# BERRY ESSEEN THEOREMS FOR SEQUENCES OF EXPANDING MAPS 

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#### Abstract

We prove Berry-Esseen theorems for sums $S_{n}=\sum_{j=0}^{n-1} f_{j} \circ T_{j-1} \circ \cdots \circ T_{1} \circ T_{0}$ where $f_{j}$ are functions with uniformly bounded "variation" and $T_{j}$ is a sequence of expanding maps. Using symbolic representations similar result follow for maps $T_{j}$ in a small $C^{1}$ neighborhood of an Axiom A map and Hölder continuous functions $f_{j}$. All of our results are already new for a single map $T_{j}=T$ and a sequence of different functions $\left(f_{j}\right)$.


## 1. Introduction

1.1. Non-autonomous dynamical systems. A great discovery made in the last century is that deterministic systems could exhibit random behavior. One of the most notable results in this direction is the fact that ergodic averages of deterministic systems could satisfy the Central Limit Theorem (CLT). Since then statistical properties of autonomous hyperbolic dynamical systems have been studied extensively. However, many systems appearing in nature are time dependent due to an interaction with the outside world. In the context of dynamical systems this leads to the study of dynamics formed by a composition of different transformations rather than a single one. Such systems are often called sequential/time-dependent/non-autonomous dynamical systems.

Many powerful tools developed for studying autonomous systems are unavailable in non autonomous setting. In particular, the spectral approach developed by Nagaev [71] and extended to dynamical systems setting by Guivarch and Hardy 41], provides a powerful tool for obtaining asymptotics expansions in limit theorems for dynamical systems. It turns out that complex Perron Frobenius Theorem proven by Hafouta and Kifer in [46] (building on a previous work of Rugh [77] and Dubois [34]) provides a convenient tool for asymptotic computations of the characteristic functions in the setting of Markov chains and dynamical systems. This theorem has already found multiple applications to limit theorems [27, 33, 35, 46, 47, 48, 49, 51]. The goal of the present paper is to study the rate of convergence in the CLT (aka Berry-Esseen theorems) for non-autonomous dynamical systems without making any assumptions on the growth of variance of the underlying partial sum. In a forthcoming work [28] local limit theorems will be considered.

A particular case of a sequential dynamical systems are random dynamical systems. Ergodic theory of random dynamical systems has attracted a lot of attention in the past decades, see [58, 65, 22, 3, 25] and [62]. This includes, for instance, the theory of random invariant measures, entropy theory, thermodynamic formalism, multiplicative ergodic theory and many other classical topics in ergodic theory. Ergodic aspects of (non-random) sequential dynamical systems were studied for the first time in [12, 13], see also [4], [24, Sections 1-2] and [14]. We refer to [24, Section 3] and [48] for examples of expanding maps and to [8, 4, 74] for some examples of sequential hyperbolic sequences (see also Sections 4, 5 and Appendix C).

Next we discuss statistical properties of random or sequential dynamical systems. Decay of correlations for random dynamical systems were obtained in [19, 10, 7]. Large deviations were obtained in [59] and the CLT in [60]. Recently limit theorems for random dynamical systems
have attracted a lot of attention. For hyperbolic maps in a $C^{3+\varepsilon}$ neighborhood of a $C^{3+\varepsilon}$ Anosov map, the CLT was obtained in [32] (see also [6, 23]). Berry-Esseen theorems (see (1.1)) were obtained in [46, 33, 49, 52]. Concerning statistical properties of non-random expanding sequential systems, the sharp CLT was obtained in [24] (see also [2, 72]). For hyperbolic systems the CLT is proven in [8] under the assumption that variance of $S_{n}$ grows faster than $n^{2 / 3}$. Berry-Esseen theorems in the sequential setup were only obtained in [48] for some classes of maps, under the assumption that $\operatorname{Var}\left(S_{n}\right)$ grows linearly fast in $n$.
1.2. Our results. In this paper we obtain optimal CLT rates. A classical result due to Berry and Esseen [18, 37, 38] asserts that for partial sums $S_{n}=\sum_{j=1}^{n} X_{j}$ of zero mean of uniformly bounded iid random variables $X_{j}$ the following optimal uniform CLT rates hold

$$
\begin{equation*}
\Delta_{0, n}:=\sup _{t \in \mathbb{R}}\left|\mathbb{P}\left(S_{n} / \sigma_{n} \leq t\right)-\Phi(t)\right|=O\left(\sigma_{n}^{-1}\right), \quad \sigma_{n}=\left\|S_{n}\right\|_{L^{2}} \tag{1.1}
\end{equation*}
$$

where $\Phi(t)$ is the standard normal distribution function. By now the optimal convergence rate in the CLT was obtained for wide classes of stationary Markov chains [71, 55] and other weakly dependent random processes including chaotic dynamical systems [76, 41, 55, 44, 56, 57], random dynamical systems [46, 33, 49] uniformly bounded stationary sufficiently fast $\phi$-mixing sequences [75], $U$-statistics [20, 39] and locally dependent random variables [11, 9, 21]. However, in all of these processes the variance of $S_{n}$ is of linear order in the number of summands $n$. To the best of our knowledge, the only case where optimal rate was obtained without any growth rates on the variances is for additive functional of uniformly elliptic inhomogeneous Markov chains [27]. In the present paper we obtain optimal CLT rates (in various forms) for Birkhoff sums generated by a sequence of expanding maps and sufficiently regular functions. This was obtained in 48] for Hölder continuous functions when $\operatorname{Var}\left(S_{n}\right) \geq c n$ for some $c>0$. Here we consider more general maps and more general functions (e.g. BV). Most importantly we will not assume any kind of growth rates for $\operatorname{Var}\left(S_{n}\right)$.

Our results are applicable to wide classes of expanding transformations including smooth expanding maps, piecewise expanding maps, Markov maps on the interval and sequential subshifts of finite type. The latter also has applications to additive functionals of finite state uniformly elliptic inhomogenuous Markov chains, where the main novelty here is that the functionals are allowed to depend on the entire path of the chain. Our main results also have applications to sequence of maps in a $C^{1}$ neighborhood of a given $C^{2}$-Axiom A map (see Appendix C). In 32 limit theorems were studied for random dynamical systems in $C^{3+\varepsilon}$ neighborhood of a $C^{3+\varepsilon}$ Anosov maps and Birkhoff sums formed by random $C^{2+\varepsilon}$ functions. Compared with [32] we can perturb more general maps in a weaker norm, consider functions which are only Hölder continuous and treat non-random systems. We would like to emphasize that when starting with sequential SRB measures the closeness of the maps is only needed here to overcome the difficulty that the variance of the underlying partial sum $S_{n}$ does not have to grow linearly fast. We refer to Remark C. 6 for a short discussion on this matter.

What we will actually prove is stronger than uniform rates. In Theorem 3.1(i) we will show that for every $p>0$,

$$
\begin{equation*}
\Delta_{p, n}:=\sup _{t \in \mathbb{R}}\left(1+|t|^{p}\right)\left|\mathbb{P}\left(S_{n} / \sigma_{n} \leq t\right)-\Phi(t)\right|=O\left(\sigma_{n}^{-1}\right) \tag{1.2}
\end{equation*}
$$

The fact that we can take can positive $p$ allows applications to optimal CLT rates in $L^{p}$ (see Theorem3.1(ii)), to Gaussian expectation estimates (Theorem3.1(iii)) and to optimal CLT rates in the Wasserstein distances (Theorem 3.2).

The proof of (1.2) consists of several ingredients. First, using a general result from 52 it is enough to verify a logarithmic growth assumption introduced in [27]. This is done in Proposition 7.1. To prove this proposition we use ideas from [27] together with an appropriate martingale-coboundary decomposition. This decomposition is needed in order to obtain the moment estimates in Proposition 3.3, which are proven using the Burkholder inequality and ideas from [24]. Martingale-coboundary decomposition uses a version of real Ruelle-Perron-Frobenius (RPF), which is proven in Theorem 2.4. The second tool needed to prove Proposition 7.1 is an extension of Theorem 2.4 to complex operators (Theorem 7.3). This result generalizes the complex RPF theorem in 46, 48 and is proven using simpler arguments (see Appendix D) which result in conditions which are easier to verify.

Our setup includes a reference measure $m_{0}$ (e.g. Lebesgue or a time 0 Gibbs measure) and a notion of variation of functions (e.g. variation on $[0,1]$ or Hölder continuity). Then our result hold true with respect to any initial measure $\mu_{0}$ which is absolutely continuous with respect to the reference measure, with bounded density with finite variation. For autonomous systems the fact that the weak limit is preserved when changing densities is called Eagleson's theorem [36] (see also [45, 80]). Eagleson's theorem concerns the asymptotic behavior, and to show that the rates are persevered it is reasonable to require that the density is sufficiently regular. In [53] non-stationary versions of Eagleson's theorem were discussed. When applied to the setup of this paper, the results in [53, §3.3.1] show that optimal uniform CLT rates are preserved under a change of density with bounded variation if the variance of the underlying sum grows linearly fast in $n$ (under one of the measures). Here we show that the non-uniform optimal rate holds for all the measures in the above class without any growth assumptions on the variance.
1.3. The layout of the paper. Section 2 describes the classes of maps we consider. Section 3 presents our main results. Sections 4 and 5 are devoted to examples. In the former section we present specific examples fitting into our abstract framework: expanding and piecewise expanding maps, subshifts of finite type, etc. In the next section we show that our abstract setting includes the setups of [24] and [48] as special cases. Section 6 discusses the moment estimates which play a key role in our analysis. Section 7 is devoted to the analysis of of the characteristic functions near the origin. This analysis plays a key role in the proof of our main results given in Section 7.3. In a forthcoming work, including [28], we combine the results of Section 7 with the estimates of the characteristic function away from the origin to get further asymptotic results related to limit theorems.

The examples in Sections 4 and the proof of the main results rely on many non-autonomous versions of known results for autonomous dynamical systems. Additionally, we find it more reader friendly to include a presentation of our results for hyperbolic maps in a separate section. For these reasons the paper has four appendixes. Appendix $A$ is devoted to mixing properties of sequential Gibbs measures. It generalizes the fact that for a topologically mixing expanding system on a compact manifold there is a unique absolutely continuous invariant measure, which is mixing exponentially fast for Hölder or BV observables. Appendix B concerns extension of classical results for subshift of finite types to their non-autonomous counterparts, including the theory of sequential Gibbs measures. Appendix Cis about non-autonomous hyperbolic systems formed by a sequence of maps $\left(T_{j}\right)$ in a small $C^{1}$-neighborhood of a given Axiom A map. We show that such systems are sequentially Hölder conjugated to the Axiom A map, which will yield that all our results hold true for these systems. Lastly, Appendix D contains a short proof of the complex sequential Ruelle-Perron-Frobenius Theorem following the arguments of [55]. The assumptions of our Theorem D. 2 are easier to verify in specific examples than the assumptions of the corresponding results in [46, 48].

## 2. Preliminaries

2.1. Setup. Our setup here will be a sequential (uniform) version of [19]. Let $\left(X_{j}, \mathcal{B}_{j}, m_{j}\right)_{j=0}^{\infty}$ be a sequence of probability spaces. For a measurable function $h: X_{j} \rightarrow \mathbb{C}$ denote $\|h\|_{p}=$ $\|h\|_{p, j}=\|h\|_{L^{p}\left(X_{j}, \mathcal{K}_{j}, m_{j}\right)}$. We suppose that there are notions of variation $v_{j}: L^{1}\left(X_{j}, m_{j}\right) \rightarrow[0, \infty]$ satisfying
(V1) $v_{j}(t h)=|t| v_{j}(h)$, for every $t \in \mathbb{C}$ and $h: X_{j} \rightarrow \mathbb{C}$;
(V2) $v_{j}(g+h) \leq v_{j}(g)+v_{j}(h)$ for every $g, h: X_{j} \rightarrow \mathbb{C}$;
(V3) $\|h\|_{\infty} \leq\|h\|_{1}+C_{\mathrm{var}} v_{j}(h)$ for some constant $C_{\text {var }}>0$ (independent of $j$ );
(V4) the functions $\mathbf{1}$ taking the constant value 1 have finite variation and $\sup v_{j}(\mathbf{1})<\infty$;
(V5) there is a constant $C>0$ such that for all $j$ and $f, g: X_{j} \rightarrow \mathbb{C}$ we have $v_{j}(f g) \leq C\left(\|f\|_{\infty} v_{j}(g)+\|g\|_{\infty} v_{j}(f)\right)$;
(V6) if $h$ is a positive function with bounded variation bounded below by a constant $c>0$ then $v_{j}(1 / h) \leq C\left(c, v_{j}(h)+\|h\|_{1, j}\right)$, where $C(x, y)$ is a function on $\mathbb{R}_{+} \times \mathbb{R}_{+}$which is increasing in both variables $x$ and $y$;
(V7) the space of bounded functions with bounded variation is dense in $C\left(X_{j}\right)$ (the space of continuous bounded functions on $X_{j}$ ).
Example 2.1. (1) $X_{j}$ are metric spaces such that $\sup \operatorname{diam}\left(X_{j}\right)<\infty$ and $v_{j}$ is the Hölder constant corresponding to some exponent $\alpha \in(0,1]$ independent of $j$.
(2) $X_{j}$ are Riemannian manifolds such that sup $\operatorname{diam}\left(X_{j}\right)<\infty$ and $v_{j}(g)=\sup |D g|$.
(3) $X_{j}=[0,1], m_{j}=$ Lebesgue and each $v_{j}(g)$ is the usual variation of a function $g$.
(4) $X_{j}=X$ is bounded subset of $\mathbb{R}^{d}, d>1$ and all $m_{j}$ coincide with the normalized Lebsegue measure on $X$. Moreover,

$$
v_{j}(g)=\sup _{0<\varepsilon \leq \varepsilon_{0}} \varepsilon^{-\alpha} \int_{X} \operatorname{Osc}\left(g, B_{\varepsilon}(x)\right) d x
$$

for some constants $\varepsilon_{0}>0$ and $\alpha \in(0,1]$, where $B_{\varepsilon}(x)$ is the ball of radius $\varepsilon$ around a point $x$ and for a set $A, \operatorname{Osc}(g, A)=\sup _{x_{1}, x_{2} \in A \cap X}\left|g\left(x_{1}\right)-g\left(x_{2}\right)\right|$.

Next, given a measurable function $h: X_{j} \rightarrow \mathbb{C}$, denote

$$
\|h\|_{B V}=\|h\|_{B V, j}=\|h\|_{1, j}+v_{j}(h) .
$$

Then $\|\cdot\|_{B V}$ is a complete norm on the space of bounded functions $B_{j}$ with finite variation. Note that $\|\cdot\|_{B V, j}$ are equivalent to the norms $\|g\|_{B V, \infty, j}=\|g\|_{\infty, j}+v_{j}(g)$, uniformly in $j$.

Let $T_{j}: X_{j} \rightarrow X_{j+1}, j \geq 0$ be a sequence of measurable maps, such that

$$
\begin{equation*}
\sup _{j} \sup _{h: v_{j+1}(h) \leq 1} v_{j}\left(h \circ T_{j}\right)<\infty . \tag{2.1}
\end{equation*}
$$

We also assume that the maps are absolutely continuous, that is, $\left(T_{j}\right)_{*} m_{j} \ll m_{j+1}$. Let $\mathcal{L}_{j}$ denote the transfer operator of $T_{j}$ with respect to the measures $m_{j}$ and $m_{j+1}$. Namely, if $\kappa_{j}=\rho_{j} d m_{j}$ for some $\rho_{j} \in L^{1}\left(X_{j}, m_{j}\right)$ then $\mathcal{L}_{j} \rho_{j}: X_{j+1} \rightarrow \mathbb{R}$ is the Radon-Nikodym derivative of the measure $\left(T_{j}\right)_{*} \kappa_{j}$. Then $\mathcal{L}_{j}$ is the unique linear operator satisfying the duality relation:

$$
\begin{equation*}
\int\left(f \circ T_{j}\right) g d m_{j}=\int f\left(\mathcal{L}_{j} g\right) d m_{j+1} \tag{2.2}
\end{equation*}
$$

for all bounded measurable functions $g$ on $X_{j}$ and $f$ on $X_{j+1}$. Denote

$$
\mathcal{L}_{j}^{n}=\mathcal{L}_{j+n-1} \circ \cdots \circ \mathcal{L}_{j+1} \circ \mathcal{L}_{j} .
$$

We make three assumptions which are sequential versions of the assumptions in [19].
(LY1) $\sup _{j}\left\|\mathcal{L}_{j}\right\|_{B V}<\infty$
(LY2) There are constants $\rho \in(0,1), K \geq 1$ and $N \in \mathbb{N}$ such that for every $j$ and a real function $h \in B_{j}$ we have

$$
v_{j+N}\left(\mathcal{L}_{j}^{N} h\right) \leq \rho v_{j}(h)+K\|h\|_{1} .
$$

(SC) For $a>0$ let $\mathcal{C}_{j, a}=\left\{h \in L^{1}\left(m_{j}\right): h \geq 0, v_{j}(h) \leq a m_{j}(h)\right\}$. Then for every $a$ there are $n(a) \geq 1$ and $\alpha(a)>0$ such that for all $j, n \geq n(a)$ and $h \in \mathcal{C}_{a, j}$

$$
\text { ess-inf } \mathcal{L}_{j}^{n} h \geq \alpha(a) m_{j}(h)
$$

Note that (LY) stands for "Lasota Yorke" and (SC) stands for "sequential covering".
In Section 4 we will verify the above assumptions for particular examples.
Remark 2.2. By iterating (LY2) it follows that there is a constant $K_{0}>0$ such that for all $j$ and all $n \geq N$ and a real function $h \in B_{j}$ we have

$$
\begin{equation*}
v_{j+n}\left(\mathcal{L}_{j}^{n} h\right) \leq \rho^{1+n-N} v_{j}(h)+K_{0}\|h\|_{1} . \tag{2.3}
\end{equation*}
$$

Lemma 2.3. Under (LY1) and (LY2) we have

$$
\sup _{j} \sup _{n}\left\|\mathcal{L}_{j}^{n} \mathbf{1}\right\|_{\infty}<\infty
$$

where 1 denotes the function taking the constant value 1, regardless of its domain.
Proof. It follows from the remark and (LY1) that $\sup _{j} \sup _{n} v_{j+n}\left(\mathcal{L}_{j}^{n} \mathbf{1}\right)<\infty$. Next, by property (V3) we have $\left\|\mathcal{L}_{j}^{n} \mathbf{1}\right\|_{\infty} \leq\left\|\mathcal{L}_{j}^{n} 1\right\|_{1}+C_{v a r} v_{j+n}\left(\mathcal{L}_{j}^{n} \mathbf{1}\right)$. To complete the proof of the lemma we note that by taking $f=\mathbf{1}$ in (2.2) with $\mathcal{L}_{j}^{n}$ instead of $\mathcal{L}_{j}$ and $T_{j}^{n}$ instead of $T_{j}$ and using the positivity of $\mathcal{L}_{j}^{n}$ we have $\left\|\mathcal{L}_{j}^{n} \mathbf{1}\right\|_{1}=\|\mathbf{1}\|_{1}=1$.

Appendix A contains the following result.
Theorem 2.4. Suppose (LY1), (LY2) and (SC).
(i) There is a sequence of positive functions $h_{j}$ which is uniformly bounded in $B_{j}$ and uniformly bounded away from 0 with $m_{j}\left(h_{j}\right)=1$, and constants $C>0, \delta \in(0,1)$ such that for all $j \geq 0$ we have $\mathcal{L}_{j} h_{j}=h_{j+1}$ and for all $n \in \mathbb{N}$,

$$
\begin{equation*}
\left\|\mathcal{L}_{j}^{n}(\cdot)-m_{j}(\cdot) h_{j+n}\right\|_{B V} \leq C \delta^{n} \tag{2.4}
\end{equation*}
$$

(ii) Let $\mu_{j}:=h_{j} d m_{j}$. Then $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$. Moreover if $\tilde{\mu}_{j}=g_{j} d m_{j}$ is another sequence satisfying

$$
\begin{equation*}
\left(T_{j}\right)_{*} \tilde{\mu}_{j}=\tilde{\mu}_{j+1} \tag{2.5}
\end{equation*}
$$

then $\lim _{n \rightarrow \infty}\left\|h_{n}-g_{n}\right\|_{1}=0$. In fact, if $g_{k} \in B V$ for some $k$ then $\lim _{n \rightarrow \infty}\left\|h_{n}-g_{n}\right\|_{B V}=0$ exponentially fast.

Remark 2.5. In general, the measures $\mu_{j}$ are not unique. That is 2.5 does not imply that $\tilde{\mu}_{j}=\mu_{j}$ for all $j$. In fact, for every BV density $g_{0}: X_{0} \rightarrow \mathbb{R}$ the measures $\tilde{\mu}_{j}=\left(T_{0}^{j}\right)_{*}\left(g_{0} d m_{0}\right)$ also satisfy (2.5) but in general they differ from $\mu_{j}$ (even when $T_{j}$ does not depend on $j$ ). The point is that the density of $\tilde{\mu}_{j}$ is $g_{j}=\mathcal{L}_{0}^{j} g_{0}$. This density satisfies $\left\|g_{n}-h_{n}\right\|_{B V}=O\left(\delta^{n}\right)$ by Theorem $2.4(\mathrm{i})$. However $g_{n}$ differs from $h_{n}$ in general since $h_{n}$ corresponds to some possibly other choice of $g_{0}$.

Remark 2.6. Let $g \in B_{j}, f \in B_{j+n}$. Then for all $j, n$ we have

$$
m_{j}\left(g \cdot\left(f \circ T_{j}^{n}\right)\right)=m_{j+n}\left(\left(\mathcal{L}_{j}^{n} g\right) f\right) .
$$

Plugging (2.4) into the RHS we get $\left|m_{j}\left(g \cdot f \circ T_{j}^{n}\right)-m_{j}(g) \mu_{j+n}(f)\right| \leq C_{1}\|g\|_{B V}\|f\|_{1} \delta^{n}$ for some constant $C_{1}>0$. Hence $m_{j+n}\left(f \circ T_{j}^{n}\right) \approx \mu_{j+n}(f)$. In particular if the operators $T_{j}$ and measures $m_{j}$ are defined for all $j \in \mathbb{Z}$ so that the assumpions (LY1), (LY2) and (SC) are valid for all $j \in \mathbb{Z}$ then $\mu_{j}(f)=\lim _{n \rightarrow \infty} m_{j-n}\left(f \circ T_{j-n}^{n}\right)$, and so in this case $\mu_{j}$ is the unique equivariant family of measures such that $\sup _{j}\left\|\frac{d \mu_{j}}{d m_{j}}\right\|<\infty$.
2.2. Changing the reference measures. It is important to note that once conditions (LY1), (LY2) and (SC) hold for a given sequence of measures $m_{j}$, they must hold with a uniformly equivalent sequence of measures (and hence, all the limit theorems stated in the next section are valid for initial measures having BV densities with respect to $m_{0}$ ). This is the content of the following result.
Proposition 2.7. Let $\mu_{j}$ be a sequence of probability measures such that $\mu_{j}=h_{j} d m_{j}$, with $h_{j}$ uniformly bounded and bounded away from the origin and $\sup _{j} v_{j}\left(h_{j}\right)<\infty$.

Then (V1)-(V7) and (LY1), (LY2) and (SC) also hold when we repalce $m_{j}$ by $\mu_{j}$.
This proposition will allow us to prove limit theorems with respect to the Lebesgue measures (i.e. the case $m_{j}=$ Lebesgue), as well as with respect to the unique absolutely continuous (sequentially) invariant measures $\mu_{j}$ or any other sequence of equivalent measures. This is one of the advantages of the setup presented in this section.

Proof. First, it is clear that there is a constant $C_{1}>0$ such that for every function $g: X_{j} \rightarrow \mathbb{C}$,

$$
\begin{equation*}
C_{1}\|g\|_{L^{1}\left(m_{j}\right)} \leq\|g\|_{L^{1}\left(\mu_{j}\right)} \leq\|g\|_{L^{1}\left(m_{j}\right)} . \tag{2.6}
\end{equation*}
$$

Consequently, the $B V$ norms induced from both measures are equivalent. Moreover, since the measures $m_{j}$ and $\mu_{j}$ are equivalent we have $\|\cdot\|_{L^{\infty}\left(m_{j}\right)}=\|\cdot\|_{L^{\infty}\left(\mu_{j}\right)}$. Therefore (V1)-(V7) remain true if we replace $m_{j}$ with $\mu_{j}$.

Next, the transfer operator $L_{j}$ corresponding to the measures $\mu_{j}$ and $\mu_{j+1}$ is given by

$$
L_{j} g=\frac{\mathcal{L}_{j}\left(g h_{j}\right)}{h_{j+1}} .
$$

By (V6), $\sup v_{j}\left(1 / h_{j}\right)<\infty$. So conditions (LY1) and (LY2) also hold true with $L_{j}$ instead of $\mathcal{L}_{j}$ (possibly with different constants).

Finally, notice that the operator $\mathcal{L}_{j}$ is positive. Therefore, if $c$ is a constant such that $h_{j} \geq c>0$ then for every function $h \geq 0$,

$$
\mathcal{L}_{j}^{n}\left(h h_{j}\right) \geq c \mathcal{L}_{j}^{n}(h), \quad m_{j+1} \text { a.s. }
$$

Now, since $\left\|1 / h_{j+1}\right\|_{L^{\infty}} \geq D$ for some constant $D<\infty$, we conclude that for every non-negative function $h$ and all $n$ we have $L_{j}^{n} h \geq D c \mathcal{L}_{j}^{n} h$, a.s. Consequently, the validity of condition (SC) for the operators $L_{j}$ follows from the corresponding validity for $\mathcal{L}_{j}$, together with 2.6).

## 3. Main results

Let $f_{j}: X_{j} \rightarrow \mathbb{R}, j \geq 0$ be a sequence of measurable functions such that $\sup _{j}\left\|f_{j}\right\|_{B V}<\infty$. Set $S_{n}(x)=\sum_{j=0}^{n-1} f_{j}\left(T_{0}^{j} x\right)$ where $T_{k}^{j}=T_{k+j-1} \circ \cdots \circ T_{k+1} \circ T_{k}$.

We consider the sequence of functions $S_{n}: X_{0} \rightarrow \mathbb{R}$ as random variables on the probability space ( $X_{0}, \mathcal{B}_{0}, m_{0}$ ). Our results will be limit theorems for such sequences.

Denote $\sigma_{n}=\sqrt{\operatorname{Var}_{m_{0}}\left(S_{n}\right)}$ and $\bar{S}_{n}=S_{n}-m_{0}\left(S_{n}\right)$. Let $F_{n}(t)=\mathbb{P}_{m_{0}}\left(\bar{S}_{n} / \sigma_{n} \leq t\right)$ be the distribution function of $\bar{S}_{n} / \sigma_{n}$, and let

$$
\Phi(t)=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{t} e^{-x^{2} / 2} d x
$$

be the standard normal distribution function.
Recall that the (self-normalized) central limit theorem (CLT) means that for every real $t$

$$
\lim _{n \rightarrow \infty} F_{n}(t)=\Phi(t) .
$$

The CLT in our setup can be proven using martingale coboundary decomposition of 86.2 and applying an appropriate CLT for martingales (cf. [42, 24]). We refer to 86.2 for a characterization when $\sigma_{n} \rightarrow \infty$. In this paper we will not give a separate proof of the CLT since our main results give not only the CLT but also rate of convergence, in various metrics.

Theorem 3.1. Suppose $\sigma_{n} \rightarrow \infty$.
(i) (A non-uniform Berry-Esseen theorem). $\forall s \geq 0$ there is a constant $C_{s}$ such that

$$
\sup _{t \in \mathbb{R}}\left(1+|t|^{s}\right)\left|F_{n}(t)-\Phi(t)\right| \leq C_{s} \sigma_{n}^{-1}
$$

(ii) (A Berry-Esseen theorem in $\left.L^{p}\right)$. For all $p \geq 1$ we have $\left\|F_{n}-\Phi\right\|_{L^{p}(d x)}=O\left(\sigma_{n}^{-1}\right)$.
(iii) For all $s \geq 1$ there is a constant $C_{s}$ such that for every absolutely continuous function $h: \mathbb{R} \rightarrow \mathbb{R}$ such that $H_{s}(h):=\int \frac{\left|h^{\prime}(x)\right|}{1+|x|^{s}} d x<\infty$ we have $\left|\mathbb{E}_{m_{0}}\left[h\left(\bar{S}_{n} / \sigma_{n}\right)\right]-\int h d \Phi\right| \leq C_{s} H_{s}(h) \sigma_{n}^{-1}$.

Next, recall that the $p$-th Wasserstien distance between two probability measures $\mu, \nu$ on $\mathbb{R}$ with finite absolute moments of order $p$ is given by

$$
W_{p}(\mu, \nu)=\inf _{(X, Y) \in \mathcal{C}(\mu, \nu)}\|X-Y\|_{L^{p}}
$$

where $\mathcal{C}(\mu, \nu)$ is the class of all pairs of random variables $(X, Y)$ on $\mathbb{R}^{2}$ such that $X$ is distributed according to $\mu$, and $Y$ is distributed according to $\nu$. Combining our estimates with the main results in [52] also yields the following result.

Theorem 3.2. [A Berry-Esseen theorem in $W_{p}$ ] For every $p$ we have

$$
W_{p}\left(d F_{n}, d \Phi\right)=O\left(\sigma_{n}^{-1}\right)
$$

where $d G$ is the measure induced by a distribution function $G$.
A key ingredient in the proof of Theorems 3.1 and 3.2 is the following proposition, which we believe has its own interest.

Proposition 3.3. For every $p \geq 2$ there is a constant $C$ such that for all $j \geq 0$ and $n \in \mathbb{N}$,

$$
\left\|\bar{S}_{j, n} f\right\|_{L^{p}\left(m_{0}\right)} \leq C\left(1+\left\|\bar{S}_{j, n}\right\|_{L^{2}}\right)
$$

Moreover, $C$ depends only on $p, \sup _{j}\left\|f_{j}\right\|_{B V}$, and the constants from assumptions (V1)-(V7), (LY1), (LY2) and (SC).

## 4. Examples

Here we exhibit several classes of systems fitting in the abstract setup of Section 2 .
4.1. Piecewise expanding maps on the interval. We take $X_{j}=I=[0,1]$ and $m_{j}=$ Lebesgue for all $j$. Let $v_{j}=v$ be the usual variation of functions on $[0,1]$ :

$$
\left.v(g)=\sup _{n} \sup _{t_{0}=0<t_{1}<\ldots<t_{n}<t_{n+1}=1} \sum_{j=0}^{n} \mid g\left(t_{n+1}\right)-g\left(t_{n}\right)\right) \mid .
$$

We suppose that for each $j$ we can write $[0,1]=\bigcup_{k=1}^{d_{j}} I_{j, k}$ where $I_{j, k}, 1 \leq k \leq d_{j}$ are intervals with disjoint interiors such that $\sup d_{j}<\infty$.

We also suppose that for all $j$ and $k$ the restriction $T_{j, k}:=T_{j} \mid I_{j, k}$ is a $C^{2}$ expanding map so that $\sup _{j} \max _{1 \leq k \leq d_{j}}\left\|T_{j, k}^{\prime \prime}\right\|_{\infty}<\infty$ and $\delta:=\inf _{j} \min _{1 \leq k \leq d_{j}} \inf \left|T_{j, k}^{\prime}\right|>1$. Moreover, we assume that

$$
\inf _{j} \min _{1 \leq k \leq d_{j}}\left|I_{j, k}\right|>0 .
$$

Let $\mathcal{L}_{j}$ be the transfer operator of $T_{j}$, namely the operator given by

$$
\mathcal{L}_{j} g(x)=\sum_{k: x \in T_{j, k}\left(I_{j, k}\right)} \frac{g\left(T_{j, k}^{-1} x\right)}{T_{j}^{\prime}\left(T_{j, k}^{-1} x\right)} .
$$

Then (LY1) holds. A standard argument (see [63]) yields that if $N$ satisfies $\delta^{N}>2$ then there are constants $K \geq 1$ and $\rho \in(0,1)$ such that for all $j$ we have $\operatorname{var}\left(\mathcal{L}_{j} g\right) \leq \rho \operatorname{var}(g)+\|g\|_{L^{1}}$. Thus, (LY2) holds.

Next, in order to verify (SC) we can assume that for every interval $J \subset[0,1]$ there is $n(J) \in \mathbb{N}$ such that for every $j$ we have

$$
\begin{equation*}
T_{j}^{n(J)} J=[0,1] . \tag{4.1}
\end{equation*}
$$

Under the above condition the verification (SC) is carried out similarly to [19, §1.2].
For piecewise expanding maps satisfying the assumptions described above we get all the limit theorems described in the previous section for sums of the form $S_{n}=\sum_{j=0}^{n-1} f_{j} \circ T_{0}^{j}$ considered as random variables with respect to a measures which is absolutely continuous with respect to Lebesgue and its density is a BV function. Here, $f_{j}$ must satisfy $\sup _{j}\left\|f_{j}\right\|_{B V}<\infty$.
4.2. High dimensional piecewise expanding maps. Let $X_{j}=X$ coincide with a single compact subset of $\mathbb{R}^{k}$ for some $k>1$. Let $m$ be the normalized Lebesgue measure on $X$ and let $v$ be the variation defined in Example 2.1(iv).

We suppose that the maps $T_{j}$ have the following properties. There are constants $d \in \mathbb{N}$, $\gamma, C, \varepsilon>0$ and $s \in(0,1)$ with the following properties. For each $j$ there are disjoint sets $A_{j, i}, \tilde{A}_{j, i}, 1 \leq i \leq d_{j} \leq d$ and maps $T_{j, i}: \tilde{A}_{j, i} \rightarrow X$ such that:
(i) The sets $A_{j, 1}, \ldots, A_{j, d_{j}}$ are disjoint and $m\left(X \backslash \bigcup_{i=1}^{d_{j}} A_{j, i}\right)=0$. Moreover, $A_{j, i} \subset \tilde{A}_{j, i}$;
(ii) Each $T_{j, i}$ is a $C^{1+\gamma}$ function;
(iii) $\left.T_{j}\right|_{A_{j, i}}=T_{j, i}$ and for all $j$ and $i$ and $B_{\varepsilon}\left(T_{j, i} A_{j, i}\right) \subset T_{j, i}\left(\tilde{A}_{j, i}\right)$, where $B_{\varepsilon}(A)$ is the $\varepsilon$ neighborhood of a set $A$;
(iv) For all $j$ and $i$ the function $J_{j, i}=\operatorname{Det}\left(D T_{j, i}^{-1}\right)$ satisfies that for all $x, y \in T_{j, i}\left(A_{j, i}\right)$,

$$
\left|\frac{J_{j, i}(y)}{J_{j, i}(x)}-1\right| \leq C \operatorname{dist}(x, y)^{\gamma} ;
$$

(v) For every $x, y \in T_{j, i}\left(\tilde{A}_{j, i}\right)$ with $\operatorname{dist}(x, y) \leq \varepsilon$ we have $\operatorname{dist}\left(T_{j, i}^{-1} x, T_{j, i}^{-1} y\right) \leq s \cdot \operatorname{dist}(x, y)$;
(vi) Each $\partial A_{j, i}$ is co-dimension one embedded compact $C^{1}$-submanifold and

$$
s^{\gamma}+\frac{4 s}{1-s} Z \frac{\Gamma_{k}}{\Gamma_{k-1}}<1
$$

where $\Gamma_{k}$ is the volume of the unit ball in $\mathbb{R}^{k}$ and $Z=\sup _{j} \sup _{x} \sum_{i} \mathbb{I}\left(x \in \tilde{A}_{i, j}\right)$.
Under the above assumptions (LY1) and (LY2) with $N=1$ are satisfied (see [78, Lemma 4.1]).
Next, we also assume that for any open set $U$ there exists $n(U) \in \mathbb{N}$ such that for all $j$

$$
\begin{equation*}
T_{j}^{n(U)} U=X . \tag{4.2}
\end{equation*}
$$

Under the above condition the verification (SC) is carried out similarly to [30, Lemma 3].
One example where 4.2 holds are Markov maps. That is, we assume that for each $i, j$ the image $T_{j} A_{i j}$ is a union of some of the sets $A_{j+1, k}$. Moreover suppose that the system is uniformly mixing in the sense that $\exists \ell$ such that $T_{j}^{\ell} A_{i j}=X_{j+\ell}=X$ for each $i, j$. This condition can be verified as follows. Consider the adjacency matrix $\mathcal{A}_{j}$ such that $\mathcal{A}_{j}(i, k)=1$ if $T_{j} A_{j, i} \supset A_{j+1, k}$. Then the uniform mixing assumption means that for each $j$ all entries of $\mathcal{A}_{j+\ell-1} \cdots \mathcal{A}_{j+1} \mathcal{A}_{j}$ are positive. Now let $U$ be an open set in $X_{j}=X$. Without loss of generality we may assume that $U$ is an open ball with center $x$ and radius $r$. Given $k$ let $\mathcal{B}_{j, k}(x)$ denote the set of points $y \in X$ such that $T_{j}^{m} x$ and $T_{j}^{m} y$ belong to the same elements of our partition $\left(A_{j+m, q}\right)_{q}$ for all $m<k$. By our assumptions $\operatorname{diam}\left(\mathcal{B}_{j, k}(x)\right) \leq C_{0} s^{k}$ for some constant $C_{0}>0$, and so for sufficiently large $\bar{k}$ we have $\mathcal{B}_{j, \bar{k}}(x) \subset U$. Accordingly, $T_{j}^{\bar{k}} U$ contains one of elements of our Markov partition and then $T_{j}^{\bar{k}+\ell} U=X$.
4.3. Covering maps and sequential SFT. Suppose that each $X_{j}$ is a metric space. Let $\mathrm{d}_{j}$ be the metric on $X_{j}$ and suppose that $\operatorname{diam}\left(X_{j}\right) \leq 1$. Let $v_{j}=v_{j, \alpha}$ be the Hölder constant corresponding to some fixed exponent $\alpha \in(0,1]$.
Assumption 4.1. (Pairing). There are constants $\xi \leq 1$ and $\gamma>1$ such that for every two points $x, x^{\prime} \in X_{j+1}$ with $\mathrm{d}_{j+1}\left(x, x^{\prime}\right) \leq \xi$ we can write

$$
T_{j}^{-1}\{x\}=\left\{y_{i}(x): i \leq k\right\}, \quad T_{j}^{-1}\left(x^{\prime}\right)=\left\{y_{i}\left(x^{\prime}\right): i \leq k\right\}
$$

where $\mathbf{d}_{j}\left(y_{i}(x), y_{i}\left(x^{\prime}\right)\right) \leq \gamma^{-1} \mathbf{d}_{j+1}\left(x, x^{\prime}\right)$ for all $i$.
Moreover sup $\operatorname{deg}\left(T_{j}\right)<\infty$, where $\operatorname{deg}(T)$ is the largest number of preimgaes that a point $x$ can have under the map $T$.

Denote by $B_{j}(x, r)$ the open ball of radius $r$ around a point $x \in X_{j}$.
Assumption 4.2. (Covering). There exists $n_{0} \in \mathbb{N}$ such that for every $j$ and $x \in X_{j}$ we have

$$
\begin{equation*}
T_{j}^{n_{0}}\left(B_{j}(x, \xi)\right)=X_{j+n_{0}} \tag{4.3}
\end{equation*}
$$

Fix some $\alpha \in(0,1]$ and a sequence of functions $\phi_{j}: X_{j} \rightarrow \mathbb{R}$ such that $\sup _{j}\left\|\phi_{j}\right\|_{\alpha}<\infty$. Here $\left\|\phi_{j}\right\|_{\alpha}=\sup \left|\phi_{j}\right|+v_{j}\left(\phi_{j}\right)$ and $v_{j}\left(\phi_{j}\right)$ is the Hölder constant of $\phi_{j}$ corresponding to the exponent $\alpha$. Let $L_{j}$ be the operator which maps a function $g: X_{j} \rightarrow \mathbb{R}$ to a function $L_{j} g: X_{j+1} \rightarrow \mathbb{R}$ given
by $L_{j} g(x)=\sum_{T_{j} y=x} e^{\phi_{j}(y)} g(y)$. Then (see $\left\{5.2\right.$ there is a sequence of probability measures $\nu_{j}$ on $X_{j}$ such that $\left(\mathcal{L}_{j}\right)^{*} \nu_{j+1}=\lambda_{j} \nu_{j}$, where $\lambda_{j}>0$ is bounded and bounded away from 0 . Then we can take any measure $m_{j}$ of the form $m_{j}=u_{j} d \nu_{j}$ with $\sup _{j}\left\|u_{j}\right\|_{\alpha}<\infty$. This includes the unique sequence of measures $\mu_{j}$ which are absolutely continuous with respect to $\nu_{j}$ and $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$ (see $\$ 5.2$ ). This setup includes the following more concrete examples.
4.3.1. Smooth expanding maps. Let $M$ be $C^{2}$ compact connected Riemannian manifold, and let $X_{j}=M$ for all $j$. Let $T_{j}: M \rightarrow M$ be $C^{2}$ endomorphisms of $M$ such that

$$
\sup _{j}\left\|D T_{j}\right\|<\infty \text { and } \sup _{j}\left\|\left(D T_{j}\right)^{-1}\right\|<1
$$

Then the arguments in [61, Section 4] (see [61, (4.6)]) yield that Assumption 4.1 is in force with some $\xi>0$. Next, arguing like at the paragraph below [61, (4.19)]) we see that Assumption 4.2 holds if $n_{0}$ is large enough. Take $\phi_{j}=-\ln J_{T_{j}}$. Then by [48, Theorem 3.3 and Proposition 3.4] (see also [61, Theorem 2.2]) we see that the measures $\nu_{j}$ described after Assumptions 4.2 coincide with the normalized volume measure on $M$. Thus, we get all the limit theorems with respect to any absolutely continuous measure whose density is Hölder continuous.
4.3.2. Subshifts of finite type. Let $\mathcal{A}_{j}=\left\{0,1, \ldots, d_{j}-1\right\}$ with $\sup _{j} d_{j}<\infty$. Le $A^{(j)}$ be matrices of sizes $d_{j} \times d_{j+1}$ with $0-1$ entries. We suppose that there exists an $M \in \mathbb{N}$ such that for every $j$ the matrix $A^{(j)} \cdot A^{(j+1)} \cdots A^{(j+M)}$ has positive entries. Define

$$
\begin{equation*}
X_{j}=\left\{\left(x_{j, k}\right)_{k=0}^{\infty}: x_{j, k} \in \mathcal{A}_{j+k}, A_{x_{j, k}, x_{j, k+1}}^{(j+k)}=1\right\} \tag{4.4}
\end{equation*}
$$

Let $T_{j}: X_{j} \rightarrow X_{j+1}$ be the left shift. Consider a metric $\mathrm{d}_{j}$ on $X_{j}$ given by

$$
\mathbf{d}_{j}(x, y)=2^{-\inf \left\{k: x_{j, k} \neq y_{j, k}\right\}} .
$$

With this metric the maps $T_{j}$ satisfy Assumptions 4.1 and 4.2. In order to introduce appropriate measures $m_{0}$ note that we can extend the dynamics for negative times by defining $X_{j}$ for $j<0$ via appropriate extensions of the sequences $\left(d_{j}\right)$ and $\left(A^{(j)}\right)$. Note that such extensions are highly non-unique. Each one of these extensions gives raise to a unique time 0 Gibbs measure $\mu_{0}$ (see Appendix B). Thus, all of our results hold true when starting with any measure which is absolutely continuous with Hölder continuous density with respect to $m_{0}=\mu_{0}$.
4.3.3. Two sided SFT. Using the same notations like in the previous section but also considering negative integers $j$, we define

$$
\begin{equation*}
\tilde{X}_{j}=\left\{\left(x_{j, k}\right)_{k=-\infty}^{\infty}: x_{j, k} \in \mathcal{A}_{j+k}, \quad A_{x_{j, k}, x_{j, k+1}}^{(j+k)}=1\right\} . \tag{4.5}
\end{equation*}
$$

Let $\tilde{T}_{j}: \tilde{X}_{j} \rightarrow \tilde{X}_{j+1}$ be the left shift, and let $\tilde{T}_{j}^{n}=\tilde{T}_{j+n-1} \circ \cdots \circ \tilde{T}_{j+1} \circ \tilde{T}_{j}$. Consider the metric $\tilde{\mathrm{d}}_{j}$ on $\tilde{X}_{j}$ given by $\tilde{\mathrm{d}}_{j}(x, y)=2^{-\inf \left\{|k|: x_{j, k} \neq y_{j, k}\right\}}$. Let $\pi_{j}: \tilde{X}_{j} \rightarrow X_{j}$ be the natural projection.

By Lemma B. 2 (proven in Appendix B), given a sequence of functions $\mathbf{f}_{j}: \tilde{X}_{j} \rightarrow \mathbb{R}$ with $\sup \left\|\mathbf{f}_{j}\right\|_{\alpha}<\infty$ there is sequences of functions $f_{j}: X_{j} \rightarrow \mathbb{R}$ and $u_{j}: \tilde{X}_{j} \rightarrow \mathbb{R}$ such that $\sup _{j}^{j}\left\|f_{j}\right\|_{\alpha / 2}<\infty, \mathfrak{u}=\sup _{j}\left\|u_{j}\right\|_{\alpha / 2}<\infty$ and

$$
\mathbf{f}_{j}=f_{j} \circ \pi_{j}+u_{j+1} \circ \tilde{T}_{j}-u_{j} .
$$

Therefore, denoting $S_{n} \mathbf{f}=\sum_{j=0}^{n-1} \mathbf{f}_{j} \circ \tilde{T}_{0}^{j}$ we have $\sup _{n} \sup \left|S_{n} \mathbf{f}-S_{n} f\right|<\infty$.

Next, let $\gamma_{j}$ denote a Gibbs measure of the two sided shift at time $j$ (see Appendix B). Then $\mu_{j}=\left(\pi_{j}\right)_{*} \gamma_{j}$ are also Gibbs measures for one sided subshift and they satisfy the assumptions of Section 2 (see Appendix B for details). We view $S_{n} \mathbf{f}$ as random variables on ( $\tilde{X}_{0}$, Borel, $\gamma_{0}$ ). Then

$$
\begin{equation*}
A:=\sup _{n}\left\|S_{n} \mathbf{f}-S_{n} f\right\|_{L^{\infty}\left(\gamma_{0}\right)}<\infty . \tag{4.6}
\end{equation*}
$$

This is enough to do deduce all of our results for $S_{n} \mathbf{f}$, relying on the corresponding results for $S_{n} f$. Indeed, part (ii) of Theorem 3.1 is a direct consequence of part (i). Theorem 3.1(iii) also follows from Theorem 3.1 (i). Indeed, for every random variable $W$ with distribution function $F$ and a function $h$ satisfying $H_{s}(h)<\infty$ we have

$$
E[h(W)]-h(\infty)=-E\left[\int_{W}^{\infty} h^{\prime}(x) d x\right]=-\int_{-\infty}^{\infty} h^{\prime}(x) P(W \leq x) d x=-\int_{-\infty}^{\infty} h^{\prime}(x) F(x) d x .
$$

To show that Theorem 3.1 (i) for $S_{n} f$ implies Theorem 3.1 (i) for $S_{n} \mathbf{f}$, let

$$
F_{n}(t)=\mathbb{P}\left(\frac{S_{n} \mathbf{f}}{\sigma_{n}} \leq t\right), \quad G_{n}(t)=\mathbb{P}\left(\frac{S_{n} f}{\kappa_{n}} \leq t\right) \quad \text { where } \kappa_{n}=\left\|S_{n} f\right\|_{L^{2}} \text { and } \sigma_{n}=\left\|S_{n} \mathbf{f}\right\|_{L^{2}} .
$$

By (4.6) and the triangle inequality, $\left|\sigma_{n}-\kappa_{n}\right| \leq A$. To complete the proof fix $s \geq 0$. Then

$$
\begin{gathered}
F_{n}(t) \leq G_{n}\left(t \sigma_{n} / \kappa_{n}+A / \kappa_{n}\right) \leq \Phi\left(t \sigma_{n} / \kappa_{n}+A / \kappa_{n}\right)+C_{s}\left(1+\left|t \sigma_{n} / \kappa_{n}+A / \kappa_{n}\right|^{s}\right)^{-1} \\
\leq \Phi(t)+C(1+|t|) \sigma_{n}^{-1} e^{-c t^{2}}+\tilde{C}_{s}\left(1+|t|^{s}\right)^{-1}
\end{gathered}
$$

where in the penultimate inequality we have used that $|\Phi(x+\varepsilon)-\Phi(x)| \leq C \varepsilon e^{-x^{2} / 2}$ for every $x$ and $\varepsilon>0$ and that $\left|t \sigma_{n} / \kappa_{n}+A / \kappa_{n}-t\right| \leq C(|t|+1) / \sigma_{n}$ for some constant $C>0$. Similarly,

$$
\left.F_{n}(t) \geq G_{n}\left(t \sigma_{n} / \kappa_{n}-A / \kappa_{n}\right)\right) \geq \Phi(t)-C(1+|t|) \sigma_{n}^{-1} e^{-c t^{2}}-\tilde{C}_{s}\left(1+|t|^{s}\right)^{-1}
$$

Finally to deduce Theorem 3.2 for $S_{n} \mathbf{f}$ from the corresponding result for $S_{n} f$ let us fix some $p \geq 1$. Then by Theorem 3.2, we can couple $S_{n} f$ with a standard normal random variable $Z$ so that $\left\|S_{n} f / \sigma_{n}-Z\right\|_{L^{p}} \leq C \sigma_{n}^{-1}$. Now by Berkes-Philipp Lemma [17, Lemma A.1], we can also couple all three random variables $S_{n} f, S_{n} \mathbf{f}$ and $Z$ so that 4.6) still holds under the new probability law.

## 5. Verification of (LY1), (LY2) and (SC) for some classes of maps

In this section we will show that the conditions in Section 3 are in force for the classes of expanding maps considered in both [24] and [48].
5.1. Verification of our assumptions in the setup of [24]. The following assumption is taken form [24].

Assumption 5.1. (i) $\left(X_{j}, \mathcal{B}_{j}, m_{j}\right)$ coincide with the same probability space $(X, \mathcal{B}, m), v_{j}=v$ does not depend on $j$.
(ii) Conditions (LY1) and (LY2) hold with $m_{j}=m$ and $v_{j}=v$.
(iii) There is a constant $\delta_{0}>0$ such that

$$
\begin{equation*}
\text { ess-inf } \mathcal{L}_{0}^{n} \mathbf{1} \geq \delta_{0} \tag{5.1}
\end{equation*}
$$

(note that the condition (5.1) is denoted by (Min) in [24]). Here 1 denotes the function taking the constant value 1 on $X$.
(iv) There are constants $C_{1}>0$ and $\delta_{1} \in(0,1)$ such that for every $h \in B$ with $m(h)=0$

$$
\begin{equation*}
\left\|\mathcal{L}_{j}^{n}\right\|_{B V} \leq C_{1} \delta_{1}^{n} \tag{5.2}
\end{equation*}
$$

Lemma 5.2. Assumption 5.1 implies the covering condition (SC).
Thus, Assumption 5.1 is less general than the combination of (LY1), (LY2) and (SC).
Proof. Since (LY1) and (LY2) hold, Lemma 2.3 implies that $C=\sup _{k}\left\|\mathcal{L}_{0}^{k} 1\right\|_{\infty} \in(0, \infty)$. Hence using the positivity of the operators $\mathcal{L}_{k}$ we see that $m$-a.s. we have

$$
\delta_{0} \leq \mathcal{L}_{0}^{j+n} \mathbf{1}=\mathcal{L}_{j}^{n}\left(\mathcal{L}_{0}^{j} \mathbf{1}\right) \leq C \mathcal{L}_{j}^{n} \mathbf{1} .
$$

We thus conclude that $\inf _{j} \operatorname{ess}-\inf \mathcal{L}_{j}^{n} \mathbf{1} \geq \delta_{2}=\delta_{0} / C$.
Next, fix some $a$ and let $h \in \mathcal{C}_{a, j}$ where $\mathcal{C}_{a, j}$ comes from the condition (SC). Denote $\|\cdot\|=\|\cdot\|_{B V}$. Then by (5.2) we have

$$
\left\|\mathcal{L}_{j}^{n} h-m(h) \mathcal{L}_{j}^{n} \mathbf{1}\right\|=\left\|\mathcal{L}_{j}^{n}(h-m(h) \mathbf{1})\right\| \leq C_{1}\|h-m(h) \mathbf{1}\| \delta_{1}^{n} \leq C_{2}\|h\| \delta_{1}^{n} \leq C_{3}(a) m(h) \delta_{1}^{n}
$$

for some constant $C_{3}(a)$ which depends only on $a$ (we can take $C_{3}(a)=C_{2}(1+a)$ because $\|h\|=v(h)+m(h) \leq(1+a) m(h))$. Using that $\mathcal{L}_{j}^{n} 1 \geq \delta_{2}>0$ we conclude that

$$
\mathcal{L}_{j}^{n} h \geq \delta_{2} m_{j}(h)-C_{3}(a) m(h) \delta_{1}^{n}=\left(\delta_{2}-C_{3}(a) \delta_{1}^{n}\right) m(h) \quad(m \text {-a.s. }) .
$$

Let $n(a)$ be the smallest positive integer such that $C_{3}(a) \delta_{1}^{n(a)} \leq \frac{1}{2} \delta_{2}$. Then for $n \geq n(a)$ we have $\mathcal{L}_{j}^{n} h \geq \frac{1}{2} \delta_{2} m(h)$ ( $m$-a.s.) and so (SC) holds with $\alpha(a)=\frac{1}{2} \delta_{2}$.
5.2. Verification of our assumptions in the setup of 48. Consider maps $T_{j}$ from 4.3 . We also suppose that we can extend the sequence $\left(T_{j}\right)_{j \geq 0}$ to a two sided sequence $\left(T_{j}\right)_{j \in \mathbb{Z}}$ with the same properties. This is possible if there is a map $T_{-1}: X_{0} \rightarrow X_{0}$ such that the sequence $\left(T_{-1}^{j}\right)_{j \geq 0}$ has same paring property and covering assumption like the sequence $\left(T_{j}\right)_{j \geq 0}$ (this is the case when $X_{0}=X_{1}$ ). Indeed, in this case we can define $T_{j}=T_{-1}$ for $j<0$. The need in a two sided sequences in this context arises from the lack of a given reference measure $m_{0}$, as will be elaborated in what follows.

Fix some Hölder exponent $\alpha \in(0,1]$ and let $\phi_{j}: X_{j} \rightarrow \mathbb{R}, j \in \mathbb{Z}$ be such that $\sup _{j}\left\|\phi_{j}\right\|_{\alpha}<\infty$. Let the operator $\mathbf{L}_{j}$ be given by

$$
\mathbf{L}_{j} g(x)=\sum_{y: T_{j} y=x} e^{\phi_{j}(y)} g(y) .
$$

Then as proven ${ }^{17}$ in [48], there are strictly positive functions $\bar{h}_{j} \in B_{j}$, probability measures $\nu_{j}$ on $X_{j}$ and numbers $\lambda_{j}>0$ such that $0<\inf _{j} \inf \bar{h}_{j} \leq \sup _{j}\left\|\bar{h}_{j}\right\|_{\alpha}<\infty$ and

$$
\begin{equation*}
\left\|\mathbf{L}_{j}^{n} / \lambda_{j, n}-\nu_{j} \otimes \bar{h}_{j+n}\right\|_{\alpha} \leq C \delta^{n}, \quad \delta \in(0,1) \tag{5.3}
\end{equation*}
$$

where $\lambda_{j, n}=\prod_{k=j}^{j+n-1} \lambda_{k}$ and $(\nu \otimes h)$ is the linear operator $g \rightarrow \nu(g) h$. Let

$$
\mathcal{L}_{j}(g)=\mathbf{L}_{j}\left(g \bar{h}_{j}\right) / \lambda_{j} \bar{h}_{j+1}
$$

and let $m_{j}=\mu_{j}$ be the sequential Gibbs measures corresponding to the potentials $\left(\phi_{j}\right)$, that is $\mu_{j}=\bar{h}_{j} \nu_{j}$. Then $\mathcal{L}_{j}$ is the dual of $T_{j}$ with respect to $\mu_{j}$. Moreover, $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$ and $\mathcal{L}_{j} \mathbf{1}=\mathbf{1}$.

Proposition 5.3. The operators $\mathcal{L}_{j}$ obey conditions (LY1), (LY2) and (SC).

[^0]Proof. First, the uniform boundedness (LY1) of the operators $\mathcal{L}_{j}$ follows from the properties of the non-normalized RPF triplets (or from (5.4)). Second, the Lasota-Yorke inequality (LY2) was obtained in [46, Lemma 5.12.2]. Note that in [46, Lemma 5.12.2] the weak norm $\|\cdot\|_{L^{1}}$ is replaced with the (weak) norm $\|\cdot\|_{\infty}$. However, we have $\|g\|_{\infty} \leq \varepsilon_{r} v(g)+C_{r}\|g\|_{L^{1}}$ with $\varepsilon_{r} \rightarrow 0$ as $r \rightarrow 0$. Using this we have that the Lasota-Yorke inequality with respect to $\left(v_{j}(\cdot),\|\cdot\|_{L^{1}}\right)$ is equivalent to the Lasota-Yorke inequality with respect to $\left(v_{j}(\cdot),\|\cdot\|_{\infty}\right)$.

Third, in [48] we proved that there are constants $C>0$ and $\delta \in(0,1)$ such that for all $j, n$ and a Hölder continuous function $h: X_{j} \rightarrow \mathbb{R}$,

$$
\begin{equation*}
\left\|\mathcal{L}_{j}^{n} h-m_{j}(h) \mathbf{1}\right\|_{\alpha} \leq C \delta^{n} \tag{5.4}
\end{equation*}
$$

and so (5.2) holds. Since $\mathcal{L}_{j} \mathbf{1}=\mathbf{1}$ we have $\inf _{j, n} \inf \mathcal{L}_{j}^{n} \mathbf{1}=\delta_{0}=1$. Thus we can repeat the arguments from the previous section verbatim with time dependent $v$ and $m$ and obtain (SC).

## 6. Moment estimates

Let $T_{j}^{n}=T_{j+n-1} \circ \cdots \circ T_{j+1} \circ T_{j}$. Consider a sequence of real valued functions $f_{k} \in B_{k}, k \geq 0$ such that $\sup _{k}\left\|f_{k}\right\|_{B V}<\infty$. Our goal is to obtain limit theorems for the sequence $S_{n}=\sum_{j=0}^{n-1} f_{j} \circ T_{0}^{j}$ considered as random variables on the probability space $\left(X, \mathcal{B}, m_{0}\right)$. It what follows it will be convenient to consider $\left(f_{j}\right)$ as two sided sequence by setting $f_{j}=0$ for $j<0$.
6.1. The pulled back measures and their transfer operators. Consider the sequence of measures $\tilde{m}_{j}=\left(T_{0}^{j}\right)_{*} m_{0}$ on $\left(X_{j}, \mathcal{B}_{j}\right)$. Then $\tilde{m}_{0}=m_{0}$ and for all $j>0$ and each bounded measurable function $G: X_{j} \rightarrow \mathbb{C}$,

$$
\tilde{m}_{j}(G)=m_{0}\left(G \circ T_{0}^{j}\right)=m_{j}\left(G \cdot \mathcal{L}_{0}^{j} \mathbf{1}\right) .
$$

Remark 6.1. Note that in the case when $m_{j}=\mu_{j}$ is an equivariant sequence (i.e. $\left(T_{j}\right)_{*} \mu_{j}=$ $\left.\mu_{j+1}\right)$ then $\tilde{m}_{j}=\mu_{j}$.

Let the operator $\tilde{\mathcal{L}}_{j}$ be the dual of $T_{j}$ with respect to the measure $\tilde{m}_{j}$ and $\tilde{m}_{j+1}$ defined by the duality relation

$$
\begin{equation*}
\int f \cdot\left(\tilde{\mathcal{L}}_{j} g\right) d \tilde{m}_{j+1}=\int\left(f \circ T_{j}\right) \cdot g \cdot d \tilde{m}_{j} \tag{6.1}
\end{equation*}
$$

for all functions such that both integrals are well defined. Then $\tilde{\mathcal{L}}_{j} \mathbf{1}=\mathbf{1}$ because $\left(T_{j}\right)_{*} \tilde{m}_{j}=\tilde{m}_{j+1}$. Let

$$
\tilde{\mathcal{L}}_{j}^{n}=\tilde{\mathcal{L}}_{j+n-1} \circ \cdots \circ \tilde{\mathcal{L}}_{j+1} \circ \tilde{\mathcal{L}}_{j}
$$

Proposition 6.2. There are constants $C_{0}>0, J \in \mathbb{N}$ and $\delta_{0} \in(0,1)$ such that for all $j \geq 0$ and $n \geq 1$ such that $j+n \geq J$ we have

$$
\left\|\tilde{\mathcal{L}}_{j}^{n}(\cdot)-\tilde{m}_{j}(\cdot) \mathbf{1}\right\|_{B V} \leq C_{0} \delta_{0}^{n} .
$$

Proof. Recall that the functions $h_{j}$ in Theorem 2.4 satisfy $\eta:=\inf _{j} \operatorname{ess}-\inf h_{j}>0$. Thus by (2.4) we see that ess-inf $\mathcal{L}_{j}^{n} \mathbf{1} \geq \eta-C \delta^{n}$. Consequently, if we take $\delta_{0}=\frac{1}{2} \eta$ and $J \in \mathbb{N}$ such that $C \delta^{J} \leq \frac{1}{2} \eta$ we see that ess-inf $\mathcal{L}_{0}^{m} \mathbf{1} \geq \delta_{0}$ for all $m \geq J$. Next, a direct calculation shows that $\tilde{\mathcal{L}}_{j}^{n} g=\frac{\mathcal{L}_{j}^{n}\left(g \cdot \mathcal{L}_{0}^{j} \mathbf{1}\right)}{\mathcal{L}_{0}^{j+n} \mathbf{1}}$. Now the proposition follows from Theorem 2.4. noting that $\left\|\mathcal{L}_{0}^{m} \mathbf{1}-h_{m}\right\|_{B V}$ decays exponentially fast and using that, due to (V6), $\sup _{j}\left\|1 / h_{j}\right\|_{B V}<\infty$.

### 6.2. A martingale coboundary representation.

Lemma 6.3. For every sequence of functions $f_{j} \in B_{j}$ such that sup $\left\|f_{j}\right\|_{B V}<\infty$ there are functions $M_{j}=M_{j}(f) \in B_{j}$ and $u_{j}=u_{j}(f) \in B_{j}$ such that

$$
\begin{equation*}
\tilde{f}_{j}:=f_{j}-m_{0}\left(f_{j} \circ T_{0}^{j}\right)=M_{j}+u_{j+1} \circ T_{j}-u_{j}, \quad j \geq J \tag{6.2}
\end{equation*}
$$

Moreover, $\sup _{j}\left\|u_{j}\right\|_{B V}<\infty$ and $\left(M_{j} \circ T_{0}^{j}\right)_{j}$ is a reverse martingale difference with respect to the reverse filtration $\mathcal{A}_{j}=\left(T_{0}^{j}\right)^{-1} \mathcal{B}_{j}$ (on the probability space $\left(X_{0}, \mathcal{B}_{0}, m_{0}\right)$ ). Furthermore, $\sup \left\|u_{j}\right\|_{B V}$ is bounded above by a constant which depends only on the constants $C$ and $\delta$ from $\stackrel{j}{\text { Theorem }} 2.4$ and on $\sup _{j}\left\|f_{j}\right\|_{B V}$.
Remark 6.4. Note that by (V8) we also get that sup $\left\|M_{j}\right\|_{B V}<\infty$.
Proof of Lemma 6.3. Define $\tilde{\mathcal{L}}_{j}=0$ for $j<0$ and $\tilde{f}_{j}=0$ for $j<J$. Set

$$
\begin{equation*}
u_{j}=\sum_{k=1}^{\infty} \tilde{\mathcal{L}}_{j-k}^{k} \tilde{f}_{j-k}=\sum_{k=1}^{j} \tilde{\mathcal{L}}_{j-k}^{k} \tilde{f}_{j-k} \tag{6.3}
\end{equation*}
$$

Since $\tilde{m}_{s}\left(\tilde{f}_{s}\right)=0$ for all $s \geq 0$ by Proposition 6.2, $u_{j} \in B_{j}$ and $\sup _{j}\left\|u_{j}\right\|_{B V}<\infty$. Set $M_{j}=\tilde{f}_{j}+u_{j}-u_{j+1} \circ T_{j}$. It remains to show that $M_{j} \circ T_{0}^{j}$ is indeed a reverse martingale difference. To prove that we notice that $\mathbb{E}\left[M_{j} \circ T_{0}^{j} \mid \mathcal{A}_{j+1}\right]=\tilde{\mathcal{L}}_{j}\left(M_{j}\right) \circ T_{0}^{j+1}$. On the other hand, a direct calculation using (6.3) shows that $\tilde{\mathcal{L}}_{j}\left(M_{j}\right)=0$.
6.3. On the divergence of the variance. The first step in proving a central limit theorem is to show that the individual summands are negligible in comparison with the variance of the sum. In particular, we need to know when the variance is bounded. In this section we prove the following result.

Theorem 6.5. The following conditions are equivalent.
(1) $\liminf _{n \rightarrow \infty} \operatorname{Var}_{m_{0}}\left(S_{n}\right)<\infty$.
(2) $\sup _{n \in \mathbb{N}} \operatorname{Var}_{m_{0}}\left(S_{n}\right)<\infty$
(3) We can write $f_{j}=m_{0}\left(f_{j} \circ T_{0}^{j}\right)+M_{j}+u_{j+1} \circ T_{j}-u_{j}$ with $u_{j}, M_{j} \in B_{j}$ such that $M_{j} \circ T_{0}^{j}$ is a reverse martingale difference on $\left(X_{0}, \mathcal{B}_{0}, m_{0}\right)$ with respect to the reverse filtration $T_{0}^{-j} \mathcal{B}_{j}$, $\sup _{j}\left\|u_{j}\right\|_{B V}<\infty, \sup _{j}\left\|M_{j}\right\|_{B V}<\infty$, and $\sum_{j} \operatorname{Var}_{m_{0}}\left(M_{j} \circ T_{0}^{j}\right)<\infty$.

Proof. First, it is enough to prove the theorem when $m_{0}\left(f_{j} \circ T_{0}^{j}\right)=0$ for all $j$. In this case, by Lemma 6.3 we can write

$$
f_{j}=m_{0}\left(f_{j} \circ T_{0}^{j}\right)+M_{j}+u_{j+1} \circ T_{j}-u_{j}=M_{j}+u_{j+1} \circ T_{j}-u_{j}
$$

with $u_{j}$ and $M_{j}$ like in (3), except that in general the sum of the variances of $M_{j}$ might not converge. Notice now that

$$
\begin{equation*}
\left\|S_{n}-S_{n} M\right\|_{L^{2}\left(m_{0}\right)} \leq\left\|S_{n}-S_{n} M\right\|_{L^{\infty}\left(m_{0}\right)} \leq 2 \sup _{j}\left\|u_{j}\right\|_{L^{\infty}\left(m_{j}\right)}:=U<\infty \tag{6.4}
\end{equation*}
$$

where $S_{n} M=\sum_{j=0}^{n-1} M_{j} \circ T_{0}^{j}$.
Now assume (1), and let $n_{k}$ be an increasing sequence such that $n_{k} \rightarrow \infty$ and $\sigma_{n_{k}}=\left\|S_{n_{k}}\right\|_{L^{2}} \leq C$ for some constant $C>0$. Then by (6.4), $\left\|S_{n_{k}} M\right\|_{L^{2}\left(m_{0}\right)}^{2} \leq(C+U)^{2}<\infty$. However, since $M_{j} \circ T_{0}^{j}$ is a reverse martingale, we have

$$
\sum_{j=0}^{n_{k}-1} \operatorname{Var}_{m_{0}}\left(M_{j} \circ T_{0}^{j}\right)=\left\|S_{n_{k}} M\right\|_{L^{2}\left(m_{0}\right)}^{2} \leq(C+U)^{2}
$$

Now, since $V_{n}:=\left\|S_{n} M\right\|_{L^{2}}^{2}=\sum_{j=0}^{n-1} \operatorname{Var}_{m_{0}}\left(M_{j} \circ T_{0}^{j}\right)$ is increasing we conclude that the summability condition in (3) holds. This shows that (1) implies (3).

Next, (2) clearly implies (1). Thus, to complete the proof it is enough to show that (3) implies (2), but this also follows from (6.4) since the latter yields $\left\|S_{n}\right\|_{L^{2}\left(m_{0}\right)}^{2} \leq\left(V_{n}+U\right)^{2}<\infty$.
6.4. Quadratic variation and moment estimates. Recall that the (unconditioned) quadratic variation difference of the reverse martingale difference $M_{j} \circ T_{0}^{j}$ is given by $q_{j}(M):=$ $M_{j}^{2} \circ T_{0}^{j}$. Henceforth we denote $Q_{j}=M_{j}^{2}$ and let

$$
S_{j, n}=S_{j, n} f=\sum_{k=j}^{j+n-1} f_{k} \circ T_{j}^{k}, \quad \bar{S}_{j, n}=S_{j, n}-\tilde{m}_{j}\left(S_{j, n}\right)=S_{j, n}-m_{0}\left(S_{j, n} \circ T_{0}^{j}\right)
$$

$S_{j, n} M$ and $S_{j, n} Q$ are defined similarly.
Proposition 6.6. There is a constant $C$ which depends only on the constants from Theorem 2.4 and on $\sup \left\|f_{j}\right\|_{B V}$ such that for all $j \geq J$ (where $J$ comes from Proposition 6.2) we have $\nabla^{\operatorname{Var}_{m_{0}}}\left(S_{j, n} Q\right)^{j} \leq C\left(1+\operatorname{Var}\left(S_{j, n} f\right)\right)$.
Proof. First, to simplify the notation let us assume that $j=J=0$. Denote

$$
g_{j}=Q_{j}-\tilde{m}_{j}\left(Q_{j}\right)=Q_{j}-\tilde{m}_{0}\left(Q_{j} \circ T_{0}^{j}\right)
$$

and $S_{n} g=\sum_{j=0}^{n-1} g_{j} \circ T_{0}^{j}$. The argument below is similar to the first part of the proof of [24, Theorem 4.1], but we provide the details to make our paper self contained. By Proposition 6.2 (using that $\tilde{m}_{0}=m_{0}$ ) we have

$$
\begin{aligned}
& \mathbb{E}_{m_{0}}\left[\left(S_{n} g\right)^{2}\right] \leq 2 \sum_{0 \leq \ell<n} \sum_{0 \leq k \leq \ell}\left|\tilde{m}_{0}\left(\left(g_{k} \circ T_{0}^{k}\right) \cdot\left(g_{\ell} \circ T_{0}^{\ell}\right)\right)\right|=2 \sum_{0 \leq \ell<n} \sum_{0 \leq k \leq \ell}\left|\tilde{m}_{k}\left(g_{k} \cdot \tilde{\mathcal{L}}_{k}^{\ell-k} g_{\ell}\right)\right| \\
& \leq C_{0} \sum_{0 \leq \ell<n} \sum_{0 \leq k \leq \ell} \tilde{m}_{k}\left(\left|g_{k}\right|\right)\left\|g_{\ell}\right\|_{B V} \delta_{0}^{\ell-k}=C_{0} \sum_{0 \leq k \leq n} \tilde{m}_{k}\left(\left|g_{k}\right|\right)\left(\sum_{k \leq \ell<n}\left\|g_{\ell}\right\|_{B V} \delta_{0}^{\ell-k}\right) \\
& \quad \leq c_{0} \sum_{0 \leq k \leq n} \tilde{m}_{k}\left(\left|g_{k}\right|\right)=c_{0} \sum_{0 \leq k \leq n} m_{0}\left(\left|g_{k} \circ T_{0}^{k}\right|\right) \leq 2 c_{0} \sum_{0 \leq k \leq n} m_{0}\left(Q_{k} \circ T_{0}^{k}\right)
\end{aligned}
$$

for some constant $c_{0}$ (the first inequality of the last line uses that sup $\left\|g_{j}\right\|_{B V}<\infty$ ). Here $\tilde{\mathcal{L}}_{j}$ are the transfer operators defined by (6.1). Observe that $m_{0}\left(Q_{k} \circ T_{0}^{k}\right)=m_{0}\left(\left(M_{k} \circ T_{0}^{k}\right)^{2}\right)$ and, because
of the orthogonality property, $\sum_{0 \leq k<n} m_{0}\left(\left(M_{k} \circ T_{0}^{k}\right)^{2}\right)=\operatorname{Var}_{m_{0}}\left(S_{n} M\right)$. Now the result follows from (6.4).

Proof of Proposition 3.3. It is enough to prove the theorem for $j \geq J$, since to get the result for $0 \leq j<J$ we can just take $C$ large enough.

To simply the notation, we will only prove the theorem when $j=J=0$, the proof when $j \geq J>0$ is similar.

Notice that it is enough to prove the claim for $p=2^{m}$ for all $m$. Moreover, by replacing $f_{j}$ with $f_{j}-\mu_{0}\left(f_{j} \circ T_{0}^{j}\right)$ we can and will assume that $\mu_{0}\left(S_{n}\right)=0$ for all $n$.

We use induction on $m$. For $m=1$ the result is trivial. Suppose that the statement is true for some $m \geq 1$. In order to estimate $\left\|S_{n} f\right\|_{2^{m+1}}$ we first use that ${ }^{2}$

$$
\left\|S_{n} f\right\|_{2^{m+1}} \leq C+\left\|S_{n} M\right\|_{2^{m+1}}
$$

for some constant $C$ which depends only on the constants $C$ and $\delta$ from Theorem 2.4 and on $\|f\|:=\sup _{j}\left\|f_{j}\right\|_{B V}$ since $\left\|S_{n} f-S_{n} M\right\|_{L^{\infty}}$ is bounded in $n$. So it suffices to show that

$$
\begin{equation*}
\left\|S_{n} M\right\|_{2^{m+1}} \leq C\left(1+\left\|S_{n} f\right\|_{2}\right) \tag{6.5}
\end{equation*}
$$

for an appropriate constant $C$.
(6.5) follows from a version of Burkholder's inequality for martingales (see [69, Theorem 2.12]). Let $\mathfrak{d}_{1}, \ldots, \mathfrak{d}_{n}$ be a martingale difference with respect to a filtration $\left(\mathcal{F}_{j}\right)_{j=1}^{n}$ on a probability space. Let $D_{n}=\mathfrak{d}_{1}+\mathfrak{d}_{2}+\ldots+\mathfrak{d}_{n}$ and $E_{n}=\mathfrak{d}_{1}^{2}+\mathfrak{d}_{2}^{2}+\ldots+\mathfrak{d}_{n}^{2}$. Then, for every $p \geq 2$ there are constants $c_{p}, C_{p}>0$ depending only on $p$ such that

$$
\begin{equation*}
c_{p}\left\|E_{n}\right\|_{p / 2}^{1 / 2} \leq\left\|D_{n}\right\|_{p} \leq C_{p}\left\|E_{n}\right\|_{p / 2}^{1 / 2} . \tag{6.6}
\end{equation*}
$$

Applying (6.6) with the (reverse) martingale difference $M_{j} \circ T_{0}^{j}$ we see that

$$
\begin{equation*}
\left\|S_{n} M\right\|_{2^{m+1}} \leq a_{m}\left\|S_{n} Q\right\|_{2^{m}}^{1 / 2} \tag{6.7}
\end{equation*}
$$

where $S_{n} Q=S_{0, n} Q$ and $a_{m}$ depends only on $m$. Applying the induction hypothesis with the sequence of functions $\tilde{Q}_{j}=Q_{j}-m_{0}\left(Q_{j} \circ T_{0}^{j}\right)$ we see that there is a constant $R_{m}>0$ depending only of $m$ and the constants in the formulation of Proposition 3.3 such that

$$
\left\|S_{n} \tilde{Q}\right\|_{2^{m}} \leq R_{m}\left(1+\left\|S_{n} \tilde{Q}\right\|_{2}\right)
$$

Since $\mathbb{E}\left[S_{n} Q\right]=\operatorname{Var}\left(S_{n} M\right)$, Proposition 6.6 gives

$$
\begin{gathered}
\left\|S_{n} Q\right\|_{2^{m}} \leq\left\|S_{n} \tilde{Q}\right\|_{2^{m}}+\mathbb{E}\left[S_{n} Q\right] \leq R_{m}\left(1+C\left(1+\operatorname{Var}\left(S_{n} f\right)\right)\right)+\operatorname{Var}\left(S_{n} M\right) \\
\leq R_{m}^{\prime}\left(1+\operatorname{Var}\left(S_{n} f\right)\right)+\operatorname{Var}\left(S_{n} M\right)
\end{gathered}
$$

for some other constant $R_{m}^{\prime}$. Using that sup $\left\|S_{n} f-S_{n} M\right\|_{L^{\infty}}<\infty$ we see that there is a constant $C>0$ such that $\operatorname{Var}\left(S_{n} M\right) \leq C\left(1+\operatorname{Var}\left(S_{n}^{n} f\right)\right)$. Thus, there is a constant $R_{m}^{\prime \prime}>0$ such that

$$
\left\|S_{n} \tilde{Q}\right\|_{2^{m}} \leq R_{m}^{\prime \prime}\left(1+\operatorname{Var}\left(S_{n} f\right)\right)
$$

Now (6.5) follows from (6.7), completing the proof of the proposition.

$$
{ }^{2} \text { Where } S_{n} M=\sum_{j=0}^{n-1} M_{j} \circ T_{0}^{j}, M_{j}=M_{j}(f) .
$$

## 7. Logarithmic growth conditions

In this section we prove the following result.
Proposition 7.1. Let $f_{j}: X_{j} \rightarrow \mathbb{R}$ be a sequence of functions such that sup $\left\|f_{j}\right\|_{B V}<\infty$ and $\tilde{m}_{j}\left(f_{j}\right)=0$. Suppose $\sigma_{n}^{2}=\operatorname{Var}_{m_{0}}\left(S_{n}\right) \rightarrow \infty$. Let $\Lambda_{n}(t)=\ln m_{0}\left(e^{i t S_{n} f / \sigma_{n}}\right)$. Then for every $k \geq 3$ there are constants $\delta_{k}>0$ and $C_{k}>0$ such that

$$
\sup _{t \in\left[-\delta_{k} \sigma_{n}, \delta_{k} \sigma_{n}\right]}\left|\Lambda_{n}^{(k)}(t)\right| \leq C_{k} \sigma_{n}^{-(k-2)}
$$

where $\Lambda_{n}^{(k)}$ is the $k$-th derivative of the function $\Lambda_{n}$.
The proof uses the complex Ruelle-Perron-Frobinuous theorem, see Theorem 7.3 below.
7.1. Sequential complex RPF theorem. Take a sequence of real valued functions $f_{j} \in B_{j}$ and $\|f\|=\sup _{j}\left\|f_{j}\right\|_{B V}<\infty$. Consider the operators $\mathcal{L}_{j, z}(h)=\mathcal{L}_{j}\left(h e^{z f_{j}}\right)$ (where $z \in \mathbb{C}$ ). We need the following result.

Lemma 7.2. (i) $\sup _{j}\left\|\mathcal{L}_{j, z}\right\|_{B V} \leq C(z)$ for some continuous function $C(z)$ (of exponential order in $|z|)$.
(ii) The map $z \rightarrow \mathcal{L}_{j, z}$ is analytic and its $k$-order derivatives $\mathcal{L}_{j, t, k}$ satisfy $\sup _{j}\left\|\mathcal{L}_{j, t, k}\right\|_{B V} \leq C_{k}(z)$ for some continuous function $C_{k}(\cdot)$ (of exponential order in $|z|$ ).

Proof. Since $\left\|e^{z f_{j}}\right\|_{B V} \leq \sum_{k} \frac{|z|^{k}\left\|f_{j}^{k}\right\|_{B V}}{k!}$ and by (V5), $\left\|f_{j}^{k}\right\|_{B V} \leq C^{k}\left\|f_{j}\right\|_{B V}^{k}$ for some constant $C$, we get $\sup \left\|e^{z f_{j}}\right\|_{B V} \leq e^{c|z|}$ for some constant $c>0$.

The analyticity of the operators follows from the analytiticty of the map $z \rightarrow e^{z f_{j}}$, and the estimate on the norms of the derivatives is obtained similarly to part (i).

The following result is a consequence of the more general perturbation theorem (Theorem D. 2 in Appendix D.

Theorem 7.3. There is a constant $r_{0}>0$ such that for every complex number with $|z| \leq r_{0}$ there are sequences $\lambda(z)=\left(\lambda_{j}(z)\right), \lambda_{j}(z) \in \mathbb{C} \backslash\{0\}, h^{(z)}=\left(h_{j}^{(z)}\right), h_{j}^{(z)} \in B_{j}, m^{(z)}=\left(m_{j}^{(z)}\right), m_{j}^{(z)} \in B_{j}^{*}$ with the following properties.
(i) The maps $z \rightarrow \lambda_{j}(z), z \rightarrow h_{j}^{(z)}$ and $z \rightarrow m_{j}^{(z)}$ are analytic in $z$ and their norms $3^{3}$ are bounded uniformly in $j$.
(ii) We have $\lambda_{j}(0)=1, m_{j}^{(0)}=m_{j}, h_{j}^{(0)}=h_{j}, m_{j}^{(z)}(\mathbf{1})=m_{j}^{(z)}\left(h_{j}^{(z)}\right)=1$ and

$$
\mathcal{L}_{j, z} h_{j}^{(z)}=\lambda_{j}(z) h_{j+1}^{(z)}, \quad\left(\mathcal{L}_{j, z}\right)^{*} m_{j+1}^{(z)}=\lambda_{j}(z) m_{j}^{(z)} .
$$

(iii) There are $\delta_{3} \in(0,1)$ and $C_{3}>0$ such that

$$
\left\|\mathcal{L}_{j, z}^{n}-\lambda_{j, n}(z) m_{j}^{(z)}(\cdot) h_{j+n}^{(z)}\right\| \leq C_{3} \delta_{3}^{n}
$$

where $\mathcal{L}_{j, z}^{n}=\mathcal{L}_{j+n-1, z} \circ \cdots \circ \mathcal{L}_{j+1, z} \circ \mathcal{L}_{j, z}$ and $\lambda_{j, n}(z)=\lambda_{j+n-1}(z) \cdots \lambda_{j+1}(z) \lambda_{j}(z)$.

[^1]7.2. Proof of Proposition 7.1. Let $\lambda_{j}(z)$ be like in Theorem 7.3, and let $\Pi_{j}(z)$ be an analytic branch of $\ln \lambda_{j}(z)$ so that $\Pi_{j}(0)=1$.

Lemma 7.4. There are $\varepsilon_{0}>0$ and $A_{0}>0$ such that for every complex number $z$ with $|z| \leq \varepsilon_{0}$ and all $j, n$ with $j+n \geq J$ we have

$$
\left|\ln \mathbb{E}_{m_{0}}\left[e^{z\left(S_{n+j} f-S_{j} f\right)}\right]-\sum_{k=j}^{j+n-1} \Pi_{j}(z)\right| \leq A_{0}
$$

Proof. Since $m_{0}=\tilde{m}_{0}$ we have $\mathbb{E}_{m_{0}}\left[e^{z\left(S_{n+j} f-S_{j} f\right)}\right]=\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, z}^{n} \mathbf{1}\right)$ where

$$
\begin{equation*}
\tilde{\mathcal{L}}_{j, z}^{n}(g)=\tilde{\mathcal{L}}_{j}^{n}\left(e^{z S_{j, n} f} g\right)=\frac{\mathcal{L}_{j, z}^{n}\left(g \mathcal{L}_{0}^{j} \mathbf{1}\right)}{\mathcal{L}_{0}^{j+n} \mathbf{1}} \tag{7.1}
\end{equation*}
$$

Using Theorem 7.3 (iii), and the facts that $m_{j}\left(\mathcal{L}_{0}^{j} \mathbf{1}\right)=1, \operatorname{essinf}\left(\mathcal{L}_{0}^{j+n} \mathbf{1}\right) \geq c_{0}>0$ and $\sup _{k}\left\|\mathcal{L}_{0}^{k}\right\|_{B V}<\infty$ we get

$$
\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, z}^{n} \mathbf{1}\right)=\lambda_{j, n}(z) U_{j, n}(z)+O\left(\delta_{3}^{n}\right)
$$

where $\delta_{3} \in(0,1)$ and $U_{j, n}(z)=m_{j}^{(z)}\left(\mathcal{L}_{0}^{j} \mathbf{1}\right) \tilde{m}_{j+n}\left(H_{j+n, z}\right), H_{j+n, z}=\frac{h_{j+n}^{(z)}}{\mathcal{L}_{0}^{j+n} \mathbf{1}}$. Next the function $G(z)=\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, z}^{n} \mathbf{1}\right)-\lambda_{j, n}(z) U_{j, n}(z)$ is analytic. Since $\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, 0}^{n} \mathbf{1}\right)=\lambda_{j, n}(0) U_{j, n}(0)=1$ we have $G(0)=0$. Applying the maximal modulus principle to the function $\frac{G(z)}{z}$ we get

$$
\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, z}^{n} \mathbf{1}\right)=\lambda_{j, n}(z) U_{j, n}(z)+z O\left(\delta_{3}^{n}\right)
$$

namely, the error term is $O\left(z \delta_{3}^{n}\right)$ and not only $O\left(\delta_{3}^{n}\right)$. Next

$$
\tilde{m}_{j+n}\left(H_{j+n, z}\right)=m_{j+n}\left(H_{j+n, z} \mathcal{L}_{0}^{j+n} \mathbf{1}\right)=m_{j+n}\left(h_{j+n}^{(z)}\right)
$$

Since $\lambda_{j}(0)=1$ and $\lambda_{j}(z)$ are uniformly bounded and analytic in $z$, we can write the above in the following form

$$
\tilde{m}_{j+n}\left(\tilde{\mathcal{L}}_{j, z}^{n} \mathbf{1}\right)=\lambda_{j, n}(z)\left(U_{j, n}(z)+z O\left(\delta_{4}^{n}\right)\right), \quad \delta_{4} \in(0,1)
$$

The functions $U_{j, n}(z)$ are analytic, uniformly bounded in $z$ and they take the value one at $z=0$. Thus, we can take the logarithm of both sides to conclude that, if $|z|$ is small enough then

$$
\ln \mathbb{E}_{m_{0}}\left[e^{z\left(S_{n+j} f-S_{j} f\right)}\right]=\ln \lambda_{j, n}(z)+V_{j, n}(z)
$$

where $V_{j, n}(z)=\ln U_{j, n}(z)=\ln \left(1+O(z)+z O\left(\delta_{4}^{n}\right)\right)=O(1)$.
Using the Cauchy integral formula we derive the following result.
Corollary 7.5. Let $\tilde{\Lambda}_{j, n}(z)=\ln \mathbb{E}_{m_{0}}\left[e^{z\left(S_{j+n} f-S_{j} f\right)}\right]$. Then there exists $\varepsilon>0$ such that for every $s \geq 1$ there is a constant $C_{s}>0$ such that for every complex $z$ with $|z| \leq \varepsilon$ and all $j, n$ with $j+n \geq J$ we have

$$
\begin{equation*}
\left|\tilde{\Lambda}_{j, n}^{(s)}(z)-\sum_{k=j}^{j+n-1} \Pi_{k}^{(s)}(z)\right| \leq C_{s} \tag{7.2}
\end{equation*}
$$

where $g^{(s)}$ denotes the $s$-th derivative of a function $g$.
Remark 7.6. Clearly, 7.2 also holds when $j+n<J$ since then the number of summands is uniformly bounded.

Set

$$
\begin{equation*}
\Pi_{j, n}(z)=\sum_{k=j}^{j+n-1} \Pi_{k}(z) \tag{7.3}
\end{equation*}
$$

Lemma 7.7. Let $B$ be a constant and let $k \geq 2$. Then if $B$ is sufficiently large there are constants $D$ and $r_{0}$ depending only on $B$ and $k$ so that for every $t \in\left[-r_{0}, r_{0}\right]$ and each $j, n$ such that $B \leq \operatorname{Var}\left(S_{j+n} f-S_{j} f\right) \leq 2 B$ we have

$$
\left|\Pi_{j, n}^{(k)}(i t)\right| \leq D
$$

Proof. Applying [27, Lemma 43] with $S=S_{j+n} f-S_{j} f$ we see that there is $r=r(B)$ such that if $t \in[-r, r]$ then

$$
\left|\tilde{\Lambda}_{j, n}^{(k)}(i t)\right| \leq D_{k} \mathbb{E}_{m_{0}}\left[|S|^{k}\right]
$$

for some constant $D_{k}$ which depends only on $k$. Using also Corollary 7.5 we derive that

$$
\left|\Pi_{j, n}^{(k)}(i t)\right| \leq D_{k} \mathbb{E}_{m_{0}}\left[|S|^{k}\right]
$$

Next, by Proposition 3.3 we have $\mathbb{E}_{m_{0}}\left[|S|^{k}\right] \leq C(k, B)$. These estimates prove the lemma.
Proof of Proposition 7.1. Fix some $k \geq 3$. Since $\sigma_{n}=\left\|S_{n}\right\|_{L^{2}\left(m_{0}\right)} \rightarrow \infty$, using the martingale coboundary representation from Lemma 6.3, given $B>0$ large enough we can decompose $\{0, \ldots, n-1\}$ into a disjoint union of intervals $I_{1}, \ldots, I_{k_{n}}$ in $\mathbb{Z}$ so that $I_{j}$ is to the left of $I_{j+1}$ and

$$
\begin{equation*}
B \leq \operatorname{Var}_{m_{0}}\left(S_{I_{j}}\right) \leq 2 B \tag{7.4}
\end{equation*}
$$

where $S_{I}=\sum_{j \in I} f_{j} \circ T_{0}^{j}$ for every interval $I$. Now, by Lemma 6.3 there is a constant $t^{4} A>0$ independent of $B$ such that $\left|\left\|S_{n} f\right\|_{L^{2}}-\left(\sum_{k=J}^{n-1} \operatorname{var}_{m_{0}}\left(M_{k} \circ T_{0}^{k}\right)\right)^{1 / 2}\right| \leq A$ and for each $j>J$ we have $\left|\left\|S_{I_{j}}\right\|_{L^{2}}-\left(\sum_{k \in I_{j}} \operatorname{var}_{m_{0}}\left(M_{k} \circ T_{0}^{k}\right)\right)^{1 / 2}\right| \leq A$.

Hence, if we also assume that $B>(4 A)^{2}$ then it follows that

$$
\begin{equation*}
C_{1} \leq k_{n} / \sigma_{n}^{2} \leq C_{2} \tag{7.5}
\end{equation*}
$$

for some constants $C_{1}, C_{2}>0$ which depend only on $B$. Next, let $\Pi_{I}(z)=\sum_{k \in I} \Pi_{k}(z)$. Then by Lemma 7.7 there are constants $r_{k}>0$ and $D_{k}$ such that $\sup _{j} \sup _{t \in\left[-r_{k}, r_{k}\right]}\left|\Pi_{I_{j}}^{(k)}(i t)\right| \leq D_{k}$. Hence,

$$
\begin{equation*}
\sup _{t \in\left[-r_{k}, r_{k}\right]}\left|\Pi_{0, n}^{(k)}(i t)\right| \leq D_{k} k_{n} \leq C_{2} D_{k} \sigma_{n}^{2} \tag{7.6}
\end{equation*}
$$

Combining this with Corollary 7.5 and taking into account that $\sigma_{n} \rightarrow \infty$ we see that

$$
\sup _{t \in\left[-r_{k}, r_{k}\right]}\left|\tilde{\Lambda}_{n}^{(k)}(i t)\right| \leq \tilde{D}_{k} \sigma_{n}^{2}
$$

for some constant $\tilde{D}_{k}$, and the proof of the proposition is complete.

[^2]7.3. Proof of Theorems 3.1 and 3.2. Let $\Lambda_{n}(t)=m_{0}\left(e^{i t S_{n} / \sigma_{n}}\right)$. Applying Theorems 5 and 9, and Corollary 11 from [52] to $W_{n}=S_{n} / \sigma_{n}$, we see that in order to prove Theorems 3.1 and 3.2 it is enough to show that for every $u \geq 3$ there are constants $\delta_{u}, C_{u}>0$ such that the function $\Lambda_{n}(t)$ is well defined, differentiable $u$ times on $\left[-\delta_{u} \sigma_{n}, \delta_{u} \sigma_{n}\right]$ and $\sup _{|t| \leq \sigma_{n} \delta_{u}}\left|\Lambda_{n}^{(u)}(t)\right| \leq C_{u} \sigma_{n}^{-(u-2)}$. However, this is exactly Proposition 7.1 .

## Appendix A. Proof of Theorem 2.4

A.1. Proof of Theorem $2.4(\mathbf{i})$. Let $\varepsilon_{0}>0$ be such that $\eta:=\rho+\varepsilon_{0}<1$ and let $a>0$ be a constant such that $a \varepsilon_{0}>K_{0}$, where $\rho$ and $K_{0}$ come from (2.3). Then it follows from (2.3) that for every $n \geq N$ and all $j \geq 0$ we have

$$
\mathcal{L}_{j}^{n} \mathcal{C}_{j, a} \subset \mathcal{C}_{j+n, a \rho^{1+N-n}+K_{0}} \subset \mathcal{C}_{j+n, \eta a}
$$

Given $C>0$ consider the cone

$$
\mathcal{K}_{j}(C)=\left\{g \in B_{j}, \quad g>0, \quad \text { esssup } g \leq C \operatorname{essinf} g\right\}
$$

where the essential supremum and infimum are with respect to $m_{j}$. Denote by $\mathrm{d}_{j}$ the projective Hilbert metric associated with the cone $\mathcal{C}_{j, a}$ (see [16, 66]). Then by [19, (1.1)] the projective diameter of $\mathcal{K}_{j}(C) \cap \mathcal{C}_{j, \eta a}$ inside $\mathcal{C}_{j, a}$ does not exceed a constant $R=R(a, \eta, V, C)$ which depends only on $a, \eta, C$ and $V:=\sup v_{j}(\mathbf{1})$. Next, combining (V3), (SC), (2.3) and the relation $\left(\mathcal{L}_{j}^{n}\right)^{*} m_{j+n}=m_{j}$, we see that there are constants $M \geq N$ and $C$ (depending on $a$ but not on $j$ ) such that for every $j$ and $n \geq M$ we have

$$
\mathcal{L}_{j}^{n} \mathcal{C}_{j, a} \subset \mathcal{K}_{j+n}(C)
$$

We conclude that for every $j$ and $n \geq M$ the projective diameter of $\mathcal{L}_{j}^{n} \mathcal{C}_{j, a}$ inside $\mathcal{C}_{j+n, a}$ does not exceed $R$.

The next step of the proof is the following result.
Lemma A.1. (a) There is a constant $A>0$ such that for every $a, j$ and $h \in \mathcal{C}_{j, a}$,

$$
\|h\| \leq A m_{j}(h)
$$

(b) If $a>V$ then there is a constant $r_{0}>0$ such that every real valued $g \in B_{j}$ can be written in the form $g=g_{1}-g_{2}$ where $g_{i} \in \mathcal{C}_{j, a}$ and $\left\|g_{1}\right\|+\left\|g_{2}\right\| \leq r_{0}\|g\|$.

Proof. To prove (a) we note that if $h \in \mathcal{C}_{j, a}$ then $\|h\|=m_{j}(h)+v_{j}(h) \leq(1+a) m_{j}(h)$ so we can take $A=1+a$.

In order to prove (b), given $g \in B_{j}$ we take $C_{0}=C_{0}(g)=\|g\|_{\infty}+\frac{1+a}{a-V}\|g\|_{B V}$. It is immediate to check that $g+C_{0} \in \mathcal{C}_{j, a}$. Since $C_{0} \leq C\|g\|_{B V}$ for some constant $C$ which does not depend on $j$ we have $g=\left(g+C_{0}\right)-C_{0}$ and, with $\|\cdot\|=\|\cdot\|_{B V}$,

$$
\left\|g+C_{0}\right\|+\left\|C_{0}\right\| \leq\|g\|+2\left\|C_{0}\right\| \leq\|g\|\left(1+2 C_{0}+2 C_{0} V\right) \leq r\|g\|
$$

for some constant $r=r(a, V, C)$. Thus we can take $g_{1}=\left(g+C_{0}\right)$ and $g_{2}=C_{0}$ (the constant function). Note that $g_{2} \in \mathcal{C}_{j, a}$ since $a>V$.

Based on Lemma A.1, the proof of proof of Theorem 2.4(ii) is completed like in [46, Chapters 4 and 5]. In fact, here the situation is much simpler, and for readers' convenience we provide the details. First, by taking $f=1$ in 2.2 and letting $g$ vary we see that $\left(\mathcal{L}_{j}^{n}\right)^{*} m_{j+n}=m_{j}$ for all $j$ and $n$. Next, fix some $a$ large enough and take two sequences of functions $g_{n}, f_{n} \in \mathcal{C}_{n, a}$. Denote $G_{j, n}=\mathcal{L}_{j}^{n} g_{j}$ and $F_{j, n}=\mathcal{L}_{j}^{n} f_{j}$. Then, since the projective diameter of the image of $\mathcal{L}_{k}^{s} \mathcal{C}_{a, k}$ inside
$\mathcal{C}_{a, k+s}$ does not exceed $R$ if $s \geq M$, we see by [66, Theorem 1.1] that for every $j$ and $n \geq M$ we have

$$
\mathrm{d}_{j}\left(F_{j, n}, G_{j, n}\right) \leq(c(R))^{[(n-1) / M]}
$$

where $c(R)=\tanh (R / 4) \in(0,1)$.
For a positive function $g$, define the normalized iterates by

$$
\overline{\mathcal{L}}_{j}^{n} g=\frac{\mathcal{L}_{j}^{n} g}{m_{j+n}\left(\mathcal{L}_{j}^{n} g\right)}=\frac{\mathcal{L}_{j}^{n} g}{m_{j}(g)} .
$$

Then by [46, Theorem A.2.3] and Lemma A.1(1) we have

$$
\begin{equation*}
\left\|\overline{\mathcal{L}}_{j}^{n} f_{j}-\overline{\mathcal{L}}_{j}^{n} g_{j}\right\|_{B V} \leq \frac{1}{2} A(c(R))^{[(n-1) / M]} . \tag{A.1}
\end{equation*}
$$

Next, fix some function $h_{0} \in \mathcal{C}_{a, 0}$ such that $m_{0}\left(h_{0}\right)=1$ and for $j>0$ define $h_{j}=\mathcal{L}_{0}^{j} h_{j}$. Then $h_{j+1}=\mathcal{L}_{j} h_{j}$ and $m_{j}\left(h_{j}\right)=1$ for all $j \geq 0$. Therefore,

$$
\begin{equation*}
\overline{\mathcal{L}}_{j}^{n} h_{j}=h_{j+n} \tag{A.2}
\end{equation*}
$$

for all $j \geq 0$ and $n>0$. Moreover, $h_{j} \in \mathcal{C}_{a, j}$ (due to the equivariance of the cones). By taking $g_{j}=h_{j}$ in A.1) and using A.2 we see that

$$
\left\|\overline{\mathcal{L}}_{j}^{n} f_{j}-h_{j+n}\right\|_{B V} \leq \frac{1}{2} A(c(R))^{[(n-1) / M]} .
$$

Multiplying by $m_{j}\left(f_{j}\right)$ we get that, with $\delta=(c(R))^{1 / M}$, for all $f \in \mathcal{C}_{a, j}$ we have

$$
\left\|\mathcal{L}_{j}^{n}(f)-m_{j}(f) h_{j+n}\right\| \leq m_{j}(f) B \delta^{n}
$$

where $B$ is some constant. Using Lemma A.1(b) we obtain (2.4) with $C=2 r_{0} B$.
A.2. Proof of Theorem 2.4 (ii). To see why $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$, take a measurable function $g: X_{j} \rightarrow \mathbb{R}$ and write

$$
\left[\left(T_{j}\right)_{*} \mu_{j}\right](g)=\mu_{j}\left(g \circ T_{j}\right)=m_{j}\left(h_{j} g \circ T_{j}\right)=m_{j+1}\left(\left(\mathcal{L}_{j} h_{j}\right) \cdot g\right)=m_{j+1}\left(h_{j+1} g\right)=\mu_{j+1}(g)
$$

To prove asymptotic uniqueness of $\mu_{j}$, let $\tilde{\mu}_{j}=g_{j} d m_{j}$ be another sequence of probability measures such that $\left(T_{j}\right)_{*} \tilde{\mu}_{j}=\tilde{\mu}_{j+1}$. We show that $\lim _{j \rightarrow \infty}\left\|h_{j}-g_{j}\right\|_{1}=0$. We claim that

$$
\begin{equation*}
\mathcal{L}_{j} g_{j}=g_{j+1}, \quad m_{j+1}-\text { a.e. } \tag{A.3}
\end{equation*}
$$

Indeed given a function $f: X_{j+1} \rightarrow \mathbb{R}$ we have

$$
m_{j+1}\left(\left(\mathcal{L}_{j} g_{j}\right) f\right)=m_{j}\left(g_{j}\left(f \circ T_{j}\right)\right)=\tilde{\mu}_{j}\left(f \circ T_{j}\right)=\tilde{\mu}_{j+1}(f)=m_{j+1}\left(g_{j+1} f\right)
$$

proving A.3). Iterating A.3) we see that $g_{n}=\mathcal{L}_{0}^{n} g_{0}, m_{n}$-a.e. Next, let $\varepsilon>0$ and let $q_{0}: X_{0} \rightarrow \mathbb{R}$ be a BV function such that $\left\|q_{0}-g_{0}\right\|_{1}<\varepsilon$. Normalizing $q_{0}$ if needed we can always assume that $m_{0}\left(q_{0}\right)=1$. By Theorem 2.4(i) we see that

$$
\left\|\mathcal{L}_{0}^{n} q_{0}-h_{n}\right\|_{B V} \leq C\left\|q_{0}\right\|_{B V} \delta^{n} \rightarrow 0, \text { as } n \rightarrow \infty
$$

On the other hand,

$$
\left\|\mathcal{L}_{0}^{n} q_{0}-g_{n}\right\|_{1}=\left\|\mathcal{L}_{0}^{n} q_{0}-\mathcal{L}_{0}^{n} g_{0}\right\|_{1} \leq\left\|\mathcal{L}_{0}^{n}\left(\left|q_{0}-g_{0}\right|\right)\right\|_{1}=\left\|g_{0}-d_{0}\right\|<\varepsilon
$$

Therefore, by taking $n$ large enough we get that $\left\|h_{n}-g_{n}\right\|_{1}<\varepsilon$, and the proof of Theorem 2.4(ii) is complete.

## Appendix B. Gibbs measures for SFT.

In this section we denote by $T_{j}: X_{j} \rightarrow X_{j+1}$, a one sided topologically mixing sequential subshift of finite type and by $\tilde{T}_{j}: \tilde{X}_{j} \rightarrow \tilde{X}_{j+1}, j \in \mathbb{Z}$ the corresponding two sided shift. For a point $\left(x_{j}, x_{j+1}, \ldots\right) \in X_{j}$ and $m \in \mathbb{N}$ let

$$
\left[x_{j}, x_{j+1}, \ldots, x_{j+m-1}\right]=\left\{\left(x_{j+k}^{\prime}\right)_{k \geq 0} \in X_{j}: x_{j+k}^{\prime}=x_{j+k}, \forall k<m\right\}
$$

be the corresponding cylinder of length $m$. Cylinders in $\tilde{X}_{j}$ are denoted similarly.
Remark B.1. The reason we consider here negative indexes $j$ is to provide a complete theory which includes uniqueness of Gibbs measures (see Section B.2) and relations with two sided shift via a non-stationary version of Sinai's lemma (see Lemma B.2). If we begin with a one sided shift then there are many Gibbs measures corresponding to a given sequence of potentials, each of which corresponds to an extension of the shift to negative indexes and an extension of the sequence of potentials to a two sided sequence. Considering negative indexes $j$ also allows us to uniquely define two sided shifts which have applications to non-autonomous hyperbolic systems (see Appendix C).
B.1. A sequential Sinai's Lemma. Let $\pi_{j}: \tilde{X}_{j} \rightarrow X_{j}$ be given by

$$
\pi_{j}\left(\left(x_{j+k}\right)_{k \in \mathbb{Z}}\right)=\left(x_{j+k}\right)_{k \geq 0} .
$$

Lemma B.2. Fix some $\alpha \in(0,1]$ and let $\psi_{j}: \tilde{X}_{j} \rightarrow \mathbb{R}$ be uniformly Hölder continuous with exponent $\alpha$. Then there are uniformly Hölder continuous functions $u_{j}: \tilde{X}_{j} \rightarrow \mathbb{R}$ with exponent $\alpha / 2$ and $\phi_{j}: X_{j} \rightarrow \mathbb{R}$ such that

$$
\psi_{j}=u_{j}-u_{j+1} \circ \tilde{T}_{j}+\phi_{j} \circ \pi_{j} .
$$

Moreover if $\left\|\psi_{j}\right\|_{\alpha} \rightarrow 0$ then $\left\|u_{j}\right\|_{\alpha / 2} \rightarrow 0$.
Proof. The proof is a modification of the proof of [15, Lemma 1.6]. For each $j$ and $t$ take a point $a^{(j, t)}=\left(a_{j+k}^{(j, t)}\right)_{k} \in \tilde{X}_{j}$ such that $a_{j}^{(j, t)}=t$. For $y=\left(y_{j+k}\right)_{k} \in \tilde{X}_{j}$ define

$$
r_{j}(y)=\left(y_{j+k}^{*}\right)_{k \in \mathbb{Z}} \in \tilde{X}_{j}
$$

where $y_{j+k}^{*}=y_{j+k}$ if $k \geq 0$ and $y_{j+k}^{*}=a_{j+k}^{\left(j, y_{j}\right)}$ if $k \leq 0$. Let

$$
\begin{equation*}
u_{j}(y)=\sum_{k=0}^{\infty}\left(\psi_{j+k}\left(\tilde{T}_{j}^{k} y\right)-\psi_{j+k}\left(\tilde{T}_{j}^{k} r_{j}(y)\right)\right) . \tag{B.1}
\end{equation*}
$$

To see that the RHS of (B.1) converges note that since $y$ and $r_{j}(y)$ have the same values at the coordinates indexed by $j+k$ for $k \geq 0$ (i.e. with indexes to the right of $j$ ) we have

$$
\begin{equation*}
\left|\psi_{j+k}\left(\tilde{T}_{j}^{k} y\right)-\psi_{j+k}\left(\tilde{T}_{j}^{k} r_{j}(y)\right)\right| \leq v_{\alpha}\left(\psi_{j+k}\right) 2^{-\alpha k} \tag{B.2}
\end{equation*}
$$

and so $\sup \left|u_{j}\right| \leq \sum_{k=0}^{\infty} v_{\alpha}\left(\psi_{j+k}\right) 2^{-\alpha k}$. Thus, $\sup _{j} \sup _{\tilde{X}_{j}}\left|u_{j}\right|<\infty$. Moreover, if $v_{\alpha}\left(\psi_{j}\right) \rightarrow 0$ then $\sup _{\tilde{x}}\left|u_{j}\right| \rightarrow 0$.
$\tilde{X}_{j}$
Next, we claim that the function $\psi_{j}-u_{j}+u_{j+1} \circ \tilde{T}_{j}$ depends only on the coordinates indexes by $j+k$ for $k \geq 0$ (so it has the form $\phi_{j} \circ \pi_{j}$ ). Indeed,

$$
u_{j}-u_{j+1} \circ \tilde{T}_{j}=\psi_{j}+\sum_{k=0}^{\infty}\left(\tilde{T}_{j+k} \circ \tilde{T}_{j}^{k} \circ r_{j}-\psi_{j+k+1} \circ \tilde{T}_{j+1}^{k} \circ r_{j+1} \circ \tilde{T}_{j}\right) .
$$

Note the the sum on the above right hand side depends only on the coordinates indexed by $j+k$ for $k \geq 0$, and the claim follows.

In view of the above, in order to complete the proof of the lemma it is enough to obtain appropriate bounds on the Hölder constants of the functions $u_{j}$ (corresponding to the exponent $\alpha / 2)$. For that end, let $y$ and $y^{\prime}$ in $\tilde{X}_{j}$ be such that $y_{j+k}=y_{j+k}^{\prime}$ for every $|k| \leq n$ for some $n>0$. Using (B.2) we have

$$
\begin{gathered}
\left|u_{j}(y)-u_{j}\left(y^{\prime}\right)\right| \leq \sum_{k=0}^{[n / 2]}\left|\psi_{j+k}\left(\tilde{T}_{j}^{k} y\right)-\psi_{j+k}\left(\tilde{T}_{j}^{k} y^{\prime}\right)\right|+\sum_{k=0}^{[n / 2]}\left|\psi_{j+k}\left(\tilde{T}_{j}^{k} r_{j}(y)\right)-\psi_{j+k}\left(\tilde{T}_{j}^{k} r_{j}\left(y^{\prime}\right)\right)\right| \\
+2 \sum_{k>[n / 2]} v_{\alpha}\left(\psi_{j+k}\right) 2^{-\alpha k}:=I_{1}+I_{2}+I_{3}
\end{gathered}
$$

To show that $u_{j}$ is Hölder continuous with exponent $\alpha / 2$ (uniformly in $j$ ) we use that

$$
\left|\psi_{j+k}\left(\tilde{T}_{j}^{k} y\right)-\psi_{j+k}\left(\tilde{T}_{j}^{k} y^{\prime}\right)\right| \leq \sup _{s} v_{\alpha}\left(\psi_{s}\right) 2^{-(n-k) \alpha}
$$

and similarly with $r_{j}(y)$ and $r_{j}\left(y^{\prime}\right)$ instead of $y$ and $y^{\prime}$, respectively. So $I_{1}+I_{2} \leq C 2^{-n \alpha / 2}$. Moreover, we note that $I_{3} \leq C 2^{-n \alpha / 2}$.

Next, if $\left\|\psi_{j}\right\|_{\alpha} \rightarrow 0$ then $I_{1}+I_{2} \leq \sum_{k=0}^{[n / 2]} v_{\alpha}\left(\psi_{j+k}\right) 2^{-(n-k) \alpha} \leq \varepsilon_{j} 2^{-n \alpha / 2}$ with $\varepsilon_{j} \rightarrow 0$. Similarly, $I_{3} \leq \varepsilon_{j} 2^{n \alpha / 2}$ for $\varepsilon_{j}$ with the same properties.

## B.2. Sequential Gibbs measures for one sided shifts.

Definition B.3. Let $\phi_{j}: X_{j} \rightarrow \mathbb{R}$ be a sequences of functions such that sup $\left\|\phi_{j}\right\|_{\alpha}<\infty$ for some $\alpha \in(0,1]$. We say that a sequence of probability measures $\mu_{j}$ on $X_{j}$ is a sequential Gibbs family for $\left(\phi_{j}\right)$ if:
(i) For all $j$ we have $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$
(ii) There is a constant $C>1$ and a sequence of positive numbers $\left(\lambda_{j}\right)$ such that for all $j$, every point $\left(x_{j+k}\right)_{k}$ in $X_{j}$ and every $r>0$ we have

$$
C^{-1} e^{S_{j, r} \phi(x)} / \lambda_{j, r} \leq \gamma_{j}\left(\left[x_{j}, \ldots, x_{j+r-1}\right]\right) \leq C e^{S_{j, r} \phi(x)} / \lambda_{j, r}
$$

where $S_{j, r} \phi(x)=\sum_{s=0}^{r-1} \phi_{j+s}\left(T_{j}^{s} x\right)$ and $\lambda_{j, r}=\prod_{k=j}^{j+r-1} \lambda_{k}$.
We say that two sequences $\left(\alpha_{j}\right)$ and $\left(\beta_{j}\right)$ of positive numbers are equivalent if there is a sequence $\left(\zeta_{j}\right)$ of positive numbers which is bounded and bounded away from 0 such that for all $j$ we have $\alpha_{j}=\zeta_{j} \beta_{j} / \zeta_{j+1}$.

We need the following simple result.
Lemma B.4. Two positive sequences $\left(\alpha_{j}\right)$ and $\left(\beta_{j}\right)$ are equivalent if and only if there is a constant $C>0$ such that for all $j$ and $n$ we have $C \leq \alpha_{j, n} / \beta_{j, n} \leq C^{-1}$.

Proof. It is clear that there exists such a constant $C$ if $\left(\alpha_{j}\right)$ and $\left(\beta_{j}\right)$ are equivalent. On other hand, suppose that such a constant exists. Then, for $j \geq 0$ define $\zeta_{j}=\beta_{0, j} / \alpha_{0, j}$. Clearly, $\alpha_{j} / \beta_{j}=\zeta_{j} / \zeta_{j+1}$. For $j<0$ we define similarly $\zeta_{j}=\alpha_{j+1,|j|} / \beta_{j+1,|j|}$.

Theorem B.5. For every sequence of functions $\phi_{j}: X_{j} \rightarrow \mathbb{R}$ such that sup $\left\|\phi_{j}\right\|_{\alpha}<\infty$ for some $\alpha \in(0,1]$ there exists unique Gibbs measures $\mu_{j}$. Moreover the sequence $\left(\lambda_{j}\right)$ is unique up to equivalence.

We will see below that the measures $\mu_{j}$ can be expressed in terms of the strong limits of the normalized operators $\mathcal{L}_{j}^{n}$ associated with the sequence $\left(\phi_{j}\right)$ defined in B.4). The rate of convergence is exponential, see (B.5) below.

Proof. Existence of $\mu_{j}$ was proven in [48]. Let us recall the main arguments. First, we define the operator $L_{j}$ which maps a function $g: X_{j} \rightarrow \mathbb{R}$ to a function $L_{j} g: X_{j+1} \rightarrow \mathbb{R}$ given by

$$
L_{j} g(x)=\sum_{T_{j} y=x} e^{\phi_{j}(y)} g(y) .
$$

Then, in [48] it was shown that there is a positive sequence $\left(\lambda_{j}\right)$ which is bounded and bounded away from the origin, a sequence of positive functions $h_{j}$ with uniformly bounded Hölder norms (corresponding to the same exponent $\alpha$ of $\phi_{j}$ ), which are also uniformly bounded below by a positive constant, and a sequence of probability measures $\nu_{j}$ on $X_{j}$ such that $\nu_{j}\left(h_{j}\right)=1$ and for all $j, n$ and a Hölder continuous function $g$ on $X_{j}$,

$$
\begin{equation*}
\left\|\left(\lambda_{j, n}\right)^{-1} L_{j}^{n} g-\nu_{j}(g) h_{j+n}\right\|_{\alpha} \leq C_{0}\|g\|_{\alpha} \delta^{n} \tag{B.3}
\end{equation*}
$$

for some constants $C>0$ and $\delta \in(0,1)$ which do not depend on $j, n$ and $g$. Here

$$
L_{j}^{n}=L_{j+n-1} \circ \cdots \circ L_{j+1} \circ L_{j} .
$$

Then the Gibbs measures constructed in [48] are given by $\mu_{j}=h_{j} d \nu_{j}$. Moreover, the operator $\mathcal{L}_{j}$ given by

$$
\begin{equation*}
\mathcal{L}_{j} g(x)=\frac{L_{j}\left(g h_{j}\right)}{\lambda_{j} h_{j+1}} \tag{B.4}
\end{equation*}
$$

is the dual of the Koopman operator corresponding to $T_{j}$ with respect to $\mu_{j}$ and $\mu_{j+1}$ and (because of B.3)),

$$
\begin{equation*}
\left\|\mathcal{L}_{j}^{n} g-\mu_{j}(g) \mathbf{1}\right\|_{\alpha} \leq C\|g\|_{\alpha} \delta^{n} \tag{B.5}
\end{equation*}
$$

for some $C>0$. In the derivation of (B.5) we used that $\sup _{j}\left\|1 / h_{j}\right\|_{\alpha}<\infty$.
Now, let us suppose that there is another Gibbs measure $\tilde{\mu}_{j}$ associated with a sequence $\tilde{\lambda}_{j}$. Namely, $\left(T_{j}\right)_{*} \tilde{\mu}_{j}=\tilde{\mu}_{j+1}$ and for every point $x \in X_{j}$,

$$
\left.(\tilde{C})^{-1} e^{S_{j, r} \phi(x)} / \tilde{\lambda}_{j, r} \leq \tilde{\mu}_{j}\left(\left[x_{j}, x_{j+1}, \ldots, x_{j+r-1}\right]\right]\right) \leq \tilde{C} e^{S_{j, r} \phi(x)} / \tilde{\lambda}_{j, r}
$$

for some constant $\tilde{C}>0$. Let us first show that the sequences $\left(\lambda_{j}\right)$ and $\left(\tilde{\lambda}_{j}\right)$ are equivalent. Indeed, for all $j$ and $n$, by the Gibbs property of both $\mu_{j}$ and $\tilde{\mu}_{j}$,

$$
C^{-1} \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} e^{S_{j, n} \phi\left(x_{y}\right)} / \lambda_{j, n} \leq \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} \mu_{j}\left(\left[y_{j}, \ldots, y_{j+n-1}\right]\right) \leq C \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} e^{S_{j, n} \phi\left(x_{y}\right)} / \lambda_{j, n}
$$

and

$$
(\tilde{C})^{-1} \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} e^{S_{j, n} \phi\left(x_{y}\right)} / \tilde{\lambda}_{j, n} \leq \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} \tilde{\mu}_{j}\left(\left[y_{j}, \ldots, y_{j+n-1}\right]\right) \leq \tilde{C} \sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} e^{S_{j, n} \phi\left(x_{y}\right)} / \tilde{\lambda}_{j, n}
$$

where the point $x_{y}$ is an arbitrary point inside the cylinder $\left[y_{j}, \ldots, y_{j+n-1}\right]$. On the other hand,

$$
1=\sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} \mu_{j}\left(\left[y_{j}, \ldots, y_{j+n-1}\right]\right)=\sum_{\left[y_{j}, \ldots, y_{j+n-1}\right]} \tilde{\mu}_{j}\left(\left[y_{j}, \ldots, y_{j+n-1}\right]\right) .
$$

Hence,

$$
(C \tilde{C})^{-1} \leq \frac{\lambda_{j, n}}{\tilde{\lambda}_{j, n}} \leq C \tilde{C}
$$

and the equivalence of the sequences $\left(\lambda_{j}\right)$ and $\left(\tilde{\lambda}_{j}\right)$ follows from Lemma B. 4 .
Next, we show that $\tilde{\mu}_{j}=\mu_{j}$ for all $j$. We first need the following result. Let $H_{j, \alpha}$ denote the space of Hölder continuous functions on $X_{j}$ with the Hölder exponent $\alpha$, equipped with the usual $\alpha$-Hölder norm $\|\cdot\|_{\alpha}$. Let $H_{j, \alpha}^{*}$ denote its dual, and denote by $\|\cdot\|_{\alpha}$ the (operator) norm on the dual, as well. Let $\mathcal{L}_{j}^{*}: H_{j+1, \alpha}^{*} \rightarrow H_{j, \alpha}^{*}$ be the dual operators of $\mathcal{L}_{j}$. Then it follows from (B.5) that for every $j, n$ and $\kappa_{j+n} \in H_{j+n, \alpha}^{*}$,

$$
\begin{equation*}
\left\|\left(\mathcal{L}_{j}^{n}\right)^{*} \kappa_{j+n}-\kappa_{j+n}\left(\mathbf{1}_{j+n}\right) \mu_{j}\right\|_{\alpha} \leq C\left\|\kappa_{j+n}\right\|_{\alpha} \delta^{n} \tag{B.6}
\end{equation*}
$$

where $\mathbf{1}_{k}$ is the function on $X_{k}$ which takes constant value 1 . In particular, if we take $\kappa_{j+n}=\tilde{\mu}_{j+n}$ then, since $\tilde{\mu}_{j+n} \mathbf{1}_{j+n}=1$, we see that

$$
\begin{equation*}
\left(\mathcal{L}_{j}^{n}\right)^{*} \tilde{\mu}_{j+n} \rightarrow \mu_{j} \text { as } n \rightarrow \infty . \tag{B.7}
\end{equation*}
$$

Next, we claim that

$$
\begin{equation*}
\left(\mathcal{L}_{j}^{n}\right)^{*} \tilde{\mu}_{j+n}=\tilde{\mu}_{j} . \tag{B.8}
\end{equation*}
$$

Once the claim is proven, combining it with B.7 we obtain $\tilde{\mu}_{j}=\mu_{j}$, as required.
Now, because of the Gibbs properties of both $\mu_{j}$ and $\tilde{\mu}_{j}$ and since $\left(\lambda_{j}\right)$ and $\left(\tilde{\lambda}_{j}\right)$ are equivalent we see that $\tilde{\mu}_{j} \ll \mu_{j}$. Let $p_{j}=\frac{d \tilde{\mu}_{j}}{d \mu_{j}}$ denote the corresponding Radon Nikodym derivative. In order to prove ( $\overline{\mathrm{B} .8}$, we will show that for every $j$,

$$
\begin{equation*}
p_{j+1} \circ T_{j}=p_{j}, \quad \mu_{j}-\text { a.s. } \tag{B.9}
\end{equation*}
$$

(B.9) implies that $p_{j+n} \circ T_{j}^{n}=p_{j}$. Therefore, for every bounded measurable function $g: X_{j} \rightarrow \mathbb{R}$,

$$
\left(\mathcal{L}_{j}^{n}\right)^{*} \tilde{\mu}_{j+n}(g)=\tilde{\mu}_{j+n}\left(\mathcal{L}_{j}^{n} g\right)=\mu_{j+n}\left(p_{j+n} \mathcal{L}_{j}^{n} g\right)=\mu_{j}\left(g \cdot\left(p_{j+n} \circ T_{j}^{n}\right)\right)=\mu_{j}\left(g \cdot p_{j}\right)=\tilde{\mu}_{j}(g)
$$

and B.8 follows (note that in the third equality above we have used that $\mathcal{L}_{j}^{n}$ is the transfer operator of $T_{j}^{n}$ with respect to $\mu_{j}$ and $\mu_{j+n}$ ).

To complete the proof of the theorem we need to prove (B.9). By identifying both measures $\mu_{j}$ and $\tilde{\mu}_{j}$ as measure on the two sided shift $\tilde{X}_{j}$ (see next section), we can assume that $\mu_{j}$ and $\tilde{\mu}_{j}$ are mapped by $\tilde{T}_{j}$ to $\mu_{j+1}$ and $\tilde{\mu}_{j+1}$, respectively. In what follows we will abuse the notation and denote the lifted measures by $\mu_{j}$ and $\tilde{\mu}_{j}$. The identification of the function $p_{j}$ to $\tilde{X}_{j}$ will also be denoted by $p_{j}$, namely, we write $p_{j}\left(\ldots, x_{j-1}, x_{j}, x_{j+1}, \ldots\right)=p_{j}\left(x_{j}, x_{j+1}, \ldots\right)$. Now, since $\left(\tilde{T}_{j}\right)_{*} \mu_{j}=\mu_{j+1}$ and $\left(\tilde{T}_{j}\right)_{*} \tilde{\mu}_{j}=\tilde{\mu}_{j+1}$ for every bounded measurable function $f: \tilde{X}_{j+1} \rightarrow \mathbb{R}$ we have

$$
\mu_{j}\left(p_{j}\left(f \circ \tilde{T}_{j}\right)\right)=\tilde{\mu}_{j}\left(f \circ \tilde{T}_{j}\right)=\tilde{\mu}_{j+1}(f)=\mu_{j+1}\left(p_{j+1} f\right)=\mu_{j}\left(\left(p_{j+1} \circ \tilde{T}_{j}\right) \cdot\left(f \circ \tilde{T}_{j}\right)\right)
$$

Since $T_{j}$ is invertible we can replace $f \circ T_{j}$ with a general bounded measurable function $g$ on $X_{j}$ and (B.9) follows.

## B.3. Gibbs measure for two sided shifts.

Definition B.6. Let $\psi_{j}: \tilde{X}_{j} \rightarrow \mathbb{R}$ be a sequences of functions such that $\sup \left\|\psi_{j}\right\|_{\alpha}<\infty$ for some $\alpha \in(0,1]$. A sequence of probability measures $\gamma_{j}$ on $\tilde{X}_{j}$ is a Gibbs family for $\left(\psi_{j}\right)$ if:
(i) For all $j$ we have $\left(\tilde{T}_{j}\right)_{*} \gamma_{j}=\gamma_{j+1}$.
(ii) There is a constant $C>1$ and a sequence of positive numbers $\left(\lambda_{j}\right)$ so that for every point $\left(y_{j+k}\right)_{k}$ in $\tilde{X}_{j}$ we have

$$
\left.C^{-1} e^{S_{j, r} \psi(y)} \lambda_{j, r} \leq \gamma_{j}\left(\left[y_{j}, y_{j+1}, \ldots, y_{j+r-1}\right]\right]\right) \leq C e^{S_{j, r} \psi(y)} \lambda_{j, r}
$$

where $S_{j, r} \psi(y)=\sum_{s=0}^{r-1} \phi_{j+s}\left(\tilde{T}_{j}^{s} y\right)$ and $\lambda_{j, r}=\prod_{k=j}^{j+r-1} \lambda_{k}$.
Proposition B.7. Every uniformly Hölder continuous sequence $\psi_{j}$ (with respect to some exponent $\alpha \in(0,1])$ admits a unique sequence of Gibbs measures $\gamma_{j}$.
Proof. Let $\phi_{j}$ be the functions obtained in Lemma B. 2 . Let $\mu_{j}$ be the unique sequential Gibbs measure on the one sided shift space $X_{j}$ corresponding to the functions $\left(\phi_{j}\right)$. Since $\left(T_{j-k}\right)_{*} \mu_{j-k}=$ $\mu_{j}$ using Kolmogorov's extension theorem we can extend $\mu_{j}$ to the space of two sided sequences $\tilde{X}_{j}$. Let us denote this measure by $\gamma_{j}$. It is clear that $\left(\tilde{T}_{j}\right)_{*} \gamma_{j}=\gamma_{j+1}$. To prove the second condition in the definition of a sequential Gibbs measure we notice that for every point $x \in \tilde{X}_{j}$,

$$
\gamma_{j}\left(\left[x_{j}, \ldots, x_{j+r-1}\right]\right)=\mu_{j}\left(\left[x_{j}, \ldots, x_{j+r-1}\right]\right) \asymp \lambda_{j, n} e^{S_{j, r} \phi \circ \pi_{j}(x)}
$$

where $\left(\lambda_{j}\right)$ is the sequence associated with the Gibbs measure $\mu_{j}$. Now, by Lemma B.2 we have that

$$
e^{S_{j, r} \phi \circ \pi_{j}(x)} \asymp e^{S_{j, r} \psi(x)}
$$

and hence $\gamma_{j}$ is a sequential Gibbs measure corresponding to the functions $\psi_{j}$, and $\left(\lambda_{j}\right)$ is the corresponding associated sequence.

To prove uniqueness of $\gamma_{j}$ constructed above, we note that if $\tilde{\gamma}_{j}$ is another Gibbs measure then the measure $\tilde{\mu}_{j}$ defined by the restriction of $\gamma_{j}$ to the $\sigma$-algebra generated by the coordinates with indexes $j+k, k \geq 0$ project ${ }^{5}$ to a Gibbs measure on $X_{j}$. Hence $\tilde{\mu}_{j}=\mu_{j}$, where $\mu_{j}$ is the unique sequential Gibbs measure corresponding to the functions $\phi_{j}$. On the other hand, since $\left(\tilde{T}_{j}\right)_{*} \tilde{\gamma}_{j}=\tilde{\gamma}_{j+1}$, for every $r>0$ the restriction of $\tilde{\gamma}_{j}$ to the $\sigma$-algebra generated by the coordinates indexed by $j+k$ for $k \geq-r$ coincides with the restriction of $\tilde{\gamma}_{j-r}$ to the $\sigma$-algebra generated by the coordinates indexed by $j-r+k$ for $k \geq 0$, which, as explained above, projects on $X_{j-r}$ to $\mu_{j-r}$. Hence, for every $r$ the measures $\gamma_{j}$ and $\tilde{\gamma}_{j}$ agree on the $\sigma$-algebra generated by the coordinates indexed by $j+k$ for $k \geq-r$ (as both coincide with $\mu_{j-r}$ ). By taking $r \rightarrow-\infty$ we conclude that $\tilde{\gamma}_{j}=\gamma_{j}$.

## Appendix C. Small perturbations of hyperbolic maps

C.1. Hyperbolic sets. Let $M$ be a compact $C^{2}$ Riemannian manifold equipped with its Borel $\sigma$-algebra $\mathcal{B}$. Denote by $\mathrm{d}(\cdot, \cdot)$ the induced metric. Let $T: M \rightarrow M$ be a $C^{2}$ diffeomorphism.
Definition C.1. A compact $T$ invariant subset $\Lambda \subset M$ is called a hyperbolic set for $T$ if there exists an open set $V$ with compact closure, constants $\lambda \in(0,1)$ and $\alpha_{0}, A_{0}, B_{0}>0$ and subbudnles $\Gamma^{s}$ and $\Gamma^{u}$ of the tangent bundle $T \Lambda$ such that:
(i) The set $\left\{x \in M: \operatorname{dist}(x, \Lambda)<\alpha_{0}\right\}$ is contained in a open subset $U$ of $V$ such that $T U \subset V$ and $\left.T\right|_{U}$ is a diffeomorphism with $\sup _{x \in U} \max \left(\left\|D_{x} T\right\|,\left\|D_{x} T^{-1}\right\|\right) \leq A_{0}$;
(ii) $T \Lambda=\Gamma^{s} \oplus \Gamma^{u}, D T\left(\Gamma^{s}\right)=\Gamma^{s}, D T\left(\Gamma^{u}\right)=\Gamma^{u}$ and the minimal angle between $\Gamma^{s}$ and $\Gamma^{u}$ is bounded below by $\alpha$;
(iii) For all $n \in \mathbb{N}$ we have

$$
\left\|D_{x} T^{n} v\right\| \leq B_{0} \lambda^{n}\|v\| \forall v \in \Gamma_{x}^{s} \quad \text { and } \quad\left\|D_{x} T^{-n} v\right\| \leq B_{0} \lambda^{n}\|v\| \quad \forall v \in \Gamma_{x}^{u} .
$$

[^3]Definition C.2. A hyperbolic set is called
(i) locally maximal if the set $U$ above could be chosen so that $\Lambda=\bigcap_{n \in \mathbb{Z}} T^{n} U$ (that is, $\Lambda$ is the largest hyperbolic set contained in $U$ );
(ii) hyperbolic attractor, if in addition, $U$ could be chosen so that $T U \subset U$ (in the case when $M=\Lambda T$ is called Anosov).

We say that $\Lambda$ is the basic hyperbolic set if it is infinite locally maximal hyperbolic set such that $\left.T\right|_{\Lambda}$ is topologically transitive.

Henceforth, we assume that $\Lambda$ is topologically mixing ${ }^{6}$ basic hyperbolic set.
A powerful tool for studying hyperbolic maps is given by a symbolic representations. Namely, every topologically mixing basic set $\Lambda$ admits a Markov partition (see [79, Chapter 10]) which gives raise to a semiconjugacy $\pi: \Sigma \rightarrow M$ where $\Sigma$ is a topologically mixing subshift of a finite type.
C.2. Structural stability. Now, consider a sequence of maps $\mathcal{T}=\left(T_{j}: M \rightarrow M\right)_{j \in \mathbb{Z}}$. Denote by $\mathrm{d}_{1}(f, g)$ the $C^{1}$ distance between $f$ and $g$. We have the following result.

Theorem C.3. If $\delta_{1}(\mathcal{T}):=\sup _{1}\left(T, T_{j}\right)$ is small enough then there is a sequence of sets $\Lambda_{j} \subset M$ and homeomorphisms $h_{j}: \Lambda \stackrel{j}{\rightarrow} \Lambda_{j}$ (that we think of as a "sequential conjugacy") such that $h_{j}$ and $h_{j}^{-1}$ are uniformly Hölder continuous,

$$
\begin{equation*}
T_{j} \Lambda_{j}=\Lambda_{j+1} \text { and } T_{j} \circ h_{j}=h_{j+1} \circ T . \tag{C.1}
\end{equation*}
$$

Moreover $\sup _{j}\left\|h_{j}-\mathrm{Id}\right\|_{C^{0}} \rightarrow 0$ as $\delta_{1}(\mathcal{T}) \rightarrow 0 .{ }^{7}$
The sets $\left(\Lambda_{j}\right)$ are sequentially hyperbolic for the sequence $\mathcal{T}$ in the following sense. They are compact, satisfy $T_{j} \Lambda_{j}=\Lambda_{j+1}$ and there exist constants $\lambda^{\prime} \in(0,1)$ and $\alpha_{1}, A_{1}, B_{1}>0$ and sequences of subbudnles $\Gamma_{j}^{s}=\left\{\Gamma_{j, x}^{s}: x \in \Lambda_{j}\right\}$ and $\Gamma_{j}^{u}=\left\{\Gamma_{j, x}^{u}: x \in \Lambda_{j}\right\}$ of the tangent bundle $T \Lambda_{j}$ such that, for each $j$ :
(i) The set $\left\{x \in M: \mathrm{d}\left(x, \Lambda_{j}\right)<\alpha_{1}\right\}$ is contained in an open subset $U_{j}$ of $V$ such that $T_{j} U_{j} \subset V$ and $\left.T_{j}\right|_{U_{j}}$ is a diffeomorphism satisfying

$$
\sup _{j} \sup _{x \in U_{j}} \max \left(\left\|D_{x} T_{j}\right\|,\left\|D_{x} T_{j}^{-1}\right\|\right) \leq A_{1}
$$

(ii) $T \Lambda_{j}=\Gamma_{j}^{s} \oplus \Gamma_{j}^{u}, D T_{j}\left(\Gamma_{j}^{s}\right)=\Gamma_{j+1}^{s}, D T_{j}\left(\Gamma_{j}^{u}\right)=\Gamma_{j+1}^{u}$ and the minimal angle between $\Gamma_{j}^{s}$ and $\Gamma_{j}^{u}$ is bounded below by $\alpha_{1}$;
(iii) For every $n \in \mathbb{N}$ and all $j$ we have $\left\|D_{x} T_{j}^{n} v\right\| \leq B_{1} \cdot\left(\lambda^{\prime}\right)^{n}\|v\|$ for every $v \in \Gamma_{j, x}^{s}$ and $\left\|D_{x} T_{j}^{-n} v\right\| \leq B_{1} \cdot\left(\lambda^{\prime}\right)^{n}\|v\|$ for every $v \in \Gamma_{j, x}^{u}$ where $T_{j}^{-n}=\left(T_{j-n}^{n}\right)^{-1}$.
(iv) $T_{j} U_{j} \subset U_{j+1}$ and $\bigcap_{n=0}^{\infty} T_{j-n}^{n} U_{j-n}=\Lambda_{j}$.

[^4]Theorem C.3 is proven in [62, Example 4.2.3], [40, Example 2.5] and [64, Theorem 1.1]) except the Hölder continuity of $h_{j}$ and $h_{j}^{-1}$ was not discussed there. The proof of Hölder continuity is quite standard but for completeness it is included in \$C.4. We note that in [62, 40, 64] the smallness of $\left\|T_{j}-T\right\|_{C^{1}}$ was not uniform in $j$, which led to non-uniform in $j$ hyperbolicity. However, when $\delta_{1}(\mathcal{T})$ is uniformly small then the arguments in the proofs yield the uniform hyperbolicity.
C.3. Limit Theorems. Let $\pi_{j}=h_{j} \circ \pi$. Then $\pi_{j}$ provides a semiconjugacy between the sequence $\mathcal{T}$ and the subshift $\Sigma$ describing the symbolic dynamics of $T$. Given a sequence of Hölder functions $\phi_{j}$ on $\Lambda_{j}$ let $\psi_{j}=\phi_{j} \circ \pi_{j}$. Note that $\psi_{j}$ are Hölder continuous due to Hölder continuity of $\pi_{j}$. Let $\nu_{j}$ be the Gibbs measures for $\left\{\psi_{j}\right\}$ which exist due to the results of Appendix B. Define measures $\mu_{j}$ on $\Lambda_{j}$ by $\mu_{j}(A)=\nu_{j}\left(\pi_{j}^{-1} A\right)$. Given a sequence of Holder functions $f_{j}$ on $\Lambda_{j}$ let $g_{j}=f_{j} \circ \pi_{j}$. Thus $S_{n} f=S_{n} g \circ \pi_{0}^{-1}$ and so $\left(S_{n} f\right)(x)$ when $x$ is distributed according to $\mu_{0}$ has the the distribution as $\left(S_{n} g\right)(\omega)$ when $\omega$ is distributed according to $\nu_{0}$. We thus obtain

Corollary C.4. Theorems 3.1 and 3.2 are valid for $\left(S_{n} f\right)(x)$ where $x$ is distributed according to $\mu_{0}$.

We also have the following result (see [40, Theorem 4.3]).
Theorem C. 5 (Sequential SRB measures). Suppose that $\Lambda$ is a hyperbolic attractor. Then there is a sequence of probability measures $\mu_{j}$ on $\Lambda$ such that $\left(T_{j}\right)_{*} \mu_{j}=\mu_{j+1}$ and

$$
\mu_{j}=\lim _{n \rightarrow \infty}\left(T_{j-n}^{n}\right)_{*}\left(\rho_{j-n} d \mathrm{Vol}\right)
$$

for every uniformly bounded sequence of probability densities $\rho_{n}$ on $\Gamma_{n}$. The measures $\left(\mu_{j}\right)$, $j \in \mathbb{Z}$ are the unique family of equivariant measures such that the conditional measures of $\mu_{j}$ on the unstable manifolds (see $\$$ C.4) at time $j$ are absolutely continuous with respect to the Riemannian volume on these submanifolds.

Moreover $\mu_{j}$ can be obtained by the construction described above corresponding to the sequence of functions $\phi_{j}=-\ln J\left(T_{j} \mid \Gamma_{j}^{u}\right)$, where $J(\cdot)$ stands for the Jacobian matrix.
Therefore $W_{n}=\sum_{j=0}^{n-1} f_{j}\left(T_{0}^{j} x\right)$ satisfies all the limit theorems in Section 3] when $x$ is distributed according to a measure having a Hölder density with respect to $\mu_{0}$.

Remark C.6. When starting with a time zero sequential SRB measure Theorems 3.1 and 3.2 hold for uniformly hyperbolic sequences $\left(T_{j}\right)$ without the assumption that the maps $T_{j}$ are close, if $\liminf _{n \rightarrow \infty} \frac{1}{n} \operatorname{Var}\left(S_{n}\right)>0$. Indeed, in this setup we can use the functional analytic approach of Bakhtin [8] together with the perturbative approach in Section D. The reason this works well only when $\lim \inf _{n \rightarrow \infty} \frac{1}{n} \operatorname{Var}\left(S_{n}\right)>0$ is that when the variance grows sublinearly fast our proofs rely on a block decomposition technique. This involves a sufficiently smooth martingale coboundary representations, a tool which is not available in the setup of [8] since the transfer operators acts on spaces of distributions and not functions. See also 43] for a similar functional approach, which has applications to sufficiently close non-autonomous hyperbolic sequences (see [43, Section 7]).
C.4. Hölder continuity of the conjugacies. To prove the Hölder continuity of $h_{j}$ and $h_{j}^{-1}$ we need some background.

Local stable and unstable manifolds. For $\varepsilon$ small enough, and $x \in \Lambda_{j}$ define $W_{j}^{s}(x, \varepsilon)$ to be the set of all points $y \in \Lambda_{j}$ such that $\mathrm{d}\left(T_{j}^{n} x, T_{j}^{n} y\right) \leq \varepsilon$ for all $n$ and $\mathrm{d}\left(T_{j}^{n} x, T_{j}^{n} y\right) \rightarrow 0$. Similarly, we define $W_{j}^{u}(x, \varepsilon)$ to be the set of all points $y \in \Lambda_{j}$ such that

$$
\mathrm{d}\left(\left(T_{j-n}^{n}\right)^{-1} x,\left(T_{j-n}^{n}\right)^{-1} y\right) \leq \varepsilon
$$

for all $n$ and $\mathrm{d}\left(\left(T_{j-n}^{n}\right)^{-1} x,\left(T_{j-n}^{n}\right)^{-1} y\right) \rightarrow 0$. Then (see [40, 62] and [70]) $W_{j}^{s}(x, \varepsilon)$ and $W_{j}^{u}(x, \varepsilon)$ are manifolds and tangent space of $W_{j}^{s}(x, \varepsilon)$ at $x$ is $\Gamma_{x, j}^{s}$ while the tangent space of $W_{j}^{u}(x, \varepsilon)$ at $x$ is $\Gamma_{x, j}^{u}$. Moreover, there are constants $C>0$ and $\delta \in(0,1)$ such that for every $j$,

$$
\begin{equation*}
\mathrm{d}\left(T_{j}^{n} x, T_{j}^{n} y\right) \leq C \delta^{n} \text { for all } y \in W_{j}^{s}(x, \varepsilon) \tag{C.2}
\end{equation*}
$$

and

$$
\begin{equation*}
\mathrm{d}\left(\left(T_{j-n}^{n}\right)^{-1} x,\left(T_{j-n}^{n}\right)^{-1} y\right) \leq C \delta^{n} \text { for all } y \in W_{j}^{u}(x, \varepsilon) \tag{C.3}
\end{equation*}
$$

Remark C.7. Like in the autonomous case, we have the following (see [70, Proposition 3.5]). Define $\bar{W}_{j}^{s}(x, \beta)$ to be the set of all points $y \in \Lambda_{j}$ such that $\mathrm{d}\left(T_{j}^{n} x, T_{j}^{n} y\right) \leq \beta$ for all $n \geq 0$. Similarly, let $\bar{W}_{j}^{u}(x, \beta)$ be the set of all points $y \in \Lambda_{j}$ such that $\mathrm{d}\left(\left(T_{j-n}^{n}\right)^{-1} x,\left(T_{j-n}^{n}\right)^{-1} y\right) \leq \beta$ for all $n \geq 0$. Then for every $\beta<\frac{1}{2} \varepsilon$,

$$
\bar{W}_{j}^{s}(x, \beta) \subset W_{j}^{s}(x, \varepsilon) \quad \text { and } \quad \bar{W}_{j}^{u}(x, \beta) \subset W_{j}^{u}(x, \varepsilon) .
$$

This essentially means that, like in the autonomous case, up to replacing $\varepsilon$ with $\frac{1}{2} \varepsilon$, the local stable/unstable manifolds can be defined using only the condition about the $\varepsilon$-closeness of the forward/backward orbits.

Let us also denote by $W^{s}(x, \varepsilon)$ and $W^{u}(x, \varepsilon)$ the local stable and unstable manifolds of $x$ with respect to $T$ (then all the above properties hold true).

Proof of Hölder continuity. The proof is a minor modification of the proof of [54, Proposition 19.1.2], but for readers' convenience we provide the details. We only prove that $h_{j}$ is Hölder continuous, the proof that $h_{j}^{-1}$ is Hölder continuous is analogous, see below. Fix some $\varepsilon$ small enough (in a way that will be determined later). We say that $x, y \in \Lambda$ are $s$-equivalent if $y \in W^{s}(x, \varepsilon)$. Similarly, we say that they are $u$-equivalent if $y \in W^{u}(x, \varepsilon)$. Then, using local coordinates and that the angles between the stable and unstable directions are uniformly bounded below we see that if $\varepsilon$ is small enough then there is a constant $K_{1}$ such that

$$
\begin{equation*}
\mathrm{d}(x, y)^{2}+\mathrm{d}(y, z)^{2} \leq K_{1} \mathrm{~d}(x, z)^{2} \tag{C.4}
\end{equation*}
$$

for every $x, y, z \in \Lambda$ such that $x$ is $s$-equivalent to $y$ and $y$ is $u$-equivalent to $z$, and $\mathrm{d}(x, z) \leq \varepsilon$. In view of (C.4) in order to show that $h_{j}$ is Hölder continuous it suffices to show that restrictions of $h_{j}$ to both unstable and stable manifolds are uniformly Hölder (see e.g. [54, Proposition 19.1.1]). We will consider $h_{j} \mid W^{u}(x, \varepsilon)$, the result for $h_{j} \mid W^{s}(x, \varepsilon)$ follows by replacing $T$ by $T^{-1}$.

Let us prove that $h_{j} \mid W^{u}(x, \varepsilon)$ are uniformly Hölder continuous. Fix some $\varepsilon_{0}>0$. By uniform continuity of $h_{j}$ proven in [64], there exists $0<\delta_{0}<\varepsilon$ such that for every $j$ and every $x, y \in \Lambda$ with $\mathrm{d}(x, y)<\delta_{0}$ we have $\mathrm{d}\left(h_{j}(x), h_{j}(y)\right)<\varepsilon_{0}$. Let $x, y \in \Lambda$ be such that $y \in W^{u}(x, \varepsilon)$. Denote $\rho=\mathrm{d}(x, y)$ and let $K>1$ be a Lipschitz constant for $T$. Assuming that $\rho \leq K^{-2} \delta_{0}$ there is a unique natural number $n$ such that $K^{n} \rho<\delta_{0} \leq K^{n+1} \rho$. Then $\mathrm{d}\left(T^{k} x, T^{k} y\right) \leq K^{n} \rho<\delta_{0}$ for all $k \leq n$ and so $\mathrm{d}\left(h_{j+n} T^{n} x, h_{j+n} T^{n} y\right)<\varepsilon_{0}$. Since $y \in W^{u}(x, \varepsilon)$ and $\mathrm{d}\left(T^{k} x, T^{k} y\right)<\delta_{0}<\varepsilon$ for $0 \leq k \leq n$ we have $T^{n} y \in W^{u}(x, \varepsilon)$. Using Remark C. 7 and the equicontinuity of $\left(h_{k}\right)$, we see that $h_{j+n} T^{n} y \in W_{j+n}^{u}\left(h_{j+n} T^{n} x, \varepsilon^{\prime}\right)$ where $\varepsilon^{\prime}=\varepsilon^{\prime}(\varepsilon) \rightarrow 0$ as $\varepsilon \rightarrow 0$ (we can take $\varepsilon^{\prime}=$ $\sup _{k}\left(\sup \left\{\mathrm{~d}\left(h_{k}(a), h_{k}(b)\right): \mathrm{d}(a, b) \leq \varepsilon\right\}\right)$. Thus, by fixing $\varepsilon$ small enough and using (C.3) (with the
above $\varepsilon^{\prime}$ instead of $\varepsilon$ ) we see that there are constants $C>0$ and $\delta \in(0,1)$ such that

$$
\begin{gathered}
\mathrm{d}\left(h_{j}(x), h_{j}(y)\right)=\mathrm{d}\left(\left(T_{j}^{n}\right)^{-1} h_{j+n} T^{n} x,\left(T_{j}^{n}\right)^{-1} h_{j+n} T^{n} y\right) \leq C \delta^{n} \varepsilon_{0} \\
=C \varepsilon_{0} K^{-c n}=C \varepsilon_{0} K^{c}\left(K^{-n-1}\right)^{c} \leq\left(C \varepsilon_{0} K^{c} \delta_{0}^{-c}\right) \rho^{c}=\left(C \varepsilon_{0} K^{c} \delta_{0}^{-c}\right) \mathrm{d}(x, y)^{c}
\end{gathered}
$$

where $c=\frac{|\ln \delta|}{\ln K}>0$.
This shows Hölder continuity of $h_{j}$. The proof that $h_{j}^{-1}$ is Hölder continuous is similar except we use the equivalence relations induced by $W_{j}^{u}(x, \varepsilon)$ and by $W_{j}^{s}(x, \varepsilon)$.

## Appendix D. Perturbation theorem

D.1. The statement. Let $\left(B_{j},\|\cdot\|_{j}\right)_{j \geq 0}$ be a sequence of Banach spaces and
$A_{j}: B_{j} \rightarrow B_{j+1}$ be a sequence of bounded linear operators. We assume here that there are $\left(\lambda_{j}, h_{j}, \nu_{j}\right) \in(\mathbb{C} \backslash\{0\}) \times B_{j} \times B_{j}^{*}$ where $B_{j}^{*}$ denotes the dual space of $B_{j}$ so that

$$
\begin{equation*}
A_{j} h_{j}=\lambda_{j} h_{j+1}, \quad\left(A_{j}\right)^{*} \nu_{j+1}=\lambda_{j} \nu_{j} \tag{D.1}
\end{equation*}
$$

We will also assume that $\operatorname{dim}\left(B_{j}\right)>1$ and that $\nu_{j}\left(h_{j}\right)=1$.
Let $\mathbf{C}$ be a complex Banach space. Let $\varepsilon>0$. Denote by $B_{\mathbf{C}}(0, \varepsilon)$ the open ball in $\mathbf{C}$ around 0 with radius $\varepsilon$. For each $t \in B_{\mathbf{C}}(0, \varepsilon)$ let $A_{j}^{(t)}: B_{j} \rightarrow B_{j+1}$ be a sequence of bounded operators such that $A_{j}^{(0)}=A_{j}$. Set $A_{j}^{t, n}=A_{j+n-1}^{(t)} \circ \cdots \circ A_{j}^{(t)}$. In the sequel, for any sequence of operators $R_{j}: B_{j} \rightarrow B_{j+1}$ we will denote $R_{j}^{n}=R_{j+n-1} \circ \cdots \circ R_{j+1} \circ R_{j}$. Write $\mathbb{C}^{\prime}=\mathbb{C} \backslash\{0\}$.
Assumption D.1. (i) $\inf _{j}\left|\lambda_{j}\right|>0$.
(ii) There are constants $C_{0}, \delta_{0}>0$ such that for all $j$ and $g \in B_{j}$ so that $\nu_{j}(g)=0$ we have

$$
\begin{equation*}
\left\|A_{j}^{n} g\right\| \leq C_{0}\|g\| \delta_{0}^{n} \tag{D.2}
\end{equation*}
$$

Moreover, $\liminf _{n \rightarrow \infty} \inf _{j}\left|\lambda_{j, n}\right|^{1 / n}>\delta_{0}$, where $\lambda_{j, n}=\prod_{k=j}^{j+n-1} \lambda_{k}$.
(iii) The maps $t \rightarrow A_{j}^{(t)}$ are analytic in some neighborhood $\mathcal{U}$ of the origin (which does not depend on $j$ ) and $\sup _{t \in \mathcal{U}}\left\|A_{j}^{(t)}\right\|$ are uniformly bounded in $j$.

The main result in this section is as follows.
Theorem D. 2 (A sequential perturbation theorem). Under Assumption D.1, for every $\delta_{1}>\delta_{0}$ there exist $r_{1}, C_{1}>0$ such that for every $j \in \mathbb{N}$ and $t \in B_{\mathbf{B}}\left(0, r_{1}\right)$ we have the following.
(i) There are triplets $\left(\lambda_{j}(t), h_{j}^{(t)}, \nu_{j}^{(t)}\right) \in \mathbb{C}^{\prime} \times B_{j} \times B_{j}^{*}$, which are uniformly bounded in $j, t$ so that

$$
A_{j}^{(t)} h_{j}^{(t)}=\lambda_{j}(t) h_{j+1}^{(t, y)}, \quad\left(A_{j}^{(t)}\right)^{*} \nu_{j+1}^{(t)}=\lambda_{j}(t) \nu_{k, j}^{(t, y)} .
$$

Moreover, $\lambda_{j}(0)=\lambda_{j}, h_{j}^{(0)}=h_{j}, \nu_{j}^{(0)}=\nu_{j}$, and the triplets are analytic functions of $t$ and their norm is uniformly bounded. Furthermore, $\nu_{j}\left(h_{j}^{(t)}\right)=\nu_{j}^{(t)}\left(h_{\ell, j}\right)=\nu_{j}^{(t)}\left(h_{j}^{(t)}\right)=1$. If instead of analyticity of $t \mapsto A_{j}^{(t)}$ we assume that $\mathbf{C}$ is a real Banach space and $t \mapsto A_{j}^{(t)}$ is $C^{k}$ for some $k \in \mathbb{N}$ with uniform bounds on the $C^{k}$ norms, then the above triplets are $C^{k}$ functions of $t$ with uniformly bounded $C^{k}$ norms.
(ii) Consider the operators $P_{j}^{(t)}: B_{j} \rightarrow B_{j+1}$ given by $P_{j}^{(t)} g=\lambda_{j} \nu_{j}^{(t)}(g) h_{j+1}^{(t)}$. Then

$$
P_{j}^{t, n}=\lambda_{j, n}(t) \nu_{j}^{(t)}(\cdot) h_{j+n}^{(t)}
$$

and denoting $E_{j}^{(t)}=A_{j}^{(t)}-P_{j}^{(t)}$ we have $A_{j}^{t, n}=P_{j}^{t, n}+E_{j}^{t, n}$ and

$$
\begin{equation*}
\left\|E_{j}^{t, n}\right\| \leq C_{1} \delta_{1}^{n} \tag{D.3}
\end{equation*}
$$

D.2. Preparations for the proof of Theorem D.2, Our first result shows that we can always consider two sided sequences instead of one sided ones.
Lemma D.3. We can extend both sequences $\left(B_{j},\|\cdot\|_{j}\right)_{j \geq 0}$ and $\left(A_{j}\right)_{j \geq 0}$ to two sided sequences such that (D.1) and Assumption D.1 still hold.

Proof. For $j<0$ we take $B_{j}=B_{0}$. Then we take a nonzero $\nu_{-1} \in B_{0}^{*}$ and set $\nu_{j}=\nu_{-1}$ for $j<-1$. We also define $\lambda_{j}=1+\delta_{0}$ for $j<0$. Finally, we take a nonzero $h_{-1} \in B_{0}$ such that $\nu_{-1}\left(h_{-1}\right)=1$ and define $h_{j}=h_{-1}$ for all $j<0$. For $j<0$, define $A_{j}(g)=\left(1+\delta_{0}\right) \nu_{-1}(g) h_{j+1}=\lambda_{j} \nu_{j}(g) h_{j+1}$. It is clear that (D.1) holds with the new two sided sequences. Assumption D.1 is in force since for $j<0$, if $\nu_{j}(g)=0$ then $A_{j}(g)=0$ and so $A_{j}^{n} g=0$.
Henceforth we assume that we have a two sided sequence satisfying (D.1) and Assumption D.1.
Remark D.4. If the operators $A_{j}$ are defined for all $j \in \mathbb{Z}$ then the proof of Theorem D.2 (which proceeds by applying the Implicit Function Theorem) shows that the projections $P_{j}$ constructed in the theorem are unique. The uniqueness does not hold if the operators are only defined for $j \in \mathbb{N}$ (cf. the discussion after Theorem 2.4).
Lemma D.5. (i) Let the operator $Q_{j}: B_{j} \rightarrow B_{j}$ be given by $Q_{j}(g)=\nu_{j}(g) h_{j}$. Then for every $j$ and $g \in B_{j}$ we have $\nu_{j}\left(Q_{j} g\right)=\nu_{j}(g)$.
(ii) Consider the operator $P_{j}: B_{j} \rightarrow B_{j+1}$ given by

$$
P_{j} g=\lambda_{j} \nu_{j}(g) h_{j+1} .
$$

Then $P_{j}$ coincides with $A_{j}$ on $\eta_{j}:=\operatorname{span}\left\{h_{j}\right\}$ and

$$
A_{j+1} P_{j}=P_{j+1} A_{j}=P_{j+1} P_{j}, \text { and } P_{j}^{n}=\lambda_{j, n} \nu_{j}(\cdot) h_{j+n}
$$

Therefore, with $E_{j}=A_{j}-P_{j}$ we have

$$
E_{j}^{n}=A_{j}^{n}-P_{j}^{n}=A_{j}^{n}-\lambda_{j, n} \nu_{j}(\cdot) h_{j+n} .
$$

In particular, $A_{j}^{n}=P_{j}^{n}+E_{j}^{n}$.
Proof. (i) $\nu_{j}\left(\nu_{j}(g) h_{j}\right)=\nu_{j}(g) \nu_{j}\left(h_{j}\right)=\nu_{j}(g)$.
(ii) We have $P_{j} h_{j}=\lambda_{j} \nu_{j}\left(h_{j}\right) h_{j}=\lambda_{j} h_{j+1}$. So $A_{j}$ and $P_{j}$ coincide on $\eta_{j}$. Next, the identity $P_{j}^{n}=\lambda_{j, n} \nu_{j}(\cdot) h_{j+n}$ follows by induction on $n$. For $n=1$ is it just the definition of $P_{j}$, and to prove the induction step, if $P_{j}^{n}=\lambda_{j, n} \nu_{j}(\cdot) h_{j+n}$ then

$$
P_{j}^{n+1}(g)=P_{j+n}\left(P_{j}^{n} g\right)=\lambda_{j, n} \nu_{j}(g) P_{j+n}\left(h_{j+n}\right)=\lambda_{j, n} \nu_{j}(g) \lambda_{j+n} h_{j+n+1}=\lambda_{j, n+1} \nu_{j}(g) h_{j+n+1} .
$$

Since $\left(A_{j}\right)^{*} \nu_{j+1}=\lambda_{j} \nu_{j}$ we have $A_{j+1} P_{j}=P_{j+1} A_{j}=P_{j+1} P_{j}$, because all three expression above coincide with the operator

$$
g \rightarrow \lambda_{j} \lambda_{j+1} \nu_{j}(g) h_{j+2}=P_{j}^{2}(g) .
$$

Using this the proof that $E_{j}^{n}=A_{j}^{n}-P_{j}^{n}$ proceeds by induction on $n$.
Lemma D.6. Suppose that Assumption D.1 holds and let $C$ be a constant so that $\sup _{j}\left\|Q_{j}\right\| \leq C$. Then for all $g \in B_{j}$,

$$
\begin{equation*}
\left\|A_{j}^{n} g-\lambda_{j, n} \nu_{j}(g) h_{j+n}\right\|=\left\|A_{j}^{n}-P_{j}^{n}\right\| \leq(C+1) C_{0}\|g\| \delta_{0}^{n} . \tag{D.4}
\end{equation*}
$$

Proof. Set $R_{j} g=g-Q_{j} g$. By Lemma D.5, $\nu_{j}\left(R_{j} g\right)=0$. So by (D.2) applied to $R_{j} g$

$$
\left\|A_{j}^{n} g-\lambda_{j, n} \nu_{j}(g) h_{j+n}\right\|=\left\|A_{j}^{n}\left(R_{j} g\right)\right\| \leq(C+1) C_{0}\|g\| \delta_{0}^{n} .
$$

Lemma D.7. Let $R_{j}: B_{j} \rightarrow B_{j+1}$ be a sequence of operators. Suppose that there are constants $\delta_{0}, C_{0}>0$ such that for all $j, n$ we have

$$
\left\|R_{j}^{n}\right\| \leq C_{0} \delta_{0}^{n}
$$

Then for every $\delta_{1}>\delta_{0}$ there exist constants $\varepsilon_{0}>0$ and $C_{1}>0$ with the following property. For any other family $S_{j}: B_{j} \rightarrow B_{j+1}$ of linear operators satisfying sup $\left\|R_{j}-S_{j}\right\|<\varepsilon_{0}$ we have

$$
\left\|S_{j}^{n}\right\| \leq C_{1} \delta_{1}^{n} .
$$

Proof. Let $\delta_{1}>\delta_{0}$ and let $n_{0}$ be so that $C_{0} \delta_{0}^{n_{0}}<\frac{1}{2} \delta_{1}^{n_{0}}$. Let $\varepsilon_{0}>0$ be so small such that

$$
\sup _{j}\left\|R_{j}-S_{j}\right\|<\varepsilon_{0}
$$

implies $\sup _{j}\left\|S_{j}^{n_{0}}\right\|<\delta_{1}^{n_{0}}$. It is indeed possible to find such $\varepsilon_{0}$ since

$$
\left\|S_{j}^{n_{0}}-R_{j}^{n_{0}}\right\| \leq C\left(C_{0}, \delta_{0}, n_{0}\right) \varepsilon_{0}
$$

for some constant $C\left(C_{0}, \delta_{0}, n_{0}\right)$ which depends only on $C_{0}, \delta$ and $n_{0}$.
Now, given $n \in \mathbb{N}$ let us write $n=k n_{0}+\ell$ for some $0 \leq \ell<n_{0}$. Then
$\left\|S_{j}^{n}\right\| \leq \prod_{u=0}^{k-1}\left\|S_{j+u n_{0}}^{n_{0}}\right\| \cdot\left\|S_{j+k n_{0}}^{\ell}\right\| \leq\left(\varepsilon_{0}+C_{0} \delta_{0}\right)^{\ell} \delta_{1}^{-\ell} \delta_{1}^{n}$.

## D.3. Proof of Theorem D.2.

Proof. We first prove the existence of triplets. Denote

$$
\mathbf{H}_{j}=\left\{g \in B_{j}: \nu_{j}(g)=0\right\}
$$

Since $A_{j}^{*} \nu_{j+1}=\lambda_{j} \nu_{j}$ we have $A_{j} \mathbf{H}_{j} \subset \mathbf{H}_{j+1}$. Also, for every $g_{j} \in \mathbf{H}_{j}$ we have $\nu_{j}\left(h_{j}+g_{j}\right)=1$ and

$$
\nu_{j+1}\left(A_{j}\left(h_{j}+g_{j}\right)\right)=\lambda_{j}\left(A_{j}^{*} \nu_{j+1}\right)\left(h_{j}+g_{j}\right)=\lambda_{j} \nu_{j}\left(h_{j}+g_{j}\right)=\lambda_{j} .
$$

Consider the space $\mathbf{H}$ of sequences $\bar{g}=\left(g_{j}\right)$ so that $g_{j} \in \mathbf{H}_{j}$ for all $j$, equipped with the norm $\|\bar{g}\|:=\sup _{j}\left\|g_{j}\right\|<\infty$. Define a function $F: \mathbf{C} \times \mathbf{H} \rightarrow \mathbf{H}$ by

$$
(F(t, \bar{g}))_{j}=A_{j-1}^{(t)}\left(h_{j-1}+g_{j-1}\right)-\nu_{j}\left(A_{j-1}^{(t)}\left(h_{j-1}+g_{j-1}\right)\right)\left(h_{j}+g_{j}\right) .
$$

Then $F_{k}(0,0)=0$, and $F$ is analytic in $t$ (in the case $C^{k}$ dependence on $t$ this map is differentiable $k$ times).

We claim that the partial derivative $\Phi:=(\partial F / \partial \mathbf{H})(0,0)$ is an isomorphism. Indeed, a direct computation shows that $(\Phi \bar{g})_{j}=A_{j-1} g_{j-1}-\lambda_{j} g_{j}$. Assume first that there is $\bar{g} \in \mathbf{H}$ so that

$$
\begin{equation*}
A_{j-1} g_{j-1}=\lambda_{j} g_{j} \tag{D.5}
\end{equation*}
$$

When the restrictions $A_{j}: \mathbf{H}_{j} \rightarrow \mathbf{H}_{j+1}$ satisfy

$$
\begin{equation*}
\left\|A_{j}^{n} g_{j}\right\| \leq C_{0}\left\|g_{j}\right\| \delta_{0}^{n} \tag{D.6}
\end{equation*}
$$

since on $\mathbf{H}_{j}$ the functionals $\nu_{j}$ vanish. Iterating (D.5) we see that $A_{j}^{n} g_{j}=\lambda_{j, n} g_{j+n}$. Plugging in $j-n$ instead of $j$ in (D.6) we get that $g_{j}=\frac{A_{j-n}^{n} g_{j-n}}{\lambda_{j-n, n}}$ and so

$$
\left\|g_{j}\right\| \leq C_{0}\|\bar{g}\| \frac{\delta_{0}^{n}}{\left|\lambda_{j, n}\right|} \leq C^{\prime} \varepsilon^{n} \rightarrow 0
$$

for some $\varepsilon \in(0,1)$ and a constant $C^{\prime}$, where in the second inequality uses Assumption D.1(ii). Hence $\bar{g}=0$.

Now we show that $\Phi$ is surjective. Take $\bar{g} \in \mathbf{H}$. Set $h_{j}=-\sum_{n=0}^{\infty} \frac{A_{j-n}^{n} g_{j-n}}{\lambda_{j-n, n}}$. This series converges uniformly in $j$ since $\left\|\frac{A_{j-n}^{n} g_{j}}{\lambda_{j-n, n}}\right\| \leq C^{\prime}\|\bar{g}\| \varepsilon^{n}$ where $\varepsilon$ and $C^{\prime}$ are like in the previous estimates. It is immediate that

$$
A_{j-1} h_{j-1}-\lambda_{j} h_{j}=\lambda_{j} g_{j}
$$

Applying the implicit function theorem in Banach spaces (e.g. [5, Theorem 3.2]) we get that there is $\varepsilon_{0}>0$ so that for every $t \in B_{\mathbf{C}}\left(0, \varepsilon_{0}\right)$ there is $\bar{g}^{(t)}=\left(g_{j}^{(t)}\right)_{j} \in \mathbf{H}$ so that

$$
F\left(t, \bar{g}^{(t)}\right)=0 .
$$

Moreover, the function $t \rightarrow \bar{g}^{(t)}$ is analytic and so by possibly reducing $\varepsilon_{0}$ we can assume that it is bounded (in the case of $C^{k}$ dependence on $t$ we get that this map is of class $C^{k}$ ).

Now we construct the families $\nu_{j}^{(t)}$. First, we obtain from (D.4) that

$$
\begin{equation*}
\left\|\left(A_{j}^{n}\right)^{*}-\lambda_{j, n} h_{j+n}^{*} \nu_{j}\right\| \leq C_{0} \delta_{0}^{n} \tag{D.7}
\end{equation*}
$$

where $h_{j+n}^{*}(\mu)=\mu\left(h_{j+n}\right)$. Let $\mathbf{E}_{j} \subset B_{j}^{*}$ be the space of all functionals $\mu_{j}$ such that $\mu_{j}\left(h_{j}\right)=0$. Then $\left(A_{j}\right)^{*} \mathbf{E}_{j+1} \rightarrow \mathbf{E}_{j}$ since

$$
\left(A_{j}\right)^{*} \mu_{j+1}\left(h_{j}\right)=\mu_{j+1}\left(A_{j} h_{j}\right)=\lambda_{j} \mu_{j+1}\left(h_{j+1}\right)=0
$$

Let $\mathbf{E}$ be the Banach space of sequences $\bar{\mu}=\left(\mu_{j}\right)$ such that $\mu_{j} \in \mathbf{E}_{j}$ for all $j$ and $\|\bar{\mu}\|:=\sup _{j}\left\|\mu_{j}\right\|<\infty$. Define

$$
(G(t, \bar{\mu}))_{j}=\left(A_{j}^{(t)}\right)^{*}\left(\nu_{j+1}+\mu_{j+1}\right)-\lambda_{j}(t)\left(\nu_{j}+\mu_{j}\right)
$$

where $\lambda_{j}(t)=\nu_{j}\left(h_{j}^{t}\right)$. Then $G(0,0)=0$ and $G$ is a well defined function on $\left\{t:\|t\|<\varepsilon_{0}\right\}$ for some $\varepsilon_{0}>0$. Moreover, $G$ is analytic in $t$. We consider $G$ as a function from a neighborhood of $(0,0) \in \mathbf{C} \times \mathbf{E}$ to $\mathbf{E}$.

Now, a direct computation shows that the derivative of $G$ at $(0,0)$ in the direction $\mathbf{E}$ is the operator given by

$$
\Psi:=(\partial G / d \mathbf{E})(0,0) \bar{\mu})_{j}=\left(A_{j}\right)^{*} \mu_{j+1}-\lambda_{j} \mu_{j} .
$$

Let us show that $\Psi$ is injective. If $\bar{\mu}$ belongs to its kernel then $\left(A_{j}\right)^{*} \mu_{j+1}=\lambda_{j} \mu_{j}$, and so $\left(A_{j}^{n}\right)^{*} \mu_{j+n}=\lambda_{j, n} \mu_{j}$. Using (D.7) we get that $\left\|\mu_{j}\right\| \leq C \varepsilon^{n}$ for some constants $\varepsilon \in(0,1)$ and $C>0$. Therefore $\bar{\mu}=0$.

Next, $\Psi$ is also surjective since for any $\bar{\mu} \in \mathbf{E}$ we have $\left(A_{j}\right)^{*} \kappa_{j+1}-\lambda_{j} \kappa_{j}=\lambda_{j} \mu_{j}$ where $\kappa_{j}=-\sum_{n=0}^{\infty} \frac{\left(A_{j+1}^{n}\right)^{*} \mu_{j+n+1}}{\lambda_{j, n}}$ (the convergence of this series follows from (D.7)).

Hence, by the implicit function theorem there is an analytic function $t \mapsto \bar{\kappa}^{(t)}=\left(\kappa_{j}^{(t)}\right)_{j}$ with values in $\mathbf{E}$ so that

$$
\left(A_{j}^{(t)}\right)^{*}\left(\nu_{k j+1}+\kappa_{j+1}^{(t)}\right)-\lambda_{j}(t)\left(\nu_{j}+\kappa_{j}^{(t)}\right)=0 .
$$

Take $\nu_{j}^{(t)}=\nu_{j}+\kappa_{j}^{(t)}$ and notice that $\nu_{j}^{(t)}\left(h_{j}^{(t)}\right)$ does not depend on $j$. Indeed, we have

$$
\lambda_{j}(t) \nu_{j+1}^{(t)}\left(h_{j+1}^{(t)}\right)=\nu_{j+1}^{(t)}\left(\lambda_{j}(t) h_{j+1}^{(t)}\right)=\nu_{j+1}^{(t)}\left(A_{j}^{(t)} h_{j}^{(t)}\right)=\left(A_{j}^{(t)}\right)^{*} \nu_{j+1}^{(t)}\left(h_{j}^{(t)}\right)=\lambda_{j}(t) \nu_{j}^{(t)}\left(h_{j}^{(t)}\right) .
$$

We conclude that there is a function $c(t)$ such that $\nu_{j}^{(t)}\left(h_{j}^{(t)}\right)=c(t)$ for every $j \geq 0$. The functions $c(t)$ are bounded and analytic and $c(0)=1$. Thus, by decreasing $\|t\|$ if necessary we can assume that $|c(t)| \geq \frac{1}{2}$. Next, by replacing $h_{j}^{(t)}$ with $h_{j}^{(t)} / c(t)$ (or $\nu_{j}^{(t)}$ by $\nu_{j}^{(t)} / c(t)$ ) we can just assume
that $c(t)=1$. Like in the construction of $g^{(t)}$, in the case $t \rightarrow\left(A_{j}^{(t)}\right)$ is of class $C^{k}$ all the above functions of $t$ are differentiable $k$ times.

Finally, the exponential convergence follows by Lemma D.7 applied with $R_{j}=A_{j}-P_{j}$ and $S_{j}=A_{j}^{(t)}-P_{j}^{(t)}$ (when $t$ is close enough to 0 ) where $P_{j}^{(t)}=\lambda_{j}(t) \nu_{j}^{(t)}(\cdot) h_{j+1}^{(t)}$ taking into account that by Lemma D.5. $S_{j}^{n}=A_{j}^{t, n}-P_{j}^{t, n}=A_{j}^{t, n}-\lambda_{j, n}(t) \nu_{j}^{(t)}(\cdot) h_{j+n}^{(t)}$.

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[^0]:    ${ }^{1}$ This requires us to have a two sides sequence.

[^1]:    ${ }^{3}$ Here if $\gamma$ is a map from $\left\{|z| \leq r_{0}\right\}$ to a Banach space we take $\|\gamma\|=\sup _{|z| \leq r_{0}}\|\gamma(z)\|$.

[^2]:    ${ }^{4}$ We can take $A=2 \sup _{j}\left\|u_{j}\right\|_{L^{\infty}}$, where $u_{j}$ are the coboundaries from 6.2.

[^3]:    ${ }^{5}$ By "projects" we mean that if the restriction is denoted by $\gamma_{j}^{*}$ then $\left(\pi_{j}\right)_{*} \gamma_{j}^{*}=\mu_{j}$.

[^4]:    ${ }^{6}$ Topological mixing assumption can be made without a loss of generality. Indeed (see e.g. [79, Chapter 8]) an arbitrary basic set $\Lambda$ can be decomposed as $\Lambda=\bigcup_{j=1}^{p} \Lambda_{j}$ so that $T \Lambda_{j}=\Lambda_{j+1} \bmod p$ where $\Lambda_{j}$ are topologically mixing basic hyperbolic sets for $T^{p}$. Then we could apply the results discussed below to $\left(T^{p}, \Lambda_{j}\right)$.
    ${ }^{7}$ Note that in Theorem C. 3 we can also consider one sided sequences $\left(T_{j}\right)_{j \geq 0}$ because they can be extended to two sided ones. The reason we consider two sided sequences is because the definition of hyperbolicity requires considering negative times to define the unstable subspaces.

