \ldots, a_{n-1} \in \alpha\right)$.

If $\gamma \subset \alpha$ is a subset of the partition $\alpha$, and $Y=\bigcup_{a \in \gamma} a$, the induced map $T_{Y}$ : $Y \rightarrow Y$ is defined as the first return map from $Y$ to $Y$, i.e. $T_{Y}=T^{\varphi_{Y}}$, where $\varphi_{Y}(x)=$ $\inf \left\{n \geqslant 1 \mid T^{n}(x) \in Y\right\}$ is the return time to $Y$. Let $\delta=\left\{\left[a, \xi_{1}, \ldots, \xi_{n-1}, Y\right] \mid a \in\right.$ $\left.\gamma, \xi_{1}, \ldots, \xi_{n-1} \in \alpha-\gamma,\left[a, \xi_{1}, \ldots, \xi_{n-1}, Y\right] \neq \emptyset\right\}$ : this is a partition of $Y$, for which $T_{Y}$ is a Markov map (with the measure $m_{Y}=m_{\mid Y}$ ). The cylinders for
this partition will be denoted by $\left[d_{0}, \ldots, d_{l-1}\right]_{Y}$ (with $d_{0}, \ldots, d_{l-1} \in \delta$ ). If $d=$ $\left[a, \xi_{1}, \ldots, \xi_{n-1}, Y\right] \in \delta$, its image is $T_{Y} d=T \xi_{n-1} \cap Y$-hence, it is $\gamma$-measurable.

For $x, y \in Y$, we will write $s(x, y)$ for the separation time of $x$ and $y$ under $T_{Y}$, i.e. $s(x, y)=\inf \left\{n \in \mathbb{N} \mid T_{Y}^{n} x\right.$ and $T_{Y}^{n} y$ are not in the same element of $\left.\delta\right\}$. This separation time is extended to the whole space $X$ : for $x, y \in X$, let $x^{\prime}$ and $y^{\prime}$ be their first returns to $Y$. If $T^{i} x$ and $T^{i} y$ stay in the same element of the partition $\alpha$ until the first return to $Y$, set $s(x, y)=s\left(x^{\prime}, y^{\prime}\right)+1$. Otherwise, $s(x, y)=0$.

If $\theta<1$, we will say that a function $f$ on $Y$ (resp. $X$ ) is locally $\theta$-Hölder if there exists a constant $C$ such that $\forall x, y \in Y$ (resp. $X$ ) with $s(x, y) \geqslant 1$, we have $|f(y)-f(x)| \leqslant C \theta^{s(x, y)}$. The smallest such constant is called the $\theta$-Hölder constant of $f$. In fact, slightly abusing notation, we shall say that a measurable function is locally $\theta$-Hölder if there exists a locally $\theta$-Hölder version of this function.

The transfer operator $\hat{T}_{Y}$ associated to $T_{Y}$ and acting on integrable functions is defined by $\int f \cdot\left(g \circ T_{Y}\right) \mathrm{d} m_{Y}=\int\left(\hat{T}_{Y} f\right) \cdot g \mathrm{~d} m_{Y}$. It can be written as $\hat{T}_{Y} f(x)=$ $\sum_{T_{Y} y=x} g_{m_{Y}}(y) f(y)$, where the weight $g_{m_{Y}}$ is defined on an element $d$ of $\delta$ by $g_{m_{Y}}=\frac{\mathrm{d} m_{Y}}{\mathrm{~d}\left(m_{Y} \circ\left(T_{Y}\right)_{d d}\right)}$, i.e. it is the inverse of the jacobian of $T_{Y}$. We will say that the distortion is locally Hölder if $\log g_{m_{Y}}$ is locally $\theta$-Hölder for some $\theta<1$.

Finally, we will say that $T_{Y}$ has the big image property if there exists a constant $\eta>0$ such that $\forall d \in \delta, m\left[T_{Y} d\right] \geqslant \eta$. This is in particular the case when the partition $\gamma$ is finite.

### 1.2.2. Main results

For the following theorems, $(X, \mathcal{B}, T, m, \alpha)$ will be a topologically mixing probability preserving Markov map. We assume that $\gamma \subset \alpha$ is such that the induced map $T_{Y}$ on $Y=\bigcup_{a \in \gamma} a$ has a locally $\theta$-Hölder distortion for some $\theta$, and that it has the big image property. The function $\varphi=\varphi_{Y}$ will denote the return time from $Y$ to itself.

We will write $I(X, Y)$ for the set of functions $f: X \rightarrow \mathbb{R}$ which induce "nice" functions on $Y$. More precisely, $f \in I(X, Y)$ if

1. $\forall N \in \mathbb{N}$, there exist constants $C_{N}$ and $0<\theta_{N}<1$ such that, on $Y_{N}=$ $\bigcup_{0}^{N-1} T^{i}(Y)$, we have $|f| \leqslant C_{N}$ and, if $x, y \in Y_{N}$ satisfy $s(x, y) \geqslant 1$, then $|f(x)-f(y)| \leqslant C_{N} \theta_{N}^{s(x, y)}$.
2. Writing $f_{Y}(x)=\sum_{i=0}^{\varphi(x)-1} f\left(T^{i} x\right)$ for the map induced by $f$ on $Y$, there exists $\theta<1$ such that $\sum_{d \in \delta} m[d] D_{\theta} f_{Y}(d)<\infty$, where $D_{\theta} f_{Y}(d)$ is the least $\theta$-Hölder constant of $f_{Y}$ restricted to the element $d$ of the partition $\delta$.
3. There exists $\theta<1$ and $0<\eta \leqslant 1$ such that $\sum_{d \in \delta} \varphi(d) m[d] D_{\theta} f_{Y}(d)^{\eta}<\infty$.

Note that, increasing $\theta$ if necessary, we can assume that the distortion of $T_{Y}$ is $\theta$ Hölder and that $\sum_{d \in \delta} m[d] D_{\theta} f_{Y}(d)<\infty$ and $\sum_{d \in \delta} \varphi(d) m[d] D_{\theta} f_{Y}(d)^{\eta}<\infty$, for the same $\theta<1$.

For example, if $f$ is bounded and locally $\theta$-Hölder, and the induced map $f_{Y}$ is also locally $\theta$-Hölder, then these conditions are satisfied, for $\eta=1$ (note that $\sum \varphi(d) m[d]=m(X)<\infty$, by Kac's Formula). The weaker conditions above are useful in the applications, for example to unbounded observables.

If $f$ is bounded and locally $\theta$-Hölder, and $m[\varphi=n]=O\left(1 / n^{\beta+1}\right)$ for some $\beta>1$, then $f \in I(X, Y)$ : the first condition is clear, the second one is given by Kac's Formula (since $D_{\theta} f_{Y}(d) \leqslant C \varphi(d)$ ), and the third one is satisfied for any $\eta<\min (1, \beta-1)$.

Theorem 1.1 (Central Limit Theorem). Let $f \in I(X, Y)$ be integrable with $\int f \mathrm{~d} m=0$ and let the induced function $f_{Y}$ be in $L^{2}\left(\mathrm{~d} m_{Y}\right)$.

Then there exists a constant $\sigma^{2} \geqslant 0$ such that $\frac{1}{\sqrt{n}} \sum_{i=0}^{n-1} f \circ T^{i}$ converges in distribution to a normal law $\mathcal{N}\left(0, \sigma^{2}\right)$. Moreover, $\sigma^{2}=0$ if and only if there exists a measurable function $\chi: X \rightarrow \mathbb{R}$ satisfying $f=\chi \circ T-\chi$.

For the next theorem, we will need the notion of stable law. The law of the random variable $X$ is said to be stable if there exist i.i.d. random variables $X_{k}$ and constants $A_{n} \in \mathbb{R}$ and $B_{n}>0$ such that

$$
\frac{\sum_{i=0}^{n-1} X_{i}-A_{n}}{B_{n}} \rightarrow X \text { in distribution. }
$$

In this case, we say that $X_{0}$ is attracted to $X$, or that the law of $X_{0}$ belongs to the domain of attraction of $X$. The stable laws are completely classified (see for example [Fel66]) and depend on many parameters, the most important of which is certainly the index $p \in(0,2]$. The case $p=2$ corresponds to the normal law, while if $p \in(1,2)$ the stable laws are in $L^{1}$ but not in $L^{2}$. In fact, for $p \in(0,1) \cup(1,2)$, the index can be characterized as follows: there exist constants $c_{1} \geqslant 0$ and $c_{2} \geqslant 0$ with $c_{1}+c_{2}>0$, satisfying $P[X>x]=\left(c_{1}+o(1)\right) x^{-p}$ and $P[X<-x]=$ $\left(c_{2}+o(1)\right) x^{-p}$. The case $p=1$ is problematic and will not be included in the following discussion.

Following [AD01], for $p \in(0,1) \cup(1,2], c>0$ and $\beta \in[-1,1]$, we will denote by $X_{p, c, \beta}$ the law whose characteristic function is

$$
E\left(e^{i t X_{p, c, \beta}}\right)=e^{-c|t|^{p}\left(1-i \beta \operatorname{sgn}(t) \tan \left(\frac{p \pi}{2}\right)\right)}
$$

This is indeed a stable law of index $p$. When $p \in(1,2]$, it is centered, i.e. $X_{p, c, \beta}$ is of zero expectation. Note that, when $p=2$, the value of $\beta$ is irrelevant.

We recall also that a function $L:(0, \infty) \rightarrow \mathbb{R}$ is slowly varying if $\forall x>$ $0, \lim _{y \rightarrow \infty} \frac{L(y x)}{L(y)}=1$. Then the random variables $Z$ attracted to stable laws of index $p$ can be described as follows ([Fel66]):

- For $p \in(0,1) \cup(1,2), P[Z>x]=\left(c_{1}+o(1)\right) x^{-p} L(x)$ and $P[Z<-x]=$ $\left(c_{2}+o(1)\right) x^{-p} L(x)$ for some $c_{1}, c_{2} \geqslant 0$ with $c_{1}+c_{2}>0$, and a slowly varying function $L$.
- For $p=2$, either $Z \in L^{2}$, or $P[Z>x]=\left(c_{1}+o(1)\right) x^{-2} l(x)$ and $P[Z<$ $-x]=\left(c_{2}+o(1)\right) x^{-2} l(x)$, for constants $c_{1}, c_{2} \geqslant 0$ with $c_{1}+c_{2}>0$, and a slowly varying function $l$ such that $u^{-1} l(u)$ is not integrable at $+\infty$. In this case, we will write $L(x)=\int_{-x}^{x} u^{2} \mathrm{~d} P_{Z}(u)$, which is an unbounded slowly varying function.

When $Z \notin L^{2}$, if the normalizing constants $B_{n}$ and $A_{n}$ are defined by

$$
n L\left(B_{n}\right)=B_{n}^{p}, \quad A_{n}= \begin{cases}0 & 0<p<1 \\ n E(Z) & 1<p \leqslant 2\end{cases}
$$

then, writing $Z_{0}, Z_{1}, \ldots$ for independent random variables having the same distribution as $Z$, we have the convergence

$$
\frac{\sum_{i=0}^{n-1} Z_{i}-A_{n}}{B_{n}} \rightarrow X_{p, c, \beta}
$$

where

$$
c=\left\{\begin{array}{ll}
\left(c_{1}+c_{2}\right) \Gamma(1-p) \cos \left(\frac{p \pi}{2}\right) & p \in(0,1) \cup(1,2) \\
\frac{1}{2} & p=2
\end{array}, \quad \beta=\frac{c_{1}-c_{2}}{c_{1}+c_{2}} .\right.
$$

In the deterministic case, if $f \in I(X, Y)$, we can consider the function $f_{Y}$ on $Y$. Since $m(Y)<1$, its distribution $\mu$ (defined by $\mu(A)=m\left\{y \in Y \mid f_{Y}(y) \in A\right\}$ ) is not a probability measure on $\mathbb{R}$, but we can replace it by the modified distribution $\tilde{\mu}=\mu+(1-m[Y]) \delta_{0}$. We will say with a little abuse of notation that the distribution of $f_{Y}$ is attracted to a stable law if the distribution $\tilde{\mu}$ is in the domain of attraction of this stable law.

Theorem 1.2 (Convergence to stable laws). Let $f \in I(X, Y)$. Assume that the induced function $f_{Y}$ on $Y$ is not in $L^{2}$, and that its distribution is in the domain of attraction of a stable law of index $p \in(0,1) \cup(1,2]$, as described previously. If $p>1$, assume moreover that $f$ is integrable and that $\int f \mathrm{~d} m=0$.

Then $\frac{\sum_{i=0}^{n-1} f \circ T^{i}}{B_{n}}$ converges in distribution to the stable law $X_{p, c, \beta}$ where $B_{n}, p, c$ and $\beta$ are as in the random case above.

The hypothesis $\int f=0$ when $p>1$ ensures that the $A_{n}$ term vanishes.
Note that, in Theorem 1.1, the constant $\sigma^{2}$ depends on the interactions at different times between $f$ and $f \circ T^{n}$, while in Theorem 1.2, the limit distribution depends only on the initial distribution (of $f_{Y}$ ).

### 1.3. Application to one-dimensional maps

Our main theorems can be applied to one-dimensional maps with a neutral fixed point with prescribed behavior. For the sake of simplicity, we will restrict ourselves to a specific case, but the results can in fact be proved for a wide class of maps.

For $0<\alpha<1$, we consider the map from [0, 1] to itself introduced in [LSV99] (this paper contains also historical references) and defined by

$$
T(x)=\left\{\begin{array}{cl}
x\left(1+2^{\alpha} x^{\alpha}\right) & \text { if } 0 \leqslant x \leqslant 1 / 2  \tag{1}\\
2 x-1 & \text { if } 1 / 2<x \leqslant 1
\end{array}\right.
$$

This map has a neutral fixed point at 0 , and is expanding elsewhere. It is known that it has a unique absolutely continuous invariant measure $\mathrm{d} m$, whose density $h$ is Lipschitz on any interval of the form ( $\varepsilon, 1]$ ([LSV99, Lemma 2.3]).

For Hölder observables on [0, 1], Young ([You99]) showed that $\operatorname{Cor}\left(f, g \circ T^{n}\right)=$ $\int f \cdot g \circ T^{n} \mathrm{~d} m-\int f \mathrm{~d} m \int g \mathrm{~d} m=O\left(1 / n^{1 / \alpha-1}\right)$. This rate is in fact optimal for a wide class of functions ([Sar02]). It implies a central limit theorem when $\alpha<1 / 2$.

For $\alpha>1 / 2$, the above rate is not summable. However, extending the methods of [Sar02], we showed in [Gou02] that, for functions $f$ vanishing in a neighborhood of the fixed point and with zero integral, the correlations decay like $O\left(n^{-1 / \alpha}\right)$, which gives again a central limit theorem. For functions with zero average and equal to a nonzero constant in a neighborhood of the fixed point, the correlations decay exactly like $C / n^{1 / \alpha-1}$ ([Gou02]), which is not summable.

Our main theorems yield (see also a list of other consequences after the proof of Theorem 1.3):

Theorem 1.3. Let $1 / 2<\alpha<1$, and $T:[0,1] \rightarrow[0,1]$ be the corresponding Liverani-Saussol-Vaienti map (1). Let h be the density of its unique absolutely continuous invariant probability d . Let $f:[0,1] \rightarrow \mathbb{R}$ be Hölder, with $\int f \mathrm{~d} m=0$.

If $f(0) \neq 0$, then $\frac{S_{n} f}{n^{\alpha}}$ converges in distribution to the stable law $X_{1 / \alpha, c, \operatorname{sgn}(f(0))}$ with $c=\frac{h(1 / 2)}{4(\alpha /|f(0)|)^{1 / \alpha}} \Gamma(1-1 / \alpha) \cos \left(\frac{\pi}{2 \alpha}\right)$.

If $f(0)=0$, assume moreover that $|f(x)| \leqslant C x^{\gamma}$ with $\gamma>\alpha-1 / 2$. Then there exists a constant $\sigma^{2} \geqslant 0$ such that $\frac{S_{n} f}{\sqrt{n}}$ tends in distribution to $\mathcal{N}\left(0, \sigma^{2}\right)$. Moreover, $\sigma^{2}=0$ if and only if there exists a measurable function $\chi$ with $f=\chi \circ T-\chi$.

This result does not address the question of the summability of correlations for a function $f$ with $f(0)=0$. Hu announced recently this summability ( $[\mathrm{Hu}]$ ), which gives another proof of the central limit theorem.

Proof. We construct a Markov structure, starting from $x_{0}=1$ and setting $x_{n+1}=$ $T^{-1}\left(x_{n}\right) \cap[0,1 / 2]$. The map $T$ is then Markov for the partition composed of the $I_{n}:=\left(x_{n+1}, x_{n}\right), n \in \mathbb{N}$. We will induce on $Y=(1 / 2,1)$. Setting $y_{n+1}=$ $T^{-1}\left(x_{n}\right) \cap(1 / 2,1]$, the induced map $T_{Y}$ is Markov for the partition given by the intervals $J_{n}=\left(y_{n+1}, y_{n}\right), n \geqslant 1$. The distortion control on $T_{Y}$ is classical (see for example [LSV99]), it remains to study the properties of $f$ and $f_{Y}$. We will write $s(x, y)$ for the separation time of $x, y \in[0,1]$ under $T_{Y}$.

As the induced map on $Y$ is uniformly expanding, with a factor $\lambda>1$, we obtain $|x-y| \leqslant C \lambda^{-s(x, y)}$. As $f$ is $v$-Hölder for some $v$, we get $|f(x)-f(y)| \leqslant$ $D|x-y|^{\nu} \leqslant E\left(\lambda^{-\nu}\right)^{s(x, y)}$, with $\theta:=\lambda^{-\nu}<1$. Thus, $f$ is bounded and locally Hölder. [LSV99] proves that $m\left[J_{n}\right] \sim \frac{C}{n^{1+1 / \alpha}}$ for some constant $C$. Consequently, $f \in I(X, Y)$ (see the examples following the definition of $I(X, Y)$ ).

We assume now that $f(0)=0$ and show that $f_{Y} \in L^{2}(Y, \mathrm{~d} m)$. We may suppose that $\gamma<\alpha$. As $\mathrm{d} m$ and dLeb are equivalent on $Y$, it is sufficient to show that $f_{Y} \in L^{2}(Y, \mathrm{dLeb})$. If $x \in J_{n}$, then for $1 \leqslant i \leqslant n-1, T^{i} x \in\left(x_{n+1-i}, x_{n-i}\right)$, whence $\left|f\left(T^{i} x\right)\right| \leqslant C\left|T^{i} x\right|^{\gamma} \leqslant C\left|x_{n-i}\right|^{\gamma}$. This implies that $\left|f_{Y}(x)\right| \leqslant|f(x)|+$ $C \sum_{j=1}^{n-1}\left|x_{j}\right|^{\gamma} \leqslant C+C \sum_{j=1}^{n-1} \frac{1}{j^{\gamma / \alpha}}$, since $x_{n} \sim \frac{1}{2}(\alpha n)^{-1 / \alpha}$ by [Sar02, Corollary 1]. Hence, $\left|f_{Y}\right| \leqslant C n^{1-\gamma / \alpha}$ on $J_{n}$, with $\operatorname{Leb}\left(J_{n}\right) \asymp x_{n}-x_{n+1} \asymp 1 / n^{1+1 / \alpha}$. Thus, $\int\left|f_{Y}\right|^{2} \leqslant C \sum \frac{n^{2-2 \gamma / \alpha}}{n^{1+1 / \alpha}}$. This is summable when $\gamma>\alpha-\frac{1}{2}$. We have proved that $f_{Y} \in L^{2}$, whence we can use Theorem 1.1 to conclude.

We assume finally that $f(0)>0$ (the case $f(0)<0$ is analogous). Write $f=f(0)+\tilde{f}$, where $\tilde{f}$ is $v$-Hölder and satisfies $\tilde{f}(0)=0$. As in the case $f(0)=0$, we get that, for $x \in J_{n},\left|\tilde{f}_{Y}(x)\right| \leqslant|\tilde{f}(x)|+C \sum_{j=1}^{n-1}\left|x_{j}\right|^{v}$, whence $\left|\tilde{f}_{Y}\right| \leqslant D n^{1-\nu / \alpha}$ on $J_{n}$. Take $q>1 / \alpha$, close to $1 / \alpha$.

$$
\int\left|\tilde{f}_{Y}\right|^{q} \mathrm{dLeb} \leqslant C \sum \frac{n^{q-q v / \alpha}}{n^{1+1 / \alpha}}<\infty
$$

i.e. $\tilde{f}_{Y} \in L^{q}(Y, \mathrm{~d} m)$ since $m$ and Leb are equivalent on $(1 / 2,1]$. Writing $u(x)=$ $n f(0)$ on $J_{n}$, we have $f_{Y}=\tilde{f}_{Y}+u$ on $Y$, and

$$
\begin{aligned}
m[u>n f(0)] & =m\left[\bigcup_{n+1}^{\infty} J_{k}\right] \\
& =m\left(1 / 2, x_{n} / 2+1 / 2\right) \sim h(1 / 2) \frac{x_{n}}{2} \sim h(1 / 2) \frac{1}{4}(\alpha n)^{-1 / \alpha} .
\end{aligned}
$$

Thus, we obtain $m[u>x] \sim \frac{h(1 / 2)}{4(\alpha x / f(0))^{1 / \alpha}}=\frac{c_{1}}{x^{1 / \alpha}}$. Adding $\tilde{f}_{Y} \in L^{q}$ does not change the asymptotics since $m\left[\left|\tilde{f}_{Y}\right|>x\right] \leqslant x^{-q} \int\left|\tilde{f}_{Y}\right|^{q}=o\left(x^{-1 / \alpha}\right)$. We have proved that $m\left[f_{Y}>x\right]=\left(c_{1}+o(1)\right) x^{-p}$ for $p=1 / \alpha$. In the same way, we get $m\left[f_{Y}<-x\right]=o\left(x^{-p}\right)$. This enables us to use Theorem 1.2, which concludes the proof.

We enumerate other consequences of Theorems 1.1 and 1.2 in other one-dimensional cases:

1. For $\alpha \in(0,1 / 2)$, the function $f_{Y}$ is always in $L^{2}$ and the condition $\gamma>\alpha-\frac{1}{2}$ becomes empty. Thus, we get another proof of the classical central limit theorem, for any Hölder function $f$.
2. For $\alpha=1 / 2$, the condition on $\gamma$ is also empty, and we get the following result for a Hölder $f$ with zero average:

- If $f(0)=0$, there is a central limit theorem.
- If $f(0) \neq 0$, for example $f(0)>0$, then we get as above $m\left[f_{Y}>x\right]=$ $\left(c_{1}+o(1)\right) x^{-2}$ for $c_{1}=\frac{h(1 / 2) \sqrt{2 f(0)}}{4}$. We are in the nonstandard domain of attraction of the normal law (i.e. $f_{Y}$ is attracted to a normal law but is not in $L^{2}$ ), for $l(x)=1$, whence $L(x) \sim \frac{c_{1}}{2} \log x$. The constant $B_{n}$ is then $\frac{\sqrt{c_{1}}}{2} \sqrt{n \log n}$, and we get the convergence

$$
\frac{\sum_{i=0}^{n-1} f \circ T^{i}}{\frac{\sqrt{c_{1}}}{2} \sqrt{n \log n}} \rightarrow \mathcal{N}(0,1)
$$

3. For $\alpha \in(0,1)$, we can also look at $f(x)=x^{-\beta}+\kappa$ with $\beta>0$, where $\kappa$ is taken so that $\int f=0$. Note that $f$ is integrable with respect to the invariant measure $m$ if and only if $\alpha+\beta<1$ (since the density of $m$ behaves like $x^{-\alpha}$ close to 0 ). A small computation shows that $D_{\theta} f_{Y}\left(J_{n}\right) \asymp n^{\beta / \alpha}$, whence $\sum n m\left[J_{n}\right] D_{\theta} f_{Y}\left(J_{n}\right)<\infty$ as soon as $\alpha+\beta<1$. Finally, on $J_{n}, f_{Y} \sim C n^{\beta / \alpha+1}$, whence $m\left[f_{Y}>x\right] \sim C x^{-\frac{1}{\alpha+\beta}}$. Thus, when $\alpha+\beta<1 / 2$, we obtain a usual
central limit theorem, while when $\alpha+\beta=1 / 2$ we have a central limit theorem with normalization $\frac{1}{\sqrt{n \log n}}$, and for $1 / 2<\alpha+\beta<1$ we obtain a convergence to a stable law for the normalization $\frac{1}{n^{\alpha+\beta}}$. So, even when $\alpha<1 / 2$, we can have convergence to a stable law if the observable is unbounded.
Note that Theorems 1.1 and 1.2 can not be used directly in this case, since the function $f$ is not bounded on $\bigcup_{0}^{N-1} T^{i}(Y)$. However, what is important in the proof is that we can take an increasing sequence $Z_{q}$ with $\bigcup Z_{q}=X$ such that the induction on $Z_{q}$ has good properties. In this case, we can take $Z_{q}=\left(x_{q}, 1\right)$ and the proof is exactly the same. In an equivalent way, we can construct an inverse tower $\bigcup_{0}^{N-1} T^{-i}(Y)$ and work with this inverse tower, the proofs being almost identical.
4. For fixed points such as $x+x^{\alpha} \log x$ with $\alpha \in(0,1)$, we also get other slowly varying functions, whence convergence with other normalizing factors.

### 1.4. Strategy of the proof

Results of convergence to normal laws or stable laws are already known when it is possible to obtain a spectral gap for the transfer operator, for example for uniformly expanding dynamics (see for example [GH88] and [AD01]). In the following, this case will be referred to as the uniform case. Such results apply in particular to the induced maps, and they are obtained by perturbing the transfer operators.

The strategy is then to translate the results on the induced map to the whole space. For this, we use first return transfer operators, which have been used by Sarig in [Sar02] (see also [Iso00]). We will perturb these operators, and transfer the results in the uniform case to these first return transfer operators, through the abstract Theorem 2.1 which is in fact the main technical part of the proof.

This gives information on the behavior of the Birkhoff sums, but only on the part of the space on which we have induced. We then have to induce on larger and larger parts of the space, to recover more and more information, which will complete the proof.

In Section 2, we will prove the abstract result referred to above, Theorem 2.1. In Section 3, we will establish estimates in the uniform case in our context, which is slightly more general than the results in the literature. Finally, we will prove the main theorems in Section 4.

## 2. An abstract perturbative theorem

In the following, $\mathbb{D}$ will denote $\{z \in \mathbb{C}||z|<1\}$, and $\overline{\mathbb{D}}=\{z \in \mathbb{C}| | z \mid \leqslant 1\}$.

### 2.1. The result

The goal of this section is to prove the following theorem:
Theorem 2.1. Let $\mathcal{L}$ be a Banach space and $R_{n} \in \operatorname{Hom}(\mathcal{L}, \mathcal{L})$ be operators on $\mathcal{L}$ with $\left\|R_{n}\right\| \leqslant r_{n}$ for a sequence $r_{n}$ such that $a_{n}=\sum_{k>n} r_{k}$ is summable. Write $R(z)=\sum R_{n} z^{n}$ for $z \in \overline{\mathbb{D}}$. Assume that 1 is a simple isolated eigenvalue of $R(1)$
and that $I-R(z)$ is invertible for $z \in \overline{\mathbb{D}}-\{1\}$. Let $P$ denote the spectral projection of $R(1)$ for the eigenvalue 1 , and assume that $P R^{\prime}(1) P=\mu P$ for some $\mu>0$.

Let $R_{n}(t)$ be an operator depending on $t \in[-\alpha, \alpha]$, continuous at $t=0$, with $R_{n}\left(\underline{0)}=R_{n}\right.$ and $\left\|R_{n}(t)\right\| \leqslant C r_{n} \forall t \in[-\alpha, \alpha]$, for some constant $C>0$. For $z \in \overline{\mathbb{D}}$ and $t \in[-\alpha, \alpha]$, write

$$
R(z, t)=\sum_{n=1}^{\infty} z^{n} R_{n}(t)
$$

This is a continuous perturbation of $R(z)$. For $t$ small and $z$ close to $1, R(z, t)$ is close to $R(1)$, whence it admits an eigenvalue $\lambda(z, t)$ close to 1 . Assume that $\lambda(1, t)=1-(c+o(1)) M(|t|)$ for $c \in \mathbb{C}$ with $\operatorname{Re}(c)>0$, and some continuous function $M: \mathbb{R}_{+} \rightarrow \mathbb{R}_{+}$vanishing only at 0 . Then

1. There exists $\varepsilon_{0}>0$ such that $\forall|t|<\varepsilon_{0}, I-R(z, t)$ is invertible for all $z \in \mathbb{D}$. We can write $(I-R(z, t))^{-1}=\sum T_{n, t} z^{n}$.
2. Furthermore, there exist functions $\varepsilon(t)$ and $\delta(n)$ tending to 0 when $t \rightarrow 0$ and $n \rightarrow \infty$ such that $\forall|t|<\varepsilon_{0}, \forall n \in \mathbb{N}^{*}$, we have

$$
\begin{equation*}
\left\|T_{n, t}-\frac{1}{\mu}\left(1-\frac{c}{\mu} M(|t|)\right)^{n} P\right\| \leqslant \varepsilon(t)+\delta(n) . \tag{2}
\end{equation*}
$$

For example, when $M(t)=t^{p}$, this gives the convergence $T_{n, t / n^{1 / p}} \rightarrow \frac{e^{-c \mid t t^{p} / \mu}}{\mu} P$.
Remark. The assumption that 1 is a simple isolated eigenvalue of $R(1)$ means that, for $\lambda$ close to 1 but different from $1, \lambda I-R(1)$ is invertible (i.e. 1 is isolated in the spectrum of $R(1))$, and that there is a codimension 1 subspace $E \subset \mathcal{L}$ and a non zero vector $\psi \in \mathcal{L}$ such that $\mathcal{L}=E \oplus \mathbb{C} \psi,(I-R(1)) \psi=0$, and $I-R(1): E \rightarrow E$ is invertible (i.e. 1 is a simple eigenvalue of $R(1)$ ).

Remark. There is an analogue of this theorem when the perturbation takes place on $[0, \alpha]$ (with exactly the same proof). Thus, for a perturbation where $\lambda(1, t)=$ $1-a|t|^{p}+i b \operatorname{sgn}(t)|t|^{p}+o\left(|t|^{p}\right)$ for $t \in[-\alpha, \alpha]$, we get once again the same result, using twice this theorem, on $[-\alpha, 0]$ and $[0, \alpha]$.

In the application of Theorem 2.1 to the proof of Theorems 1.1 and 1.2, the operators $R_{n}$ will describe the returns to the basis $Y$, and will be easily understood, as well as their perturbations $R_{n}(t)$. On the other hand, $T_{n}$ will describe all the iterates at time $n$, and $T_{n, t}$ will be closely related to the characteristic function $E\left(e^{i t S_{n} f}\right)$ (see Section 4 for more details). Thus, (2) will enable us to describe precisely $E\left(e^{i t S_{n} f}\right)$, and this information will be sufficient to get Theorems 1.1 and 1.2.

### 2.2. Wiener's Lemma in Banach algebras

The classical Wiener's Lemma says that if $f: S^{1} \rightarrow \mathbb{C}$ has summable Fourier coefficients and is everywhere nonzero, then the Fourier coefficients of $1 / f$ are also summable (see for example [Kah70]). We will use an analogue of this result for functions taking their values in non-commutative Banach algebras.

Let $\mathcal{A}$ be a Banach algebra with unit (its norm will be written $\left\|\|_{\mathcal{A}}\right.$ ). We will denote by $A(\mathcal{A})$ the set of continuous functions $f: S^{1} \rightarrow \mathcal{A}$ whose Fourier coefficients, defined by

$$
c_{n}(f)=\frac{1}{2 \pi} \int_{0}^{2 \pi} e^{-i n \theta} f\left(e^{i \theta}\right) \mathrm{d} \theta
$$

are summable in norm. Then $f$ coincides with the sum $\sum_{n \in \mathbb{Z}} c_{n}(f) z^{n}$ of its Fourier series (see [Gou02, Proposition 2.6]). In particular, if $f, g \in A(\mathcal{A})$, then $c_{n}(f g)=$ $\sum_{k \in \mathbb{Z}} c_{n+k}(f) c_{n-k}(g)$, whence $f g \in A(\mathcal{A})$. Moreover, $A(\mathcal{A})$ is a Banach algebra for the norm $\|f\|_{A(\mathcal{A})}=\sum\left\|c_{n}(f)\right\|_{\mathcal{A}}$.

For $f: S^{1} \rightarrow \mathcal{A}$, we will also write $\|f\|_{\mathcal{A}}=\sup \left\{\|f(z)\|_{\mathcal{A}} \mid z \in S^{1}\right\}$. We have of course $\|f\|_{\mathcal{A}} \leqslant\|f\|_{A(\mathcal{A})}$, but the other inequality is false.

The following result can be found in [BP42] (I thank Omri Sarig for this reference).
Theorem 2.2. Let $f \in A(\mathcal{A})$ be such that for all $z \in S^{1}, f(z)$ is invertible in the Banach algebra $\mathcal{A}$. Then, setting $g(z)=f(z)^{-1}$, we also have $g \in A(\mathcal{A})$.

Let $\Omega$ be an open subset of $\mathbb{C}$ and $h: \Omega \rightarrow \mathbb{C}$ an holomorphic function. If $x \in \mathcal{A}$ is such that its spectrum $\sigma(x)$ is included in $\Omega$, it is possible to define $h(x) \in \mathcal{A}$ using the Cauchy formula, by $h(x):=\frac{1}{2 i \pi} \int_{\gamma} \frac{h(u)}{u I-x} \mathrm{~d} u$, where $\gamma$ is a path around $\sigma(x)$ in $\Omega$. This is independent of $\gamma$ (see e.g. [DS57, VII.3]). For example, if $h(z)=z^{n}$ for $n \in \mathbb{Z}$, then $h(x)=x^{n}$ (for the multiplication in the algebra $\mathcal{A}$ ). For another example, take $\mathcal{A}=\operatorname{Hom}(\mathcal{L}, \mathcal{L})$ where $\mathcal{L}$ is a Banach space, and $x \in \mathcal{A}$ for which 1 is isolated in $\sigma(x)$. Then, if $h$ is equal to 1 in a small neighborhood of 1 , and to 0 outside of this neighborhood, then $h(x)$ is the spectral projection associated to 1 .

Theorem 2.3. Let $\Omega$ be an open subset of $\mathbb{C}$ and $h: \Omega \rightarrow \mathbb{C}$ be holomorphic. Let $f \in A(\mathcal{A})$ be such that $\forall z \in S^{1}, \sigma(f(z)) \subset \Omega$. If $g(z)=h(f(z))$, then we also have $g \in A(\mathcal{A})$.

The previous Wiener's Lemma is a particular case of this result, for $h(z)=1 / z$.
Proof. As the spectrum is semi-continuous, there exists a path $\gamma$ in $\Omega$ around $\sigma(f(z))$ for every $z \in S^{1}$. Thus, $g(z)=\frac{1}{2 i \pi} \int_{\gamma} \frac{h(u)}{u I-f(z)} \mathrm{d} u$. For $u \in \gamma$ fixed, $M_{u}: z \mapsto \frac{h(u)}{u I-f(z)}$ is in $A(\mathcal{A})$ by Theorem 2.2. As the inversion is continuous in the Banach algebra $A(\mathcal{A}), u \mapsto M_{u}$ is continuous on the compact $\gamma$. Since $g=\int_{\gamma} M_{u} \mathrm{~d} u$, this concludes the proof.
Note that this result applies in particular to spectral projections.

### 2.3. Study of $(I-R(z, t))^{-1}$ at $t=0$

In this section, we will describe the behavior of the coefficients of $(I-R(z))^{-1}$, and in particular show that they tend to $\frac{1}{\mu} P$. We will use techniques introduced by Sarig in [Sar02]. The main technical difference is that in the proof of Theorem 2.4 (an analogue of his "first main lemma"), we will replace his estimates using $C^{1+\alpha}$ regularity by applications of Wiener's Lemma.

The following result is interesting in itself, and its proof will be instrumental in the proof of Theorem 2.1.

Theorem 2.4. Let $R_{n}$ be operators on a Banach space $\mathcal{L}$ such that $a_{n}=\sum_{k>n}\left\|R_{k}\right\|$ is summable. Write $R(z)=\sum_{n=1}^{\infty} R_{n} z^{n}$ and assume that 1 is a simple isolated eigenvalue of $R(1)$, while $I-R(z)$ is invertiblefor $z \in \overline{\mathbb{D}}-\{1\}$. Write $P$ for the spectral projection of $R(1)$ for the eigenvalue 1, and assume also that $P R^{\prime}(1) P=\mu P$ for a nonzero constant $\mu$.

Then the function $\left(\frac{I-R(z)}{1-z}\right)^{-1}$ can be continuously extended to $\overline{\mathbb{D}}$, and its Fourier coefficients on $S^{1}$ are summable. Writing $(I-R(z))^{-1}=\sum_{n} T_{n} z^{n}$, we have $\sum_{n}\left\|T_{n}-T_{n-1}\right\|<\infty$.

Proof. Write $S(z)=\frac{I-R(z)}{1-z}$ and $\mathcal{A}=\operatorname{Hom}(\mathcal{L}, \mathcal{L})$.
Step 1. $S(z)^{-1}$ can be continuously extended to $\overline{\mathbb{D}}$.
The problem is the extension to 1 . For $z \in \overline{\mathbb{D}}$ close to $1, R(z)$ has a unique eigenvalue $\lambda(z)$ close to 1 . We write $P(z)$ for the corresponding spectral projection, and $Q(z)=I-P(z)$. For $z$ close to 1 but different from 1,

$$
\begin{equation*}
S(z)^{-1}=\frac{1-z}{1-\lambda(z)} P(z)+(1-z)(I-R(z) Q(z))^{-1} Q(z) \tag{3}
\end{equation*}
$$

As $I-R(z) Q(z)$ is everywhere invertible on $\overline{\mathbb{D}}$, the second term has a continuous extension to 1 . As $(I-R(1)) P(1)=0$,

$$
\begin{equation*}
\frac{1-\lambda(z)}{1-z} P(z)=P(z) \frac{R(1)-R(z)}{1-z} P(z)+P(z)(I-R(1)) \frac{P(z)-P(1)}{1-z} \tag{4}
\end{equation*}
$$

Note that

$$
\frac{R(z)-R(1)}{z-1}=\sum_{n=0}^{\infty}\left(\sum_{k=n+1}^{\infty} R_{k}\right) z^{n}
$$

As $\sum_{k>n}\left\|R_{k}\right\|$ is a summable sequence, the above sum converges in norm, which implies that its limit at 1 , denoted by $R^{\prime}(1)$, is well defined.

If $\delta>0$ is small enough, then

$$
P(z)=\frac{1}{2 i \pi} \int_{|u-1|=\delta} \frac{1}{u I-R(z)} \mathrm{d} u .
$$

Thus,

$$
\begin{equation*}
\frac{P(z)-P(1)}{z-1}=\frac{1}{2 i \pi} \int_{|u-1|=\delta} \frac{1}{u I-R(z)} \frac{R(z)-R(1)}{z-1} \frac{1}{u I-R(1)} \mathrm{d} u \tag{5}
\end{equation*}
$$

As $\frac{R(z)-R(1)}{z-1}$ converges at 1 , we get the convergence of $\frac{P(z)-P(1)}{z-1}$ at 1 , to a limit $P^{\prime}(1)$.

In Equation (4), when $z$ tends to 1, the right hand term tends to $P(1) R^{\prime}(1) P(1)-$ $P(1)(I-R(1)) P^{\prime}(1)=P(1) R^{\prime}(1) P(1)$ since $P(1)(I-R(1))=0$. Recall that $P(1) R^{\prime}(1) P(1)=\mu P$ for some nonzero $\mu$.

Let $\xi$ be a bounded linear functional on $\operatorname{Hom}(\mathcal{L}, \mathcal{L})$ such that $\xi(P(1))=1$. We apply $\xi$ to both members of Equation (4) and let $z$ go to 1 , which gives that

$$
\begin{equation*}
\frac{1-\lambda(z)}{1-z} \rightarrow \mu \tag{6}
\end{equation*}
$$

As $\mu \neq 0$, this shows that, in Equation (3), both terms on the right hand side converge at 1 .

Step 2. Construction of a function $\tilde{R}(z)$ on $S^{1}$, coinciding with $R(z)$ for $z$ close to 1 , such that it has everywhere a simple eigenvalue $\tilde{\lambda}(z)$ close to 1 , but different from 1 if $z \neq 1$, and with $\frac{\tilde{R}(z)-\tilde{R}(1)}{z-1} \in A(\mathcal{A})$.

We construct $\tilde{R}$ in three steps, first defining two approximations $F, G$ and then gluing them together to get $\tilde{R}$.

Fix some $\gamma>0$, very small. Let $\varphi+\psi$ be a $C^{\infty}$ partition of unity on $S^{1}$ associated to the sets $\{\theta \in[0, \gamma)\}$ and $\{\theta \in(\gamma-\eta, \pi / 2]\}$ where $\theta$ is the angle on the circle (for some very small $0<\eta<\gamma$ ). We define $F(z)=\varphi(z) R(z)+\psi(z) R\left(e^{i \gamma}\right)$ on $\{\theta \in[0, \pi / 2]\}: F$ is equal to $R$ on $\{\theta \in[0, \gamma-\eta]\}$ and to $R\left(e^{i \gamma}\right)$ on $\{\theta \in[\gamma, \pi / 2]\}$. In particular, the spectrum of $F(z)$ is "almost the same" as the spectrum of $R(1)$ if $\gamma$ is small enough.

We define in the same way $F$ on $\{\theta \in[-\pi / 2,0]\}$, equal to $R\left(e^{-i \gamma}\right)$ on $\{\theta \in$ $[-\pi / 2,-\gamma]\}$ and to $R$ on $\{\theta \in[-\gamma+\eta, 0]\}$.

Finally, we construct $F$ on the remaining half-circle by symmetrizing, i.e. $F\left(e^{i(\pi / 2+a)}\right)=F\left(e^{i(\pi / 2-a)}\right)$, to ensure that everything fits well.

Provided $\gamma$ is small enough, there is a well defined eigenvalue close to 1 for every $F(z)$, depending continuously on $z$, which we denote by $\rho(z)$. The problem would be solved if $\rho(z) \neq 1$ for $z \neq 1$, which is not the case since $\rho(-1)=$ $\rho(1)=1$. Consequently, we have to perturb $\rho$ a little. Let $v$ be a $C^{\infty}$ function on $\{\theta \in[\pi / 2,3 \pi / 2]\}$, arbitrarily close to $\rho$ and which does not take the value 1 . On $\{\theta \in[\pi / 2+\eta, 3 \pi / 2-\eta]\}$, we define $G(z)=\frac{v(z)}{\rho(z)} F(z):$ its eigenvalue close to 1 is $v(z) \neq 1$. Finally, we glue $F$ and $G$ together on $\{\theta \in[\pi / 2, \pi / 2+\eta]\}$ and $\{\theta \in[3 \pi / 2-\eta, 3 \pi / 2]\}$ with a partition of unity, as above. As the spectrum of $F\left(e^{i \pi / 2}\right)=R\left(e^{i \gamma}\right)$ does not contain 1 , the gluing will not give an eigenvalue equal to 1 if we choose $\eta$ small enough and $\nu$ close enough to $\rho$.

Note finally that, since $\frac{R(z)-R(1)}{z-1} \in A(\mathcal{A})$, we also have $\frac{\tilde{R}(z)-\tilde{R}(1)}{z-1} \in A(\mathcal{A})$ : to obtain $\tilde{R}(z)$, we have multiplied by $C^{\infty}$ functions, which are in $A(\mathbb{C})$.
Step 3. Writing $\tilde{S}(z)=\frac{I-\tilde{R}(z)}{1-z}$, then $\tilde{S}(z)^{-1} \in A(\mathcal{A})$.
By construction, $\frac{\tilde{R}(z)-\tilde{R}(1)}{z-1} \in A(\mathcal{A})$, and $\tilde{R}(z) \in A(\mathcal{A})$. Equation (5) (with tildes everywhere) implies that $\frac{\tilde{P}(z)-\tilde{P}(1)}{z-1} \in A(\mathcal{A})$ (use first Wiener's Lemma Theorem 2.2 on $u I-\tilde{R}(z)$, then multiply and integrate, as in the proof of Theorem 2.3). In particular, $\tilde{P}(z) \in A(\mathcal{A})$.

Let $\xi$ be a bounded linear functional on $\mathcal{A}$ satisfying $\xi(P(1))=1$ and $\forall z \in$ $S^{1}, \xi(\tilde{P}(z)) \neq 0$ (take $\gamma$ small enough in the construction of $\tilde{R}$ to ensure this). Then $\xi(\tilde{P}(z)) \in A(\mathbb{C})$, thus $1 / \xi(\tilde{P}(z)) \in A(\mathbb{C})$ by Wiener's Lemma 2.2. Applying $\xi$ to both sides of Equation (4) (with tildes) and dividing by $\xi(\tilde{P}(z))$, we obtain that $\frac{1-\tilde{\lambda}(z)}{1-z} \in A(\mathbb{C})$. As this function is everywhere non-vanishing, we can use once more Wiener's Lemma and get that $\frac{1-z}{1-\tilde{\lambda}(z)} \in A(\mathbb{C})$.

Finally, we use Equation (3) (with tildes everywhere). Wiener's Lemma ensures that $(I-\tilde{R}(z) \tilde{Q}(z))^{-1} \in A(\mathcal{A})$ since $I-\tilde{R}(z) \tilde{Q}(z)$ is in $A(\mathcal{A})$ and pointwise invertible. Thus, the right hand side is in $A(\mathcal{A})$, which ends the proof.
Step 4. $S(z)^{-1} \in A(\mathcal{A})$.
Let $\varphi+\psi$ be a $C^{\infty}$ partition of unity on $S^{1}$, such that $\tilde{R}=R$ on the support of $\varphi$. Then

$$
\left(\frac{I-R(z)}{1-z}\right)^{-1}=\varphi(z)\left(\frac{I-\tilde{R}(z)}{1-z}\right)^{-1}+\psi(z)\left(\frac{I-R(z)}{1-z}\right)^{-1}
$$

The first term on the right hand side is in $A(\mathcal{A})$ according to Step 3. For the second term, we can modify $\frac{I-R(z)}{1-z}$ outside of the support of $\psi$ so that it is everywhere defined and invertible, using the same techniques as in the construction of $\tilde{R}$. Wiener's Lemma gives that its inverse is in $A(\mathcal{A})$, which concludes the proof.

Step 5. conclusion.
$I-R(z)$ is a power series that is invertible everywhere on $\mathbb{D}$. According to [DS57, Lemma VII.6.13], its pointwise inverse is also analytic on $\mathbb{D}$, whence it is a power series, that can be written as $\sum T_{n} z^{n}$. Multiplying by $1-z$, we get $S(z)^{-1}=$ $\sum\left(T_{n}-T_{n-1}\right) z^{n}$. Thus, if $r<1$, we have $T_{n}-T_{n-1}=\frac{1}{2 \pi r^{n}} \int e^{-i n \theta} S\left(r e^{i \theta}\right)^{-1} \mathrm{~d} \theta$. As $S(z)^{-1}$ is continuous on the whole disk $\overline{\mathbb{D}}$, we obtain by letting $r$ tend to 1 that $T_{n}-T_{n-1}$ is the $n$th Fourier coefficient of $S(z)^{-1}$ on $S^{1}$. But we already know that $S(z)^{-1} \in A(\mathcal{A})$, which proves that $\sum\left\|T_{n}-T_{n-1}\right\|<\infty$.

Remark. Theorem 2.4 is in fact the main technical ingredient in the proof of Sarig in [Sar02], but he had to assume $a_{n} \sim \frac{1}{n^{\beta}}$ with $\beta>2$. Thus, it is possible using Theorem 2.4 to extend Sarig's results on lower bounds for the decay of correlations (Theorem 2 and Corollaries 1 and 2 in [Sar02]), even when the mass of the set of points returning at time $n$ is not of the order $1 / n^{\beta}$. For example, it is not hard to generalize his results to maps of the interval with a very neutral fixed point, such as $x\left(1+x \log ^{2} x\right)$. With a little work, this shows that all the upper bounds on the decay of correlations obtained by Holland in [Hol02] are in fact optimal.

The following lemma will be needed later on in the proof of Theorem 2.1.
Lemma 2.5. Under the same hypotheses as in Theorem 2.4, we have $\left(\frac{I-R(z)}{1-z}\right)^{-1}=$ $\frac{1}{\mu} P+(1-z) A(z)$, where $A(z)=\sum_{n=0}^{\infty} A_{n} z^{n}$ with $A_{n} \rightarrow 0$ as $n \rightarrow \infty$.

Proof. We use the method of Sarig in [Sar02], starting from our analogue of his first main lemma, Theorem 2.4. More precisely, Sarig constructs a polynomial $R_{B}(z)$, approximating $R(z)$, such that $\frac{R(1)-R_{B}}{1-z}$ and $\frac{1}{1-z}\left(\frac{R(1)-R_{B}}{1-z}-R^{\prime}(1)\right)$ are polynomials. Moreover, $I-R_{B}(z)$ is invertible for $z \in \mathbb{D}$, and $\left(\frac{I-R_{B}(z)}{1-z}\right)^{-1}=$ $\frac{1}{\mu} P+(1-z) B(z)$ where $B(z)=\sum B_{n} z^{n}$ and $B_{n}=o\left(\kappa^{n}\right)$ for some $\kappa<1$. The proof of this result ([Sar02, Lemma 6]) relies only on his first main lemma, which we have already proved in our context.

Then, he uses the following perturbative expansion of $S(z)^{-1}=\left(\frac{I-R(z)}{1-z}\right)^{-1}$ around $S_{B}(z)^{-1}=\left(\frac{I-R_{B}(z)}{1-z}\right)^{-1}$ :

$$
S(z)^{-1}=S_{B}(z)^{-1}+S_{B}(z)^{-1}\left(S_{B}(z)-S(z)\right) S(z)^{-1}
$$

To prove the lemma, we only have to show that $S_{B}(z)^{-1}\left(S_{B}(z)-S(z)\right) S(z)^{-1}$ can be written as $(1-z) C(z)$ for some $C(z)=\sum C_{n} z^{n}$ with $C_{n} \rightarrow 0$ (indeed, $A_{n}=B_{n}+C_{n}$ ).

We first study $\frac{S_{B}-S}{1-z}$. We have

$$
S_{B}-S=\left(\frac{R(1)-R_{B}}{1-z}-R^{\prime}(1)\right)+\left(R^{\prime}(1)-\frac{R(1)-R}{1-z}\right)=I+I I
$$

As $I /(1-z)$ is a polynomial (by construction of $R_{B}$ ), we can forget this term. For the other term,

$$
I I=\sum_{k=0}^{\infty}\left(1-z^{k}\right) \sum_{n=k+1}^{\infty} R_{n} .
$$

Thus, writing $D_{n}=\sum_{k=n+1}^{\infty} \sum_{l=k+1}^{\infty} R_{l}$, we have $\frac{I I}{1-z}=\sum D_{n} z^{n}$. The summability of $a_{k}=\sum_{l=k+1}^{\infty}\left\|R_{l}\right\|$ implies that $D_{n} \rightarrow 0$. We have shown that

$$
\frac{S_{B}-S}{1-z}=\sum E_{n} z^{n}
$$

where $E_{n} \rightarrow 0$.
We already know that $S_{B}(z)^{-1}$ and $S(z)^{-1}$ have summable coefficients. Thus, the lemma will be proved if we show that, if $E_{n} \rightarrow 0$ and $F_{n}$ is summable, then the coefficient $(E F)_{n}$ of $z^{n}$ in $\left(\sum E_{n} z^{n}\right)\left(\sum F_{n} z^{n}\right)$ tends to 0 as $n \rightarrow \infty$.

We have $(E F)_{n}=\sum_{k=0}^{n} E_{k} F_{n-k}$. We fix $\varepsilon>0$, and $N$ such that if $k \geqslant N$, $\left\|E_{k}\right\| \leqslant \varepsilon$. If $K$ is greater than $\left\|E_{k}\right\|$ and $\sum\left\|F_{k}\right\|$, we get for $n \geqslant N$

$$
\left\|(E F)_{n}\right\| \leqslant \sum_{k=0}^{N-1}\left\|E_{k}\right\|\left\|F_{n-k}\right\|+\sum_{k=N}^{n} \varepsilon\left\|F_{n-k}\right\| \leqslant K \sum_{l=n-N}^{\infty}\left\|F_{l}\right\|+K \varepsilon
$$

For $n$ large enough, this is less than $2 K \varepsilon$.

### 2.4. Invertibility of $I-R(z, t)$

From this point on, and until the end of the proof of Theorem 2.1, the notations and hypotheses of all the results will be those of Theorem 2.1.

In this section, we prove the invertibility of $I-R(z, t)$ when $(z, t) \neq(1,0)$ and $t$ is small enough. The problem occurs for $z$ close to 1 , where $R(z, t)$ has a well defined eigenvalue $\lambda(z, t)$ close to 1 . Let us first study this eigenvalue.

Lemma 2.6. In a neighborhood of $(1,0)$, it is possible to write $\lambda(z, t)=1+(z-$ 1) $(\mu+a(z, t))-(c+b(t)) M(|t|)$ where the functions $a$ and $b$ tend to 0 respectively at $(1,0)$ and 0 .

Proof. As

$$
\frac{R(z, t)-R(1, t)}{z-1}=\sum_{n=0}^{\infty} z^{n}\left(\sum_{k>n} R_{k}(t)\right)
$$

the summability assumption implies that $(z, t) \mapsto \frac{R(z, t)-R(1, t)}{z-1}$ is continuous at $t=0$. Writing

$$
\begin{aligned}
& \frac{P(z, t)-P(1, t)}{z-1} \\
& \quad=\frac{1}{2 i \pi} \int_{|u-1|=\delta} \frac{1}{u I-R(z, t)} \frac{R(z, t)-R(1, t)}{z-1} \frac{1}{u I-R(1, t)} \mathrm{d} u,
\end{aligned}
$$

we also get the continuity of $(z, t) \mapsto \frac{P(z, t)-P(1, t)}{z-1}$ at $t=0$. Thus, applying a bounded linear functional to the equation

$$
\begin{aligned}
\frac{\lambda(z, t)-\lambda(1, t)}{z-1} P(z, t)= & \frac{R(z, t)-R(1, t)}{z-1} P(z, t) \\
& +(R(1, t)-\lambda(1, t)) \frac{P(z, t)-P(1, t)}{z-1},
\end{aligned}
$$

we get also the continuity of $F(z, t)=\frac{\lambda(z, t)-\lambda(1, t)}{z-1}$. Moreover, $F(z, 0) \rightarrow \mu$ when $z \rightarrow 1$ (Equation (6)), which makes it possible to write $F(z, t)=\mu+a(z, t)$ with $a(z, t)$ vanishing at $(1,0)$. Moreover, by assumption, $\lambda(1, t)=1-(c+b(t)) M(|t|)$ where $b$ vanishes at 0 . Hence,

$$
\begin{aligned}
\lambda(z, t) & =\lambda(1, t)+(z-1) F(z, t) \\
& =1-(c+b(t)) M(|t|)+(z-1)(\mu+a(z, t))
\end{aligned}
$$

Proposition 2.7. There exists $\varepsilon_{0}>0$ such that $\forall(z, t) \in\left(\overline{\mathbb{D}} \times\left[-\varepsilon_{0}, \varepsilon_{0}\right]\right)-\{(1,0)\}$, $I-R(z, t)$ is invertible.

Proof. Assume that there is a neighborhood $V$ of $(1,0)$ such that, in $V-\{(1,0)\}$, $I-R(z, t)$ is invertible. We can assume that $V=U \times\left[-\varepsilon_{1}, \varepsilon_{1}\right]$. For $w \in \overline{\mathbb{D}}-U$, $I-R(w)$ is invertible, thus $I-R(z, t)$ is invertible in a small neighborhood of $(w, 0)$, which can be taken of the form $U_{w} \times\left[-\varepsilon_{w}, \varepsilon_{w}\right]$. By compactness, $\overline{\mathbb{D}}-U$ is
covered by a finite number of the $U_{w}$, say $U_{w_{1}}, \ldots, U_{w_{n}}$. To conclude, just take $\varepsilon_{0}=\min \left(\varepsilon_{1}, \varepsilon_{w_{1}}, \ldots, \varepsilon_{w_{n}}\right)$.

So, we just have to construct a neighborhood $V$ as above. We have to show that the eigenvalue $\lambda(z, t)$ of $R(z, t)$ is different from 1 when $(z, t)$ is close to but different from ( 1,0 ). This essentially comes from the asymptotic expansion of $\lambda(z, t)$ given by Lemma 2.6, but we will write it more carefully.

Let $K$ be such that $\frac{\mu}{4}>\frac{3|c|}{2 K}$. The neighborhood $V$ will be the set of $(z, t)$ such that $|b(t)| \leqslant \operatorname{Re}(c) / 2$ and $|a(z, t)| \leqslant \min \left(\frac{\mu}{4}, \frac{\operatorname{Re}(c)}{2 K}\right)$.

We fix $(z, t) \in V \cap(\overline{\mathbb{D}} \times \mathbb{R})$. Assuming that $\lambda(z, t)=1$, we show that $(z, t)=$ $(1,0)$. There are two possibilities, namely $|\operatorname{Im}(z-1)| \geqslant|z-1| / 2$ or $|\operatorname{Re}(z-1)| \geqslant$ $|z-1| / 2$. The latter case, being easier, is left to the reader. We can assume for example that $\operatorname{Im}(z-1) \geqslant|z-1| / 2$.

As $\operatorname{Re}(\mu(z-1)) \leqslant 0$ since $\mu>0$ and $z \in \overline{\mathbb{D}}$, we have

$$
\begin{aligned}
1= & \operatorname{Re}(\lambda(z, t))=1+\operatorname{Re}(\mu(z-1))+\operatorname{Re}(a(z, t)(z-1)) \\
& -(\operatorname{Re}(c)+\operatorname{Re}(b(t)) M(|t|) \\
\leqslant & 1+|a(z, t)||z-1|-(\operatorname{Re}(c)-|b(t)|) M(|t|) \\
\leqslant & 1+\frac{\operatorname{Re}(c)}{2 K}|z-1|-\frac{\operatorname{Re}(c)}{2} M(|t|) .
\end{aligned}
$$

Thus, $|z-1| \geqslant K M(|t|)$. The imaginary part then satisfies

$$
\begin{aligned}
0=\operatorname{Im}(\lambda(z, t)) & \geqslant(\mu-|a(z, t)|) \operatorname{Im}(z-1)-(|c|+|b(t)|) M(|t|) \\
& \geqslant \frac{\mu}{2} \frac{|z-1|}{2}-\frac{3|c|}{2} \frac{|z-1|}{K}
\end{aligned}
$$

The definition of $K$ implies that $|z-1|=0$, which in turn gives $t=0$ since $|z-1| \geqslant K M(|t|)$ and $M$ vanishes only at 0 .

In particular, for $|t|<\varepsilon_{0}, I-R(z, t)$ is a power series invertible everywhere on $\mathbb{D}$, whence its inverse can be written as $\sum T_{n, t} z^{n}$ (see Step 5 of the proof of Theorem 2.4). Moreover, for $t \neq 0$, everything can be continuously extended to $\partial \mathbb{D}$, which implies that the $T_{n, t}$ are the Fourier coefficients of the function $(I-R(z, t))^{-1}$ (considered as a function from $S^{1}$ to $\mathcal{A}$ ).

### 2.5. Convergence of $\tilde{R}(z, t)$

In this section, we will work exclusively on $S^{1}$ ( $z$ will thus denote a point in $S^{1} \subset \mathbb{C}$ and not in $\mathbb{D}$ ).

To apply Fourier series methods to $R(z, t)$, we will need it to have a well defined eigenvalue close to 1 , which is not a priori true outside of a neighborhood of $(1,0)$.
Construction of $\tilde{R}(z, t)$
We use the function $\tilde{R}(z)$ that has been constructed in Step 2 of the proof of Theorem 2.4. We set $\tilde{R}(z, t)=\tilde{R}(z)-\left(R(z)-\sum_{n=1}^{\infty} z^{n} R_{n}(t)\right)$. The map $t \mapsto \tilde{R}(\cdot, t)$ is then continuous at $t=0$ (for the $A(\mathcal{A})$-norm, whence for the supremum norm),
and $\tilde{R}(z, t)$ has a well defined eigenvalue $\tilde{\lambda}(z, t)$ close to 1 . Moreover, for $z$ in a neighborhood of $1, \tilde{\lambda}(z, t)=\lambda(z, t)$, which is different from 1 when $(z, t) \neq(1,0)$, and outside of this neighborhood $\tilde{\lambda}(z, t)$ is close to $\tilde{\lambda}(z, 0)$ and is thus different from 1 if $t$ is small enough.

The main result of this section will be the following proposition (recall that the hypotheses are those of Theorem 2.1):

Proposition 2.8. We have

$$
\left\|\left(\frac{I-\tilde{R}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right)^{-1}-\left(\frac{I-\tilde{R}(z)}{1-z}\right)^{-1}\right\|_{A(\mathcal{A})} \underset{t \rightarrow 0}{\longrightarrow} 0
$$

Proof. Write $\tilde{P}(z, t)$ for the spectral projection associated to the eigenvalue $\tilde{\lambda}(z, t)$, and $\tilde{Q}(z, t)=I-\tilde{P}(z, t)$. We then have

$$
I-\tilde{R}(z, t)=(1-\tilde{\lambda}(z, t)) \tilde{P}(z, t)+(I-\tilde{R}(z, t) \tilde{Q}(z, t)) \tilde{Q}(z, t)
$$

Thus,

$$
\begin{aligned}
& \left(\frac{I-\tilde{R}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right)^{-1}=\frac{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}{1-\tilde{\lambda}(z, t)} \tilde{P}(z, t) \\
& +\left(1-\left(1-\frac{c}{\mu} M(|t|)\right) z\right)(I-\tilde{R}(z, t) \tilde{Q}(z, t))^{-1} \tilde{Q}(z, t)
\end{aligned}
$$

As $\tilde{R}(z, t) \rightarrow \tilde{R}(z)$ as $t \rightarrow 0$, Wiener's Lemma Theorem 2.3 gives the convergence of $\tilde{P}(z, t)$ and $\tilde{Q}(z, t)$ respectively to $\tilde{P}(z)$ and $\tilde{Q}(z)$ (in $A(\mathcal{A})$ ). We also obtain (again by Wiener's Lemma) the convergence of $(I-\tilde{R}(z, t) \tilde{Q}(z, t))^{-1}$ to $(I-\tilde{R}(z) \tilde{Q}(z))^{-1}$, since $I-\tilde{R}(z) \tilde{Q}(z)$ is invertible for every $z \in S^{1}$.

It only remains to prove that $\frac{1-\left(1-\frac{c}{\mu} M(|t|) z\right.}{1-\tilde{\lambda}(z, t)} \rightarrow \frac{1-z}{1-\tilde{\lambda}(z)}($ in $A(\mathbb{C}))$, which is given by the following lemma and Wiener's Lemma (since inversion is continuous in the Banach algebra $A(\mathbb{C})$ ).

Lemma 2.9. We have

$$
\left\|\frac{1-\tilde{\lambda}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}-\frac{1-\tilde{\lambda}(z)}{1-z}\right\|_{A(\mathbb{C})} \xrightarrow[t \rightarrow 0]{ } 0
$$

Proof. We will usually work with the circle $S^{1}$ embedded in $\mathbb{C}$, but it will sometimes be convenient to consider the circle as $\mathbb{T}=\mathbb{R} / 2 \pi \mathbb{Z}$, in which case the variable will be denoted by $\theta$.

We recall some results on Fourier series. For $|\theta| \leqslant \pi$ and small $\varepsilon>0$, write $\Delta_{\varepsilon}(\theta)=\sup \left(0,1-\frac{|\theta|}{\varepsilon}\right)$ : this function is supported in $\{\theta \in[-\varepsilon, \varepsilon]\}$. We write also $V_{\varepsilon}=2 \Delta_{2 \varepsilon}-\Delta_{\varepsilon}$ : this function is equal to 1 on $\{\theta \in[-\varepsilon, \varepsilon]\}$ and vanishes outside of $\{\theta \in[-2 \varepsilon, 2 \varepsilon]\}$. We will use $V_{\varepsilon}$ and $1-V_{\varepsilon}$ as a $C^{0}$ partition of unity. By [Kah70, page 56], we have $\left\|V_{\varepsilon}\right\|_{A(\mathbb{C})} \leqslant 3$. Note that $V_{\varepsilon}$ is piecewise $C^{1}$, its derivative being bounded by $1 / \varepsilon$.

If $g: \mathbb{T} \rightarrow \mathbb{C}$ is continuous and piecewise $C^{1}$, then ([Kah70, page 56 , Equation (1)])

$$
\begin{equation*}
\|g\|_{A(\mathbb{C})} \leqslant\left|c_{0}(g)\right|+\left(\frac{\pi}{12} \int_{-\pi}^{\pi}\left|g^{\prime}(\theta)\right|^{2} \mathrm{~d} \theta\right)^{1 / 2} \tag{7}
\end{equation*}
$$

Finally, if $f: S^{1} \rightarrow \mathbb{C}$ is in $A(\mathbb{C})$ and satisfies $f(1)=0$, then ([Kah70, page 56])

$$
\begin{equation*}
\left\|f V_{\varepsilon}\right\|_{A(\mathbb{C})} \xrightarrow[\varepsilon \rightarrow 0]{ } 0 \tag{8}
\end{equation*}
$$

Fix $|t|<\varepsilon_{0}$ and write $\psi_{t}=V_{G(|t|)}$ for $G(t)=M(t)^{1 / 3}$, and $\varphi_{t}=1-\psi_{t}$, on $S^{1}$ or $\mathbb{T}$. Setting $A_{t}(z)=\frac{1-\tilde{\lambda}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}$, we have to check that $A_{t}(z)$ tends (in $A(\mathbb{C}))$ to $B(z)=\frac{1-\tilde{\lambda}(z)}{1-z}$. We will prove that $A_{t}(z) \psi_{t}(z)-B(z) \psi_{t}(z) \rightarrow 0$, and $A_{t}(z) \varphi_{t}(z)-B(z) \varphi_{t}(z) \rightarrow 0$.

In this proof, we denote by $C(z)$ a function whose norm in $A(\mathbb{C})$ is bounded by a constant. For example, as $\left\|V_{t}\right\|_{A(\mathbb{C})} \leqslant 3$, we sometimes replace $V_{t}(z)$ by $C(z)$. As $\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}=1-\frac{c}{\mu} M(|t|) \sum_{n=1}^{\infty}\left(1-\frac{c}{\mu} M(|t|)\right)^{n-1} z^{n}$, we also have $\left\|\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right\|_{A(\mathbb{C})} \leqslant 1+\frac{\frac{|c|}{\mu} M(|t|)}{1-\left|1-\frac{c}{\mu} M(|t|)\right|}$, which is bounded when $t \rightarrow 0$ (since $\operatorname{Re} c>0)$. Thus, we may also replace $\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}$ by $C(z)$.

Using exactly the same equations as in the proof of Lemma 2.6 (with tildes everywhere) and Wiener's Lemma, we prove that $K(z, t)=\frac{\tilde{\lambda}(z, t)-\tilde{\lambda}(1, t)}{z-1}-\frac{\tilde{\lambda}(z)-\tilde{\lambda}(1)}{z-1}$ tends to 0 in $A(\mathbb{C})$ when $t \rightarrow 0$.

Step 1 (close to $z=1$ ). $A_{t}(z) \psi_{t}(z)-B(z) \psi_{t}(z)$ tends to 0 in $A(\mathbb{C})$-norm.
We proved in the third step of the proof of Theorem 2.4 that $\frac{\tilde{\lambda}(z)-\tilde{\lambda}(1)}{z-1} \in A(\mathbb{C})$. As this quotient tends to $\mu$ when $z \rightarrow 1$, we can write $\frac{\tilde{\lambda}(z)-\tilde{\lambda}(1)}{z-1}=\mu+F(z)$ where $F \in A(\mathbb{C})$ and $F(1)=0$. We also write $\tilde{\lambda}(1, t)=1-(c+b(t)) M(|t|)$ (since $\tilde{\lambda}(1, t)=\lambda(1, t)$, as $\tilde{R}(z)=R(z)$ in a neighborhood of 1$)$. This gives

$$
\tilde{\lambda}(z, t)=1-(c+b(t)) M(|t|)+(z-1)(K(z, t)+F(z)+\mu) .
$$

Thus,

$$
\begin{aligned}
A_{t}(z)= & \frac{1-\tilde{\lambda}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}=\mu+\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}(F(z) \\
& +K(z, t)+c M(|t|))+\frac{M(|t|)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} b(t)
\end{aligned}
$$

Therefore,

$$
\begin{aligned}
A_{t}(z) \psi_{t}(z)= & \mu \psi_{t}(z)+C(z) F(z) \psi_{t}(z)+C(z)(K(z, t)+M(|t|)) \\
& +\frac{M(|t|)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} b(t) \psi_{t}(z) .
\end{aligned}
$$

A small computation shows that $\frac{M(|t|)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}$ has a bounded norm in $A(\mathbb{C})$. Moreover, $\mu=B(z)-F(z)$, and the term $F(z) \psi_{t}(z)$ can be included in $C(z) F(z) \psi_{t}(z)$. Finally,

$$
A_{t}(z) \psi_{t}(z)-B(z) \psi_{t}(z)=C(z) F(z) \psi_{t}(z)+C(z)(M(|t|)+K(z, t)+b(t)) .
$$

In the rightmost term, everything tends to 0 as $t \rightarrow 0$. Moreover, Equation (8) ensures that $F(z) \psi_{t}(z) \rightarrow 0$. This concludes the first step.

Step 2 (away from $z=1$ ). $A_{t}(z) \varphi_{t}(z)-B(z) \varphi_{t}(z)$ tends to 0 in $A(\mathbb{C})$-norm.
We start from

$$
\begin{aligned}
\frac{1-\tilde{\lambda}(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}-\frac{1-\tilde{\lambda}(z)}{1-z}= & \frac{1-\tilde{\lambda}(z)}{1-z}\left(\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}-1\right) \\
& +\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} \frac{\tilde{\lambda}(z)-\tilde{\lambda}(z, t)}{1-z}
\end{aligned}
$$

We will see that each term on the right, multiplied by $\varphi_{t}$, tends to 0 in $A(\mathbb{C})$ when $t \rightarrow 0$.

We will use that, on the support of $\varphi_{t}$, we have $|z-1| \geqslant C G(|t|)$, which implies that $\left|1-\left(1-\frac{c}{\mu} M(|t|)\right) z\right| \geqslant C G(|t|)-D M(|t|) \geqslant C G(|t|)$ (for a smaller constant $C)$ since $M(t)=o(G(t))$. In the following, $C$ will denote a generic constant.

## The first term

As $\frac{1-\tilde{\lambda}(z)}{1-z} \in A(\mathbb{C})$, it is enough to prove the convergence of $f=\left(\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}-1\right)$ $\varphi_{t}(z)$ to 0 . We will use Equation (7).

We have

$$
|f(z)| \leqslant \frac{\frac{|c|}{\mu} M(|t|)}{C G(t)} \leqslant C M(|t|)^{2 / 3} .
$$

In particular, $\left|c_{0}(f)\right| \leqslant C M(|t|)^{1 / 3}$.
For the derivative, writing $f(\theta)=h(\theta) \varphi_{t}(\theta)$, we have $f^{\prime}=h^{\prime} \varphi_{t}+h \varphi_{t}^{\prime}$. As $|h| \leqslant C M(|t|)^{2 / 3}$ and $\left|\varphi_{t}^{\prime}\right| \leqslant 1 / G(|t|)$, we get $\left|h \varphi_{t}^{\prime}\right| \leqslant C M(|t|)^{1 / 3}$.

To control $h^{\prime}$, we write

$$
h(\theta)=-\frac{\frac{c}{\mu} M(|t|)}{1-\frac{c}{\mu} M(|t|)}+\frac{\frac{c}{\mu} M(|t|)}{1-\frac{c}{\mu} M(|t|)} \frac{1}{1-\left(1-\frac{c}{\mu} M(|t|)\right) e^{i \theta}} .
$$

Thus,

$$
h^{\prime}(\theta)=i \frac{c M(|t|)}{\mu} \frac{1}{\left(1-\left(1-\frac{c}{\mu} M(|t|)\right) e^{i \theta}\right)^{2}} .
$$

On the support of $\varphi_{t}$, the modulus of the denominator on the right is $\geqslant(C G(|t|))^{2}$, which gives $\left|h^{\prime}\right| \leqslant C M(|t|)^{1 / 3}$.

We have shown that $\left|f^{\prime}\right| \leqslant C M(|t|)^{1 / 3}$. As $\left|c_{0}(f)\right| \leqslant C M(|t|)^{1 / 3}$, Equation (7) gives that $\|f\|_{A(\mathbb{C})} \leqslant C M(|t|)^{1 / 3}$.

The second term
As $\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}$ is bounded in $A(\mathbb{C})$, it is enough to show that $g(z)=\frac{\tilde{\lambda}(z, t)-\tilde{\lambda}(z)}{1-z}$ $\varphi_{t}(z)$ tends to 0 in $A(\mathbb{C})$.

Writing again $K(z, t)=\frac{\tilde{\lambda}(z, t)-\tilde{\lambda}(1, t)}{z-1}-\frac{\tilde{\lambda}(z)-\tilde{\lambda}(1)}{z-1}$ (this difference tends to 0 when $t \rightarrow 0$ ), we have

$$
g(z)=K(z, t) \varphi_{t}(z)+(\tilde{\lambda}(1)-\tilde{\lambda}(1, t)) \frac{\varphi_{t}(z)}{1-z} .
$$

The first term $K(z, t) \varphi_{t}(z)$ tends to 0 . For the second term, if $h(z)=\frac{\varphi_{t}(z)}{1-z}$, we have $\left|c_{0}(h)\right| \leqslant \frac{1}{C G(t t \mid)}$ and

$$
\left|h^{\prime}(\theta)\right|=\left|i \frac{\varphi_{t}(\theta)}{\left(1-e^{i \theta}\right)^{2}}+\frac{1}{1-e^{i \theta}} \varphi_{t}^{\prime}(\theta)\right| \leqslant \frac{1}{C^{2} G(|t|)^{2}}+\frac{1}{C G(|t|)} \frac{1}{G(|t|)}
$$

Equation (7) gives that $\|h\|_{A(\mathbb{C})} \leqslant \frac{C}{G(|t|)^{2}}=O\left(M(|t|)^{-2 / 3}\right)$. As $\tilde{\lambda}(1)-\tilde{\lambda}(1, t)=$ $O(M(|t|))$, we finally get $\left\|(\tilde{\lambda}(1)-\tilde{\lambda}(1, t)) \frac{\varphi_{t}(z)}{1-z}\right\|_{A(\mathbb{C})}=O\left(M(|t|)^{1 / 3}\right)$, which concludes the proof of Step 2 of the proof of Lemma 2.9.

### 2.6. Conclusion of the proof of Theorem 2.1

Proposition 2.10. We have

$$
\left\|\left(\frac{I-R(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right)^{-1}-\left(\frac{I-R(z)}{1-z}\right)^{-1}\right\|_{A(\mathcal{A})} \underset{t \rightarrow 0}{\longrightarrow} 0
$$

Proof. Let $\chi_{1}, \chi_{2}$ be a partition of unity such that $R=\tilde{R}$ on the support of $\chi_{1}$, and set $A(z, t)=\left(\frac{I-R(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right)^{-1}$.

Proposition 2.8 shows that $\chi_{1}(z) A(z, t) \rightarrow \chi_{1}(z) A(z, 0)$. For the term $\chi_{2}(z)$ $A(z, t)$, we can safely modify $R$ on a small neighborhood of 1 so that $I-R(z)$ is everywhere invertible (this does not change $\chi_{2}(z) A(z, t)$ ). Wiener's Lemma and the continuity of inversion then give that $\chi_{2}(z) A(z, t) \rightarrow \chi_{2}(z) A(z, 0)$.

Proof of Theorem 2.1. We now conclude the proof of Theorem 2.1, i.e. we construct functions $\varepsilon(t)$ and $\delta(n)$ satisfying Equation (2).

Proposition 2.10 shows that it is possible to write

$$
\left(\frac{I-R(z, t)}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}\right)^{-1}=\left(\frac{I-R(z)}{1-z}\right)^{-1}+F_{t}(z)
$$

where $F_{t}(z) \in A(\mathcal{A})$ tends to 0 when $t \rightarrow 0$. Moreover, Lemma 2.5 gives that

$$
\left(\frac{I-R(z)}{1-z}\right)^{-1}=\frac{1}{\mu} P+(1-z) A(z)
$$

where $A(z)=\sum A_{n} z^{n}$ with $A_{n} \rightarrow 0$. We get

$$
\begin{aligned}
& \sum T_{n, t} z^{n}=(I-R(z, t))^{-1}=\frac{1}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} \frac{1}{\mu} P \\
& \quad+\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} A(z)+\frac{1}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} F_{t}(z)
\end{aligned}
$$

In the first term, the coefficient of $z^{n}$ is $\frac{1}{\mu}\left(1-\frac{c}{\mu} M(|t|)\right)^{n} P$ and corresponds to (2). We have to check that the other terms are smaller than $\varepsilon(t)+\delta(n)$.

Let us denote by $\delta(n)$ the maximum for $t \in\left[-\varepsilon_{0}, \varepsilon_{0}\right]$ of the coefficient $c_{n}(t)$ of $z^{n}$ in $\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} A(z)$. We will prove that $\delta(n) \rightarrow 0$ as $n \rightarrow \infty$. We have

$$
\frac{1-z}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}=1-\frac{c}{\mu} M(|t|) \sum_{n=1}^{\infty}\left(1-\frac{c}{\mu} M(|t|)\right)^{n-1} z^{n} .
$$

Thus,

$$
c_{n}(t)=A_{n}-\frac{c}{\mu} M(|t|) \sum_{k=0}^{n-1}\left(1-\frac{c}{\mu} M(|t|)\right)^{n-k-1} A_{k}
$$

Note that, since $\operatorname{Re}(c)>0$ and $M(t) \rightarrow 0$ when $t \rightarrow 0$, we have $\left|1-\frac{c}{\mu} M(|t|)\right| \leqslant$ $1-\frac{\operatorname{Re} c}{2 \mu} M(|t|)$ when $t$ is small enough. Let $C$ be greater than the $\left\|A_{n}\right\|, n \in \mathbb{N}$. We fix $\varepsilon>0$, and $N$ such that $\forall k \geqslant N,\left\|A_{k}\right\| \leqslant \varepsilon$. Then, if $n \geqslant N$,

$$
\begin{aligned}
\left\|c_{n}(t)\right\| \leqslant & \left\|A_{n}\right\|+\frac{|c|}{\mu} M(|t|) \sum_{k=0}^{N-1}\left(1-\frac{\operatorname{Re} c}{2 \mu} M(|t|)\right)^{n-k-1}\left\|A_{k}\right\| \\
& +\frac{|c|}{\mu} M(|t|) \sum_{k=N}^{n-1}\left(1-\frac{\operatorname{Re} c}{2 \mu} M(|t|)\right)^{n-k-1} \varepsilon \\
\leqslant & \varepsilon+N \frac{|c|}{\mu} C M(|t|)\left(1-\frac{\operatorname{Re} c}{2 \mu} M(|t|)\right)^{n-N}+E \varepsilon .
\end{aligned}
$$

As $M(|t|)\left(1-\frac{\operatorname{Re} c}{2 \mu} M(|t|)\right)^{n-N}$ tends uniformly to 0 on $\left[-\varepsilon_{0}, \varepsilon_{0}\right]$ (if $\varepsilon_{0}$ is small enough) when $n \rightarrow \infty$, we get that for $n$ large enough, $\left\|c_{n}(t)\right\| \leqslant(2+E) \varepsilon$, and $\delta(n) \leqslant(2+E) \varepsilon$.

Let us finally bound the coefficients of $\frac{1}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z} F_{t}(z)$, which will give the $\varepsilon(t)$. The Fourier coefficients of $\frac{1}{1-\left(1-\frac{c}{\mu} M(|t|)\right) z}$ are $\leqslant 1$, which implies that
the coefficient of $z^{n}$ in the product is at most $\sum_{k=0}^{n}\left\|F_{t}(z)_{k}\right\| \leqslant\left\|F_{t}\right\|_{A(\mathcal{A})}$. Since $\left\|F_{t}\right\|_{A(\mathcal{A})} \rightarrow 0$ when $t \rightarrow 0$, we can take $\varepsilon(t)=\left\|F_{t}\right\|_{A(\mathcal{A})}$ to conclude.

## 3. Results in the uniform case

In Sections 3.1 and 3.2, we state some technical estimates on Markov maps, and we use them in Sections 3.3 and 3.4 to derive an expansion of the eigenvalue of a perturbed transfer operator, which will subsequently be used in the proofs of Theorems 1.1 and 1.2.

We work in the setting of Markov maps of Section 1.2.1, and we will use freely the definitions introduced here.

### 3.1. Some technical estimates

In this section, $(Z, \mathcal{B}, T, m, \delta)$ will be a Markov map preserving a measure $m$ of finite mass. Write $t(x, y)$ for the separation time of the points $x$ and $y$ (under the iteration of $T$ ) and assume that the distortion of $T$ is $\tau$-Hölder for some $\tau<1$. For $f: Z \rightarrow \mathbb{C}$, we will denote by $D_{\tau} f$ the least $\tau$-Hölder constant of $f$ for points in the same elements of the partition, i.e. $D_{\tau} f=\inf \{c \mid \forall x, y \in Z, t(x, y) \geqslant$ $\left.1 \Rightarrow|f(x)-f(y)| \leqslant c \tau^{t(x, y)}\right\}$. Let also $\mathcal{L}_{\tau}$ be the space of bounded functions on $Z$ with $D_{\tau} f<\infty$. It is a Banach space for the norm $\|f\|_{\mathcal{L}_{\tau}}=\|f\|_{\infty}+$ $D_{\tau} f$.

We will need a property slightly weaker than the big image property from Section 1.2.1. We say that $Z$ is a bounded tower if there exists $q \in \mathbb{N}$, a subset $\delta_{0} \subset \delta$ (corresponding to the basis of the tower) and a function $R: \delta_{0} \rightarrow\{0, \ldots, q-1\}$, such that $Z$ is isomorphic to the set $\left\{(x, i) \mid x \in a \subset \delta_{0}, i<R(a)\right\}$, with the partition into the sets $a \times\{i\}$ for $a \in \delta_{0}$ and $i<R(a)$. We require $T$ to map $a \times\{i\}$ isomorphically onto $a \times\{i+1\}$ if $i+1<R(a)$, while if $i+1=R(a)$ then $T(a \times\{i\}) \subset \bigcup_{d \in \delta_{0}} d$, i.e. the points get out of the tower through the top and fall down to the basis. We will write $\Delta_{k}=\{(x, k) \mid x \in a, R(a)>k\}$, i.e. this is the set of points at height $k$ in the tower.

We will say that $T$ has the bounded tower big image property if $Z$ is a bounded tower, and the returns to the basis have measure bounded away from 0, i.e. $\forall a \in \delta$, $T a \subset \Delta_{0} \Rightarrow m[T a]>\eta$ for a constant $\eta>0$. Note that, when $q=1$, i.e. there are no floors in the tower, we recover the usual big image property.

In the rest of Section 3.1, $T$ will have the bounded tower big image property and the distortion $\log g_{m}$ will be $\tau$-Hölder.

Lemma 3.1 (Distortion Lemma). Writing $g_{m}^{(n)}(x)=g_{m}(x) \cdots g_{m}\left(T^{n-1} x\right)$ (this is the inverse of the jacobian of $\left.T^{n}\right)$, there exists a constant $C$ such that $t(x, y) \geqslant$ $n \Rightarrow\left|1-\frac{g_{m}^{(n)}(x)}{g_{m}^{(n)}(y)}\right| \leqslant C \tau^{t(x, y)-n}$.

In particular, there exists a constant $D$ such that, if $d=\left[d_{0}, \ldots, d_{n-1}\right]$ is a cylinder of length $n$ and $x \in \underline{d}$, then $D^{-1} g_{m}^{(n)}(x) \leqslant \frac{m[d]}{m\left[T^{n} d\right]} \leqslant D g_{m}^{(n)}(x)$.

Proof. Let $x, y$ be such that $t(x, y) \geqslant n$. Then

$$
\begin{aligned}
\left|\log \left(g_{m}^{(n)}(x)\right)-\log \left(g_{m}^{(n)}(y)\right)\right| & \leqslant \sum_{i=0}^{n-1}\left|\log \left(g_{m}\left(T^{i} x\right)\right)-\log \left(g_{m}\left(T^{i} y\right)\right)\right| \\
& \leqslant \sum_{i=0}^{n-1} D_{\tau}\left(\log \left(g_{m}\right)\right) \tau^{t(x, y)-i} \\
& \leqslant \frac{D_{\tau}\left(\log \left(g_{m}\right)\right)}{1-\tau} \tau^{t(x, y)-n+1}
\end{aligned}
$$

Taking exponentials, we get the first statement of the lemma.
In particular, if $\underline{d}$ is a cylinder of length $n$ and $x, y \in \underline{d}$, we have $\frac{g_{m}^{(n)}(x)}{g_{m}^{(n)}(y)} \leqslant D$ for some $D$. Using that $1 / g_{m}^{(n)}(y)$ it the jacobian of $T^{n}$, we integrate on $y \in \underline{d}$ and get that $g_{m}^{(n)}(x) m\left[T^{n} \underline{d}\right] \leqslant \operatorname{Dm}[\underline{d}]$. The other inequality is analogous.

If $\psi$ is a function on $Z$ and $\hat{T}$ denotes the transfer operator associated to $T$, then $\hat{T}^{n} \psi(x)=\sum_{T^{n}(y)=x} g_{m}^{(n)}(y) \psi(y)$. There is at most one preimage of $x$ in every non-empty cylinder $\underline{d}=\left[d_{0}, \ldots, d_{n-1}\right]$, which we denote by $d_{0} \ldots d_{n-1} x$. Set $M_{\underline{d}} \psi(x)=g_{m}^{(n)}\left(d_{0} \ldots d_{n-1} x\right) \psi\left(d_{0} \ldots d_{n-1} x\right)$ if this point is defined, and 0 otherwise. Then $\hat{T}^{n}=\sum M_{\underline{d}}$ where the sum is over all non-empty cylinders $\underline{d}$ of length $n$.

Lemma 3.2. There exists a constant $B$ such that $\forall \underline{d}=\left[d_{0}, \ldots, d_{n-1}\right], \forall \psi$ : $Z \rightarrow \mathbb{C}$,

$$
\left\|M_{\underline{d}} \psi\right\|_{\mathcal{L}_{\tau}} \leqslant B \frac{m[\underline{d}]}{m\left[T^{n} \underline{d}\right]}\left(\tau^{n} D_{\tau} \psi+\frac{1}{m[\underline{d}]} \int_{[\underline{d}]}|\psi| \mathrm{d} m\right) .
$$

Proof. Let $x$ and $y$ be in the same partition element $a$. If $a$ is not in the image of $T\left(d_{n-1}\right),\left(M_{\underline{d}} \psi\right)_{\mid a}=0$ and there is nothing to prove. Otherwise, write $x^{\prime}=$ $d_{0} \ldots d_{n-1} x$, and $y^{\prime}=d_{0} \ldots d_{n-1} y$.

First note that, for $z^{\prime} \in \underline{d}$, we have $\left|\psi\left(y^{\prime}\right)-\psi\left(z^{\prime}\right)\right| \leqslant D_{\tau} \psi \tau^{n}$, which gives, after integrating on $z^{\prime}$, that $\left|\psi\left(y^{\prime}\right)\right| \leqslant D_{\tau} \psi \tau^{n}+\frac{1}{m[\underline{d}]} \int_{\underline{d}}|\psi| \mathrm{d} m$. The inequality $g_{m}^{(n)}\left(y^{\prime}\right) \leqslant D \frac{m[d]}{m\left[T^{n} d\right]}$ from Lemma 3.1 gives then the claimed bound on $\left\|M_{\underline{d}} \psi\right\|_{\infty}$.

Using Lemma 3.1, we have

$$
\begin{aligned}
\left|M_{\underline{d}} \psi(x)-M_{\underline{d}} \psi(y)\right| \leqslant & \left|g_{m}^{(n)}\left(x^{\prime}\right)\right|\left|\psi\left(x^{\prime}\right)-\psi\left(y^{\prime}\right)\right| \\
& +\left|\psi\left(y^{\prime}\right)\right|\left|g_{m}^{(n)}\left(y^{\prime}\right)\right|\left|1-\frac{g_{m}^{(n)}\left(x^{\prime}\right)}{g_{m}^{(n)}\left(y^{\prime}\right)}\right| \\
\leqslant & D \frac{m[\underline{d}]}{m\left[T^{n} \underline{d}\right]} D_{\tau} \psi \tau^{t(x, y)+n}+\left|\psi\left(y^{\prime}\right)\right| D \frac{m[\underline{d}]}{m\left[T^{n} \underline{d}\right]} C \tau^{t(x, y)}
\end{aligned}
$$

which gives the conclusion, using the aforementioned bound on $\left|\psi\left(y^{\prime}\right)\right|$.

Corollary 3.3. If $T$ has the bounded tower big image property and is ergodic, then the associated transfer operator $\hat{T}$ acts continuously on $\mathcal{L}_{\tau}$, and it has a simple isolated eigenvalue at 1 , the eigenspace being the constant functions.

Proof. The proof will use Lemma 3.2 and the fact that $\hat{T}^{n}=\sum M_{\underline{d}}$, where the sum is over all cylinders of length $n$.

We first show the continuity of $\hat{T}$. Decompose $\hat{T}=1_{\Delta_{0}} \hat{T}+\left(1-1_{\Delta_{0}}\right) \hat{T}$, where the first term sees the points on the basis and the second one in the floors. If $x$ is in the basis, the elements $a$ of the partition such that $x \in T a$ return to the basis, which implies that they have a big image. We can thus sum the estimates given by Lemma 3.2 for cylinders of length 1 , and forget the $m[T \underline{d}]$, to get that $\left\|1_{\Delta_{0}} \hat{T} \psi\right\|_{\mathcal{L}_{\tau}} \leqslant B\left(\tau\|\psi\|_{\mathcal{L}_{\tau}}+\|\psi\|_{1}\right)$. On the other hand, if $x$ is not in the basis and $x^{\prime}$ denotes the point just below, then $\hat{T} \psi(x)=\psi\left(x^{\prime}\right)$. This implies that $\left\|\left(1-1_{\Delta_{0}}\right) \hat{T}\right\|_{\mathcal{L}_{\tau}} \leqslant 1$, and concludes the proof of the continuity of $\hat{T}$.

For $n \geqslant q$ (where $q$ is the height of the tower), we can use the same argument on the basis and get that

$$
\left\|\left(\hat{T}^{n} \psi\right)_{\mid \Delta_{0}}\right\|_{\mathcal{L}_{\tau}} \leqslant B\left(\tau^{n}\|\psi\|_{\mathcal{L}_{\tau}}+\|\psi\|_{1}\right) .
$$

If $x \in \Delta_{k}$, let $x^{\prime}$ denote the corresponding point in the basis of the tower. Then $\hat{T}^{n} \psi(x)=\hat{T}^{n-k} \psi\left(x^{\prime}\right)$, and the inequality on the basis gives

$$
\left\|\left(\hat{T}^{n} \psi\right)_{\mid \Delta_{k}}\right\|_{\mathcal{L}_{\tau}} \leqslant B\left(\tau^{n-k}\|\psi\|_{\mathcal{L}_{\tau}}+\|\psi\|_{1}\right) .
$$

As the tower is bounded, we finally get that

$$
\left\|\hat{T}^{n} \psi\right\|_{\mathcal{L}_{\tau}} \leqslant \frac{B}{\tau^{q}}\left(\tau^{n}\|\psi\|_{\mathcal{L}_{\tau}}+\|\psi\|_{1}\right) .
$$

This is a so-called Doeblin-Fortet inequality ([ITM50]), since the inclusion $\mathcal{L}_{\tau} \rightarrow$ $L^{1}$ is compact. Thus, by Hennion's Theorem ([Hen93]) the essential spectral radius of $\hat{T}$ is bounded by $\tau$, and 1 is an isolated eigenvalue of $\hat{T}$. The constant functions are eigenfunctions, and ergodicity of $T$ shows that they are the only eigenfunctions ([Aar97, Theorem 1.4.8]). Moreover, there is no nilpotent part since $\left\|\hat{T}^{n}\right\|$ is bounded.

The following lemma will be useful later, applied to unbounded functions in $L^{2}$. We recall that, if $d$ is an element of the partition $\delta$, then $D_{\tau} h(d)$ is the least $\tau$-Hölder constant of $h$ restricted to $d$.

Lemma 3.4. Assume that $T$ has the bounded tower big image property. Let $h \in$ $L^{1}(m)$ satisfy $\sum_{d \in \delta} m[d] D_{\tau} h(d)<\infty$. Assume that $h$ is bounded on the set of points whose image is not in the basis of the tower, and that $h$ is uniformly $\tau$-Hölder on this set. Then $\hat{T} h \in \mathcal{L}_{\tau}$.

Proof. For $\underline{d}=\left[d_{0}\right]$ of length 1, the estimates given by Lemma 3.2 depend in fact only on $D_{\tau} \psi\left(d_{0}\right)$. Thus, summing these estimates on the basis, and using the big image property, we get

$$
\left\|(\hat{T} h)_{\mid \Delta_{0}}\right\|_{\mathcal{L}_{\tau}} \leqslant B\left(\sum_{d \in \delta} \tau m[d] D_{\tau} h(d)+\|h\|_{1}\right)<\infty .
$$

If $x$ is in the floors, and $x^{\prime}$ is just below $x$, then $\hat{T} h(x)=h\left(x^{\prime}\right)$. As $h$ is bounded and Hölder on the set of such points $x^{\prime}$ by hypothesis, we get that $\hat{T} h$ is in $\mathcal{L}_{\tau}$.

### 3.2. Some technical estimates: the perturbative case

In this section, $(Z, \mathcal{B}, T, m, \delta)$ will satisfy the same assumptions as in the previous section, i.e. $T$ is a Markov map, preserving the finite measure $m$, with Hölder distortion. Moreover, $T$ has the bounded tower big image property. Additionally, we will use a function $h: Z \rightarrow \mathbb{C}$ which, on the set of points whose image is not in the basis, will be bounded and uniformly $\tau$-Hölder, and satisfying $\sum_{d \in \delta} m[d] D_{\tau} h(d)<\infty$.

Let $\underline{d}=\left[d_{0}\right]$ be a non-empty cylinder in $Z$ of length 1 . For small $t$, we define a perturbation $M_{\underline{d}}(t)$ of $M_{\underline{d}}$ by

$$
M_{\underline{d}}(t) \psi(x)= \begin{cases}e^{i t h\left(d_{0} x\right)} g_{m}\left(d_{0} x\right) \psi\left(d_{0} x\right) & \text { if } d_{0} x \text { is defined }  \tag{9}\\ 0 & \text { otherwise }\end{cases}
$$

Lemma 3.5. Take $0<\eta \leqslant 1$. Then there exists a constant $B$ such that $\forall \underline{d}=\left[d_{0}\right]$, $\forall t \in[-1,1]$,

$$
\left\|M_{\underline{d}}(t)-M_{\underline{d}}\right\|_{\mathcal{L}_{\tau} \eta} \leqslant B \frac{m[\underline{d}]}{m[T \underline{d}]}\left(|t|^{\eta} D_{\tau} h(\underline{d})^{\eta}+\frac{1}{m[\underline{d}]} \int_{\underline{d}}\left|e^{i t h}-1\right| \mathrm{d} m\right) .
$$

Note that the norm in this lemma is the $\mathcal{L}_{\tau^{\eta}}$ norm.
Proof. Let $\psi \in \mathcal{L}_{\tau^{\eta}}$. If $x, y$ are in the same element of the partition, we have to study $\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(x)-\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(y)$. If $d_{0} x$ is not defined, then $d_{0} y$ is not defined either, and there is nothing to prove. Otherwise, denote $x^{\prime}=d_{0} x$ and $y^{\prime}=d_{0} y$. Then

$$
\begin{aligned}
& \left|\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(x)-\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(y)\right| \\
& \quad \leqslant\left|e^{i t h\left(x^{\prime}\right)}-1\right|\left|M_{\underline{d}} \psi(x)-M_{\underline{d}} \psi(y)\right|+\left|M_{\underline{d}} \psi(y)\right|\left|e^{i t h\left(x^{\prime}\right)}-e^{i t h\left(y^{\prime}\right)}\right| .
\end{aligned}
$$

Take $C$ such that $\left|e^{i s}-1\right| \leqslant C|s|^{\eta}$ for all real $s$. Then

$$
\left|e^{i t h\left(x^{\prime}\right)}-e^{i t h\left(y^{\prime}\right)}\right| \leqslant C|t|^{\eta}\left|h\left(x^{\prime}\right)-h\left(y^{\prime}\right)\right|^{\eta} \leqslant C|t|^{\eta}\left(D_{\tau} h(\underline{d}) \tau^{t(x, y)+1}\right)^{\eta}
$$

If $\chi=e^{i t h}-1$, we see that $\left|\chi\left(x^{\prime}\right)\right| \leqslant\left|\chi\left(y^{\prime}\right)\right|+C|t|^{\eta} D_{\tau} h(\underline{d})^{\eta}$ on the cylinder $\underline{d}$, whence $\left|e^{i t h\left(x^{\prime}\right)}-1\right| \leqslant \frac{1}{m[\underline{d}]} \int_{\underline{d}}\left|e^{i t h}-1\right| \mathrm{d} m+C|t|^{\eta} D_{\tau} h(\underline{d})^{\eta}$. The estimates on $M_{\underline{d}} \psi$ from Lemma 3.2 (in the space $\mathcal{L}_{\tau^{\eta}}$ ) imply that

$$
\begin{aligned}
& \left|\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(x)-\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi(y)\right| \\
& \leqslant\left(\frac{1}{m[\underline{d}]} \int_{\underline{d}}\left|e^{i t h}-1\right| \mathrm{d} m+C|t|^{\eta} D_{\tau} h(\underline{d})^{\eta}\right) \\
& \quad \times B \tau^{\eta t(x, y)} \frac{m[\underline{d}]}{m[T \underline{d}]}\left(\tau^{\eta} D_{\tau^{\eta}} \psi+\frac{1}{m[\underline{d}]} \int_{[\underline{d}]}|\psi| \mathrm{d} m\right) \\
& \quad+B \frac{m[\underline{d}]}{m[T \underline{d}]}\left(\tau^{\eta} D_{\tau^{\eta}} \psi+\frac{1}{m[\underline{d}]} \int_{[\underline{d}]}|\psi| \mathrm{d} m\right) C|t|^{\eta} D_{\tau} h(\underline{d})^{\eta} \tau^{\eta t(x, y)} .
\end{aligned}
$$

This gives the estimate on $D_{\tau^{\eta}}\left(\left(M_{\underline{d}}(t)-M_{\underline{d}}\right) \psi\right)$, using $\left(\tau^{\eta} D_{\tau^{\eta}} \psi+\frac{1}{m[\underline{d}]} \int_{[\underline{d}]}|\psi| \mathrm{d} m\right)$ $\leqslant\|\psi\|_{\mathcal{L}_{\tau} \eta}$. The bound on the supremum norm is analogous.

Corollary 3.6. Assume that $T$ has the bounded tower big image property. Writing $R_{t}=\hat{T}\left(e^{i t h} \cdot\right)$, then $t \mapsto R_{t}$ is continuous at 0 , and

$$
\left\|R_{t}-R_{0}\right\|_{\mathcal{L}_{\tau}}=O\left(|t|+E\left(\left|e^{i t h}-1\right|\right)\right)
$$

In particular, when $h \in L^{1},\left\|R_{t}-R_{0}\right\|_{\mathcal{L}_{\tau}}=O(|t|)$.
The notation $E$ denotes the expectation, i.e. the integral with respect to the measure $\mathrm{d} m$.

Proof. We will use that $R_{t}=\hat{T}\left(e^{i t h}\right)=\sum M_{\underline{d}}(t)$, where the sum is over the cylinders of length 1.

To prove continuity of $t \mapsto R_{t}$ at $t=0$, we write $R_{t}=1_{\Delta_{0}} R_{t}+\left(1-1_{\Delta_{0}}\right) R_{t}$. On the basis, summing the estimates of Lemma 3.5 (with $\eta=1$ ) and using the big image property, we get

$$
\left\|1_{\Delta_{0}} R_{t}-1_{\Delta_{0}} R_{0}\right\|_{\mathcal{L}_{\tau}} \leqslant B\left(|t| \sum m[\underline{d}] D_{\tau} h(\underline{d})+\int\left|e^{i t h}-1\right|\right) .
$$

In the floors, if $x^{\prime}$ is the point just below $x$, then $R_{t} \psi(x)=e^{i t h\left(x^{\prime}\right)} \psi\left(x^{\prime}\right)$, which yields that $\left\|\left(1-1_{\Delta_{0}}\right) R_{t}-\left(1-1_{\Delta_{0}}\right) R_{0}\right\|$ is bounded by the $\mathcal{L}_{\tau}$-norm of $e^{i t h}-1$ restricted to the set of points not returning to the basis at the next step. As $h$ is uniformly bounded and Hölder on these points, we get that this $\mathcal{L}_{\tau}$-norm is $O(|t|)$. This proves that $\left\|R_{t}-R_{0}\right\|=O\left(|t|+\int\left|e^{i t h}-1\right|\right)$. By the dominated convergence theorem, the integral tends to 0 when $t \rightarrow 0$, and this proves the continuity of $R_{t}$.

Moreover, when $h$ is integrable, then $\int\left|e^{i t h}-1\right| \leqslant|t| \int|h|$.

### 3.3. The classical method: central limit theorem case

In this section, we recall classical results on the central limit theorem, obtained in the uniform case by perturbing the transfer operator. For bounded $h$, the method dates back to [Nag57], and is developed in [RE83] and [GH88]. The result we prove here is slightly more complicated to obtain, since the function $h$ we will consider is generally unbounded, so that the perturbation of the transfer operator is not analytic, and we can not use an a priori expansion of the eigenvalue. In view of applications, we give the result for Markov maps with the bounded tower big image property.
Theorem 3.7. Let $(Z, \mathcal{B}, T, m, \delta)$ be a Markov map, preserving a finite ergodic measure m. Assume that $\log g_{m}$ is $\tau$-Hölder for some $\tau<1$, and that $T$ has the bounded tower big image property.

Let $h: Z \rightarrow \mathbb{C}$ with $\sum_{d \in \delta} m[d] D_{\tau} h(d)<\infty$ and $h \in L^{2}$, with $\int h \mathrm{~d} m=0$. Assume that, restricted to the set of points that do not return to the basis of the tower at the next iteration, $h$ is bounded and uniformly $\tau$-Hölder.

Denote by $R$ the transfer operator associated to $T$ (acting on $\mathcal{L}_{\tau}$ ), and $R_{t} \psi=$ $R\left(e^{i t h} \psi\right)$. Then, for small t, $R_{t}$ has an eigenvalue $\lambda(t)$, close to 1 , satisfying $\lambda(t)=$ $1-\frac{\sigma^{2}}{2 m(Z)} t^{2}+o\left(t^{2}\right)$. Write $a=(I-R)^{-1}(R h)$ (where we consider $I-R$ acting on the space of functions of $\mathcal{L}_{\tau}$ with 0 average, on which it is invertible) and $u=(I-R)^{-1}(h)$. Then

$$
\sigma^{2}=\int h^{2} \mathrm{~d} m+2 \int a h \mathrm{~d} m=\int R\left(u^{2}\right)-(R u)^{2} \mathrm{~d} m
$$

Moreover, $\sigma^{2}=0$ if and only if there exists a function $\psi \in \mathcal{L}_{\tau}$ with $h=\psi \circ T-\psi$.
Note that Lemma 3.4 ensures that $R h \in \mathcal{L}_{\tau}$, which implies that it makes sense to consider $a=(I-R)^{-1}(R h)$, since $I-R$ is invertible on the functions of $\mathcal{L}_{\tau}$ of zero integral, according to Corollary 3.3. However, $u=(I-R)^{-1}(h)$ is a priori not defined, since $h \notin \mathcal{L}_{\tau}$. In fact, we simply set $u=h+a$, and indeed $(I-R) u=h-R h+(I-R) a=h$.
Proof. We can replace $m$ by $m / m(Z)$ and assume that $m$ is a probability measure (the greater generality in the statement of the theorem will be useful in the applications).

Corollary 3.3 gives that $R=\hat{T}$ is continuous on $\mathcal{L}_{\tau}$, and 1 is a simple isolated eigenvalue. According to Corollary 3.6, $t \mapsto R_{t}$ is continuous at 0 . As simple isolated eigenvalues depend continuously on the operator, $R_{t}$ has a unique eigenvalue $\lambda(t)$ close to 1 for $t$ small. Write $P_{t}$ for the corresponding spectral projection, and $\xi_{t}$ for the eigenfunction (with $\int \xi_{t}=1$ ).

Since eigenvalues and eigenfunctions depend holomorphically on operators (Lipschitz would be enough), $\left\|\xi_{t}-1\right\|=O\left(\left\|R_{t}-R\right\|\right)=O(|t|)$ using Corollary 3.6. In the same way, $\lambda(t)=1+O(t)$.

Note for future use that

$$
\begin{equation*}
E\left(e^{i t h}\right)=1-\frac{t^{2}}{2} \int h^{2} \mathrm{~d} m+o\left(t^{2}\right) \tag{10}
\end{equation*}
$$

Indeed, if we consider $\Phi(t)=E\left(e^{i t h}\right)$ as the characteristic function of the random variable $h$, it is classical that if $h \in L^{2}$, then $\Phi$ is $C^{2}$ with $\Phi(0)=1, \Phi^{\prime}(0)=i E(h)$ and $\Phi^{\prime \prime}(0)=-E\left(h^{2}\right)([$ Fel66, Corollary to Lemma XV.4.2]).

As $\lambda(t) \xi_{t}=R_{t} \xi_{t}$, we get after integration that

$$
\begin{align*}
\lambda(t) & =\int R_{t} \xi_{t}=\int\left(R_{t}-R\right)\left(\xi_{t}-1\right)+\int\left(R_{t}-R\right)(1)+\int R \xi_{t} \\
& =\int\left(R_{t}-R\right)\left(\xi_{t}-1\right)+E\left(e^{i t h}-1\right)+1 \tag{11}
\end{align*}
$$

As $R_{t}-R=o(1)$ and $\xi_{t}-1=O(t)$, we get that $\lambda(t)=1+o(t)$ (using Equation (10) to estimate $E\left(e^{i t h}-1\right)$ ).

Thus,

$$
\begin{aligned}
\frac{\xi_{t}-\xi_{0}}{t} & =\frac{\lambda(t) \xi_{t}-\xi_{0}}{t}+o(1)=\frac{R_{t} \xi_{t}-R_{0} \xi_{0}}{t}+o(1) \\
& =\left(R_{t}-R_{0}\right) \frac{\xi_{t}-\xi_{0}}{t}+R_{0} \frac{\xi_{t}-\xi_{0}}{t}+\frac{R_{t}-R_{0}}{t} \xi_{0}+o(1) .
\end{aligned}
$$

But $\left(R_{t}-R_{0}\right)\left(\xi_{0}\right)=R\left(e^{i t h}-1\right)$. Moreover, $\frac{\xi_{t}-\xi_{0}}{t}$ is bounded and $R_{t}-R_{0}$ tends to 0 , whence $\left(R_{t}-R_{0}\right) \frac{\xi_{t}-\xi_{0}}{t}=o(1)$. Thus,

$$
(I-R) \frac{\xi_{t}-\xi_{0}}{t}=R\left(\frac{e^{i t h}-1}{t}\right)+o(1) .
$$

The sequence $\left(\xi_{t}-\xi_{0}\right) / t$ is bounded in $\mathcal{L}_{\tau}$. By compactness, we can extract a subsequence converging in $L^{2}$ to a function ia, which will still be in $\mathcal{L}_{\tau}$ (we have multiplied by $i$ for convenience). In $L^{2}, \frac{e^{i t h}-1}{t} \rightarrow i h$ by the dominated convergence theorem (as $\left|\frac{e^{i t h}-1}{t}\right|^{2} \leqslant|h|^{2}$ integrable), which implies that $(I-R) a=R(h)$. Moreover, according to Lemma 3.4, $R h \in \mathcal{L}_{\tau}$. We also have $\int R h=\int h=0$. As $I-R$ is invertible on the set of functions in $\mathcal{L}_{\tau}$ with zero integral, we get $a=(I-R)^{-1}(R h)$. Thus, $\left(\xi_{t}-\xi_{0}\right) / t$ has a unique cluster value, whence it converges (in $L^{2}$ ) to ia $=i(I-R)^{-1}(R h)$.

By continuity of the product $L^{2} \times L^{2} \rightarrow L^{1}$, we obtain the convergence of $\int \frac{e^{i t h}-1}{t} \frac{\xi_{t}-\xi_{0}}{t}$ to $\int(i h) \cdot(i a)$. Thus,

$$
\begin{aligned}
\int\left(R_{t}-R\right)\left(\xi_{t}-1\right) & =\int R\left(\left(e^{i t h}-1\right)\left(\xi_{t}-1\right)\right) \\
& =\int\left(e^{i t h}-1\right)\left(\xi_{t}-1\right)=-t^{2} \int h a+o\left(t^{2}\right)
\end{aligned}
$$

Equation (11) then gives that

$$
\begin{aligned}
\lambda(t) & =1+E\left(e^{i t h}-1\right)+\int\left(R_{t}-R\right)\left(\xi_{t}-1\right) \\
& =1-\frac{t^{2}}{2} \int h^{2}-t^{2} \int h a+o\left(t^{2}\right)
\end{aligned}
$$

We transform this result a little to obtain the other claimed expression for $\lambda(t)$. Recall that $h=u-R u$ and $a=u-h$.

$$
\begin{aligned}
-\lambda_{\mid t=0}^{\prime \prime} & =\int h^{2}+2 \int a h=\int h^{2}+2(u-h) h \\
& =\int(u-R u)^{2}+\int 2 R u \cdot(u-R u) \\
& =\int(u+R u)(u-R u)=\int u^{2}-\int(R u)^{2}=\int R\left(u^{2}\right)-(R u)^{2} .
\end{aligned}
$$

We have $R\left(u^{2}\right)(x)=\sum_{T y=x} g_{m}(y) u^{2}(y)$ and $(R u)^{2}(x)=\left(\sum_{T y=x} g_{m}(y) u(y)\right)^{2}$. Moreover, $\sum_{T y=x} g_{m}(y)=1$ since $R 1=1$. The convexity of $w \mapsto w^{2}$ then gives that $(R u)^{2}(x) \leqslant R\left(u^{2}\right)(x)$, which implies that $-\lambda^{\prime \prime} \geqslant 0$. We can thus write $\lambda^{\prime \prime}=-\sigma^{2}$.

Finally, if $\sigma^{2}=0$, then there is equality in the convexity inequality, whence all the $u(y)$ are equal (since $g_{m}$ is everywhere nonzero as $\log g_{m}$ is defined everywhere). Writing $\psi=R u$ (with $\psi \in \mathcal{L}_{\tau}$ since $R u=R a+R h$ ) we get $u=\psi \circ T$. As $h=u-R u$, this gives $h=\psi \circ T-\psi$.

### 3.4. The classical method: stable laws case

The following theorem has essentially been proved by Aaronson and Denker in [AD01].

Theorem 3.8. Let $(Z, \mathcal{B}, T, m, \delta)$ be a Markov map preserving the finite ergodic measure $m$. Assume that $\log g_{m}$ is $\tau$-Hölder for some $\tau<1$ and that $T$ has the bounded tower big image property.

Let $p \in(0,1) \cup(1,2)$ and $h$ be a function on $Z$ with $\sum_{d \in \delta} m[d] D_{\tau} h(d)<\infty$, such that, on the elements of the partition not returning to the basis of the tower at the next iteration, $h$ is bounded and uniformly Hölder. Moreover, assume that $m[h>x]=\left(c_{1}+o(1)\right) x^{-p} L(x)$ and $m[h<-x]=\left(c_{2}+o(1)\right) x^{-p} L(x)$ for constants $c_{1}, c_{2} \geqslant 0$ with $c_{1}+c_{2}>0$, and a slowly varying function L. Finally, assume $\int h=0$ if $p>1$.

Let $R$ be the transfer operator associated to $T$, acting on $\mathcal{L}_{\tau}$, and $R_{t} \psi=$ $R\left(e^{i t h} \psi\right)$. Then, for small enough $t, R_{t}$ has a unique eigenvalue $\lambda(t)$ close to 1 , and this eigenvalue satisfies

$$
\begin{aligned}
\lambda(t)= & 1-\frac{c}{m(Z)}|t|^{p} L(1 /|t|)+i \frac{c}{m(Z)}|t|^{p} L(1 /|t|) \beta \operatorname{sgn}(t) \tan \left(\frac{p \pi}{2}\right) \\
& +o\left(|t|^{p} L(1 /|t|)\right)
\end{aligned}
$$

for constants $\beta=\frac{c_{1}-c_{2}}{c_{1}+c_{2}}$ and $c=\left(c_{1}+c_{2}\right) \Gamma(1-p) \cos \left(\frac{p \pi}{2}\right)$.
Proof. Replacing $m$ by $m / m(Z)$, we can assume that $m$ is a probability measure.
This theorem is essentially Theorem 5.1 in [AD01], with some differences in the hypotheses: we use the bounded tower big image property, whence their proof has to be slightly modified, as we explain now.

Corollary 3.3 gives that 1 is a simple eigenvalue of $R$ acting continuously on $\mathcal{L}_{\tau}$. Corollary 3.6 ensures that $t \mapsto R_{t}$ is continuous at 0 , which implies that $R_{t}$ also has a simple isolated eigenvalue $\lambda(t)$ close to 1 . The proof of Theorem 5.1 in [AD01] applies then literally if we can prove that, for $p>1$, the eigenfunction $\xi_{t}$ of $R_{t}$ (normalized to have integral 1) satisfies $\left\|\xi_{t}-1\right\|=O(|t|)$.

As $\xi_{t}$ is an eigenfunction of $R_{t}$ and eigenfunctions depend holomorphically on operators, $\left\|\xi_{t}-1\right\|=O\left(\left\|R_{t}-R\right\|\right)$. Corollary 3.6 gives that this is a $O(|t|)$, which concludes the proof.

An analogous theorem for $p=2$ and $h \notin L^{2}$ may be proved using [AD98].

## 4. Proof of the main theorems

In this section, we will prove Theorems 1.1 and 1.2 using the results of Sections 2 and 3 . $(X, \mathcal{B}, T, m, \alpha)$ will be a probability preserving topologically mixing Markov map, such that the induced map on $Y \subset X$ has the big image property and Hölder distortion. The function $\varphi=\varphi_{Y}$ will denote the first return time from $Y$ to itself.

Note first that, since $T$ is topologically mixing, $T_{Y}$ is topologically transitive. Theorem 4.6.3 in [Aar97] implies that $T_{Y}$ is ergodic, whence $T$ is also ergodic.

### 4.1. Construction of an extension

In order to change variables between $\{x \in Y \mid \varphi(x)>i\}$ and its image by $T^{i}$, we have to construct an extension of the system since it is possible that two different points in $Y$ are sent on the same image in $X$ before they return to $Y$.

We use a tower extension, sometimes called a Kakutani tower (the towers in Section 3 will actually correspond to making a finite height cutoff of the Kakutani tower). It is built as follows: we set $X^{\prime}=\{(x, i) \mid x \in Y, 0 \leqslant i<\varphi(x)\}$. $Y$ is identified with $Y^{\prime}:=\{(x, 0) \mid x \in Y\} \subset X^{\prime}$. The space $X^{\prime}$ is endowed with a measure $m^{\prime}$, equal to $m$ on $Y^{\prime}$ and then lifted in the tower, i.e. if $A \subset Y$ and $\varphi>i$ on $A$, then $m^{\prime}(A \times\{i\})=m(A)$.

We define $T^{\prime}$ on $X^{\prime}$ by $T^{\prime}(x, i)=(x, i+1)$ if $i<\varphi(x)-1$, and $\left(T^{\varphi(x)} x, 0\right)$ otherwise. We have a projection $\pi: X^{\prime} \rightarrow X$ defined by $\pi(x, i)=T^{i} x$ (which is not necessarily injective). The following diagram is then commutative, and all applications are measure preserving:


All this is classical material, but I have not been able to locate a satisfying reference in the literature. Thus, we give for completeness a proof of the fact that the map $\pi$ is measure-preserving.

Proposition 4.1. $\forall K \subset X$, we have $m^{\prime}\left(\pi^{-1} K\right)=m(K)$.
Proof. Let $A_{0}=K-Y, B_{0}=K \cap Y$. We define inductively $A_{k}=T^{-1}\left(A_{k-1}\right)-Y$ and $B_{k}=T^{-1}\left(A_{k-1}\right) \cap Y$ : the points of $A_{k}$ have not been treated up to time $k$, while the points in $B_{k}$ are treated at time $k$. Then we have $\pi^{-1}(K)=\{(x, i) \mid i \in$ $\left.\mathbb{N}, x \in B_{i}\right\}$. Thus, $m^{\prime}\left(\pi^{-1} K\right)=\sum m\left(B_{k}\right)$.

As $m(K)=m\left(A_{0}\right)+m\left(B_{0}\right)$ and, by construction, $m\left(A_{k-1}\right)=m\left(T^{-1} A_{k-1}\right)=$ $m\left(A_{k}\right)+m\left(B_{k}\right)$, we get that for any $n, m(K)=m\left(B_{0}\right)+\ldots+m\left(B_{n}\right)+m\left(A_{n}\right)$. To conclude, we just have to show that $m\left(A_{n}\right) \rightarrow 0$ when $n \rightarrow \infty$.

Let $C_{n}=\left\{x \mid \forall i \leqslant n, T^{i} x \notin Y\right\}$. Then $A_{n} \subset C_{n}$. Moreover, $C_{n}$ is decreasing. Let $C=\bigcap C_{n}$. It suffices to show that $m(C)=0$, since this will prove that $m\left(C_{n}\right) \rightarrow 0$. As $C \subset T^{-1} C$ and $T$ is measure preserving, $C=T^{-1} C \bmod 0$. By ergodicity of $T, m(C)=0$ or 1 . As $C \cap Y=\emptyset$ and $m(Y)>0$, we obtain $m(C)=0$.

If $\chi$ is a function on $X$, we define $\chi^{\prime}$ on $X^{\prime}$ by $\chi^{\prime}=\chi \circ \pi$. Then $\int_{X^{\prime}} \chi^{\prime} \mathrm{d} m^{\prime}=$ $\int_{X} \chi \mathrm{~d} m$.

To construct a partition on $X^{\prime}$, we consider first the partition $\delta$ on $Y$ for which the induced map $T_{Y}$ is Markov. Then, we lift it in the tower, i.e. the partition at height $i$ is the same as the partition of the projection of this floor on the basis. Then $T^{\prime}$ is Markov for this new partition $\alpha^{\prime}$. Note that the partition for the induced map $T_{Y^{\prime}}^{\prime}$ is then simply the restriction of the partition $\alpha^{\prime}$ to the basis of the tower.

Moreover, the inverse of the jacobian of $T_{Y^{\prime}}^{\prime}$ is the same as the inverse of the jacobian of $T_{Y}$, whence it is locally $\theta$-Hölder (note that the separation times are the same in $Y$ and in $Y^{\prime}$ ).

Finally, $T^{\prime}$ is still topologically mixing: if $a \in \delta$, then one of its iterates $T^{n} a$ contains an element $b$ of $\alpha$. If $c \in \delta$, it is contained in some $c^{\prime} \in \alpha$, and there exists $N$ such that $\forall p \geqslant N, c^{\prime} \subset T^{p}\left(T^{n} a\right)$. This proves that $\forall p \geqslant n+N, c \subset T^{p}(a)$. This easily implies topological mixing for $T^{\prime}$.

In fact, all the hypotheses on ( $X, T, f, f_{Y}$ ) are carried over to $\left(X^{\prime}, T^{\prime}, f^{\prime}, f_{Y}^{\prime}\right)$. If we prove the central limit theorem or the convergence to a stable law on $X^{\prime}$, this will give the same result on $X$ since $\left(S_{n} f\right)^{\prime}=S_{n}\left(f^{\prime}\right)$. Thus, it is sufficient to prove the result for $X^{\prime}$.

From this point on, we will thus work in a tower. It is Markov, and the returns to the basis have a big image (but are not necessarily surjective). We will first prove Theorem 1.1 on the Central Limit Theorem, and we will indicate in a last section the modifications to be done for Theorem 1.2 (the stable case).

### 4.2. Local result

The hypotheses of the main theorems are tailor-made so that the induction on the basis $Y$ of the tower has good properties. However, we will have to induce on larger parts of the space. In this section, we describe properties of the induced maps on bounded parts of the tower. In this section, the hypotheses will be those of Theorem 1.1.

### 4.2.1. The space

Let $q \geqslant 1$. We write $Z=Z_{q}$ for the union of the first $q$ floors of the Kakutani tower from Section 4.1. We will induce on $Z$. We will write $\delta$ for the partition on $Z$ for which the induced map $T_{Z}$ is Markov - this is in fact simply the restriction of the original partition $\alpha$ to $Z$, because of the particular combinatorics of the tower. For $k<q$, we will also denote by $\Delta_{k}$ the $k^{\text {th }}$ floor of the tower. Note that the hypothesis $f \in I(X, Y)$ guarantees that the function $f$ is bounded and Hölder on $Z$.

If $s$ is the separation time with respect to the basis $Y$ of the tower, and $t$ is the separation time for $T_{Z}$, then an iteration of $T_{Y}$ corresponds to at most $q$ iterations of $T_{Z}$, which implies that $t(x, y) \leqslant q s(x, y)$. Thus, if $\psi$ satisfies $|\psi(x)-\psi(y)| \leqslant$ $C \theta^{s(x, y)}$, it will satisfy $|\psi(x)-\psi(y)| \leqslant C \tau^{t(x, y)}$ for $\tau=\theta^{1 / q}$.

We write $g_{m_{Z}}=\frac{\mathrm{d} m_{Z}}{\mathrm{~d} m_{Z} \circ T_{Z}}$. For $(x, i) \in Z$, if $i<\varphi(x)-1$ then $g_{m_{Z}}(x)=1$, and if $i=\varphi(x)-1$ then $g_{m_{Z}}(x, i)=g_{m_{Y}}(x, 0)$ (the inverse of the jacobian of $T_{Y}$ with respect to $m_{Y}$ ). As $\log g_{m_{Y}}$ is locally $\theta$-Hölder (for the separation time $s$ ), we obtain that $\log g_{m_{Z}}$ is locally $\tau$-Hölder (for the separation time $t$ ).

Since $T$ is ergodic, so is $T_{Z}$ ([Aar97, Proposition 1.5.2]). Moreover, $T_{Z}$ has the bounded tower big image property. Thus, the transfer operator $\hat{T}_{Z}$ acts continuously on $\mathcal{L}_{\tau}(Z)$ according to Corollary 3.3, and has a simple isolated eigenvalue at 1.

### 4.2.2. First return transfer operators

Following [Sar02], we define operators $R_{n}$ and $T_{n}: R_{n}$ is a first return transfer operator, i.e. it sees only the first returns to $Z$ at time $n$, while $T_{n}$ sees all returns at time $n$. They are defined by $R_{n} \psi=1_{Z} \hat{T}^{n}\left(1_{\left\{\varphi_{Z}=n\right\}} \psi\right)$ (where $\varphi_{Z}$ is the first return time from $Z$ to itself) and $T_{n} \psi=1_{Z} \hat{T}^{n}\left(1_{Z} \psi\right)$. They can also be written in the following way: if $x \in Z$, then $R_{n} \psi(x)=\sum g_{m}^{(n)}(y) \psi(y)$ where the sum is over all points $y$ in $Z$ such that $T y, \ldots, T^{n-1} y \notin Z$ and $T^{n} y=x$, while $T_{n} \psi(x)=\sum g_{m}^{(n)}(y) \psi(y)$ where the sum is over all points $y \in Z$ with $T^{n} y=x$.

If $y \in Z$ with $T^{n} y=x$, we can consider all its iterates between 0 and $n$ that fall into $Z$. To go from one of these iterates to the following corresponds to iterating one of the $R_{k} \mathrm{~s}$. Thus, we get $T_{n}=\sum_{k_{1}+\ldots+k_{l}=n} R_{k_{1}} \ldots R_{k_{l}}$ : this is the renewal equation, which will make it possible to understand the $T_{n}$ if the $R_{n}$ are well understood.

Lemma 4.2. The operators $R_{n}$ and $T_{n}$ act continuously on $\mathcal{L}_{\tau}(Z)$, and $\left\|R_{n}\right\|_{\mathcal{L}_{\tau}(Z)}=$ $O\left(m\left[\varphi_{Z}=n\right]\right)$ (where $\varphi_{Z}$ is the return time from $Z$ to itself $)$.
Proof. As $\hat{T}_{Z}$ acts continuously on $\mathcal{L}_{\tau}$ and $R_{1} \psi=\hat{T}_{Z}\left(1_{\left\{\varphi_{Z}=1\right\}} \psi\right)$, the operator $R_{1}$ is continuous.

For $n \geqslant 2$, we estimate $\left\|R_{n}\right\|$ using the notations of Section 3.1 (applied to $Z$ and $T_{Z}$ ):

$$
R_{n}=\sum_{\underline{d}=\left[d_{0}\right] z, d_{0}=\left[a_{0}, \ldots, a_{n-1}, Z\right]} M_{\underline{d}} .
$$

For $n \geqslant 2$, the cylinders appearing in $R_{n}$ correspond to points coming back to $Z$ in $n$ steps: they get out of the tower $Z$ through the top, spend $n-1$ iterates in the higher
floors of the Kakutani tower, and then get back to the basis of the tower, with a big image. Thus, we can forget the terms $m\left[T_{Z} d_{0}\right]$ for these cylinders in the estimates of Lemma 3.2, and we get $\left\|R_{n}\right\| \leqslant B(1+\tau) \sum m\left[d_{0}\right]=B(1+\tau) m\left[\varphi_{Z}=n\right]$.

The $R_{n}$ are continuous and $T_{n}=\sum_{k_{1}+\ldots+k_{l}=n} R_{k_{1}} \ldots R_{k_{l}}$, whence $T_{n}$ is also continuous.

Lemma 4.3. For $z \in \overline{\mathbb{D}}$, write $R(z)=\sum R_{n} z^{n}$. Then $R(1)$ acting on $\mathcal{L}_{\tau}(Z)$ has a simple isolated eigenvalue equal to 1, and the corresponding spectral projection is given by $\psi \mapsto \frac{1}{m[Z]} \int_{Z} \psi \mathrm{~d} m$.

Moreover, for $z \neq 1, I-R(z)$ is invertible.
Note that, since $\left\|R_{n}\right\|=O\left(m\left[\varphi_{Z}=n\right]\right)$ is summable by Kac's Formula, the series $R(z)$ is converging for all $z \in \overline{\mathbb{D}}$.

Proof. Let $z \in \overline{\mathbb{D}}$. We will show a Doeblin-Fortet inequality for $R(z)^{s}$ when $s \geqslant q$. Let $\psi \in \mathcal{L}_{\tau}(Z)$. Then

$$
R(z)^{s}=\sum_{\underline{d}=\left[d_{0}, \ldots, d_{s-1}\right] z} z^{l(\underline{d})} M_{\underline{d}}
$$

where $l(\underline{d})$ is the sum of the lengths of the $d_{i} \mathrm{~s}$ seen as cylinders in $X$. In particular, $l(\underline{d}) \geqslant s$, whence $|z|^{l(d)} \leqslant|z|^{s}$. On the basis, summing the inequalities given by Lemma 3.2 (and using the big image property for the returns to the basis), we obtain that $\left\|\left(R(z)^{s} \psi\right)_{\mid \Delta_{0}}\right\|_{\mathcal{L}_{\tau}} \leqslant B|z|^{s}\left(\tau^{s}\|\psi\|_{\mathcal{L}_{\tau}}+\int|\psi|\right)$.

On $\Delta_{k}$ the set of points at height $k, R(z)^{s} \psi(x)=z^{k} R(z)^{s-k} \psi\left(x^{\prime}\right)$ where $x^{\prime}$ is the point corresponding to $x$ but in the basis. Thus, $\left\|\left(R(z)^{s} \psi\right)_{\mid \Delta_{k}}\right\| \leqslant B|z|^{s}$ $\left(\tau^{s-k}\|\psi\|_{\mathcal{L}_{\tau}}+\int|\psi|\right)$.

Summing on all the floors, we have proved that

$$
\forall s \geqslant q, \forall \psi \in \mathcal{L}_{\tau},\left\|R(z)^{s} \psi\right\|_{\mathcal{L}_{\tau}} \leqslant \frac{B}{\tau^{q}}|z|^{s}\left(\tau^{s}\|\psi\|_{\mathcal{L}_{\tau}}+\int|\psi|\right) .
$$

As the injection from $\mathcal{L}_{\tau}$ to $L^{1}$ is compact, this is a Doeblin-Fortet inequality, which implies that the essential spectral radius of $R(z)$ is $\leqslant \tau|z|<|z|$. Moreover, this inequality also gives that the spectral radius of $R(z)$ is $\leqslant|z|$.

If $|z|<1$, then $R(z)$ has spectral radius $\leqslant|z|<1$, whence $I-R(z)$ is invertible.

If $z=1$, as $R(1)$ counts the first returns to $Z$, it is not hard to check that $R(1)=\hat{T}_{Z}$ is the transfer operator associated to $T_{Z}$. Hence, Corollary 3.3 ensures that it has a simple isolated eigenvalue at 1 .

If $|z|=1$ but $z \neq 1$, we have to show that $I-R(z)$ is invertible. We write $z=e^{i t}$ for some $0<t<2 \pi$. Since the essential spectral radius of $R(z)$ is $<1$, it is enough to show that 1 is not an eigenvalue of $R(z)$. Suppose on the contrary that $R(z) a=a$ for some nonzero $a \in \mathcal{L}_{\tau}(Z)$.

For $\psi, \xi \in L^{2}\left(m_{Z}\right)$, write $\langle\psi, \xi\rangle=\int \bar{\psi} \xi \mathrm{d} m_{Z}$. Define the operator $W$ : $L^{\infty}\left(m_{Z}\right) \rightarrow L^{\infty}\left(m_{Z}\right)$ by $W \psi=e^{-i t \varphi_{Z}} \psi \circ T_{Z}$. As $R(z) \xi=R(1)\left(e^{i t \varphi_{Z}} \xi\right)$, the
operator $W$ satisfies

$$
\begin{aligned}
\langle\psi, R(z) \xi\rangle & =\int \bar{\psi} R(z) \xi=\int \bar{\psi} R(1)\left(e^{i t \varphi_{Z}} \xi\right)=\int \bar{\psi} \circ T_{Z} e^{i t \varphi_{Z}} \xi \\
& =\int \overline{W \psi} \cdot \xi=\langle W \psi, \xi\rangle
\end{aligned}
$$

We show that $a$ is an eigenfunction of $W$ for the eigenvalue 1 :

$$
\begin{aligned}
\|W a-a\|_{2}^{2} & =\|W a\|_{2}^{2}-2 \operatorname{Re}\langle W a, a\rangle+\|a\|_{2}^{2} \\
& =\|W a\|_{2}^{2}-2 \operatorname{Re}\langle a, R(z) a\rangle+\|a\|_{2}^{2} \\
& =\|W a\|_{2}^{2}-2 \operatorname{Re}\langle a, a\rangle+\|a\|_{2}^{2}=\|W a\|_{2}^{2}-\|a\|_{2}^{2} .
\end{aligned}
$$

As $T_{Z}$ preserves the measure $m_{Z}$, we have $\|W a\|_{2}^{2}=\int|a|^{2} \circ T_{Z}=\int|a|^{2}=\|a\|_{2}^{2}$, which gives $\|W a-a\|_{2}^{2}=0$. Hence, the function $W a-a$ is zero $m_{Z}$-almost everywhere. As $a \in \mathcal{L}_{\tau}$ and $m_{Z}$ is nonzero on every cylinder, the function $a$ is continuous, whence $W a-a=0$ everywhere, i.e. $e^{-i t \varphi_{Z}} a \circ T_{Z}=a$.

Let $b$ be the restriction of $a$ to the basis $Y$ of the tower. Then the previous equality implies that $e^{-i t \varphi_{Y}} b \circ T_{Y}=b$. Taking the modulus, ergodicity of $T_{Y}$ gives that $|b|$ is constant almost everywhere, hence everywhere by continuity. As $b \not \equiv 0$, this constant is nonzero, and we get $e^{-i t \varphi_{Y}}=b / b \circ T_{Y}$. We can apply Theorem 3.1. in [AD01] since $T_{Y}$ has the big image property (this is not the case for $T_{Z}$, which is why we have to restrict $a$ to the basis of the tower). Writing $\beta$ for the induced partition on the basis, this theorem gives that $b$ is $\beta^{*}$-measurable, where $\beta^{*}$ is the smallest partition such that $\forall d \in \beta, T_{Y} d$ is contained in an atom of $\beta^{*}$. In particular, $b$ is constant on each set of $\beta$.

Let $d \in \beta$. On $[d], b$ is equal to a constant $c$. As $T$ is topologically mixing, there exists $N$ such that, $\forall n \geqslant N,[d] \subset T^{n}[d]$. Let $n \geqslant N$, and $x \in[d]$ be such that $T^{n} x \in[d]$. Let $T^{k_{1}} x, T^{k_{2}} x, \ldots, T^{k_{p}} x$ be the successive returns of $x$ to $Y$, with $k_{p}=n$. Then $T^{n} x=T_{Y}^{p} x$ and $n=\sum_{k=0}^{p-1} \varphi_{Y}\left(T_{Y}^{k} x\right)$. Thus,
$e^{-i t n}=e^{-i t \sum_{k=0}^{p-1} \varphi_{Y}\left(T_{Y}^{k} x\right)}=\frac{b(x)}{b\left(T_{Y} x\right)} \frac{b\left(T_{Y} x\right)}{b\left(T_{Y}^{2} x\right)} \cdots \frac{b\left(T_{Y}^{p-1} x\right)}{b\left(T_{Y}^{p} x\right)}=\frac{b(x)}{b\left(T^{n} x\right)}=\frac{c}{c}=1$.

This is true for any $n \geqslant N$. Taking for example $n=N$ and $N+1$ and quotienting, we obtain $e^{i t}=1$, which is a contradiction and concludes the proof.

Lemma 4.4. Writing $R^{\prime}(1)=\sum_{n=1}^{\infty} n R_{n}$ and $P$ for the spectral projection associated to the eigenvalue 1 of $R(1)$, we have $P R^{\prime}(1) P=\mu P$ for $\mu=\frac{1}{m[Z]}>0$.

Proof. Since $P f=\frac{1}{m[Z]} \int_{Z} f \mathrm{~d} m$, we have $P R_{n} P=\frac{m\left[\varphi_{Z}=n\right]}{m[Z]} P$, whence

$$
P R^{\prime}(1) P=\left(\sum_{n=1}^{\infty} n m\left[\varphi_{Z}=n\right]\right) \frac{1}{m[Z]} P=\frac{1}{m[Z]} P
$$

by Kac's Formula.

### 4.2.3. Perturbation of the transfer operators

Recall that we are in the setting of Theorem 1.1: $f$ is a function on $X$ belonging to $I(X, Y)$ such that the induced function $f_{Y}=\sum_{0}^{\varphi_{Y}-1} f \circ T^{k}$ is square integrable on $Y$.

Recall also that $\varphi_{Z}$ is the first return time to our truncated tower $Z$, and define a function $f_{Z}$ on $Z$ by $f_{Z}(x)=\sum_{k=0}^{\varphi_{Z}(x)-1} f\left(T^{k} x\right)$. If $x$ does not go out of the bounded tower $Z$ through the top, we have $f_{Z}(x)=f(x)$, while otherwise $f_{Z}(x)$ is the sum of the $f(y)$ for $y$ above $x$ in the tower. Since $f \in L^{1}(m)$, we also have $f_{Z} \in L^{1}\left(m_{Z}\right)$ and $\int_{Z} f_{Z}=\int_{X} f=0$. This function $f_{Z}$ is interesting since we can study the Birkhoff sums of $f$ for $T$ by looking at the Birkhoff sums of $f_{Z}$ for $T_{Z}$ : if $x \in Z$ and $T^{n} x \in Z$, and if $k$ is the number of returns of $x$ to $Z$ between times 0 and $n$, then $\sum_{i=0}^{n-1} f\left(T^{i} x\right)=\sum_{j=0}^{k-1} f_{Z}\left(T_{Z}^{j} x\right)$.

We show that $f_{Z} \in L^{2}(Z)$ (recall that we are proving the Central Limit Theorem, i.e. we assume that $f_{Y} \in L^{2}(Y)$ ). Let $\Delta=\Delta_{q-1}$ be the highest floor of $Z$, and $\Lambda$ its projection on the basis. Then $f_{Z}$ is bounded on $Z-\Delta$. Moreover, if $x \in \Delta$ and $x^{\prime}$ is the corresponding point in $\Lambda$, then $f_{Y}\left(x^{\prime}\right)=f_{Z}(x)+\sum_{0}^{q-2} f\left(T^{j} x^{\prime}\right)$, and the sum is bounded by $q\left\|f_{\mid Z}\right\|_{\infty}$, whence $f_{Y}\left(x^{\prime}\right)=f_{Z}(x) \pm C$. As $f_{Y_{\mid \Lambda}}$ is in $L^{2}$, we obtain also $f_{Z \mid \Delta} \in L^{2}$.

We now perturb the first return transfer operators, setting $R_{n}(t)=R_{n}\left(e^{i t f z}.\right)$ and $T_{n, t}=T_{n}\left(e^{i t S_{n} f}\right.$.). Since the Birkhoff sums of $f$ for $T$ and of $f_{Z}$ for $T_{Z}$ are equal, the renewal equation still holds for these operators, i.e. $T_{n, t}=$ $\sum_{k_{1}+\ldots+k_{l}=n} R_{k_{1}}(t) \ldots R_{k_{l}}(t)$. This equation can also be written as $\sum T_{n, t} z^{n}=$ $\left(I-\sum R_{n}(t) z^{n}\right)^{-1}$.

For technical reasons, we will let the operators act on $\mathcal{L}_{\tau^{\eta}}(Z)$ instead of $\mathcal{L}_{\tau}(Z)$, where $\eta$ is as in the definition of $I(X, Y)$. Since $\tau^{\eta} \geqslant \tau$, the distortion is also $\tau^{\eta}$-Hölder. Thus, the results of the previous paragraphs still apply. In particular, Lemma 4.2 implies that $\left\|R_{n}\right\|_{\mathcal{L}^{\eta}}=O\left(m\left[\varphi_{Z}=n\right]\right)$.

Lemma 4.5. There exist constants $r_{n}$ and $\varepsilon_{0}>0$ such that, $\forall n \geqslant 1, \forall t \in$ $\left[-\varepsilon_{0}, \varepsilon_{0}\right],\left\|R_{n}(t)\right\|_{\mathcal{L}_{\tau} \eta} \leqslant r_{n}$, and satisfying $\sum n r_{n}<\infty$.

Moreover, for every $n \geqslant 1$, the map $t \mapsto R_{n}(t)$ is continuous at $t=0$.

Proof. Let $n \geqslant 2$. We have $R_{n}(t)=\sum M_{\underline{d}}(t)$ where the sum is over all $\underline{d}=\left[d_{0}\right]_{Z}$ with $d_{0}=\left[a_{0}, \ldots, a_{n-1}, Z\right]$ (we use here the operators $M_{\underline{d}}(t)$ introduced in (9), for the function $h=f_{Z}$ ). The returns to $Z$ do not take place at the first iteration, so they have to be returns to the basis, whence they have a big image. Summing the estimates given by Lemma 3.5, we obtain

$$
\begin{aligned}
\left\|R_{n}(t)\right\|_{\mathcal{L}_{\tau} \eta} \leqslant & \left\|R_{n}\right\|_{\mathcal{L}_{\tau^{\eta}}} \\
& +B\left(\sum_{d \in \delta, \varphi_{Z}(d)=n} m[d] D_{\tau} f_{Z}(d)^{\eta}+\int_{\left\{\varphi_{Z}=n\right\}}\left|e^{i t f_{Z}}-1\right| \mathrm{d} m\right)
\end{aligned}
$$

$$
\begin{aligned}
\leqslant & C m\left[\varphi_{Z}=n\right]+B \sum_{d \in \delta, \varphi_{Z}(d)=n} m[d] D_{\tau} f_{Z}(d)^{\eta} \\
& +2 B m\left[\varphi_{Z}=n\right]=: r_{n} .
\end{aligned}
$$

We know that $\sum n m\left[\varphi_{Z}=n\right]=m(X)<\infty$, by Kac's formula. Thus, to show that $\sum n r_{n}<\infty$, it is sufficient to prove that $\sum_{d \in \delta} \varphi_{Z}[d] m[d] D_{\tau} f_{Z}(d)^{\eta}<\infty$. Note that this inequality holds for $f_{Y}$ by assumption, since $f \in I(X, Y)$.

Let $A$ be the set of elements of the partition of $Z$ that do not return to the basis of $Z$ at the next iteration, and $B$ the set of those that come back to the basis. If $d \in A$, then $\varphi_{Z}=1$ on $d$, and $f_{Z}=f$ is uniformly $\tau$-Hölder, whence

$$
\sum_{d \in A} \varphi_{Z}[d] m[d] D_{\tau} f_{Z}(d)^{\eta} \leqslant C \sum_{d \in A} m[d] \leqslant C m[Z]<\infty .
$$

Let $d \in B$ and $x, y \in d$. Let $x^{\prime}$ and $y^{\prime}$ be the points in the basis corresponding to $x$ and $y$, and $d^{\prime}$ be the corresponding partition element in the basis: $x=T^{i} x^{\prime}$ and $y=T^{i} y^{\prime}$ for some $i<q$. Then $f_{Y}\left(x^{\prime}\right)=\sum_{0}^{i-1} f\left(T^{j} x^{\prime}\right)+f_{Z}(x)$, and we have the same expression for $y$, whence

$$
\begin{aligned}
\left|f_{Z}(x)-f_{Z}(y)\right| & \leqslant\left|f_{Y}\left(x^{\prime}\right)-f_{Y}\left(y^{\prime}\right)\right|+\sum_{j=0}^{i-1}\left|f\left(T^{j} x^{\prime}\right)-f\left(T^{j} y^{\prime}\right)\right| \\
& \leqslant C D_{\theta} f_{Y}\left(d^{\prime}\right) \theta^{s\left(x^{\prime}, y^{\prime}\right)}+D q \theta^{s\left(x^{\prime}, y^{\prime}\right)} \leqslant\left(C D_{\theta} f_{Y}\left(d^{\prime}\right)+D q\right) \tau^{t(x, y)}
\end{aligned}
$$

using $s\left(x^{\prime}, y^{\prime}\right)=s(x, y) \geqslant t(x, y) / q$ and $\tau=\theta^{1 / q}$. Since $\varphi_{Z}(d) \leqslant \varphi_{Y}\left(d^{\prime}\right)$, we get

$$
\sum_{d \in B} \varphi_{Z}[d] m[d] D_{\tau} f_{Z}(d)^{\eta} \leqslant \sum_{d^{\prime}} \varphi_{Y}\left[d^{\prime}\right] m\left[d^{\prime}\right]\left(C D_{\theta} f_{Y}\left(d^{\prime}\right)+D q\right)^{\eta}<\infty
$$

by Kac's formula and the assumption on $f_{Y}$. This proves that $\sum n r_{n}<\infty$.
For $n=1, R_{1}(t)=R_{t}\left(1_{\left\{\varphi_{Z}=1\right\}} \cdot\right)$ and $R_{t}$ depends continuously on $t$ at $t=0$ according to Corollary 3.6 (in the space $\mathcal{L}_{\tau^{\eta}}$ ).

Finally, to prove the continuity of $t \mapsto R_{n}(t)$ at $t=0$ for $n \geqslant 2$, we sum once again the estimates of Lemma 3.5 (in the space $\mathcal{L}_{\tau^{\eta}}$ ) and we get $\left\|R_{n}(t)-R_{n}\right\|_{\mathcal{L}^{\eta}} \leqslant$ $B|t|$ since the sum $\sum m[d] D_{\tau^{\eta}} f_{Z}(d)$ is finite (this is as above a consequence of the fact that $\sum m\left[d^{\prime}\right] D_{\theta} f_{Y}\left(d^{\prime}\right)<\infty$, which implies that $\sum m\left[d^{\prime}\right] D_{\theta^{\eta}} f_{Y}\left(d^{\prime}\right)<\infty$ since $\left.D_{\theta^{\eta}} f_{Y}\left(d^{\prime}\right) \leqslant D_{\theta} f_{Y}\left(d^{\prime}\right)\right)$.

### 4.2.4. The local result

We will apply the abstract Theorem 2.1 to the $R_{n}(t)$ acting on $\mathcal{L}_{\tau^{\eta}}(Z)$. Lemmas 4.2, 4.3, 4.4 and 4.5 ensure that the spectral hypotheses of this theorem are satisfied.

The only hypothesis that remains to be checked is the behavior of the eigenvalue $\lambda(1, t)$ of $R(1, t)=\sum R_{n}(t)=\hat{T}_{Z}\left(e^{i t f z}\right.$. $)$. As $T_{Z}$ has the bounded tower big image property and $f_{Z} \in L^{2}$, Theorem 3.7 gives $\lambda(t)=1-\frac{\sigma^{2}}{2 m(Z)} t^{2}+o\left(t^{2}\right)$.

If $\sigma^{2} \neq 0$, we can use the abstract result Theorem 2.1 for $M(t)=t^{2}$ and $c=\frac{\sigma^{2}}{2 m(Z)}$, and we get:

Proposition 4.6. Assume $\sigma^{2}>0$. Then, $\forall t \in \mathbb{R}$, we have

$$
\int_{T^{-n} Z \cap Z} e^{i t S_{n} f / \sqrt{n}} \mathrm{~d} m \rightarrow m(Z)^{2} e^{-\sigma^{2} t^{2} / 2}
$$

Proof. Note that, if $t$ is fixed, $t / \sqrt{n} \in\left[-\varepsilon_{0}, \varepsilon_{0}\right]$ for $n$ large enough, whence Theorem 2.1 applies. The constant $\mu$ is equal to $1 / m(Z)$ according to Lemma 4.4.

On the one hand,

$$
\begin{aligned}
\int T_{n, t / \sqrt{n}} 1 \mathrm{~d} m & =\int 1_{Z} \hat{T}^{n}\left(e^{i t S_{n} f / \sqrt{n}} 1_{Z}\right) \mathrm{d} m=\int 1_{Z} \circ T^{n} \cdot e^{i t S_{n} f / \sqrt{n}} 1_{Z} \mathrm{~d} m \\
& =\int_{Z \cap T^{-n} Z} e^{i t S_{n} f / \sqrt{n}} \mathrm{~d} m
\end{aligned}
$$

and on the other hand,

$$
\begin{aligned}
\int T_{n, t / \sqrt{n}} 1 \mathrm{~d} m & =\int\left(1-\frac{\sigma^{2}}{2 \mu m(Z)} \frac{t^{2}}{n}\right)^{n} \frac{1}{\mu} P(1) \mathrm{d} m+o(1) \\
& =\frac{m[Z]}{\mu} e^{-\sigma^{2} t^{2} / 2}+o(1)
\end{aligned}
$$

using $P(1)=1_{Z}$, and $1 / \mu=m(Z)$ whence $\mu m(Z)=1$.
If $\sigma^{2}=0$, Theorem 2.1 can not be used, but we have nevertheless the following result:

Proposition 4.7. Assume that $\sigma^{2}=0$. Then $f$ is a coboundary on $X$, i.e. it is possible to write $f=\chi \circ T-\chi$ for some measurable function $\chi$ on $X$.

Proof. Theorem 3.7 gives that, if $\sigma^{2}=0$, then $f_{Z}$ is a coboundary, i.e. it can be written as $f_{Z}=\psi \circ T_{Z}-\psi$ for some function $\psi$ on $Z$. We have to go from $f_{Z}$ to $f$.

We extend $\psi$ to the whole space $X$ by $\psi(x)=0$ if $x \notin Z$. Then, writing $\tilde{f}=f-\psi \circ T+\psi$, we have for $x \in Z$ that

$$
\sum_{0}^{\varphi_{Z}(x)-1} \tilde{f}\left(T^{i} x\right)=f_{Z}(x)-\psi\left(T_{Z} x\right)+\psi(x)=0
$$

We then set $\delta(x)=-\sum_{0}^{\varphi_{Z}(x)-1} \tilde{f}\left(T^{i} x\right)$, where $\varphi_{Z}$ denotes the first return (resp. entry) time to $Z$ if $x \in Z$ (resp. $x \notin Z$ ). Then, if $T x \notin Z$, we have $\tilde{f}(x)=$ $\delta(T x)-\delta(x)$, while if $T x \in Z$, then $\delta(T x)=0$ and $\delta(x)=-\tilde{f}(x)$, whence $\tilde{f}(x)=\delta(T x)-\delta(x)$. Thus, $\tilde{f}=\delta \circ T-\delta$, whence $f=(\psi+\delta) \circ T-(\psi+\delta)$.

### 4.3. Proof of the Central Limit Theorem 1.1

A priori, the constant $\sigma$ associated to $Z_{q}$ depends on $q$, and should be written $\sigma_{q}$. We first prove that this is not the case:
Proposition 4.8. $\sigma_{q}$ is independent of $q$.
Proof. We fix $q>1$, and we want to prove that $\sigma_{q}=\sigma_{1}$. We will write $Z=Z_{q}$ and $f_{Z}(x)=\sum_{0}^{\varphi_{Z}(x)-1} f(x)$ (recall that $f_{Y}$ is the same map, but on the basis $Y$ and with the return time $\varphi_{Y}$ ). We will write $\tilde{\varphi}$ for the return time from $Y$ to itself for the map $T_{Z}$ induced by $T$ on $Z$. Thus, with these notations, we have for $x \in Y$ that $f_{Y}(x)=\sum_{j<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{j} x\right)$.

Theorem 3.7 gives that $\sigma_{1}^{2}=\int f_{Y}^{2}+2 \int a f_{Y}$ where $a=\left(I-\hat{T}_{Y}\right)^{-1}\left(\hat{T}_{Y} f_{Y}\right)$ and $\hat{T}_{Y}$ is the transfer operator associated to $T_{Y}$. In the same way, $\sigma_{q}^{2}=\int f_{Z}^{2}+2 \int b f_{Z}$ where $b=\left(I-\hat{T}_{Z}\right)^{-1}\left(\hat{T}_{Z} f_{Z}\right)$.

We define a function $c$ on $Z$ by $c(x, i)=a(x)+\sum_{j=0}^{i-1} f_{Z}\left(T_{Z}^{j} x\right)$ for $x$ in the basis of the tower and $i<\tilde{\varphi}(x)$. Then, if $\psi$ is a bounded function on $Z$, we will check that

$$
\begin{equation*}
\int_{Z} c \cdot\left(\psi-\psi \circ T_{Z}\right)=\int_{Z} f_{Z} \cdot \psi \circ T_{Z} \tag{12}
\end{equation*}
$$

Indeed, note that the definition of $a$ implies that for a function $\chi$ on $Y, \int_{Y} a \cdot(\chi-$ $\left.\chi \circ T_{Y}\right)=\int_{Y} f_{Y} \cdot \chi \circ T_{Y}$. Thus,

$$
\begin{aligned}
& \int_{Z} c \cdot\left(\psi-\psi \circ T_{Z}\right) \\
& \quad=\int_{Y} \sum_{i<\tilde{\varphi}(x)}\left(a(x)+\sum_{j<i} f_{Z}\left(T_{Z}^{j} x\right)\right)\left(\psi\left(T_{Z}^{i} x\right)-\psi\left(T_{Z}^{i+1} x\right)\right) \\
& \quad=\int_{Y} a(x)\left(\psi(x)-\psi\left(T_{Y} x\right)\right)+\sum_{j<i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{j} x\right)\left(\psi\left(T_{Z}^{i} x\right)-\psi\left(T_{Z}^{i+1} x\right)\right) \\
& \quad=\int_{Y} f_{Y}(x) \psi\left(T_{Y} x\right)+\sum_{j<\tilde{\varphi}(x)-1} f_{Z}\left(T^{j} x\right)\left(\psi\left(T_{Z}^{j+1} x\right)-\psi\left(T_{Y} x\right)\right) \\
& \quad=\int_{Y} \sum_{j<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{j} x\right) \psi\left(T_{Y} x\right)+\sum_{j<\tilde{\varphi}(x)-1} f_{Z}\left(T^{j} x\right)\left(\psi\left(T_{Z}^{j+1} x\right)-\psi\left(T_{Y} x\right)\right) \\
& \quad=\int_{Y} \sum_{j<\tilde{\varphi}(x)-1} f_{Z}\left(T_{Z}^{j} x\right) \psi\left(T_{Z}^{j+1} x\right)+f_{Z}\left(T_{Z}^{\tilde{\varphi}(x)-1} x\right) \psi\left(T_{Y} x\right) \\
& \quad=\int_{Y} \sum_{j<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{j} x\right) \psi\left(T_{Z}^{j+1} x\right)=\int_{Z} f_{Z}(x) \psi\left(T_{Z} x\right) .
\end{aligned}
$$

Equation (12) implies that $\left(I-\hat{T}_{Z}\right) c=\hat{T}_{Z} f_{Z}$. By definition of the function $b$, we also have $\left(I-\hat{T}_{Z}\right) b=\hat{T}_{Z} f_{Z}$, whence $c-b$ is in the kernel of $I-\hat{T}_{Z}$, i.e. it is a constant $K$, and $c=b+K$.

We now compute $\sigma_{1}^{2}$, using $\int f_{Z}=0$ :

$$
\begin{aligned}
\sigma_{1}^{2}= & \int_{Y} f_{Y}^{2}+2 f_{Y} a=\int_{Y}\left(\sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right)\right)^{2}+2 \sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right) a(x) \\
= & \int_{Y} \sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right)^{2}+2 \sum_{j<i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right) f_{Z}\left(T_{Z}^{j} x\right) \\
& +2 \sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right)\left(c\left(T_{Z}^{i} x\right)-\sum_{j<i} f_{Z}\left(T_{Z}^{j} x\right)\right) \\
= & \int_{Y} \sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right)^{2}+2 \sum_{i<\tilde{\varphi}(x)} f_{Z}\left(T_{Z}^{i} x\right) c\left(T_{Z}^{i} x\right)=\int_{Z} f_{Z}(x)^{2}+2 f_{Z}(x) c(x) \\
= & \int_{Z} f_{Z}(x)^{2}+2 f_{Z}(x)(b(x)+K)=\int_{Z} f_{Z}(x)^{2}+2 f_{Z}(x) b(x)=\sigma_{q}^{2} .
\end{aligned}
$$

Conclusion of the proof of Theorem 1.1. Let $\sigma^{2}$ be given by Proposition 4.8.
If $\sigma^{2}=0$, then Proposition 4.7 ensures that $f$ can be written as $\chi \circ T-\chi$ for some measurable function $\chi$. Hence, $S_{n} f / \sqrt{n}$ tends to 0 in probability, and the theorem is proved.

Assume now that $\sigma^{2}>0$. Let $t \in \mathbb{R}$, we want to prove that $\int_{X} e^{i t S_{n} f / \sqrt{n}} \rightarrow$ $e^{-\sigma^{2} t^{2} / 2}$. Fix $\varepsilon>0$. Let $Z=Z_{q}$ with $q$ large enough so that $m(Z) \geqslant 1-\varepsilon$. Proposition 4.6 gives $N$ such that $\forall n \geqslant N$

$$
\left|\int_{T^{-n} Z \cap Z} e^{i t S_{n} f / \sqrt{n}}-m(Z)^{2} e^{-\sigma^{2} t^{2} / 2}\right| \leqslant \varepsilon .
$$

Moreover,

$$
\begin{aligned}
\left|\int_{X} e^{i t S_{n} f / \sqrt{n}}-\int_{T^{-n} Z \cap Z} e^{i t S_{n} f / \sqrt{n}}\right| & \leqslant m\left(X-\left(Z \cap T^{-n} Z\right)\right) \\
& \leqslant m(X-Z)+m\left(X-T^{-n} Z\right) \\
& =2 m(X-Z) \leqslant 2 \varepsilon
\end{aligned}
$$

Finally,

$$
\left|e^{-\sigma^{2} t^{2} / 2}-m(Z)^{2} e^{-\sigma^{2} t^{2} / 2}\right| \leqslant 1-m(Z)^{2} \leqslant 2 \varepsilon .
$$

The above three inequalities conclude the proof of Theorem 1.1.

### 4.4. Proof of Theorem 1.2 on stable laws

The proof is almost the same as that of the Central Limit Theorem. We consider only the case $p \in(0,1) \cup(1,2)$, the case $p=2$ being analogous using [AD98].

We work on $Z=Z_{q}$, the union of the $q$ first floors of the Kakutani tower from Section 4.1. Setting $f_{Z}(x)=\sum_{0}^{\varphi_{Z}(x)-1} f(x)$, we show as in the proof of the central limit theorem that $\sum \varphi_{Z}[d] m[d] D_{\tau} f_{Z}(d)^{\eta}<\infty$ and $\sum m[d] D_{\tau} f_{Z}(d)<\infty$. To apply Theorem 3.8, we have to estimate $m\left[f_{Z}>x\right]$ and $m\left[f_{Z}<-x\right]$ for large $x$.

On the basis, $c_{1}$ is characterized by $m\left[f_{Y}>x\right]=\left(c_{1}+o(1)\right) x^{-p} L(x)$. We will show that we also have $m\left[f_{Z}>x\right]=\left(c_{1}+o(1)\right) x^{-p} L(x)$. Let $\Delta$ be the highest floor in $Z$, and $\Lambda$ its projection on the basis $Y$. Then $f_{Z}$ is bounded on $Z-\Delta$ and $f_{Y}$ is bounded on $Y-\Lambda$, whence it suffices to consider $\Delta$ and $\Lambda$. Let $x \in \Lambda$ and $y=T^{q-1} x$ its image in $\Delta$. Then $f_{Z}(y)=f_{Y}(x)+\sum_{i=0}^{q-2} f\left(T^{i} x\right)$. As the sum is bounded by $C=q\left\|f_{\mid Z}\right\|_{\infty}$, we get $f_{Z}(y)=f_{Y}(x) \pm C$. Thus, for $x$ large enough, $m\left[f_{Y}>x-C\right] \geqslant m\left[f_{Z}>x\right] \geqslant m\left[f_{Y}>x+C\right]$. As $m\left[f_{Y}>x\right]=\left(c_{1}+o(1)\right) x^{-p} L(x)$, we obtain the same estimate for $m\left[f_{Z}>x\right]$, using the slow variation of $L$. In the same way, $m\left[f_{Z}<-x\right]=\left(c_{2}+o(1)\right) x^{-p} L(x)$.

Using Theorem 3.8, we obtain that $\lambda(t)=1-a|t|^{p} L(1 /|t|)+i b \operatorname{sgn}(t)|t|^{p}$ $L(1 /|t|)+o\left(|t|^{p} L(1 /|t|)\right)$, for $a=\frac{c}{m(Z)}$ and $b=\frac{c}{m(Z)} \beta \tan \left(\frac{p \pi}{2}\right)$.

We use then the remark following Theorem 2.1, with $M(t)=t^{p} L(1 / t)$, to get

$$
\begin{aligned}
& \left\|T_{n, t}-\frac{1}{\mu}\left(1-c|t|^{p} L(1 /|t|)+i c \beta \tan \left(\frac{p \pi}{2}\right) \operatorname{sgn}(t)|t|^{p} L(1 /|t|)\right)^{n} P\right\| \\
& \quad \leqslant \varepsilon(t)+\delta(n) .
\end{aligned}
$$

If $B_{n}$ is defined by $B_{n}^{p}=n L\left(B_{n}\right)$, we have

$$
n \frac{|t|^{p}}{B_{n}^{p}} L\left(\frac{B_{n}}{|t|}\right)=|t|^{p} \frac{L\left(B_{n} /|t|\right)}{L\left(B_{n}\right)} \rightarrow|t|^{p}
$$

since $L$ is slowly varying. Thus,

$$
T_{n, t / B_{n}} \rightarrow \frac{1}{\mu} e^{-c|t|^{p}+i c \beta \tan \left(\frac{p \pi}{2}\right) \operatorname{sgn}(t)|t|^{p}} P
$$

Applying this result to the function 1 and integrating, we get

$$
\int_{T^{-n} Z \cap Z} e^{i t S_{n} f / B_{n}} \mathrm{~d} m \rightarrow m(Z)^{2} e^{-c|t|^{p}\left(1-i \beta \operatorname{sgn}(t) \tan \left(\frac{p \pi}{2}\right)\right)},
$$

the exponential being the characteristic function of the stable law $X_{p, c, \beta}$.
We then let $q \rightarrow \infty$ and conclude the proof as for the Central Limit Theorem.

## References

[Aar97] Aaronson, J.: An introduction to infinite ergodic theory. volume 50 of Mathematical Surveys and Monographs. American Mathematical Society, 1997
[AD98] Aaronson, J., Denker, M.: A local limit theorem for stationary processes in the domain of attraction of a normal distribution. Preprint, 1998
[AD01] Aaronson, J., Denker, M.: Local limit theorems for partial sums of stationary sequences generated by Gibbs-Markov maps. Stoch. Dyn. 1, 193-237 (2001)
[ADU93] Aaronson, J., Denker, M., Urbański, M.: Ergodic theory for Markov fibred systems and parabolic rational maps. Trans. Am. Math. Soc. 337, 495-548 (1993)
[BP42] Bochner, S., Phillips, R.S.: Absolutely convergent Fourier expansions for noncommutative normed rings. Ann. Math. 43, 409-418 (1942)
[DS57] Dunford, N., Schwartz, J.T.: Linear Operators, Part 1: General Theory, volume 7 of Pure and Applied Mathematics: a Series of Texts and Monographs. Interscience, 1957
[Fel66] Feller, W.: An Introduction to Probability Theory and its Applications, volume 2. Wiley Series in Probability and Mathematical Statistics. John Wiley, 1966
[FL01] Fisher, A., Lopes, A.O.: Exact bounds for the polynomial decay of correlation, $1 / f$ noise and the CLT for the equilibrium state of a non-Hölder potential. Nonlinearity. 14, 1071-1104 (2001)
[GH88] Guivarc'h, Y., Hardy, J.: Théorèmes limites pour une classe de chaînes de Markov et applications aux difféomorphismes d'Anosov. Ann. Inst. H. Poincaré Probab. Statist. 24, 73-98 (1988)
[GLJ93] Guivarc'h, Y., Le Jan, Y.: Asymptotic winding of the geodesic flow on modular surfaces and continued fractions. Ann. Sci. École Norm. Sup. 26 (4), 23-50 (1993)
[Gou02] Gouëzel, S.: Sharp polynomial bounds for the decay of correlations. To be published in Israel J. Math. 2002
[Hen93] Hennion, H.: Sur un théorème spectral et son application aux noyaux lipschitziens. Proc. Am. Math. Soc. 118, 627-634 (1993)
[Hol02] Holland, M.: Slowly mixing systems and intermittency maps. Preprint, 2002
[Hu] $\mathrm{Hu}, \mathrm{H} .:$ Rates of convergence to equilibriums and decay of correlations. Announcement in the Kyoto 2002 conference available at ndds. math.sci.hokudai.ac.jp/data/NDDS/1024888882-hu.ps
[Iso00] Isola, S.: On systems with finite ergodic degree. Preprint, 2000
[ITM50] Ionescu-Tulcea, Marinescu, G.: Théorie ergodique pour des classes d'opérations non complètement continues. Ann. Math. 47, 140-147 (1950)
[Kah70] Kahane, J.-P.: Séries de Fourier absolument convergentes, volume 50 of Ergebnisse der Mathematik und ihre Grenzgebiete. Springer-Verlag, 1970
[Lév52] Lévy, P.: Fractions continues aléatoires. Rend. Circ. Mat. Palermo 1 (2), 170-208 (1952)
[LSV99] Liverani, C., Saussol, B., Vaienti, S.: A probabilistic approach to intermittency. Ergodic Theory and Dynamical Systems 19, 671-685 (1999)
[Nag57] Nagaev S.V.: Some limit theorems for stationary Markov chains. Theor. Probab. Appl. 2, 378-406 (1957)
[Rau02] Raugi, A.: Étude d'une transformation non uniformément hyperbolique de l'intervalle [0,1]. Preprint, 2002
[RE83] Rousseau-Egele, J.: Un théorème de la limite locale pour une classe de transformations dilatantes et monotones par morceaux. Ann. Probab. 11, 772-788 (1983)
[Sar02] Sarig, O.: Subexponential decay of correlations. Inv. Math. 150, 629-653 (2002)
[You99] Young, L.-S.: Recurrence times and rates of mixing. Israel J. Math. 110, 153-188 (1999)


[^0]:    S. Gouëzel: Département de Mathématiques et Applications, École Normale Supérieure, 45 rue d'Ulm 75005 Paris, France. e-mail: Sebastien. Gouezel@ens.fr
    Mathematics Subject Classification (2000): 37A30, 37A50, 37C30, 37E05, 47A56, 60F05
    Key words or phrases: Decay of correlations - Intermittency - Countable Markov shift Central limit theorem - Stable laws - Wiener's Lemma

