Mutual Information of Wireless Channels and Block-Jacobi Ergodic Operators

Walid Hachem, Adrien Hardy, Shlomo Shamai

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Abstract

Shannon's mutual information of a random multiple antenna and multipath channel is studied in the general case where the channel impulse response is an ergodic and stationary process. From this viewpoint, the channel is represented by an ergodic self-adjoint block-Jacobi operator, which is close in many aspects to a block version of a random Schrödinger operator. The mutual information is then related to the so-called density of states of this operator. In this paper, it is shown that under the weakest assumptions on the channel, the mutual information can be expressed in terms of a matrix-valued stochastic process coupled with the channel process. This allows numerical approximations of the mutual information in this general setting. Moreover, assuming further that the channel impulse response is a Markov process, a representation for the mutual information offset in the large Signal to Noise Ratio regime is obtained in terms of another related Markov process. This generalizes previous results from Levy *et.al.* [16, 17]. It is also illustrated how the mutual information expressions that are closely related to those predicted by the random matrix theory can be recovered in the large dimensional regime.

1 Introduction and statement of the results

This paper is a contribution towards understanding the behavior of Shannon's mutual information of a general wireless ergodic channel with the help of the ergodic operator theory. A particular attention is devoted to the case where this channel is a Markov process. The large Signal to Noise Ratio (SNR) regime and the large dimensional regime are considered.

1.1 The model

Given two positive integers N and K, we consider the wireless transmission model

$$Y_n = \mathsf{F}_n S_{n-1} + \mathsf{G}_n S_n + V_n \tag{1}$$

with $n \in \mathbb{Z}$ and where:

- $(Y_n)_{n\in\mathbb{Z}}$ represents the $\mathbb{C}^N$-valued sequence of received signals.
- $(S_n)_{n\in\mathbb{Z}}$ is the $\mathbb{C}^K$-valued sequence of transmitted information symbols.
- $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ with $\mathsf{F}_n, \mathsf{G}_n \in \mathbb{C}^{N\times K}$ is a matrix representation of a random wireless channel.
- $(V_n)_{n\in\mathbb{Z}}$ is the additive noise.

Let us first give a few examples which fit this transmission model.

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The multipath single antenna fading channel. This model assumes the representation

$$y_n = \sum_{\ell=0}^{L} c_{n,\ell} s_{n-\ell} + v_n \tag{2}$$

where n is the time index, and $(y_n)_{n\in\mathbb{Z}}$, $(s_n)_{n\in\mathbb{Z}}$ and $(v_n)_{n\in\mathbb{Z}}$ are respectively the $\mathbb{C}$-valued received signal, the sent signal, and a white noise. Here L is the channel degree and $C_n = [c_{n,0}, \ldots, c_{n,L}]^\mathsf{T}$ is the channel impulse response at time n. When $L > 0$, this can be cast into the framework of (1) with

$$Y_n := \begin{bmatrix} y_{nL} \\ \vdots \\ y_{nL+L-1} \end{bmatrix}, \quad S_n := \begin{bmatrix} s_{nL} \\ \vdots \\ s_{nL+L-1} \end{bmatrix}, \quad V_n := \begin{bmatrix} v_{nL} \\ \vdots \\ v_{nL+L-1} \end{bmatrix}, \quad N := K := L,$$

and $\mathsf{F}_n, \mathsf{G}_n \in \mathbb{C}^{L\times L}$ are the upper triangular and lower triangular matrices defined by

$$\left[\begin{array}{c|c} \mathsf{F}_n & \mathsf{G}_n \end{array}\right] := \left[\begin{array}{ccc|ccc} c_{nL,L} & \cdots & c_{nL,1} & c_{nL,0} & & \\ & \ddots & \vdots & \vdots & \ddots & \\ & & c_{nL+L-1,L} & c_{nL+L-1,L-1} & \cdots & c_{nL+L-1,0} \end{array}\right]. \tag{3}$$

When $L = 0$, we set instead $N := K := 1$, $Y_n := y_n$, $S_n := s_n$, $V_n := v_n$, $\mathsf{F}_n := 0$, and $\mathsf{G}_n := c_{n,0}$.

In the multiple antenna variant of this model, the channel coefficients $c_{n,\ell}$ are $R \times T$ matrices, where R, resp. T, is the number of antennas at the receiver, resp. transmitter. In this case, the $N \times K$ matrices F_n and G_n given by Eq. (3) are block triangular matrices with $N := RL$ and $K := TL$.

The Wyner multi-cell model. Another instance of the transmission model introduced above is a generalization of the so-called Wyner multi-cell model considered in [12, 26]. Assume that the Base Stations (BS) of a wireless cellular network are arranged on a line, and that each BS receives in a given frequency slot the signals of L users. In general, most of these users are geographically closer to other BS. In this setting, the signal y_n received by the BS n is described by Eq. (2), where s_n is the signal emitted by User n and $c_{n,\ell}$ is the uplink channel carrying the signal of User $n - \ell$ to BS n.

In the two previous examples, the parameter n in Model (1) represents respectively the time and the space. Other domains such as the frequency domain can be covered, see *e.g.* [25], which deals with a time and frequency selective model. Moreover, this could even address different connected domains as the Doppler-Delay (connected via the so-called Zak transform), as in [3, 2], which lead to modulation schemes that are considered as interesting candidates for the fifth generation (5G) wireless systems, as reflected in the references [11, 5].

1.2 General assumptions

The purpose of this work is to study Shannon's mutual information between (S_n) and (Y_n) when the channel is known at the receiver. To this end, we focus on the usual setting where:

- The information sequence $(S_n)_{n\in\mathbb{Z}}$ is random i.i.d. with law $\mathcal{CN}(0, I_K)$.
- The noise $(V_n)_{n\in\mathbb{Z}}$ is i.i.d. with law $\mathcal{CN}(0, \frac{1}{\rho} I_K)$ for some $\rho > 0$.
- The random sequences $(S_n)_{n\in\mathbb{Z}}$, $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$, and $(V_n)_{n\in\mathbb{Z}}$ are independent.

We also make the following assumptions on the process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ representing the channel:

Assumption 1. The process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ is a *stationary* and *ergodic* process. Moreover,

$$\mathbb{E}\|\mathsf{F}_0\|^2 < \infty \quad \text{and} \quad \mathbb{E}\|\mathsf{G}_0\|^2 < \infty. \tag{4}$$

Here and in the following, i.i.d. means "independent and identically distributed", $\mathcal{CN}(0,\Sigma)$ stands for the law of a centered complex Gaussian vector with covariance matrix Σ, and $\rho > 0$ scales with the SNR. Note that the moment assumption (4) does not depend on the specific choice of the norm on the space of $N \times K$ complex matrices. In the remainder, we choose $\|\cdot\|$ to be the spectral norm.

Let us make precise the assumptions of stationarity and ergodicity. In the following we set for convenience

$$E := \mathbb{C}^{N\times K} \times \mathbb{C}^{N\times K}$$

and consider the measure space $\Omega := E^{\mathbb{Z}}$ equipped with its Borel σ–field $\mathscr{F} := \mathscr{B}(E)^{\otimes\mathbb{Z}}$. An element of Ω reads $\omega = (\dots,(F_{-1},G_{-1}),(F_0,G_0),(F_1,G_1),\dots)$ where (F_n,G_n) is the n^{th} coordinate of ω, with $(F_n,G_n) \in E$. The shift $T : \Omega \to \Omega$ acts by $T\omega := (\dots,(F_0,G_0),(F_1,G_1),(F_2,G_2),\dots)$. The assumption that $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$ is an ergodic stationary process, seen as a measurable map from $(\Omega,\mathscr{F})$ to itself, means that the shift T is a measure preserving and ergodic transformation with respect to the probability distribution of the process $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$.

Example 1. In the multipath single antenna fading channel model, this assumption amounts to assume the stationarity and ergodicity of the channel impulse response process $(C_n)_{n\in\mathbb{Z}}$, which is often the case to model the Doppler effect induced by the mobility of the communicating devices. For instance, the autoregressive (AR) channel model is a realistic model for representing the statistics of the channel in the time and the frequency dimensions. Given an integer $M > 0$, this model reads

$$C_n = \sum_{\ell=1}^{M} A_\ell C_{n-\ell} + U_n, \tag{5}$$

where $\{A_1,\dots,A_M\}$ is a collection of deterministic $(L+1)\times(L+1)$ matrices, and where $(U_n)_{n\in\mathbb{Z}}$ is an i.i.d. sequence. Since the matrix $[\mathsf{F}_n \ \mathsf{G}_n]$ is an arrangement of the vectors $C_{nL},\dots,C_{nL+L-1}$, see (3), it follows from (5) that the process $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$ is Markovian when $M \le L$. Further, if the polynomial $\det(I - \sum_{\ell=1}^{M} z^\ell A_\ell)$ does not vanish in the closed unit disc, it is well known that there exists a stationary and ergodic process whose law is characterized by (5), see *e.g.* [13, 21], leading to a stationary and ergodic process $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$.

1.3 Mutual information and statement of the main result

In order to define the mutual information of the channel described by (1), define for any $m \le n$ the random matrix of size $(n-m+1)N \times (n-m+2)K$,

$$\mathsf{H}_{m,n} := \begin{bmatrix} \mathsf{F}_m & \mathsf{G}_m & & & \\ & \mathsf{F}_{m+1} & \mathsf{G}_{m+1} & & \\ & & \ddots & \ddots & \\ & & & \mathsf{F}_n & \mathsf{G}_n \end{bmatrix}. \tag{6}$$

For any fixed $\rho > 0$, the *mutual information* of the channel is then given by

$$\mathcal{I}_\rho := \lim_{n-m\to\infty} \frac{1}{(n-m+1)N}\, \mathbb{E} \log\det\left(I + \rho\, \mathsf{H}_{m,n}\mathsf{H}_{m,n}^*\right). \tag{7}$$

More precisely, $\mathcal{I}_\rho$ is usually defined as an almost sure limit, namely without the expectation in the right hand side of (7), but the two quantities match due to the ergodicity assumption. It is known to represent the required mutual information per degree of freedom of our wireless channel, provided the input S_n is as in Section 1.2, see [8]. The purpose of this paper is to study this quantity. As we shall explain later, this limit exists and is finite, and does not depend on the way $n - m \to \infty$ due to the Assumption 1.

Denoting by $\mathcal{H}_K^{++}$, resp. $\mathcal{H}_K^{+}$, the cone of the Hermitian positive definite, resp. semidefinite, $K \times K$ matrices, we show that one can construct a stationary $\mathcal{H}_K^{++}$-valued process $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ defined recursively and coupled with $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ which allows a rather simple formula for the mutual information $\mathcal{I}_\rho$.

Theorem 1 (Mutual information of an ergodic channel)**.** If Assumption 1 holds true, then:

(a) There exists a unique stationary $\mathcal{H}_K^{++}$-valued process $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ satisfying

$$\mathsf{W}_n = \Big(I + \rho\, \mathsf{G}_n^* \left(I + \rho\, \mathsf{F}_n \mathsf{W}_{n-1} \mathsf{F}_n^*\right)^{-1} \mathsf{G}_n\Big)^{-1}. \tag{8}$$

In particular, the process (W_n) is ergodic.

(b) We have the representation for the mutual information:

$$\mathcal{I}_\rho = \frac{1}{N}\Big(\mathbb{E} \log\det\left(I + \rho\, \mathsf{F}_0 \mathsf{W}_{-1} \mathsf{F}_0^*\right) - \mathbb{E}\log\det \mathsf{W}_0\Big). \tag{9}$$

(c) Given *any* matrix $\mathsf{X}_{-1} \in \mathcal{H}_K^{+}$, if one defines a process $(\mathsf{X}_n)_{n\in\mathbb{N}}$ by setting

$$\mathsf{X}_n := \Big(I + \rho\, \mathsf{G}_n^* \left(I + \rho\, \mathsf{F}_n \mathsf{X}_{n-1} \mathsf{F}_n^*\right)^{-1} \mathsf{G}_n\Big)^{-1}$$

for all $n \geq 0$, then we have

$$\mathcal{I}_\rho = \lim_{n\to\infty} \frac{1}{nN} \sum_{\ell=0}^{n-1} \log\det\left(I + \rho\, \mathsf{F}_\ell \mathsf{X}_{\ell-1} \mathsf{F}_\ell^*\right) - \log\det \mathsf{X}_\ell \quad \text{a.s.}$$

The proof of Theorem 1 is provided in Section 3.

Remark 1. As we will illustrate in Section 2, Theorem 1(c) yields an estimator for $\mathcal{I}_\rho$ that is less costly numerically than the naive one, due to the dimension of the involved matrices.

Remark 2. The proof of Theorem 1 reveals that the moment assumption (4) can be weakened to

$$\mathbb{E}\log(1 + \|\mathsf{F}_0\|^2) < \infty \qquad \text{and} \qquad \mathbb{E}\log(1 + \|\mathsf{G}_0\|^2) < \infty.$$

The second moment assumption (4) is here to ensure that the received signal power is finite.

1.4 Connection to block-Jacobi operators and previous results

Recall Eq. (6). Due to Assumption 1, it is well known, see [22], that there exists a deterministic probability measure μ that can defined by the fact that for each bounded and continuous function f on $[0, \infty)$,

$$\frac{1}{(n-m+1)N} \operatorname{tr} f(\mathsf{H}_{m,n}\mathsf{H}_{m,n}^*) \xrightarrow[n-m\to\infty]{} \int f(\lambda)\,\mu(d\lambda) \quad \text{a.s.} \tag{10}$$

(here, f is of course extended by functional calculus to the semi-definite positive matrices). The measure μ is intimately connected with the so-called *ergodic self-adjoint block-Jacobi (or block-tridiagonal) operator* $\mathsf{H}\mathsf{H}^*$, where H is the random linear operator acting on the Hilbert space $\ell^2(\mathbb{Z}, \mathbb{C})$, and defined by its doubly-infinite matrix representation in the canonical basis $(e_k)_{k\in\mathbb{Z}}$ of this space as

$$\mathsf{H} = \begin{bmatrix} \ddots & \ddots & & & & \\ & \mathsf{F}_{-1} & \mathsf{G}_{-1} & & & \\ & & \mathsf{F}_0 & \mathsf{G}_0 & & \\ & & & \mathsf{F}_1 & \mathsf{G}_1 & \\ & & & & \ddots & \ddots \end{bmatrix}, \tag{11}$$

and where H^* stands for its adjoint operator. The random positive self-adjoint operator HH^* is an ergodic operator in the sense of [22, Page 33] (see also [10]), and the measure μ is called its *density of states*. Recalling (7), it holds that

$$\mathcal{I}_\rho = \int \log(1 + \rho\lambda)\, \mu(d\lambda),$$

where this limit is finite, due to the moment assumption (4) and a standard uniform integrability argument.

Mainly motivated by models from statistical physics, a very rich literature is devoted to the spectral analysis of random Jacobi operators, in connection with the Schrödinger equation in a random environment. In this framework, the Herbert-Jones-Thouless formula [6, 22] provides a means of describing the density of states μ of an ergodic Jacobi operator, in connection with the so-called Lyapounov exponent associated with a certain sequence of matrices. In [17], Levy *et al.* develop a version of this formula that is well suited to the block-Jacobi setting. The expression of the mutual information they obtained is then used to perform a large SNR asymptotic analysis so as to obtain bounds on the mutual information.

Here we take another route and identify $\mathcal{I}_\rho$ by considering the resolvents of certain random operators built from the process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ instead of using the Herbert-Jones-Thouless formula. The expression we obtain for $\mathcal{I}_\rho$ involves the ergodic process (W_n) which is coupled with the process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ by Eq. (8). This expression appears to be more tractable than the expression based on the top Lyapounov exponent provided in [17]. We note that an expression similar to ours is obtained by Levy *et.al.* in their other paper [16] that deals with the specific case where $N = 1$ and where the process $(\mathsf{F}_n, \mathsf{G}_n)$ is i.i.d.

We moreover exploit the rather simple expression (9) obtained for $\mathcal{I}_\rho$ to study two asymptotic regimes: we first consider the large SNR regime, namely when $\rho \to \infty$, and obtain an exact representation for the constant term in the expansion. We also consider the large dimensional regime where both N and K converge to infinity in a proportional way on an example; the expression of the mutual information that we recover is then closely related to what is obtained from random matrix theory [15, 10]. Among other applications, these results can be used to analyze the behavior of the mutual information of time and frequency selective channels in the framework of the massive MIMO systems ([20]), which are destined to play a dominant role in the future wireless cellular techniques/standards.

1.5 The Markovian case and large SNR regime

First, assuming extra assumptions on the process $(\mathsf{F}_n, \mathsf{G}_n)$, we obtain a description for the constant term (or mutual information offset) in the large SNR regime. Indeed, it often happens that there exists a real number κ_∞ such that the mutual information admits the expansion as $\rho \to \infty$,

$$\mathcal{I}_\rho = \min(\tfrac{K}{N}, 1) \log \rho + \kappa_\infty + o(1),$$

see e.g. [18]. Our next task is to prove this expansion indeed holds true and to derive an expression for the offset κ_∞ when the process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ is further assumed to be a Markov process satisfying some regularity and moment assumptions. Namely, consider for any $n \in \mathbb{Z}$ the σ-field $\mathscr{F}_n := \sigma((\mathsf{F}_k, \mathsf{G}_k)\ : k \le n)$ and assume there exists a transition kernel $P : E \times \mathscr{B}(E) \to [0, 1]$ such that, for any Borel function $f : E \to [0, \infty)$,

$$\mathbb{E}\big[f(\mathsf{F}_{n+1}, \mathsf{G}_{n+1})|\mathscr{F}_n\big] = Pf((\mathsf{F}_n, \mathsf{G}_n)) := \int f(F, G)\, P\big((\mathsf{F}_n, \mathsf{G}_n), dF \times dG\big).$$

Besides $Pf((\mathsf{F}, \mathsf{G}))$, we use the common notations from the Markov chains literature and also write $P((\mathsf{F}, \mathsf{G}), A) := P\mathbf{1}_A((\mathsf{F}, \mathsf{G}))$ for any Borel set $A \in \mathscr{B}(E)$; the iterated kernel P^n stands for the Markov kernel defined inductively by $P^n f := P(P^{n-1} f)$ with the convention that $P^0 f := f$; given any $\eta \in \mathcal{M}(E)$, we let ηP be the probability measure on E defined by

$$\eta P(A) := \int P((F, G), A)\, \eta(dF \times dG), \qquad A \in \mathscr{B}(E).$$

The following assumption is formulated in the context where $N > K$.

Assumption 2. The process $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ is a Markov process with transition kernel P associated with a unique invariant probability measure $\theta \in \mathcal{M}(E)$, namely satisfying $\theta P = \theta$. Moreover,

(a) P is Feller, namely, if $f : E \to \mathbb{R}$ is continuous and bounded, then so is Pf.

(b) $\mathbb{E}\|F_0\|^2 + \mathbb{E}\|\mathsf{G}_0\|^2 < \infty$.

(c) $\mathbb{E}|\log\det(\mathsf{F}_0^*\mathsf{F}_0)| < \infty$.

(d) For every non-zero $v \in \mathbb{C}^K$, we have for θ-a.e. $(F, G) \in E$ that

$$\det(G^*F) \neq 0 \qquad \text{and} \qquad \Pi_G^{\perp} F v \neq 0\,, \tag{12}$$

where $\Pi_G^{\perp}$ stands for the orthogonal projection on the orthogonal subspace to the linear span of the columns of G.

Remark 3. Since a Markov chain $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ associated with a unique invariant probability measure is automatically ergodic, we see that Assumption 2 is stronger than Assumption 1 and thus Theorem 1 applies in this setting.

Remark 4. If one assumes $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ is a sequence of i.i.d random variables with law θ having a density on E, then it satisfies Assumption 2 (and hence Assumption 1) provided that the moment conditions Assumption 2(b)-(c) are satisfied. We also provide more sophisticated examples were Assumption 2 holds in Section 1.5.1.

Remark 5. Since $\theta = \theta P$, Assumption 2(d) equivalently says that, for θ-a.e. (F, G), (12) holds true for $P((\mathsf{F}, \mathsf{G}), \cdot)$-a.e. $(F, G) \in E$. We will use this observation at several instances in the following.

Theorem 2 (The Markov case)**.** Let $N > K$. Then, under Assumption 2, the following hold true:

(a) There exists a unique stationary process $(\mathsf{Z}_n)_{n\in\mathbb{Z}}$ on $\mathcal{H}_K^{++}$ satisfying

$$\mathsf{Z}_n = \mathsf{G}_n^*(I + \mathsf{F}_n \mathsf{Z}_{n-1}^{-1} \mathsf{F}_n^*)^{-1} \mathsf{G}_n. \tag{13}$$

(b) We have, as $\rho \to \infty$,

$$\mathcal{I}_\rho = \frac{K}{N} \log \rho + \kappa_\infty + o(1),$$

where $\log\det(\mathsf{Z}_0 + \mathsf{F}_1^*\mathsf{F}_1)$ is integrable, and

$$\kappa_\infty := \frac{1}{N} \mathbb{E} \log\det(\mathsf{Z}_0 + \mathsf{F}_1^*\mathsf{F}_1).$$

(c) Given *any* $\mathsf{X}_{-1} \in \mathcal{H}_K^{++}$, if we consider the process $(\mathsf{X}_n)_{n\in\mathbb{N}}$ defined recursively by

$$\mathsf{X}_n = \mathsf{G}_n^*(I + \mathsf{F}_n \mathsf{X}_{n-1}^{-1} \mathsf{F}_n^*)^{-1} \mathsf{G}_n\,,$$

then we have, in probability,

$$\kappa_\infty = \lim_{n\to\infty} \frac{1}{nN} \sum_{\ell=0}^{n-1} \log\det(\mathsf{X}_\ell + \mathsf{F}_{\ell+1}^* \mathsf{F}_{\ell+1}). \tag{14}$$

The proof of Theorem 2 is provided in Section 4.

Remark 6 (The case $N \le K$)**.** In the statement of Theorem 2, it is assumed that $N > K$. Let us say a few words about the case where $N < K$. In this case, assuming that $(\mathsf{F}_n, \mathsf{G}_{n-1})$ is a Markov chain, there is an analogue $(\widetilde{\mathsf{Z}}_n)$ of the process (Z_n) satisfying the recursion

$$\widetilde{\mathsf{Z}}_n = \mathsf{F}_n (I_K + \mathsf{G}_{n-1}^* \widetilde{\mathsf{Z}}_{n-1}^{-1} \mathsf{G}_{n-1})^{-1} \mathsf{F}_n^*,$$

and adapting Assumption 2 to this new setting, we can show that $\mathcal{I}_\rho = \log \rho + \tilde{\kappa}_\infty + o(1)$, where

$$\tilde{\kappa}_\infty := \frac{1}{N} \mathbb{E} \log \det(\widetilde{\mathsf{Z}}_0 + \mathsf{G}_0 \mathsf{G}_0^*).$$

This result can be obtained by adapting the proof of theorem 2 in a straightforward manner. The case $K = N$ is somehow singular and requires a specific treatment that will not be undertaken in this paper; see also the end of Section 4.1.2 for further explanations.

Remark 7. In the case where $K = 1$, $N > 1$, and the process $(\mathsf{F}_n, \mathsf{G}_n)_{n \in \mathbb{Z}}$ is i.i.d., we recover [16, Th. 2], where this result is obtained with the help of the theory of Harris Markov chains.

1.5.1 Examples where Assumption 2 is verified

In the following two examples, it is well known that the Markov process $(\mathsf{F}_n, \mathsf{G}_n)_{n \in \mathbb{Z}}$ is an ergodic process satisfying Assumptions 2-(a) and 2-(b) [21]. We shall focus on Assumptions 2-(c) and 2-(d).

Proposition 3 (AR-model)**.** For $N > K$, assume $(\mathsf{F}_n, \mathsf{G}_n)$ is the multidimensional ergodic AR process defined by the recursion

$$\begin{bmatrix} \mathsf{F}_n \\ \mathsf{G}_n \end{bmatrix} = A \begin{bmatrix} \mathsf{F}_{n-1} \\ \mathsf{G}_{n-1} \end{bmatrix} + \begin{bmatrix} U_n \\ V_n \end{bmatrix},$$

where $A \in \mathbb{C}^{2N \times 2N}$ is a deterministic matrix whose eigenvalue spectrum belongs to the open unit disk, and where $(U_n, V_n)_{n \in \mathbb{Z}}$ is an i.i.d. process on E such that $\mathbb{E}\|U_0\|^2 + \mathbb{E}\|V_0\|^2 < \infty$. If the entries of the matrix $\begin{bmatrix} U_n & V_n \end{bmatrix}$ are independent with their distributions being absolutely continuous with respect to the Lebesgue measure on $\mathbb{C}$, then Assumption 2-(d) is verified. If, furthermore, the densities of the elements of U_n and V_n are bounded, then, Assumption 2-(c) is verified.

Our second example is a particular multi-antenna version of the AR channel model of Example 1. This model is general enough to capture the Doppler effect, the correlations within each bin of the channel impulse response, as well as the power profile of these bins.

Proposition 4 (Multipath single antenna fading channel)**.** Given three positive integers L, R, and T such that $R > T$, let $(C_n)_{n \in \mathbb{Z}}$ be the $\mathbb{C}^{(L+1)R \times T}$-valued random process described by the iterative model

$$C_n = \begin{bmatrix} H_0 & & \\ & \ddots & \\ & & H_L \end{bmatrix} C_{n-1} + U_n, \tag{15}$$

where the $\{H_\ell\}_{\ell=0}^L$ are deterministic $R \times R$ matrices whose spectra lie in the open unit disk, and where $(U_n)_{n \in \mathbb{Z}}$ is an i.i.d. matrix process such that $\mathbb{E}\|U_0\|^2 < \infty$. Let F_n and G_n be the $LR \times LT$ matrices defined as in (3) with $C_n = \begin{bmatrix} c_{n,0}^\mathsf{T} & \cdots & c_{n,L}^\mathsf{T} \end{bmatrix}^\mathsf{T}$, the $c_{n,\ell}$'s being $R \times T$ matrices. If the entries of U_n are independent with their distributions being absolutely continuous with respect to the Lebesgue measure on $\mathbb{C}$, then Assumption 2-(d) is verified on the Markov process $(\mathsf{F}_n, \mathsf{G}_n)_{n \in \mathbb{Z}}$. If, furthermore, the densities of the elements of U_n are bounded, then, Assumption 2-(c) is verified.

Propositions 3 and 4 are proven in Section 4.4.

Organisation of the paper. In Section 2 we illustrate our results with numerical experiments, where we also discuss the large dimensional regime. The next sections are devoted to the proofs of our results.

Notations. will start from zero. Given a matrix A, the notations $\mathrm{rank}(A)$, $\mathrm{span}(A)$, Π_A and $\Pi_A^\perp$ refer respectively to the rank of A, its column subspace, the orthogonal projector on $\mathrm{span}(A)$, and the orthogonal projector on $\mathrm{span}(A)^\perp$. We denote as $\lambda_{\min}(A)$ the smallest eigenvalue of A when A is a Hermitian matrix. The norm $\|\cdot\|$ is the operator norm for matrices and the Euclidean norm for vectors. We write $A \geq B$ if $A - B \in \mathcal{H}_K^+$. The spaces $\mathcal{M}(S)$ and $C_\mathrm{b}(S)$ stand respectively for the space of Borel probability measures and the space of bounded continuous functions with support in a set S.

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2 Numerical illustrations

We consider here a multiple antenna version of the multipath channel desribed in the introduction, see (2)–(3). We assume the channel coefficient matrices $c_{n,\ell}$ satisfy the AR model $c_{n,\ell} = \alpha c_{n-1,\ell} + \sqrt{1-\alpha^2} a_\ell u_{n,\ell}$. Here the AR coefficient α takes the form $\alpha = \exp(-f_\mathrm{d})$, where f_d represents the Doppler frequency; For $n \in \mathbb{Z}$ and $\ell \in \{0, \ldots L\}$, the $u_{n,\ell}$'s are i.i.d. $R \times T$ random matrices with i.i.d $\mathcal{CN}(0, T^{-1})$ entries; the real vector $a = [a_0, \ldots, a_L]$ is a multipath amplitude profile vector such that $\|a\| = 1$; As is well known, the vector $[a_0^2, \ldots, a_L^2]$ represents the so called multipath variance profile.

Illustration of Theorem 1. We choose an exponential profile of the form $a_\ell \propto \exp(-0.4\ell)$. We start by comparing the mutual information estimates $\hat{\mathcal{I}}_{m,n}$ of $\mathcal{I}_\rho$ that naturally come with (7), namely by taking empirical averages of

$$\frac{1}{(n-m+1)N} \log\det\left(I + \rho\, \mathsf{H}_{m,n} \mathsf{H}_{m,n}^*\right)$$

for several realizations of $\mathsf{H}_{m,n}$, with those coming with Theorem 1(c), namely

$$\hat{\mathcal{I}}_n^{\mathrm{Th1}} := \frac{1}{nN} \sum_{\ell=0}^{n-1} \log\det\left(I + \rho\, \mathsf{F}_\ell \mathsf{X}_{\ell-1} \mathsf{F}_\ell^*\right) - \log\det \mathsf{X}_\ell$$

where, for any $n \in \mathbb{N}$,

$$\mathsf{X}_n := \left(I + \rho\, \mathsf{G}_n^* \left(I + \rho\, \mathsf{F}_n \mathsf{X}_{n-1} \mathsf{F}_n^*\right)^{-1} \mathsf{G}_n\right)^{-1}, \qquad \mathsf{X}_{-1} := I.$$

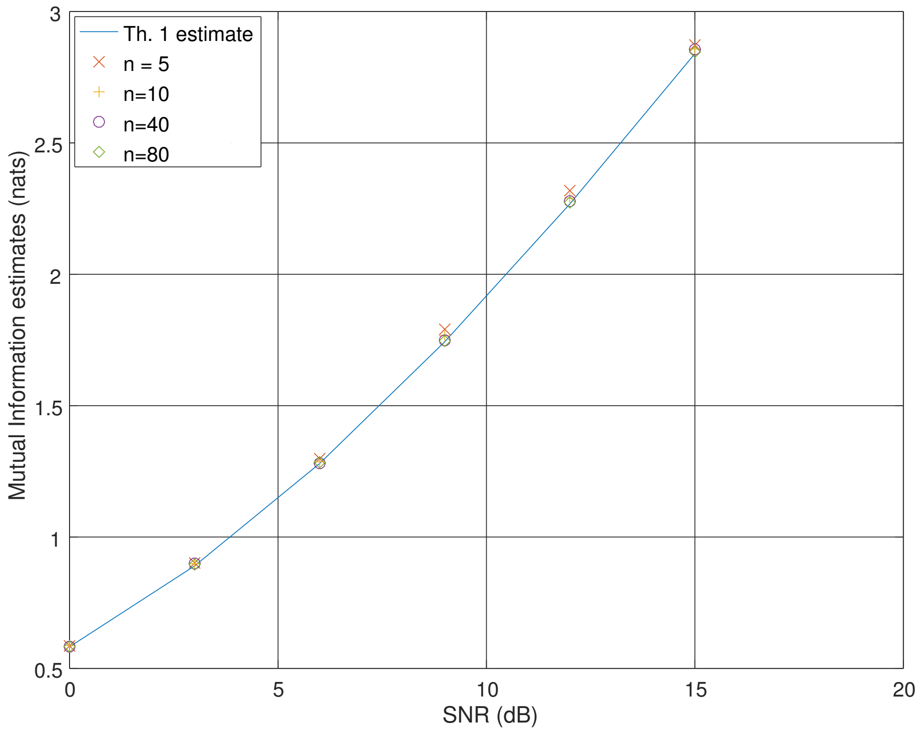

Figure 1: Plots of $\hat{\mathcal{I}}_{1,n}$ and $\hat{\mathcal{I}}^{\mathrm{Th1}}_{12000}$ w.r.t. the SNR and n. Setting: $R = T = 2$, $L = 3$, $f_{\mathrm{d}} = 0.05$. Each empirical average $\hat{\mathcal{I}}_{1,n}$ comes from 150 channel realizations.

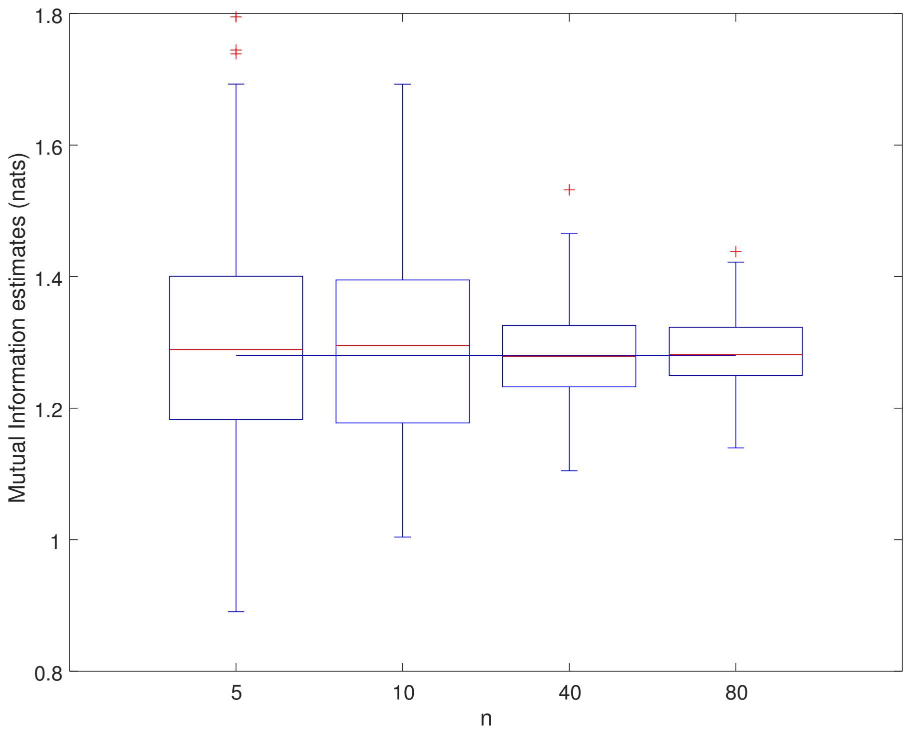

Figure 2: Boxplots of $\hat{\mathcal{I}}_{1,n}$ w.r.t. n. Same setting as for Fig. 1 with $\rho = 6$ dB. The continuous horizontal line represents $\hat{\mathcal{I}}^{\mathrm{Th1}}_{12000}$.

Figure 1 shows that the estimates of $\mathcal{I}_\rho$ obtained by doing empirical averages $\mathcal{I}_{1,n}$ are not affected by important biases. However, Figure 2 shows that the dispersion parameters associated with these estimates are still important for n as large as 80. We note that in the setting of this figure, the matrix $H_{1,n}H_{1,n}^* \in \mathbb{C}^{nRL\times nRL}$ is a 480×480 matrix when $n = 80$. On the other hand, the mutual information estimates $\hat{\mathcal{I}}_n^{\mathrm{Th1}}$ provided by Theorem 1 require much less numerical computations since they involve the inversions of $RL \times RL = 6 \times 6$ matrices.

The large random matrix regime. Next, we consider the asymptotic regime where both N and K converge to infinity in a proportional way. For a large class of processes $(\mathsf{F}_n, \mathsf{G}_n)$ it is known that, in this regime, the Density of States of the operator HH^* converges to a probability measure encountered in the field of large random matrix theory; see [15] for "Wigner analogues" of our model, and [10] for models closer to those of this paper. One important feature of this probability measure is that it depends on the probability law of the channel process only through its first and second order statistics. We illustrate herein this phenomenon on the multipath fading channel model described at the beginning of this section. For the simplicity of the presentation, we assume that the numbers of antennas R and T are fixed and equal (note that $N = K = RL$ in this case) and moreover set the AR coefficient $\alpha = 0$. If we let the channel degree L grow to infinity, then we have the following result.

Proposition 5 (large dimensional regime)**.** In the specific model described above, assume the vector a, which depends on L, satisfies $\|a\| = 1$ for every L, and that there exists a constant $a_{\max} > 0$ independent on L such that

$$\sup_L \max_{\ell\in\{0,\ldots,L\}} \sqrt{L}|a_\ell| \leq a_{\max}.$$

Then,

$$\lim_{L\to\infty} \mathcal{I}_\rho = 2\log\frac{\sqrt{4\rho+1}+1}{2} - \frac{2\rho+1-\sqrt{4\rho+1}}{2\rho}\,. \tag{16}$$

To prove this proposition, we shall show that $\mathcal{I}_\rho$ converges as $L \to \infty$ to $\int \log(1+\rho\lambda)\,\mu_{\mathrm{MP}}(d\lambda)$, where $\mu_{\mathrm{MP}}(d\lambda) = (2\pi)^{-1}\sqrt{4/\lambda - 1}\mathbb{1}_{[0,4]}(\lambda)\,d\lambda$. This is the element of the family of the celebrated Marchenko-Pastur distributions which is the limiting spectral measure of XX^* when X is a square random matrix with iid elements. We provide a proof in Section 5 which is based on Theorem 1. More sophisticated channel models can be considered, including non centered models or models with correlations along the time index n, and for which one can prove similar asymptotics, see [10].

We illustrate this result on an example, represented in Figure 3. As an instance of the statistical channel model used in the statement of Proposition 5, we assume a generalized Wyner model as described in the introduction of this paper, and we consider the regime where the network of Base Stations becomes denser and denser. By densifying the network, the number of users occupying a frequency slot will grow linearly with the number of BS. The number of interferers will grow as well. Yet, provided the BS are connected through a high rate backbone to a central processing unit which is able to perform a joint processing, the overall network capacity will grow linearly. To be more specific, we assume that the channel power gain when the mobile is at the distance d to the BS is

$$\frac{1}{10+(10d/D)^3}\mathbb{1}_{[-D/2,D/2]}(d),$$

where $D > 0$ is a parameter that has the dimension of a distance. If the BS are regularly spaced and that there are L Base Stations per D units of distance, then one channel model approaching this power decay behavior is the setting where the a_ℓ's are given by

$$a_\ell^2 \propto \frac{1}{10+|10(\ell-L/2)/L|^3}, \quad \ell \in \{0,\ldots,L\}.$$

The quantity $R\times\lim_{L\to\infty}\mathcal{I}_\rho$, where the limit is given by Proposition 5, thus represents the ergodic mutual information per mobile user. Figure 3 shows that the predictions of Proposition 5 fit with the values provided by Theorem 1 for L as small as one.

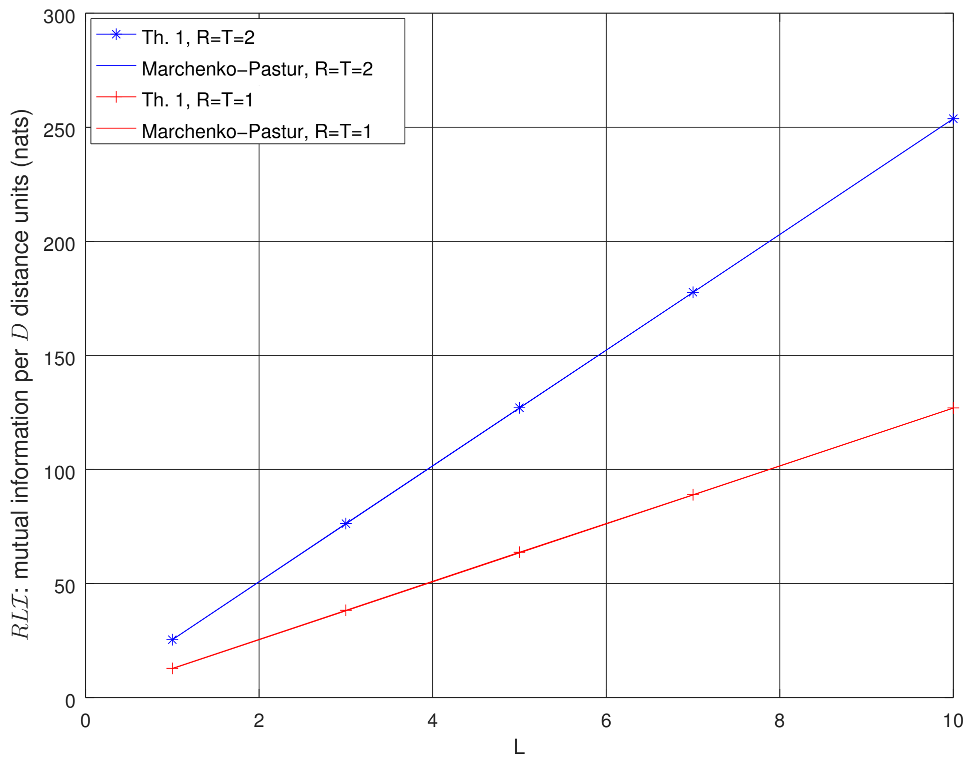

Figure 3: Aggregated mutual information *vs* density of the BS. Setting: $\rho = 6$dB.

Illustration of Theorem 2. Finally, we illustrate the asymptotic behavior of $\mathcal{I}_\rho$ in the high SNR regime as predicted by Theorem 2. In this experiment, we consider a more general model than the one described above where we replace the centered channel coefficient matrix $c_{n,\ell}$ of the model by

$$\sqrt{\frac{K_\mathrm{R}}{K_\mathrm{R}+1}} d_{n,\ell} + \sqrt{\frac{1}{K_\mathrm{R}+1}} c_{n,\ell},$$

where $d_{n,\ell} := [d_{n,\ell}(r,t)]_{r,t=0}^{R-1,T-1}$ is a determistic matrix with entries

$$d_{n,\ell}(r,t) = a_\ell \exp(2\imath\pi(r-t)\sin(\pi\ell/L)),$$

and where the nonnegative number K_R plays the role of the so-called Rice factor. We take again $a_\ell \propto \exp(-0.4\ell)$ and $\alpha = \exp(-f_\mathrm{d})$ as in the first paragraph of the section. The high SNR behavior of $\mathcal{I}_\rho$ is illustrated by Figure 4.

Keeping the same channel model, the behavior of κ_∞ in terms of the Doppler frequency f_d and the Rice factor is illustrated by Figure 5. This figure shows that the impact of f_d is marginal. Regarding K_R, the channel randomness has a beneficial effect on the mutual information for our model, assuming of course that the channel is perfectly known at the receiver.

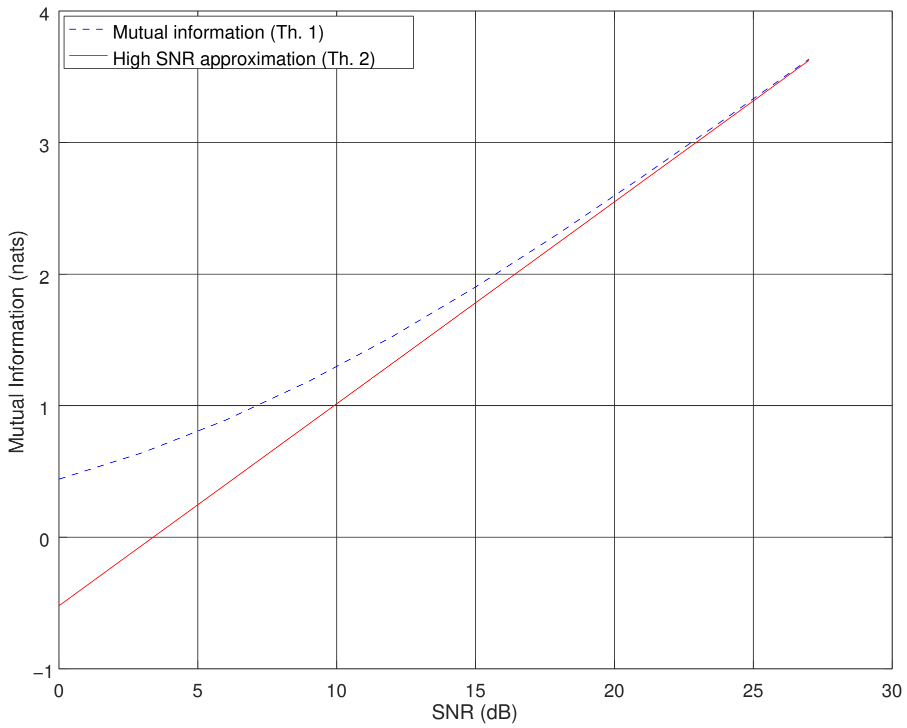

Figure 4: High SNR behavior of $\mathcal{I}_\rho$. Setting: $R = 3$, $T = 2$, $L = 3$, $f_\mathrm{d} = 0.05$, $K_\mathrm{R} = 10$.

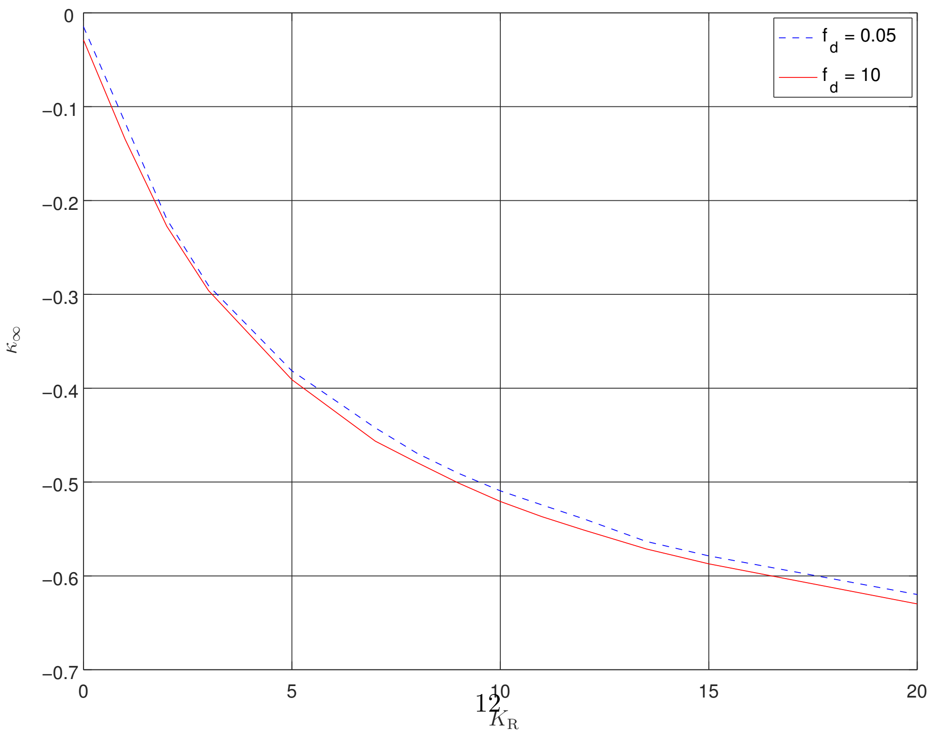

Figure 5: Behavior of κ_∞ w.r.t. f_d and K_R. Setting: $R = 3$, $T = 2$, $L = 3$.

3 Proofs of Theorem 1

In this section, we let Assumption 1 hold true.

3.1 Preparation

The idea behind the proof of Theorem 1 is to show that $\mathcal{I}_\rho$ can be given an expression that involves the resolvents of infinite block-Jacobi matrices and to manipulate these resolvents to obtain the recursion formula for W_n. We denote for any $m, n \in \mathbb{Z} \cup \{\pm\infty\}$ by $\mathsf{H}_{m,n}$ the operator on $\ell^2 := \ell^2(\mathbb{Z}, \mathbb{C})$ defined as the truncation of H, defined in (11), having the bi-infinite matrix representation

$$\mathsf{H}_{m,n} = \begin{bmatrix} \mathsf{F}_m & \mathsf{G}_m & & & \\ & \mathsf{F}_{m+1} & \mathsf{G}_{m+1} & & \\ & & \ddots & \ddots & \\ & & & \mathsf{F}_n & \mathsf{G}_n \end{bmatrix} \tag{17}$$

where the remaining entries are set to zero. Recalling the definition of the random matrix $\mathsf{H}_{m,n}$ already provided in (6) for finite $m, n \in \mathbb{Z}$, we thus identify this matrix with the associated finite rank operator acting on ℓ^2 for which we use the same notation.

Let us now introduce a convenient notation: If one considers an operator on ℓ^2 with block-matrix form $\mathsf{A} = \left[\mathsf{A}_{ij}\right]_{i,j\in\mathbb{Z}}$, where the A_{ij}'s are $Q \times Q$ matrices, then $[\mathsf{A}]_{\square_Q}$ stands for the $Q \times Q$ block A_{ii} with largest index $i \in \mathbb{Z}$ such that $\mathsf{A}_{ii} \neq 0$. For the operators of interest in this work, $[\mathsf{A}]_{\square_Q}$ will always be the bottom rightmost non-vanishing $Q \times Q$ block. Of importance in the proof will be the operators of the type $\mathsf{H}_{-\infty,n}$. This operator is closed and densely defined, thus, defining as $\mathsf{H}^*_{-\infty,n}$ is adjoint, the operator $\mathsf{H}^*_{-\infty,n}\mathsf{H}_{-\infty,n}$ is a positive self-adjoint operator [10, Sec. 4],[1, Sec. 46]. Thus, the resolvent $(I + \rho\,\mathsf{H}^*_{-\infty,n}\mathsf{H}_{-\infty,n})^{-1}$ is defined for each $\rho > 0$, and we can set

$$\mathsf{W}_n := [(I + \rho\,\mathsf{H}^*_{-\infty,n}\mathsf{H}_{-\infty,n})^{-1}]_{\square_K}. \tag{18}$$

We shall prove that the sequence (W_n) indeed satisfies the statements of Theorem 1. To do so, we will use in a key fashion the following Schur complement identities:

$$\det \begin{bmatrix} A & B \\ C & D \end{bmatrix} = \det D \times \det(A - BD^{-1}C)\,, \tag{19}$$

$$\begin{bmatrix} A & B \\ C & D \end{bmatrix}^{-1} = \begin{bmatrix} (A - BD^{-1}C)^{-1} & * \\ * & (D - CA^{-1}B)^{-1} \end{bmatrix}, \tag{20}$$

where the $*$'s can be made explicit in terms of A, B, C, D but are not of interest for our purpose.

3.2 Proof of Theorem 1(a)

We first show that W_n defined in (18) indeed satisfies the recursive equations (8), that is we prove the existence part of Theorem 1(a).

3.2.1 Existence

Proof of Theorem 1(a); existence. Introduce the truncation of $\mathsf{H}_{-\infty,n}$ defined by deleting the right-most non-zero column,

$$\widetilde{\mathsf{H}}_{-\infty,n} := \begin{bmatrix} \ddots & \ddots & \\ & \mathsf{F}_{n-1} & \mathsf{G}_{n-1} \\ & & \mathsf{F}_n \end{bmatrix},$$

so that

$$\mathsf{H}_{-\infty,n} = \left[\, \widetilde{\mathsf{H}}_{-\infty,n} \;\middle|\; \begin{matrix} 0 \\ \mathsf{G}_n \end{matrix} \,\right] =: \left[\, \widetilde{\mathsf{H}}_{-\infty,n} \;\middle|\; Q \,\right].$$

Recalling W_n's definition (18), the Schur's complement formula (20) then provides

$$\begin{aligned}\mathsf{W}_n &= \left(\begin{bmatrix} I+\rho\,\widetilde{\mathsf{H}}^*_{-\infty,n}\widetilde{\mathsf{H}}_{-\infty,n} & \rho\,\widetilde{\mathsf{H}}^*_{-\infty,n}Q \\ \rho\,Q^*\widetilde{\mathsf{H}}_{-\infty,n} & I+\rho\,\mathsf{G}_n^*\mathsf{G}_n\end{bmatrix}^{-1}\right)_{\square_K} \\ &\overset{(a)}{=} \left(I+\rho\,\mathsf{G}_n^*\mathsf{G}_n-\rho^2\,Q^*\widetilde{\mathsf{H}}_{-\infty,n}(I+\rho\,\widetilde{\mathsf{H}}^*_{-\infty,n}\widetilde{\mathsf{H}}_{-\infty,n})^{-1}\widetilde{\mathsf{H}}^*_{-\infty,n}Q\right)^{-1} \\ &\overset{(b)}{=} \left(I+\rho\,\mathsf{G}_n^*\widetilde{\mathsf{W}}_n\mathsf{G}_n\right)^{-1} \end{aligned} \tag{21}$$

where we introduced

$$\widetilde{\mathsf{W}}_n := [(I+\rho\,\widetilde{\mathsf{H}}^*_{-\infty,n}\widetilde{\mathsf{H}}_{-\infty,n})^{-1}]_{\square_N}\,.$$

Here the identity $\overset{(a)}{=}$ can be easily checked similarly to its finite dimensional counterpart, and $\overset{(b)}{=}$ is shown in, *e.g.*, [10, Lemma 7.2].

By similarly expressing $\widetilde{\mathsf{H}}_{-\infty,n}$ in terms of $\mathsf{H}_{-\infty,n-1}$ and F_n, the same computation further yields

$$\widetilde{\mathsf{W}}_n = (I+\rho\,\mathsf{F}_n\mathsf{W}_{n-1}\mathsf{F}_n^*)^{-1}$$

and thus we obtain with (21) the identity

$$\mathsf{W}_n = \left(I+\rho\,\mathsf{G}_n^*\,(I+\rho\,\mathsf{F}_n\mathsf{W}_{n-1}\mathsf{F}_n^*)^{-1}\,\mathsf{G}_n\right)^{-1}.$$

□

3.2.2 Uniqueness

Next, we establish the uniqueness of the process $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ satisfying the recursive relations (8) within the class of stationary processes, to complete the proof of Theorem 1(a).

The proof relies on a contraction argument with the distance on $\mathcal{H}_K^{++}$:

$$\mathrm{dist} : \mathcal{H}_K^{++}\times\mathcal{H}_K^{++}\to[0,\infty), \quad (X,Y)\mapsto\left[\mathrm{Tr}\log^2(XY^{-1})\right]^{1/2},$$

which is the geodesic distance associated with the Riemannian metric $g_X(A,B) := \mathrm{Tr}(X^{-1}AX^{-1}B)$ on the convex cone $\mathcal{H}_K^{++}$; we refer e.g. to [4, §1.2] or [19, §3] for further information. Convergence in dist is equivalent to convergence in the Euclidean norm. It has the following invariance properties: for any $X,Y\in\mathcal{H}_K^{++}$ and any $K\times K$ complex invertible matrix A,

$$\mathrm{dist}(X,Y) = \mathrm{dist}(AXA^*,AYA^*), \qquad \mathrm{dist}(X,Y) = \mathrm{dist}(X^{-1},Y^{-1})\,.$$

Moreover, for any $S\in\mathcal{H}_K^{+}$, we have according to [4, Prop. 1.6],

$$\mathrm{dist}(X+S,Y+S) \le \frac{\max(\|X\|,\|Y\|)}{\max(\|X\|,\|Y\|)+\lambda_{\min}(S)}\,\mathrm{dist}(X,Y)\,,$$

where $\lambda_{\min}(S)$ stands for the smallest eigenvalue of S. Combined with the invariance under conjugation of dist, and using that the $K\times K$ invertible matrices form a dense subset of the space of $K\times K$ complex matrices, this yields the following inequality which will be the key to prove the uniqueness of the process.

Lemma 6. Let $X,Y,S\in\mathcal{H}_K^{++}$ and let A be any $K\times K$ complex matrix. We have,

$$\mathrm{dist}(AXA^*+S,AYA^*+S) \le \frac{\|A\|^2\max(\|X\|,\|Y\|)}{\|A\|^2\max(\|X\|,\|Y\|)+\lambda_{\min}(S)}\,\mathrm{dist}(X,Y)\,.$$

Proof of Theorem 1(a); uniqueness. To prove the uniqueness, we assume that $N \geq K$ for simplicity, since the case $N < K$ can be treated in a similar manner. If one introduces, for any $F, G \in \mathbb{C}^{N\times K}$, the mapping $\psi_{F,G} : \mathcal{H}_K^{++} \to \mathcal{H}_K^{++}$ defined by

$$\psi_{F,G}(W) := \Big(I + \rho\, G^* \left(I + \rho\, F W F^*\right)^{-1} G\Big)^{-1} \tag{22}$$

then (8) reads $\mathsf{W}_n = \psi_{\mathsf{F}_n,\mathsf{G}_n}(\mathsf{W}_{n-1})$.

Next, split F as $F = UT$ where $U \in \mathbb{C}^{N\times K}$ is an isometry ($U^*U = I_K$) and $T \in \mathbb{C}^{K\times K}$. One can complete the columns of U so as to obtain a $N \times N$ unitary matrix $[U\, U^\perp]$, where $U^\perp \in \mathbb{C}^{N\times(N-K)}$. This yields

$$\begin{aligned}\psi_{F,G}(W) &= \Big(I + \rho\, G^* \left(I + \rho\, UTWT^*U^*\right)^{-1} G\Big)^{-1} \\ &= \Big(I + \rho\, G^*U \left(I + \rho\, TWT^*\right)^{-1} U^*G + \rho\, G^*U^\perp(U^\perp)^*G\Big)^{-1} \\ &= \iota \circ \tau_{\sqrt{\rho}\, G^*U; I+\rho\, G^*U^\perp(U^\perp)^*G} \circ \iota \circ \tau_{\sqrt{\rho}\, T; I}(W)\ ,\end{aligned} \tag{23}$$

where we introduced

$$\tau_{A;S}(X) := AXA^* + S \quad \text{and} \quad \iota(X) := X^{-1}\,.$$

Using Lemma 6 together with the invariance of dist with respect to the inversion, we obtain for any $W, W' \in \mathcal{H}_K^{++}$,

$$\begin{aligned}&\operatorname{dist}\big(\psi_{F,G}(W), \psi_{F,G}(W')\big) \\ &\leq \frac{\rho\|G\|^2}{\rho\|G\|^2 + \lambda_{\min}(I + \rho\, G^*U^\perp(U^\perp)^*G)} \frac{\|F\|^2 \max(\|W\|, \|W'\|)}{\|F\|^2 \max(\|W\|, \|W'\|) + 1} \operatorname{dist}(W, W') \\ &\leq \frac{\rho\|G\|^2}{\rho\|G\|^2 + 1} \operatorname{dist}(W, W'),\end{aligned} \tag{24}$$

where for the first inequality we used that $\|(I + \rho TWT^*)^{-1}\| \leq 1$ for any $W \in \mathcal{H}_K^+$.

Now, let $(\mathsf{W}'_n)_{n\in\mathbb{Z}}$ be any stationary process on $\mathcal{H}_K^{++}$ satisfying $\mathsf{W}'_n = \psi_{\mathsf{F}_n,\mathsf{G}_n}(\mathsf{W}'_{n-1})$ a.s. for every $n \in \mathbb{Z}$. If we let $n \geq 0$, then we have from (24) a.s. that

$$\operatorname{dist}(\mathsf{W}_n, \mathsf{W}'_n) \leq \frac{\rho\|\mathsf{G}_n\|^2}{\rho\|\mathsf{G}_n\|^2 + 1} \operatorname{dist}(\mathsf{W}_{n-1}, \mathsf{W}'_{n-1})$$

and, iterating, we obtain

$$\operatorname{dist}(\mathsf{W}_n, \mathsf{W}'_n) \leq \Big(\prod_{i=1}^n \xi_i\Big) \operatorname{dist}(\mathsf{W}_0, \mathsf{W}'_0), \qquad \xi_i := \frac{\rho\|\mathsf{G}_i\|^2}{\rho\|\mathsf{G}_i\|^2 + 1}.$$

By the ergodicity of $(\mathsf{G}_n)_{n\in\mathbb{Z}}$, we have

$$\frac{1}{n} \sum_{i=1}^n \log \xi_i \xrightarrow[n\to\infty]{\text{a.s.}} \mathbb{E} \log \xi_0 < 0$$

and thus we have proven that $\operatorname{dist}(\mathsf{W}_n, \mathsf{W}'_n) \to 0$ a.s. as $n \to \infty$. Finally, since

$$(\mathsf{W}_{n+m_1}, \ldots, \mathsf{W}_{n+m_M}) \overset{\text{law}}{=} (\mathsf{W}_{m_1}, \ldots, \mathsf{W}_{m_M})$$

for any $M-$uple of integers $(m_1, \ldots, m_M)$ and similarly for W'_n, by letting $n \to \infty$ this yields that the finite-dimensional distributions of the two stationary processes $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ and $(\mathsf{W}'_n)_{n\in\mathbb{Z}}$ are the same, and consequently these two processes have the same distribution. □

3.3 Proof of Theorem 1(b)

We start with the following lemma.

Lemma 7. For any fixed $n \in \mathbb{Z}$ and $\rho > 0$, we have

$$[(I + \rho\, \mathsf{H}_{m,n}^* \mathsf{H}_{m,n})^{-1}]_{\square_K} \xrightarrow[m\to-\infty]{} \mathsf{W}_n. \tag{25}$$

Proof. Denote by $\mathcal{K} \subset \ell^2$ the subspace of sequences with finite support. Clearly, for any fixed $n \in \mathbb{Z}$ and fixed event $\omega \in \Omega$, we have for all $x \in \mathcal{K}$,

$$\mathsf{H}_{m,n}^* \mathsf{H}_{m,n} x \xrightarrow[m\to-\infty]{} \mathsf{H}_{-\infty,n}^* \mathsf{H}_{-\infty,n} x\,,$$

where $\to$ denotes the strong convergence in ℓ^2. Now $\mathcal{K}$ is a common core for the set of operators $\{\mathsf{H}_{m,n}^* \mathsf{H}_{m,n} : m \in \{n, n-1, n-2, \ldots\}\}$ and $\mathsf{H}_{-\infty,n}^* \mathsf{H}_{-\infty,n}$, see e.g. [14, §III.5.3] or [23, Chap. VIII] for this notion. As a consequence, the convergence also holds in the strong resolvent sense, see [23, §VIII], and thus for every $x \in \ell^2$ and $\rho > 0$,

$$(I + \rho\, \mathsf{H}_{m,n}^* \mathsf{H}_{m,n})^{-1} x \xrightarrow[n\to\infty]{} (I + \rho\, \mathsf{H}_{-\infty,n}^* \mathsf{H}_{-\infty,n})^{-1} x \quad \text{a.s.}$$

from which (25) follows by definition (18) of W_n. □

Proof of Theorem 1(b). We start by writing

$$\mathsf{H}_{m,n} = \left[\begin{array}{ccc|c} & \mathsf{H}_{m,n-1} & & 0 \\ \hline 0 & \cdots & 0 \;\; \mathsf{F}_n & \mathsf{G}_n \end{array}\right] = \left[\begin{array}{c|c} \mathsf{H}_{m,n-1} & 0 \\ \hline P & \mathsf{G}_n \end{array}\right]$$

with $P := [\,0 \;\cdots\; 0\;\mathsf{F}_n]$, and use Schur's complement formula (19) to obtain,

$$\begin{aligned}
&\log\det(I + \rho\, \mathsf{H}_{m,n} \mathsf{H}_{m,n}^*) \\
&= \log\det \begin{bmatrix} I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^* & \rho\, \mathsf{H}_{n,m-1}^* P^* \\ \rho\, P \mathsf{H}_{m,n-1}^* & I + \rho\, \mathsf{F}_n \mathsf{F}_n^* + \rho\, \mathsf{G}_n \mathsf{G}_n^* \end{bmatrix} \\
&= \log\det(I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^*) \\
&\quad + \log\det\big(I + \rho\, \mathsf{F}_n \mathsf{F}_n^* + \rho\, \mathsf{G}_n \mathsf{G}_n^* - \rho^2\, P \mathsf{H}_{m,n-1}^* (I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^*)^{-1} \mathsf{H}_{n,m-1} P^*\big) \\
&= \log\det(I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^*) \\
&\quad + \log\det\big(I + \rho\, \mathsf{F}_n \mathsf{F}_n^* + \rho\, \mathsf{G}_n \mathsf{G}_n^* + \rho\, P[(I + \rho\, \mathsf{H}_{m,n-1}^* \mathsf{H}_{m,n-1})^{-1} - I] P^*\big) \\
&= \log\det(I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^*) \\
&\quad + \log\det\big(I + \rho\, \mathsf{F}_n \mathsf{F}_n^* + \rho\, \mathsf{G}_n \mathsf{G}_n^* + \rho\, \mathsf{F}_n [(I + \rho\, \mathsf{H}_{m,n-1}^* \mathsf{H}_{m,n-1})^{-1} - I]_{\square K} \mathsf{F}_n^*\big) \\
&= \log\det(I + \rho\, \mathsf{H}_{m,n-1} \mathsf{H}_{m,n-1}^*) \\
&\quad + \log\det\big(I + \rho\, \mathsf{G}_n \mathsf{G}_n^* + \rho\, \mathsf{F}_n [(I + \rho\, \mathsf{H}_{m,n-1}^* \mathsf{H}_{m,n-1})^{-1}]_{\square_K} \mathsf{F}_n^*\big)\,.
\end{aligned}$$

By iterating this manipulation after replacing $\mathsf{H}_{m,n-i}$ by $\mathsf{H}_{m,n-i-1}$ at the i^{th} step, if we set

$$\xi_{m,i} := \log\det\big(I + \rho\, \mathsf{G}_i \mathsf{G}_i^* + \rho\, \mathsf{F}_i [(I + \rho\, \mathsf{H}_{m,i-1}^* \mathsf{H}_{m,i-1})^{-1}]_{\square_K} \mathsf{F}_i^*\big)$$

for any $m \le i \le n$ with the convention that $\mathsf{H}_{m,m-1} := 0$, we have

$$\log\det\big(I + \rho\, \mathsf{H}_{m,n} \mathsf{H}_{m,n}^*\big) = \sum_{i=m}^{n} \xi_{m,i}\,. \tag{26}$$

Next, Lemma 7 yields

$$\xi_{m,i} \xrightarrow[m\to-\infty]{} \log\det\left(I + \rho\, \mathsf{G}_i \mathsf{G}_i^* + \rho\, \mathsf{F}_i \mathsf{W}_{i-1} \mathsf{F}_i^*\right). \tag{27}$$

Since $\|[(I+\rho\,\mathsf{H}_{m,i-1}^*\mathsf{H}_{m,i-1})^{-1}]_{\square_K}\| \leq 1$, we have $\xi_{m,i} \leq N\log(1+\rho\|\mathsf{F}_i\|^2+\rho\|\mathsf{G}_i\|^2)$. Thus, by the moment assumption (4), we obtain from (27) and dominated convergence that

$$\begin{aligned}\mathbb{E}\xi_{m,i} \xrightarrow[m\to-\infty]{} &\mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_i\mathsf{G}_i^*+\rho\,\mathsf{F}_i\mathsf{W}_{i-1}\mathsf{F}_i^*\right)\\ &=\mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_0\mathsf{G}_0^*+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right),\end{aligned}$$

where the equality follows from the stationarity of the process $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$. The stationarity further provides that $\mathbb{E}\xi_{m,i}$ only depends on $i-m$ and thus, for any fixed n, we obtain by Cesàro summation that

$$N\mathcal{I}_\rho = \lim_{m\to-\infty}\frac{1}{(n-m+1)}\sum_{i=m}^{n}\mathbb{E}\xi_{m,i} = \mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_0\mathsf{G}_0^*+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right). \quad (28)$$

By taking $n=0$ in the recursive relation (8), we moreover see that

$$\begin{aligned}N\mathcal{I}_\rho &= \mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_0\mathsf{G}_0^*+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right)\\ &=\mathbb{E}\log\det\left(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right)+\mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_0\mathsf{G}_0^*(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*)^{-1}\right)\\ &=\mathbb{E}\log\det\left(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right)+\mathbb{E}\log\det\left(I+\rho\,\mathsf{G}_0^*(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*)^{-1}\mathsf{G}_0\right)\\ &=\mathbb{E}\log\det\left(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right)-\mathbb{E}\log\det\mathsf{W}_0, \qquad (29)\end{aligned}$$

which proves (9). □

3.4 Proof of Theorem 1(c)

Proof of Theorem 1(c). Since the process $(\mathsf{F}_n,\mathsf{G}_n)_{n\in\mathbb{Z}}$ is assumed to be ergodic, and so does $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ by construction, we have a.s.

$$\begin{aligned}&\lim_{n\to\infty}\frac{1}{n}\sum_{\ell=0}^{n-1}\log\det\left(I+\rho\,\mathsf{F}_\ell\mathsf{W}_{\ell-1}\mathsf{F}_\ell^*\right)-\mathbb{E}\log\det\mathsf{W}_\ell\\ &\qquad=\mathbb{E}\log\det\left(I+\rho\,\mathsf{F}_0\mathsf{W}_{-1}\mathsf{F}_0^*\right)-\mathbb{E}\log\det\mathsf{W}_0=\mathcal{I}_\rho. \quad (30)\end{aligned}$$

Next, for the same reason than and with the same notations as in the proof of the uniqueness of W_n provided in Section 3.2.2, we have $\mathrm{dist}(\mathsf{X}_n,\mathsf{W}_n)\to 0$ a.s. as $n\to\infty$. Thus,

$$\frac{1}{n}\sum_{\ell=0}^{n-1}\log\det\mathsf{X}_\ell-\log\det\mathsf{W}_\ell \xrightarrow[n\to\infty]{\text{a.s.}} 0$$

as a Cesàro average. Since Lemma 6 also yields

$$\mathrm{dist}(I+\rho\,\mathsf{F}_n\mathsf{X}_{n-1}\mathsf{F}_n\,,\,I+\rho\,\mathsf{F}_n\mathsf{W}_{n-1}\mathsf{F}_n)\leq\mathrm{dist}(\mathsf{X}_{n-1},\mathsf{W}_{n-1})\xrightarrow[n\to\infty]{\text{a.s.}} 0,$$

we similarly have

$$\frac{1}{n}\sum_{\ell=0}^{n-1}\log\det(I+\rho\,\mathsf{F}_\ell\mathsf{X}_{\ell-1}\mathsf{F}_\ell^*)-\log\det(I+\rho\,\mathsf{F}_\ell\mathsf{W}_{\ell-1}\mathsf{F}_\ell^*)\xrightarrow[n\to\infty]{\text{a.s.}} 0.$$

and the result follows from this convergence along with (30). □

This completes the proof of Theorem 1.

4 Proof of Theorem 2

Assume from now that $N>K$ and that Assumption 2 holds true.

4.1 Preparation

To obtain an expansion of the type $\mathcal{I}_\rho = (K/N)\log\rho + \kappa_\infty + o(1)$ as $\rho \to \infty$, it is more convenient to work with the new variables:

$$\gamma := \frac{1}{\rho} \in (0,\infty), \qquad \mathsf{Z}_{\gamma,n} := \gamma\,\mathsf{W}_n^{-1}. \tag{31}$$

Indeed, it follows the identity (9) of Theorem 1 and the stationarity of $(\mathsf{W}_n)_{n\in\mathbb{Z}}$ that

$$\begin{aligned} N\mathcal{I}_\rho &= -\mathbb{E}\log\det\mathsf{W}_0 + \mathbb{E}\log\det(I + \rho\,\mathsf{F}_1\mathsf{W}_n\mathsf{F}_1^*) \\ &= K\log\rho + \mathbb{E}\log\det\mathsf{Z}_{\gamma,0} + \mathbb{E}\log\det(I + \mathsf{F}_1\mathsf{Z}_{\gamma,0}^{-1}\mathsf{F}_1^*) \\ &= K\log\rho + \mathbb{E}\log\det\mathsf{Z}_{\gamma,0} + \mathbb{E}\log\det(I + \mathsf{Z}_{\gamma,0}^{-1}\mathsf{F}_1^*\mathsf{F}_1) \\ &= K\log\rho + \mathbb{E}\log\det(\mathsf{Z}_{\gamma,0} + \mathsf{F}_1^*\mathsf{F}_1), \end{aligned} \tag{32}$$

which is the starting point of the asymptotic analysis $\gamma \to 0$. With this expression at hand, we would like to take the limit $\gamma \to 0$ and identify the limit

$$\kappa_\infty := \frac{1}{N}\lim_{\gamma\to 0}\mathbb{E}\log\det(\mathsf{Z}_{\gamma,0} + \mathsf{F}_1^*\mathsf{F}_1). \tag{33}$$

To study this limiting case, we start from the recursive equation (8), which reads for these new variables

$$\mathsf{Z}_{\gamma,n} = \gamma I + \mathsf{G}_n^*(I + \mathsf{F}_n\mathsf{Z}_{\gamma,n-1}^{-1}\mathsf{F}_n^*)^{-1}\mathsf{G}_n = h_{\gamma,\mathsf{F}_n,\mathsf{G}_n}(\mathsf{Z}_{\gamma,n-1}), \tag{34}$$

where, for any $\gamma \geq 0$ and $F, G \in \mathbb{C}^{N\times K}$, we define $h_{\gamma,F,G} : \mathcal{H}_K^{++} \to \mathcal{H}_K^{+}$ by

$$h_{\gamma,F,G}(Z) := \gamma I + G^*(I + FZ^{-1}F^*)^{-1}G. \tag{35}$$

Note that if $\gamma > 0$ then $h_{\gamma,F,G}(Z) \in \mathcal{H}_K^{++}$. The same holds true when $\gamma = 0$, which is now allowed, as soon as G has full rank. We now observe that one can extend this mapping to the whole of $\mathcal{H}_K^+$, provided F has full rank.

4.1.1 Extension of the mapping $h_{\gamma,F,G}$ to $\mathcal{H}_K^+$

Assume that $F \in \mathbb{C}^{N\times K}$ has full rank, namely $\mathrm{rank}(F) = K$. By setting $T := (F^*F)^{1/2}$ and $U := F(F^*F)^{-1/2}$, we have the polar decomposition $F = UT$ where $U \in \mathbb{C}^{N\times K}$ is an isometry matrix and $T \in \mathcal{H}_K^{++}$. By completing U so as to obtain a $N \times N$ unitary matrix $\begin{bmatrix} U & U^\perp \end{bmatrix}$ and setting $\Pi_F^\perp := U^\perp(U^\perp)^* = I - F(F^*F)^{-1}F^*$, which the orthogonal projection onto the orthogonal space to the linear span of the columns of F, we can write

$$h_{\gamma,F,G}(Z) \tag{36}$$

$$= \gamma I + G^*(I + FZ^{-1}F^*)^{-1}G$$

$$= \gamma I + G^*U(I + TZ^{-1}T)^{-1}U^*G + G^*\Pi_F^\perp G \tag{37}$$

$$= \gamma I + G^*UT^{-1}Z^{1/2}(I + Z^{1/2}T^{-2}Z^{1/2})^{-1}Z^{1/2}T^{-1}U^*G + G^*\Pi_F^\perp G$$

$$= \gamma I + G^*F(F^*F)^{-1}Z^{1/2}(I + Z^{1/2}(F^*F)^{-1}Z^{1/2})^{-1}Z^{1/2}(F^*F)^{-1}F^*G + G^*\Pi_F^\perp G \tag{38}$$

where for the second equality we used the matrix identity $(I + AB)^{-1} = B^{-1}(I + A^{-1}B^{-1})^{-1}A^{-1}$ with $A := TZ^{-1/2}$ and $B := Z^{-1/2}T$ for any $Z^{1/2} \in \mathcal{H}_K^+$ satisfying $(Z^{1/2})^2 = Z$. Note that the alternative expression (38) for $h_{\gamma,F,G}(Z)$ does now make sense when $Z \in \mathcal{H}_K^+$ is not invertible, provided that F has full rank. Moreover, since two Hermitian square roots of $Z \in \mathcal{H}_K^+$ are identical up to the multiplication by a unitary matrix, the right hand side of (38) does not depend on the choice for $Z^{1/2}$. In the following, we chose $Z \mapsto Z^{1/2}$ so that it is continuous (for the operator norm). Thus, by taking the right hand side of (38) as the definition of $h_{\gamma,F,G}(Z)$ in this case, we properly extended $h_{\gamma,F,G}$ to a mapping $\mathcal{H}_K^+ \to \mathcal{H}_K^+$ which is continuous, and that we continue to denote by $h_{\gamma,F,G}$. An important property of $h_{0,F,G}$ we use in what follows is:

Lemma 8. If F has full rank, then $h_{0,F,G} : \mathcal{H}_K^+ \to \mathcal{H}_K^+$ is non-decreasing.

Proof. It is clear from (35) this mapping is non-decreasing on $\mathcal{H}_K^{++}$ and this property extends to $\mathcal{H}_K^+$ since one can write $h_{0,F,G}(Z) = \lim_{\varepsilon\to 0} h_{0,F,G}(Z + \varepsilon I)$ by continuity of $h_{0,F,G}$. $\square$

4.1.2 The Markov kernel Q_γ

Equipped with the extended definition of $h_{\gamma,F,G}$ to $\mathcal{H}_K^+$, let us consider for any $\gamma \geq 0$ the Markov transition kernel $Q_\gamma : (E \times \mathcal{H}_K^+) \times \mathscr{B}(E \times \mathcal{H}_K^+) \to [0, 1]$ defined by

$$Q_\gamma f(\mathsf{F}, \mathsf{G}, \mathsf{Z}) := \int f\big(F, G, h_{\gamma,F,G}(\mathsf{Z})\big) P\big((\mathsf{F}, \mathsf{G}), dF \times dG\big)$$

for any $(\mathsf{F}, \mathsf{G}) \in E$, any $\mathsf{Z} \in \mathcal{H}_K^+$ and any Borel test function $f : E \times \mathcal{H}_K^+ \to [0, \infty)$.

Remark 8. In the following, we will use at several instances the following fact: Since $\theta = \theta P$, Assumption 2(d) yields that $G^* F$ is non-singular, and thus that both F and G have full rank, $P((\mathsf{F}, \mathsf{G}), \cdot)$-a.s. for θ-a.e. (F, G). In particular, $Q_0 f(\mathsf{F}, \mathsf{G}, \mathsf{Z})$ is properly defined for θ-a.e. (F, G), which will be enough for our purpose.

When $\gamma > 0$, if $(\mathsf{F}_n, \mathsf{G}_n, \mathsf{Z}_{\gamma,n})_{n\in\mathbb{Z}}$ denotes the Markov process defined by $\mathsf{Z}_{\gamma,n} = h_{\gamma,\mathsf{F}_n,\mathsf{G}_n}(\mathsf{Z}_{\gamma,n-1})$ with $(\mathsf{F}_n, \mathsf{G}_n)_{n\in\mathbb{Z}}$ the Markov process with transition kernel P, then by the definition of $\mathsf{Z}_{\gamma,n}$ in (31) and by Theorem 1, it follows that Q_γ has a unique invariant measure, that we denote by π_γ. The strategy of the proof of Theorem 2 is to show that Q_0 has also a unique invariant measure π_0, which will yield the existence of the process $\mathsf{Z}_n := \mathsf{Z}_{0,n}$, and we also show that $\pi_\gamma \to \pi_0$ narrowly as $\gamma \to 0$ and that one can legally take the limit $\gamma \to 0$ in (33), so as to obtain $N\mathcal{I}_\rho + K \log\gamma \to \mathbb{E}\det(\mathsf{Z}_0 + \mathsf{F}_1^*\mathsf{F}_1)$. It turns out when $N = K$ one can possibly lose the uniqueness of the invariant measure for Q_0, which makes this setting out of reach for our current approach.

4.2 Existence of a unique invariant mesure π_0 for Q_0

The key to prove the existence of an invariant measure for Q_0 is the following result.

Lemma 9. The family of probability measures on $\mathcal{H}_K^+$,

$$\mathscr{C} := \big\{\zeta Q_0^n(E \times \cdot) \,:\, \zeta \in \mathcal{M}(E \times \mathcal{H}_K^+),\ \zeta(\cdot \times \mathcal{H}_K^+) = \theta(\cdot),\ n \geq K\big\}. \tag{39}$$

is a tight subset of $\mathcal{M}(\mathcal{H}_K^{++})$.

Proof. Let us fix $\varepsilon > 0$. We first prove there exists $\eta > 0$ such that, for any $\xi \in \mathscr{C}$,

$$\xi(\lambda_{\min}(Z) \geq \eta) \geq 1 - \varepsilon, \tag{40}$$

where $\lambda_{\min}(Z)$ is the smallest eigenvalue of $Z \in \mathcal{H}_K^+$. To do so, observe from (35) that if $Z \in \mathcal{H}_K^{++}$ then so does $h_{0,F,G}(Z)$ as soon as G has full rank, which is true θ-a.s. due to Assumption 2(d). We claim that this assumption further yields that, that for all $(\mathsf{F}, \mathsf{G}, \mathsf{Z})$ satisfying $\mathrm{rank}(\mathsf{Z}) < K$, we have $Q_0((\mathsf{F}, \mathsf{G}, \mathsf{Z}),\ \mathrm{rank}(Z) > \mathrm{rank}(\mathsf{Z})) = 1$, namely at each step of the process the rank of the random matrix Z increases $Q_0((\mathsf{F}, \mathsf{G}, \mathsf{Z}), \cdot)$-a.s. To prove this, we start from

$$Q_0((\mathsf{F}, \mathsf{G}, \mathsf{Z}), \mathrm{rank}(Z) \leq \mathrm{rank}(\mathsf{Z})) = P((\mathsf{F}, \mathsf{G}),\ \mathrm{rank}(h_{0,F,G}(\mathsf{Z})) \leq \mathrm{rank}(\mathsf{Z})).$$

Recalling (38), we have $\mathrm{rank}(h_{0,F,G}(\mathsf{Z}) - G^*\Pi_F^\perp G) = \mathrm{rank}(\mathsf{Z})$ as soon as F^*G is invertible. Using the general fact that $\mathrm{rank}(A + B) \leq \mathrm{rank}(A)$ implies $\mathrm{span}(B) \subset \mathrm{span}(A)$ any $A, B \in \mathcal{H}_K^+$ and Assumption 2(d), this yields

$$Q_0((\mathsf{F}, \mathsf{G}, \mathsf{Z}), \mathrm{rank}(Z) \leq \mathrm{rank}(\mathsf{Z})) = P\big((\mathsf{F}, \mathsf{G}),\ \mathrm{span}(G^*\Pi_F^\perp G) \subset \mathrm{span}(h_{0,F,G}(\mathsf{Z}) - G^*\Pi_F^\perp G)\big)$$

for θ-a.e. (F,G). Next, we will use repeatedly that, for two matrices A and B we have $\mathrm{span}(A)\subset \mathrm{span}(B)$ if and only if $\mathrm{span}(CAD)\subset\mathrm{span}(CBD)$ for all invertible matrices C and D. If we let $\mathsf{Z}^\perp\in\mathbb{C}^{K\times K}$ be any matrix such that $\mathrm{span}(\mathsf{Z}^\perp)=\mathrm{span}(\mathsf{Z})^\perp$, we have:

$$\begin{aligned}
&\mathrm{span}(G^*\Pi_F^\perp G)\subset\mathrm{span}(h_{0,F,G}(\mathsf{Z})-G^*\Pi_F^\perp G)\\
\Leftrightarrow&\,\mathrm{span}(G^*\Pi_F^\perp G)\subset\mathrm{span}(G^*F(F^*F)^{-1}\mathsf{Z}^{1/2}(I+\mathsf{Z}^{1/2}(F^*F)^{-1}\mathsf{Z}^{1/2})^{-1}\mathsf{Z}^{1/2}(F^*F)^{-1}F^*G)\\
\Leftrightarrow&\,\mathrm{span}(G^*G-G^*F(F^*F)^{-1}F^*G)\\
&\qquad\qquad\subset\mathrm{span}(G^*F(F^*F)^{-1}\mathsf{Z}^{1/2}(I+\mathsf{Z}^{1/2}(F^*F)^{-1}\mathsf{Z}^{1/2})^{-1}\mathsf{Z}^{1/2}(F^*F)^{-1}F^*G)\\
\Leftrightarrow&\,\mathrm{span}(F^*F(G^*F)^{-1}G^*G(F^*G)^{-1}F^*F-F^*F)\subset\mathrm{span}(\mathsf{Z})\\
\Leftrightarrow&\,\mathrm{span}(F^*F(F^*\Pi_GF)^{-1}F^*F-F^*F)\subset\mathrm{span}(\mathsf{Z})\\
\Leftrightarrow&\,F^*F(F^*\Pi_GF)^{-1}F^*F\mathsf{Z}^\perp-F^*F\mathsf{Z}^\perp=0\\
\Leftrightarrow&\,F^*F\mathsf{Z}^\perp=F^*\Pi_GF\mathsf{Z}^\perp\\
\Leftrightarrow&\,F^*\Pi_G^\perp F\mathsf{Z}^\perp=0,
\end{aligned}$$

provided that F and G have full rank. Therefore, together with Assumption 2(d), we obtain

$$Q_0((\mathsf{F},\mathsf{G},\mathsf{Z}),\mathrm{rank}(Z)\le\mathrm{rank}(\mathsf{Z}))=P((\mathsf{F},\mathsf{G}),\ F^*\Pi_G^\perp F\mathsf{Z}^\perp=0)=0,$$

for θ-a.e. (F,G), and our claim follows. As a consequence, Z has full rank $(\theta\otimes\delta_0)Q_0^K((\mathsf{F},\mathsf{G}),\cdot)$-a.s. and thus there exists $\eta>0$ such that

$$(\theta\otimes\delta_0)Q_0^K((\mathsf{F},\mathsf{G}),\ \lambda_{\min}(Z)\ge\eta)\ge1-\varepsilon.$$

Next, we use that $Z\mapsto h_{0,F,G}(Z)$ and $Z\mapsto\lambda_{\min}(Z)$ are non-decreasing on $\mathcal{H}_K^+$, see Lemma 8, so that for any $\zeta\in\mathcal{M}(E\times\mathcal{H}_K^+)$ satisfying $\zeta(\cdot\times\mathcal{H}_K^+)=\theta(\cdot)$ and any $n\ge K$, we have

$$\begin{aligned}
\zeta Q_0^n(\lambda_{\min}(Z)\ge\eta)&\ge(\theta\otimes\delta_0)Q_0^n(\lambda_{\min}(Z)\ge\eta)\\
&=\big((\theta\otimes\delta_0)Q_0^{n-K}(E\times\cdot)\big)Q_0^K(\lambda_{\min}(Z)\ge\eta)\\
&\ge Q_0^K(\lambda_{\min}(Z)\ge\eta)\ge1-\varepsilon,
\end{aligned}$$

which finally proves (40).

Finally, let $C>0$ be such that $\theta(\|G\|^2>C)<\varepsilon$ and consider the compact subset $\mathcal{K}$ of $\mathcal{H}_K^{++}$ given by

$$\mathcal{K}:=\big\{Z\in\mathcal{H}_K^{++}:\ \lambda_{\min}(Z)\ge\eta,\quad\|Z\|\le C\big\}.$$

It follows from (38) that $\|h_{0,F,G}(Z)\|\le\|G\|^2$ for any $(F,G)\in E$ such that F has full rank and any $Z\in\mathcal{H}_K^+$. This provides, for any $\zeta\in\mathcal{M}(E\times\mathcal{H}_K^+)$ satisfying $\zeta(\cdot\times\mathcal{H}_K^+)=\theta(\cdot)$ and any $n\ge K$,

$$\begin{aligned}
\zeta Q_0^n(\|Z\|>C)&\le\zeta Q_0^n(\|G\|^2>C)\\
&=\theta P^n(\|G\|^2>C)\\
&=\theta(\|G\|^2>C)<\varepsilon
\end{aligned}$$

and thus $\xi(\mathcal{K})\ge1-2\varepsilon$ for any $\xi\in\mathscr{C}$. The proof of the lemma is therefore complete. □

Lemma 10. For any $\gamma\ge0$ the kernel Q_γ maps $C_{\mathrm{b}}(E\times\mathcal{H}_K^{++})$ to itself.

Proof. Let $f:E\times\mathcal{H}_K^{++}\to\mathbb{R}$ be a bounded and continuous function, and note from the definition of Q_γ that $Q_\gamma f$ is clearly bounded. To show it is continuous, let $(\mathsf{F}_k,\mathsf{G}_k,\mathsf{Z}_k)_{k\ge1}$ be a sequence converging to $(\mathsf{F}_0,\mathsf{G}_0,\mathsf{Z}_0)$ in $E\times\mathcal{H}_K^{++}$ as $k\to\infty$. If we set $g_k(F,G):=f(F,G,h_{\gamma,F,G}(\mathsf{Z}_k))$ and $\mu_k(\cdot):=P((\mathsf{F}_k,\mathsf{G}_k),\cdot)$, then this amounts to show that $\int g_k\,d\mu_k\to\int g_0\,d\mu_0$ as $k\to\infty$. Since P is Feller by Assumption 2(a), we have the narrow convergence $\mu_k\to\mu_0$. Since $(F,G)\mapsto h_{\gamma,F,G}(Z)$ is continuous on E for any $Z\in\mathcal{H}_K^{++}$ we have $g_0\in C_{\mathrm{b}}(\mathcal{H}_K^{++})$ and that $g_k\to g_0$ locally uniformly on E. Together with the tightness of (μ_k) and that $\sup_{k\in\mathbb{N}}\|g_k\|_\infty<\infty$, we obtain $\int g_k\,d\mu_k\to\int g_0\,d\mu_0$ and the proof of the lemma is complete. □

Corollary 11. Q_0 has an invariant measure in $\mathcal{M}(E\times\mathcal{H}_K^{++})$.

Proof. Let $\zeta := \theta\otimes\delta_0$ so that by Lemma 9 we have $\zeta Q_0^n \in \mathcal{M}(E\times\mathcal{H}_K^{++})$ for every $n\geq K$ and $\zeta Q_0^n \to \pi$ narrowly as $n\to\infty$ for some $\pi\in\mathcal{M}(E\times\mathcal{H}_K^{++})$, possibly up to the extraction of a subsequence. If we set, for any $n>K$,

$$\zeta\bar{Q}_{0,n} := \frac{1}{n-K}\sum_{\ell=K}^{n-1}\zeta Q_0^\ell \in \mathcal{M}(E\times\mathcal{H}_K^{++}),$$

then we also have the narrow convergence $\zeta\bar{Q}_{0,n}\to\pi$. Next, given any $f\in C_{\rm b}(E\times\mathcal{H}_K^{++})$, we write

$$\zeta\bar{Q}_{0,n}f = \frac{\zeta Q_0^K f}{n-K} + \zeta\bar{Q}_{0,n}(Q_0 f) - \frac{\zeta Q_0^n f}{n-K}.$$

Since $Q_0 f\in C_{\rm b}(E\times\mathcal{H}_K^{++})$ according to Lemma 10, by taking the limit $n\to\infty$ we obtain $\pi f = \pi Q_0 f$ and thus π is an invariant measure for Q_0. □

Lemma 12. If Q_0 has an invariant distribution, then it is unique.

Proof. If $\pi\in\mathcal{M}(E\times\mathcal{H}_K^{+})$ satisfies $\pi=\pi Q_0$ then $\pi=\pi Q_0^K$ and Lemma 9 yields that necessarily $\pi\in\mathcal{M}(E\times\mathcal{H}_K^{++})$. Let $\pi^1,\pi^2\in\mathcal{M}(E\times\mathcal{H}_K^{++})$ be two invariant distributions for Q_0. Since θ is the unique invariant distribution for P by assumption, necessarily $\pi^i(\cdot\times\mathcal{H}_K^{++})=\theta(\cdot)$. Let $\mathsf{X}_0^{\pi^1} := (\mathsf{F}_0,\mathsf{G}_0,\mathsf{Z}_0^{\pi^1})$ and $\mathsf{X}_0^{\pi^2} := (\mathsf{F}_0,\mathsf{G}_0,\mathsf{Z}_0^{\pi^2})$ be two $E\times\mathcal{H}_K^{++}$–valued random variables such that $\mathsf{X}_0^{\pi^i}\sim\pi^i$. Starting from $\mathsf{X}_0^{\pi^1}$ and $\mathsf{X}_0^{\pi^2}$, construct two Markov processes $(\mathsf{X}_n^{\pi^i} = (\mathsf{F}_n,\mathsf{G}_n,\mathsf{Z}_n^{\pi^i}))_{n\in\mathbb{N}}$ with the transition kernel Q_0 for $i=1,2$ respectively. To show that $\pi^1=\pi^2$, it will be enough to show that $\|\mathsf{X}_n^{\pi^1}-\mathsf{X}_n^{\pi^1}\|\to 0$ in probability as $n\to\infty$, or equivalently, that $\mathrm{dist}(\mathsf{Z}_n^{\pi^1},\mathsf{Z}_n^{\pi^2})\to 0$ in probability. We use similar arguments and the same notations as in Section 3.2.2.

Recalling (37) and keeping in mind that Assumption 2(d) yields that $\mathsf{Z}_n\in\mathcal{H}_K^{++}$ a.s. and that F_n has full rank a.s. for every $n\in\mathbb{N}$, we have

$$\mathsf{Z}_n^{\pi^i} = h_{0,\mathsf{F}_n,\mathsf{G}_n}(\mathsf{Z}_{n-1}^{\pi^i}) = \tau_{\mathsf{G}_n^*\mathsf{F}_n(\mathsf{F}_n^*\mathsf{F}_n)^{-1/2},\,\mathsf{G}_n^*\Pi_{\mathsf{F}_n}^{\perp}\mathsf{G}_n}\circ\iota\circ\tau_{I,(\mathsf{F}_n^*\mathsf{F}_n)^{-1}}\circ\iota(\mathsf{Z}_{n-1}^{\pi^i}).$$

Lemma 6 then yields

$$\mathrm{dist}(\mathsf{Z}_n^{\pi^1},\mathsf{Z}_n^{\pi^2}) \leq \frac{\|\mathsf{F}_n\|^2\min(\|(\mathsf{Z}_{n-1}^{\pi^1})^{-1}\|,\|(\mathsf{Z}_{n-1}^{\pi^2})^{-1}\|)}{\|\mathsf{F}_n\|^2\min(\|(\mathsf{Z}_{n-1}^{\pi^1})^{-1}\|,\|(\mathsf{Z}_{n-1}^{\pi^2})^{-1}\|)+1}\mathrm{dist}(\mathsf{Z}_{n-1}^{\pi^1},\mathsf{Z}_{n-1}^{\pi^2})$$

which implies, for any $n\geq 1$, that

$$\mathrm{dist}(\mathsf{Z}_n^{\pi^1},\mathsf{Z}_n^{\pi^2}) \leq \Big(\prod_{i=0}^{n-1}\xi_i\Big)\mathrm{dist}(\mathsf{Z}_0^{\pi^1},\mathsf{Z}_0^{\pi^2}),\quad \xi_i := \frac{\|\mathsf{F}_{i+1}\|^2\max(\|(\mathsf{Z}_i^{\pi^1})^{-1}\|,\|(\mathsf{Z}_i^{\pi^2})^{-1}\|)}{\|\mathsf{F}_{i+1}\|^2\max(\|(\mathsf{Z}_i^{\pi^1})^{-1}\|,\|(\mathsf{Z}_i^{\pi^2})^{-1}\|)+1}. \tag{41}$$

By Hölder's inequality, we have

$$\mathbb{E}\prod_{i=0}^{n-1}\xi_i \leq \prod_{i=0}^{n-1}(\mathbb{E}\xi_i^n)^{1/n} \leq \mathbb{E}\left[\left(\frac{\|\mathsf{F}_1\|^2\max(\|(\mathsf{Z}_0^{\pi^1})^{-1}\|,\|(\mathsf{Z}_0^{\pi^2})^{-1}\|)}{\|\mathsf{F}_1\|^2\max(\|(\mathsf{Z}_0^{\pi^1})^{-1}\|,\|(\mathsf{Z}_0^{\pi^2})^{-1}\|)+1}\right)^n\right].$$

By dominated convergence, the rightmost term of these inequalities converges to zero as $n\to\infty$, and thus $\prod_{i=0}^{n-1}\xi_i\to 0$ in probability. It thus follows from (41) that $\mathrm{dist}(\mathsf{Z}_n^{\pi^1},\mathsf{Z}_n^{\pi^2})\to 0$ in probability, which concludes the proof. □

4.3 The last step for the proof of Theorem 2

Proof of Theorem 2. First, Corollary 11 and Lemma 12 show that Q_0 has a unique invariant measure, that we denote by π_0, and moreover that $\pi_0 \in \mathcal{M}(E \times \mathcal{H}_K^{++})$. Kolomorogov's existence theorem then yields there exists a unique stationary Markov process $(\mathsf{F}_n, \mathsf{G}_n, \mathsf{Z}_n)_{n\in\mathbb{Z}}$ on $E \times \mathcal{H}_K^{++}$ with transition kernel Q_0, which is in particular ergodic. Moreover, $(\mathsf{Z}_n)_{n\in\mathbb{Z}}$ satisfies the equation (13) by definition of Q_0, which proves part (a) of the theorem.

To prove (b), we claim that the family $\{\pi_\gamma\}_{\gamma\in[0,1]}$ is tight in $\mathcal{M}(E\times\mathcal{H}_K^{+})$. Indeed, if $(F, G, Z) \sim \pi_\gamma$, then $Z = h_{\gamma,F,G}(Z)$ in law and, since $\|h_{\gamma,F,G}(Z)\| \le \|G\|^2 + \gamma$ and $\pi_\gamma(\cdot \times \mathcal{H}_K^{++}) = \theta(\cdot)$ is independent on γ, the claim follows. As a consequence, $\pi_\gamma \to \zeta$ narrowly for some $\zeta \in \mathcal{M}(E\times\mathcal{H}_K^{+})$ as $\gamma \to 0$ along a subsequence. By definition of π_γ, for any $f \in C_{\mathrm{b}}(E \times \mathcal{H}_K^{++})$ we have

$$\pi_\gamma f = \pi_\gamma Q_\gamma f.$$

The left hand side converges to ζf as $\gamma \to 0$ by definition of ζ, and the exact same lines of arguments as in the proof of Lemma 10 yield that the right hand side converges to $\zeta Q_0 f$, showing that $\zeta = \zeta Q_0$. Since the invariant measure π_0 of Q_0 is unique, we thus have shown that $\pi_\gamma \to \pi_0$ narrowly as $\gamma \to 0$.

We finally go back to the identity (32), which can be rewritten as

$$\begin{aligned} N\mathcal{I}_\rho + K\log\gamma &= \mathbb{E}\log\det(\mathsf{Z}_{\gamma,0} + \mathsf{F}_1^*\mathsf{F}_1) \\ &= \int \log\det(\mathsf{Z} + F^*F)P((\mathsf{F},\mathsf{G}), dF\times dG)\pi_\gamma(d\mathsf{F}\times d\mathsf{G}\times d\mathsf{Z}) \\ &= \int \log\det(\mathsf{Z} + F^*F)Q_\gamma((\mathsf{F},\mathsf{G},\mathsf{Z}), dF\times dG\times dZ)\pi_\gamma(d\mathsf{F}\times d\mathsf{G}\times d\mathsf{Z}) \\ &= \int \log\det(\mathsf{Z} + \mathsf{F}^*\mathsf{F})\pi_\gamma(\mathsf{F},\mathsf{G},\mathsf{Z}) \\ &= \mathbb{E}\log\det(\mathsf{Z}_{\gamma,1} + \mathsf{F}_1^*\mathsf{F}_1). \end{aligned} \tag{42}$$

Using Skorokhod's representation theorem, we can introduce a probability space $(\Omega', \mathscr{F}', \mathbb{P}')$ and a family of $E \times \mathcal{H}_K^{++}$-valued random variables $\{\mathsf{U}_{\gamma,i} = (\mathsf{F}_i', \mathsf{G}_i', \mathsf{Z}_{\gamma,i}') : \gamma \in [0,1],\ i \in \{0,1\}\}$ such that $\mathsf{U}_{\gamma,i} \sim \pi_\gamma$, $\mathsf{Z}_{\gamma,1}' = h_{\mathsf{F}_1',\mathsf{G}_1',\gamma}(\mathsf{Z}_{\gamma,0}')$ for every $\gamma \in [0,1]$ and $\mathsf{U}_{\gamma,i} \to \mathsf{U}_{0,i}$ as $\gamma \to 0$ $\mathbb{P}'$-a.s. This yields that, as $\gamma \to 0$,

$$\log\det(\mathsf{Z}_{\gamma,1}' + (\mathsf{F}_1')^*\mathsf{F}_1') \to \log\det(\mathsf{Z}_{1,0}' + (\mathsf{F}_1')^*\mathsf{F}_1'), \qquad \mathbb{P}'\text{-a.s.}$$

Moreover, using that $\|\mathsf{Z}_{\gamma,1}'\| = \|h_{\mathsf{F}_1',\mathsf{G}_1',\gamma}(\mathsf{Z}_{\gamma,0}')\| \le \gamma + \|\mathsf{G}_1'\|^2 \le 1 + \|\mathsf{G}_1'\|^2$ and that $\log(1 + a + b) \le \log(1 + a) + \log(1 + b)$ for any $a, b \ge 0$, we also have

$$\log\det((\mathsf{F}_1')^*\mathsf{F}_1') \le \log\det(\mathsf{Z}_{\gamma,1}' + (\mathsf{F}_1')^*\mathsf{F}_1') \le N\log(1 + \|\mathsf{G}_1'\|^2 + \|\mathsf{F}_1'\|^2)$$

and thus

$$|\log\det(\mathsf{Z}_{\gamma,1}' + (\mathsf{F}_1')^*\mathsf{F}_1')| \le H(\mathsf{F}_1', \mathsf{G}_1') := |\log\det((\mathsf{F}_1')^*\mathsf{F}_1')| + N\log(1 + \|\mathsf{G}_1'\|^2) + N\log(1 + \|\mathsf{F}_1'\|^2).$$

Since $(\mathsf{F}_1', \mathsf{G}_1')$ has law θ by construction, Assumption 2(b)-(c) yields that $\mathbb{E}H(\mathsf{F}_1', \mathsf{G}_1') < \infty$ and thus, by dominated convergence, we obtain from (42),

$$\begin{aligned} \lim_{\gamma\to0} N\mathcal{I}_\rho + K\log\gamma &= \lim_{\gamma\to0} \mathbb{E}\log\det(\mathsf{Z}_{\gamma,1} + \mathsf{F}_1^*\mathsf{F}_1) \\ &= \lim_{\gamma\to0} \mathbb{E}\log\det(\mathsf{Z}_{\gamma,1}' + (\mathsf{F}_1')^*\mathsf{F}_1') \\ &= \mathbb{E}\log\det(\mathsf{Z}_{0,1}' + (\mathsf{F}_1')^*\mathsf{F}_1') \\ &= \mathbb{E}\log\det(\mathsf{Z}_{0,0}' + (\mathsf{F}_1')^*\mathsf{F}_1') \\ &= \mathbb{E}\log\det(\mathsf{Z}_0 + \mathsf{F}_1^*\mathsf{F}_1), \end{aligned}$$

where we used a similar computation than in (42) for the fourth equality, and Theorem 2-(b) is proven.

To establish Theorem 2-(c), we follow the same strategy as in the proof of Theorem 1-(c): Since the Markov chain $(\mathsf{F}_n, \mathsf{G}_n, \mathsf{Z}_n)_{n\in\mathbb{Z}}$ is ergodic, we have

$$\kappa_\infty = \frac{1}{N}\mathbb{E}\log\det(\mathsf{Z}_0 + \mathsf{F}_1^*\mathsf{F}_1) = \lim_{n\to\infty}\frac{1}{Nn}\sum_{\ell=0}^{n-1}\log\det(\mathsf{Z}_\ell + \mathsf{F}_{\ell+1}^*\mathsf{F}_{\ell+1}) \quad \text{a.s.} \tag{43}$$

By using the same line of argument as in the proof of Lemma 12, we obtain with a bound similar to (41) and the arguments below that $\operatorname{dist}(\mathsf{X}_n, \mathsf{Z}_n) \to 0$ in probability. This implies in turn that $\operatorname{dist}(\mathsf{X}_n + \mathsf{F}_{n+1}^*\mathsf{F}_{n+1}, \mathsf{Z}_n + \mathsf{F}_{n+1}^*\mathsf{F}_{n+1}) \le \operatorname{dist}(\mathsf{X}_n, \mathsf{Z}_n) \to 0$, and thus, that $\log\det(\mathsf{X}_n + \mathsf{F}_{n+1}^*\mathsf{F}_{n+1}) - \log\det(\mathsf{Z}_n + \mathsf{F}_{n+1}^*\mathsf{F}_{n+1}) \to 0$ in probability. As a consequence, part (c) is obtained by taking a Cesàro average and (43). □

4.4 Proofs for Section 1.5.1

We shall need the following result, which follows from the fact that the zero set of a non-zero polynomial of d variables has zero measure for the Lebesgue measure of $\mathbb{R}^d$.

Lemma 13. Let X be a random complex $n\times n$ matrix whose distribution is absolutely continuous with respect to the Lebesgue measure on $\mathbb{C}^{n\times n} \simeq \mathbb{R}^{2n^2}$. Then, $\mathbb{P}(\operatorname{rank}(X) = n) = 1$.

We also need in this paragraph the following notations: Given a positive integer n, we set $[n] := \{0, \ldots, n-1\}$. Given a matrix $X \in \mathbb{C}^{m\times n}$ and two sets of indices $J_1 \subset [m]$ and $J_2 \in [n]$, we denote by X^{J_1,J_2} the $|J_1| \times |J_2|$ submatrix of X obtained by keeping the rows of X whose indices belong to J_1 and the columns of X whose indices belong to J_2. We also write for convenience $X^{J_1,\cdot} := X^{J_1,[n]}$ and $X^{\cdot,J_2} := X^{[m],J_2}$. Finally, we write $\log^-(x) = \min(\log x, 0)$ and $\log^+(x) = \max(\log x, 0)$.

Proof of Proposition 3. We start with Assumption 2-(d). Using that (U_n, V_n) and $(\mathsf{F}_k, \mathsf{G}_k)_{k\le n-1}$ are independent, it is enough to show that for any $B, D \in \mathbb{C}^{N\times K}$,

$$\mathbb{P}\left[\det((V_n + D)^*(U_n + B)) = 0\right] = 0, \tag{44}$$

$$\forall v \in \mathbb{C}^K \setminus \{0\}, \qquad \mathbb{P}\left[\Pi^\perp_{V_n+D}(U_n + B)v = 0\right] = 0. \tag{45}$$

Letting $J := [K]$ and $J^c := [N] \setminus [K]$, we have

$$\begin{aligned}&\mathbb{P}\left[\det((V_n + D)^*(U_n + B)) = 0\right]\\ &= \mathbb{P}\Big[\det((V_n^{J,\cdot} + D^{J,\cdot})^*(U_n^{J,\cdot} + B^{J,\cdot}) + (V_n^{J^c,\cdot} + D^{J^c,\cdot})^*(U_n^{J^c,\cdot} + B^{J^c,\cdot})) = 0\Big].\end{aligned} \tag{46}$$

Since U_n has a density (for Lebesgue), then for any invertible matrix $S \in \mathbb{C}^{K\times K}$, we see that $S(U_n^{J,\cdot} + B^{J,\cdot})$ has a density. Since Lemma 13 yields that the random matrix $(V_n^{J,\cdot} + D^{J,\cdot})$ is invertible a.s (it has a density), the square matrix $(V_n^{J,\cdot} + D^{J,\cdot})^*(U_n^{J,\cdot} + B^{J,\cdot})$ has a density. Recall that the convolution between an absolutely continuous probability and any probability measure is absolutely continuous. Thus, since $(U_n^{J,\cdot}, V_n^{J,\cdot})$ and $(U_n^{J^c,\cdot}, V_n^{J^c,\cdot})$ are independent, the matrix within the determinant at the right hand side of (46) has a density. Using Lemma 13 again, we obtain (44).

For any $v \in \mathbb{C}^K \setminus \{0\}$, the vector $w := (U_n + B)v$ is a random vector whose elements are independent and have probability densities. It results that for any matrix $C \in \mathbb{C}^{N\times K}$, we have $\Pi^\perp_C w \neq 0$ a.s. Thus, $\mathbb{P}\left[\Pi^\perp_{V_n+D}(U_n + B)v = 0\right] = 0$ by the Fubini-Tonelli theorem, and (45) is obtained.

We now establish the truth of Assumption 2-(c). Write $\mathsf{F}_n = \begin{bmatrix}\mathsf{f}_n^0 & \cdots & \mathsf{f}_n^{K-1}\end{bmatrix}$, where f_n^k is the k^{th} column of the matrix F_n. For $k \in [K-1]$, let $J_k = \{k+1, \ldots, K-1\}$. Applying, *e.g.*, a

Gram-Schmidt process to the successive columns $\mathsf{f}_n^0, \ldots, \mathsf{f}_n^{K-1}$, setting $\mathsf{F}_n^{\cdot,\emptyset} = 0 \in \mathbb{C}^N$, and using the obvious inequality $\log^+ x \leq x$ for $x > 0$, we get that

$$\begin{aligned}\mathbb{E}\,|\log\det \mathsf{F}_n^*\mathsf{F}_n| &= \mathbb{E}\left|\sum_{k=0}^{K-1}\log(\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k\right| \leq \mathbb{E}\sum_{k=0}^{K-1}\left|\log(\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k\right| \\ &\leq \sum_{k=0}^{K-1}\mathbb{E}\left|\log^-((\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k)\right| + \sum_{k=0}^{K-1}\mathbb{E}\|\mathsf{f}_n^k\|^2 \\ &\leq \sum_{k=0}^{K-1}\mathbb{E}\left|\log^-((\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k)\right| + C,\end{aligned}$$

where $C < \infty$ since Assumption 2-(b) is satisfied. Fix $k \in [K]$. In the remainder of the proof, "conditional" refers to a conditionning on $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1}, u_n^{k+1}, \ldots, u_n^{K-1})$. All the bounds are constants that only depend on the bound on the densities of the elements of U_n.

The vector f_n^k can be written as $\mathsf{f}_n^k = \mathsf{d}_{n-1}^k + u_n^k$, where d_{n-1}^k is $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1})$-measurable, and where u_n^k is the k^{th} column of U_n. By the assumptions on (U_n), the elements of f_n^k are conditionally independent and have bounded densities. If $k < K-1$, make a $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1}, u_n^{k+1}, \ldots, u_n^{K-1})$-measurable choice of a unit-norm vector p^k which is orthogonal to the subspace $\operatorname{span}\mathsf{F}_n^{\cdot,J_k}$, otherwise, take p^k as an arbitrary constant unit-norm vector. Since $|\log^-(\cdot)|$ is a nonincreasing function, $|\log^-((\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k)| \leq \left|\log^-(|\langle \mathsf{p}^k, \mathsf{f}_n^k\rangle|^2)\right|$. Since $\mathsf{p}^k = \left[\mathsf{p}_0^k, \ldots, \mathsf{p}_{N-1}^k\right]^\mathsf{T}$ has unit-norm, it has at least one element, say p_0^k, such that $|\mathsf{p}_0^k| \geq 1/\sqrt{N}$. Writing $\mathsf{f}_n^k = \left[\mathsf{f}_{n,0}^k, \ldots, \mathsf{f}_{n,N-1}^k\right]^\mathsf{T}$, we get that the conditional density of $\mathsf{p}_0^k\mathsf{f}_{n,0}^k$ is bounded, and by doing a simple calculation involving density convolutions, we finally obtain that $\langle \mathsf{p}^k, \mathsf{f}_n^k\rangle$ has a bounded conditional density. Now, it is easy to see that if X is a complex random variable with a density bounded by a constant C then $\mathbb{E}|\log^-(|X|^2)| \leq C\pi$. This shows that $\mathbb{E}\left|\log^-((\mathsf{f}_n^k)^*\Pi^\perp_{\mathsf{F}_n^{\cdot,J_k}}\mathsf{f}_n^k)\right| < \infty$ for each $k \in [K]$, which completes the proof. □

To prove Proposition 4, we first need the following lemma.

Lemma 14. Given any positive integers m, n, r satisfying $r \leq n \leq m$, let X be a $m \times n$ matrix with rank n, write $X = \left[Y^\mathsf{T}\ \widetilde{Y}^\mathsf{T}\right]^\mathsf{T}$ where Y is a $r \times n$ matrix, and assume that $\operatorname{rank}(Y) = r$. Then $\Pi_X^{[r],[r]} = I$ *iff* $\operatorname{span}(\widetilde{Y}) = \operatorname{span}(\widetilde{Y}A)$ for some matrix A satisfying $\operatorname{span}(A) = \ker Y$.

Proof. The formula $\Pi_X = X(X^*X)^{-1}X^*$ yields $\Pi_X^{[r],[r]} = Y(Y^*Y + \widetilde{Y}^*\widetilde{Y})^{-1}Y^*$. Performing a singular value decomposition,

$$Y = U\begin{bmatrix}\Lambda & 0\end{bmatrix}\begin{bmatrix}V_1^* \\ V_2^*\end{bmatrix},$$

with Λ the diagonal $r \times r$ matrix of singular values and V_2 satisfying $\operatorname{span}(V_2) = \ker Y$, and using Schur's complement formula (20), we obtain

$$\begin{aligned}\Pi_X^{[r],[r]} &= U\begin{bmatrix}\Lambda & 0\end{bmatrix}\left(\begin{bmatrix}\Lambda^2 & \\ & 0\end{bmatrix} + \begin{bmatrix}V_1^* \\ V_2^*\end{bmatrix}\widetilde{Y}^*\widetilde{Y}\begin{bmatrix}V_1 & V_2\end{bmatrix}\right)^{-1}\begin{bmatrix}\Lambda \\ 0\end{bmatrix}U^* \\ &= U\Lambda\left(\Lambda^2 + V_1^*\widetilde{Y}^*(I - \widetilde{Y}V_2(V_2^*\widetilde{Y}^*\widetilde{Y}V_2)^{-1}V_2^*\widetilde{Y}^*)\widetilde{Y}V_1\right)^{-1}\Lambda U^* \\ &= U\Lambda\left(\Lambda^2 + V_1^*\widetilde{Y}^*\Pi^\perp_{\widetilde{Y}V_2}\widetilde{Y}V_1\right)^{-1}\Lambda U^*.\end{aligned}$$

This expression shows that $\Pi_X^{[r],[r]} = I$ *iff* $V_1^*\widetilde{Y}^*\Pi^\perp_{\widetilde{Y}V_2}\widetilde{Y}V_1 = 0$. We then have

$$\begin{aligned}V_1^*\widetilde{Y}^*\Pi^\perp_{\widetilde{Y}V_2}\widetilde{Y}V_1 = 0 &\Leftrightarrow \operatorname{span}(\widetilde{Y}V_1V_1^*\widetilde{Y}^*) \subset \operatorname{span}(\widetilde{Y}V_2V_2^*\widetilde{Y}^*) \\ &\Leftrightarrow \operatorname{span}(\widetilde{Y}\widetilde{Y}^*) \subset \operatorname{span}(\widetilde{Y}V_2V_2^*\widetilde{Y}^*) \\ &\Leftrightarrow \operatorname{span}(\widetilde{Y}) = \operatorname{span}(\widetilde{Y}V_2),\end{aligned}$$

which is the required result. □

Proof of Proposition 4. Let us prove that Assumption 2-(d) holds. The recursive equation (15) satisfied by $(C_n)_{n\in\mathbb{Z}}$ yields, for any $\ell \in [L-1]$ and $k \in [L]$,

$$\begin{aligned}
c_{nL+\ell,k} &= H_k c_{nL+\ell-1,k} + u_{nL+\ell,k} \\
&= H_k^2 c_{nL+\ell-2,k} + H_k u_{nL+\ell-1,k} + u_{nL+\ell,k} \\
&= \cdots \\
&= H_k^{\ell+1} c_{nL-1,k} + \sum_{i=0}^{\ell} H_k^i u_{nL+\ell-i,k}
\end{aligned}$$

where $U_n =: \begin{bmatrix} u_{n,0}^{\mathsf{T}} & \cdots & u_{n,L}^{\mathsf{T}} \end{bmatrix}^{\mathsf{T}}$, the $u_{n,\ell}$'s being $R \times T$ matrices. Notice that the $c_{nL-1,k}$ and the $u_{nL+\ell-i,k}$ terms in the rightmost term above are respectively $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1})$-measurable and independent from $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1})$. Plugging these equations in the expressions for F_n and G_n, we obtain

$$\begin{aligned}
\mathsf{F}_n &= \begin{bmatrix} u_{nL,L} & \cdots & \cdots & u_{nL,1} \\ & u_{nL+1,L} + H_L u_{nL,L} & & \vdots \\ & & \ddots & \vdots \\ & & & u_{nL+L-1,L} + \sum_{i=1}^{L-1} H_L^i u_{nL+L-1-i,L} \end{bmatrix} + B_{n-1} \\
&=: Q_n + B_{n-1},
\end{aligned} \tag{47}$$

and

$$\begin{aligned}
\mathsf{G}_n &= \begin{bmatrix} u_{nL,0} & & & \\ \vdots & u_{nL+1,0} + H_0 u_{nL,0} & & \\ & & \ddots & \\ & \cdots & \cdots & u_{nL+L-1,0} + \sum_{i=1}^{L-1} H_0^i u_{nL+L-1-i,0} \end{bmatrix} + D_{n-1} \\
&=: S_n + D_{n-1},
\end{aligned}$$

where the matrices B_{n-1} and D_{n-1} are $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1})$-measurable random matrices which are block-upper triangular and block-lower triangular respectively, with $R \times T$ blocks (the exact expressions of these matrices are irrelevant). Furthermore, the matrices Q_n and S_n are independent of $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1})$. Thus, the proposition will be proven if we show that for all constant block-upper triangular matrices $B \in \mathbb{C}^{LR\times LT}$ and all constant block-lower triangular matrices $D \in \mathbb{C}^{LR\times LT}$ with $R \times T$ blocks,

$$\mathbb{P}\left[\det((S_n + D)^*(Q_n + B)) = 0\right] = 0, \tag{48}$$

$$\forall v \in \mathbb{C}^{LT} \setminus \{0\}, \qquad \mathbb{P}\left[\Pi^{\perp}_{S_n+D}(Q_n + B)v = 0\right] = 0. \tag{49}$$

The matrix $(S_n + D)^*(Q_n + B)$ is a square $LT \times LT$ block-upper triangular matrix with $T \times T$ blocks. Using Lemma 13 as in the proof of Proposition 3, one can check that all the diagonal blocks of this matrix are a.s. invertible, and (48) is proven.

To establish (49), we set $J_\ell := \{\ell R, \ldots, \ell R + R - 1\}$ and prove that

$$\forall \ell \in [L], \qquad (\Pi^{\perp}_{S_n+D})^{\cdot, J_\ell} \neq 0 \quad \text{a.s.} \tag{50}$$

Indeed, given $v = [v_0^{\mathsf{T}}, \ldots, v_{L-1}^{\mathsf{T}}]^{\mathsf{T}} \in \mathbb{C}^{LT} \setminus \{0\}$ with $v_i \in \mathbb{C}^T$, let $k := \max\{i \in [L] : v_i \neq 0\}$. An inspection of (47) reveals that

$$(Q_n + B)v = \begin{bmatrix} \vdots \\ 0 \\ u_{nL+k,L} v_k \\ 0 \\ \vdots \end{bmatrix} + a,$$

for a random vector a which is independent from $u_{nL+k,L}$. With this at hand, we see that

$$\Pi^{\perp}_{S_n+D}(Q_n + B)v = (\Pi^{\perp}_{S_n+D})^{\cdot,J_k} u_{nL+k,L} v_k + \Pi^{\perp}_{S_n+D} a\,.$$

Since $\Pi^{\perp}_{S_n+D}$ and $u_{nL+k,L}$ are independent and $u_{nL+k,L} v_k$ has a density, (49) follows from (50).

To complete the proof of that Assumption 2-(d) holds true, we now turn to the proof of (50). We use the equivalence $(\Pi^{\perp}_{S_n+D})^{\cdot,J_\ell} = 0 \Leftrightarrow (\Pi_{S_n+D})^{J_\ell,J_\ell} = I$. Let us write

$$S_n + D = \begin{bmatrix} \widetilde{Y}_1 \\ Y \\ \widetilde{Y}_2 \end{bmatrix},$$

where $Y = (S_n + D)^{J_\ell,\cdot} \in \mathbb{C}^{R\times LT}$, and set

$$\widetilde{Y} := \begin{bmatrix} \widetilde{Y}_1 \\ \widetilde{Y}_2 \end{bmatrix} \in \mathbb{C}^{(L-1)R\times LT}.$$

Since $\operatorname{rank}((\Pi_{S_n+D})^{J_\ell,J_\ell}) \le \operatorname{rank}(Y)$, then if $\operatorname{rank}(Y) < R$ we have $(\Pi_{S_n+D})^{J_\ell,J_\ell} \ne I$. Assume $\operatorname{rank}(Y) = R$. Then $\dim\ker(Y) = LT - R$. By Lemma 14, $(\Pi_{S_n+D})^{J_\ell,J_\ell} = I$ implies $\operatorname{rank}\widetilde{Y} = \dim(\widetilde{Y}(\ker Y))$. Observe that $\dim(\widetilde{Y}(\ker Y)) \le LT - R$. For $m \in [L]$, let $J'_m := \{mR, \ldots, mR + T - 1\}$ and $\widetilde{J}_\ell := \cup_{m\in[L]\setminus\{\ell\}} J'_m$. Then, $(S_n + D)^{\widetilde{J}_\ell,\cdot}$ is a submatrix of $\widetilde{Y}$. But thanks to the block-triangular stucture of $S_n + D$, one can check that $(S_n + D)^{\widetilde{J}_\ell,\cdot}$ has a block-echelon form, and its diagonal blocks $\{(S_n + D)^{J'_m,J'_m}\}_{m\ne\ell}$ are all a.s. invertible. Thus, $\operatorname{rank}(S_n + D)^{\widetilde{J}_\ell,\cdot} = (L-1)T$ a.s. Consequently, $\operatorname{rank}(\widetilde{Y}) \ge (L-1)T > LT - R \ge \dim(\widetilde{Y}(\ker Y))$ a.s. which shows that (50) holds true, and therefore, that Assumption 2-(d) is verified.

We now turn to Assumption 2-(c). Getting back to Equation (47), write

$$B_{n-1} = \begin{bmatrix} B_{n-1,0} & \times & \times \\ & \ddots & \times \\ 0 & & B_{n-1,L-1} \end{bmatrix},$$

where the $B_{n-1,\ell}$ are the $R\times T$ diagonal blocks of B_{n-1}. Defining $J := \{0,\ldots,T-1\}\cup\{R,\ldots,R+T-1\}\cup\cdots\cup\{(L-1)R,\ldots,(L-1)R+T-1\}$, we observe from Equation (47) that

$$\mathsf{F}_n^{J,\cdot} = \begin{bmatrix} u^{[T],\cdot}_{nL,L} + B^{[T],\cdot}_{n-1,0} & \times & \times \\ & \ddots & \times \\ 0 & & u^{[T],\cdot}_{nL+L-1,L} + (B_{n-1,L-1} + \sum_{i=1}^{L-1} H_L^i u_{nL+L-1-i,L})^{[T],\cdot} \end{bmatrix}$$

is a square upper block-triangular matrix with $T\times T$ blocks. Moreover, the ℓ^{th} diagonal block of this matrix is the sum of $u^{[T],\cdot}_{nL+\ell,L}$ and a $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1}, u_{nL},\ldots,u_{nL+\ell-1})$-measurable term that we denote by $\mathsf{d}_{n,\ell}$. Now, since

$$(1 + \|\mathsf{F}_n\|^2)I > \mathsf{F}_n^*\mathsf{F}_n \ge (\mathsf{F}_n^{J,\cdot})^*\mathsf{F}_n^{J,\cdot}$$

in the Hermitian semidefinite ordering, it holds that

$$LT\log(1 + \|\mathsf{F}_n\|^2) > \log\det(\mathsf{F}_n^*\mathsf{F}_n) \ge \log\det((\mathsf{F}_n^{J,\cdot})^*\mathsf{F}_n^{J,\cdot}),$$

thus,

$$\mathbb{E}|\log\det(\mathsf{F}_n^*\mathsf{F}_n)| < \mathbb{E}|\log\det((\mathsf{F}_n^{J,\cdot})^*\mathsf{F}_n^{J,\cdot})| + LT\,\mathbb{E}\|\mathsf{F}_n\|^2 \le \mathbb{E}|\log\det((\mathsf{F}_n^{J,\cdot})^*\mathsf{F}_n^{J,\cdot})| + C,$$

where $C < \infty$ since Assumption 2-(b) is verified. Moreover,

$$\begin{aligned}\mathbb{E}|\log\det((\mathsf{F}_n^{J,\cdot})^*\mathsf{F}_n^{J,\cdot})| &= \mathbb{E}\Big|\sum_{\ell=0}^{L-1}\log\det(u^{[T],\cdot}_{nL+\ell,L} + \mathsf{d}_{n,\ell})(u^{[T],\cdot}_{nL+\ell,L} + \mathsf{d}_{n,\ell})^*\Big| \\ &\le \sum_{\ell=0}^{L-1}\mathbb{E}\Big|\log\det(u^{[T],\cdot}_{nL+\ell,L} + \mathsf{d}_{n,\ell})(u^{[T],\cdot}_{nL+\ell,L} + \mathsf{d}_{n,\ell})^*\Big|,\end{aligned}$$

and the summands in this last expression can be dealt with as in the last part of the proof of Proposition 3. The main distinctive feature of the proof here is that when we deal with the ℓ^{th} summand and when it comes to manipulate the conditional densities, we need to condition on $(\mathsf{F}_{n-1}, \mathsf{G}_{n-1}, u_{nL}, \dots, u_{nL+\ell-1})$. This concludes the proof of Proposition 4. □

5 Proof of Proposition 5

The expression of Shannon's mutual information given by Theorem 1 provides a means of recovering the large random matrix regime when $K, N \to \infty$ with $K/N \to \gamma \in (0, \infty)$ in a general setting. We describe a general approach before to specify to the setting of Proposition 5, which is encoded in the following result.

Lemma 15. Under Assumption 1, if we introduce for any $m \le n$,

$$\mathring{\mathsf{H}}_{m,n} := \begin{bmatrix} \mathsf{G}_m & & & \mathsf{F}_m \\ \mathsf{F}_{m+1} & \mathsf{G}_{m+1} & & \\ & \ddots & \ddots & \\ & & \mathsf{F}_n & \mathsf{G}_n \end{bmatrix},$$

then we have as $M \to \infty$,

$$\mathcal{I}_\rho = \frac{1}{(M+1)N} \mathbb{E} \log\det(I + \rho\, \mathring{\mathsf{H}}_{0,M} \mathring{\mathsf{H}}_{0,M}^*) + \mathcal{O}(1/M) \tag{51}$$

where $\mathcal{O}(1/M)$ is uniform in K, N.

As an illustration, we now prove Proposition 5 as an easy consequence of this lemma and well known results from random matrix theory.

Proof of Proposition 5. Observe from (3) and the assumptions made on the process $(C_n)_{n\in\mathbb{Z}}$ that, for any $M \ge 1$, the $(M+1)RL \times (M+1)RL$ matrix $\mathring{H}_{0,M}$ is a square matrix having independent entries with a *doubly stochastic* variance profile. It is well known in random matrix theory that when $L \to \infty$, the empirical spectral measure of $\mathring{H}_{0,M}\mathring{H}_{0,M}^*$ converges narrowly to the Marchenko-Pastur distribution $\mu_{\text{MP}}(d\lambda) = (2\pi)^{-1}\sqrt{4/\lambda - 1}\mathbb{1}_{[0,4]}(\lambda)\, d\lambda$ a.s, see [7, 24, 9]. Using the moment condition (4), we therefore obtain, for every fixed $M \ge 1$,

$$\frac{1}{(M+1)RL} \mathbb{E} \log\det(I + \rho\, \mathring{\mathsf{H}}_{0,M} \mathring{\mathsf{H}}_{0,M}^*) \xrightarrow[L\to\infty]{} \int \log(1+\rho\lambda)\, \mu_{\text{MP}}(d\lambda).$$

One can compute, see *e.g.* [24, Th. 2.53] or [9, Th. 4.1], that this limiting integral coincides with the right hand side of (16). Letting $M \to \infty$, the proposition follows from Lemma 15.

□

We finally turn to the proof of the lemma.

Proof of Lemma 15. Using the notations of Theorem 1, we set

$$\xi_n := \log\det\left(I + \rho\, \mathsf{F}_n \mathsf{W}_{n-1} \mathsf{F}_n^*\right) - \log\det \mathsf{W}_n$$

and check, similarly as in (29), that

$$\xi_n = \log\det\left(I + \rho\, \mathsf{G}_n\, \mathsf{G}_n^* + \rho\, \mathsf{F}_n \mathsf{W}_{n-1} \mathsf{F}_n^*\right).$$

If we set for convenience

$$\mathsf{V}_n := \rho\, \mathsf{G}_n^* (I + \widetilde{\mathsf{V}}_n)^{-1} \mathsf{G}_n$$
$$\widetilde{\mathsf{V}}_n := \rho\, \mathsf{F}_n \mathsf{W}_{n-1} \mathsf{F}_n^*$$

then we have the relation $\widetilde{\mathsf{V}}_n = \rho\,\mathsf{F}_n(I+\mathsf{V}_{n-1})^{-1}\mathsf{F}_n^*$ and we moreover see that ξ_n equals to

$$
\begin{aligned}
&\log\det(I+\widetilde{\mathsf{V}}_n+\rho\,\mathsf{G}_n\mathsf{G}_n^*)\\
&=\log\det(I+\rho\,\mathsf{F}_n(I+\mathsf{V}_{n-1})^{-1}\mathsf{F}_n^*+\rho\,\mathsf{G}_n\mathsf{G}_n^*)\\
&=\log\det\left(I+\rho\begin{bmatrix}\mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\begin{bmatrix}(I+\mathsf{V}_{n-1})^{-1} & \\ & I\end{bmatrix}\begin{bmatrix}\mathsf{F}_n^*\\ \mathsf{G}_n^*\end{bmatrix}\right)\\
&=\log\det\left(I+\rho\begin{bmatrix}\mathsf{F}_n^*\\ \mathsf{G}_n^*\end{bmatrix}\begin{bmatrix}\mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\begin{bmatrix}(I+\mathsf{V}_{n-1})^{-1} & \\ & I\end{bmatrix}\right)\\
&=\log\det\left(I+\begin{bmatrix}\mathsf{V}_{n-1} & \\ & 0\end{bmatrix}+\rho\begin{bmatrix}\mathsf{F}_n^*\\ \mathsf{G}_n^*\end{bmatrix}\begin{bmatrix}\mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\right)-\log\det(I+\mathsf{V}_{n-1})\\
&=\log\det\left(I+\rho\begin{bmatrix}\mathsf{G}_{n-1}^*(I+\widetilde{\mathsf{V}}_{n-1})^{-1/2}\\ 0\end{bmatrix}\begin{bmatrix}(I+\widetilde{\mathsf{V}}_{n-1})^{-1/2}\mathsf{G}_{n-1} & 0\end{bmatrix}+\rho\begin{bmatrix}\mathsf{F}_n^*\\ \mathsf{G}_n^*\end{bmatrix}\begin{bmatrix}\mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\right)-\log\det(I+\mathsf{V}_{n-1})\\
&=\log\det\left(I+\rho\begin{bmatrix}(I+\widetilde{\mathsf{V}}_{n-1})^{-1/2}\mathsf{G}_{n-1} & 0\\ \mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\begin{bmatrix}\mathsf{G}_{n-1}^*(I+\widetilde{\mathsf{V}}_{n-1})^{-1/2} & \mathsf{F}_n^*\\ 0 & \mathsf{G}_n^*\end{bmatrix}\right)-\log\det(I+\mathsf{V}_{n-1})\\
&=\log\det\left(I+\begin{bmatrix}\widetilde{\mathsf{V}}_{n-1} & \\ & 0\end{bmatrix}+\rho\begin{bmatrix}\mathsf{G}_{n-1} & 0\\ \mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\begin{bmatrix}\mathsf{G}_{n-1}^* & \mathsf{F}_n^*\\ 0 & \mathsf{G}_n^*\end{bmatrix}\right)-\log\det(I+\widetilde{\mathsf{V}}_{n-1})-\log\det(I+\mathsf{V}_{n-1}).
\end{aligned}
$$

Using further the relation $I+\mathsf{V}_n=\mathsf{W}_n^{-1}$, we thus obtain that

$$
\xi_n+\xi_{n-1}=\log\det\left(I+\begin{bmatrix}\widetilde{\mathsf{V}}_{n-1} & \\ & 0\end{bmatrix}+\rho\begin{bmatrix}\mathsf{G}_{n-1} & 0\\ \mathsf{F}_n & \mathsf{G}_n\end{bmatrix}\begin{bmatrix}\mathsf{G}_{n-1}^* & \mathsf{F}_n^*\\ 0 & \mathsf{G}_n^*\end{bmatrix}\right).
$$

By iterating similar manipulations M times, where M will be made large in a moment, we have

$$
\sum_{i=0}^{M}\xi_{n-i}=\log\det\left(I+\mathsf{U}_{n-M}+\rho\,\hat{\mathsf{H}}_{n-M,n}\hat{\mathsf{H}}_{n-M,n}^*\right),
$$

where we introduced

$$
\mathsf{U}_m:=\begin{bmatrix}\widetilde{\mathsf{V}}_m & \\ & 0\end{bmatrix}\quad\text{and}\quad \hat{\mathsf{H}}_{m,n}:=\begin{bmatrix}\mathsf{G}_m & & & \\ \mathsf{F}_{m+1} & \mathsf{G}_{m+1} & & \\ & \ddots & \ddots & \\ & & \mathsf{F}_n & \mathsf{G}_n\end{bmatrix}.
$$

By definition of ξ_n and together with Theorem 1(b), this yields the identity

$$
\mathcal{I}_\rho=\frac{1}{(M+1)N}\,\mathbb{E}\log\det\left(I+\mathsf{U}_0+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*\right)
$$

for all positive integers M.

Next, we control the cost of eliminating U_0 from this expression. To do so, we use that $|\log\det(I+A)|\leq\operatorname{tr}(A)$ and $\operatorname{tr}(AB)\leq\|B\|\operatorname{tr}(A)$ for any positive semi-definite Hermitian matrices A,B and obtain

$$
\begin{aligned}
\left|\log\det\left(I+\mathsf{U}_0+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*\right)-\log\det\left(I+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*\right)\right| &\leq \operatorname{tr}((I+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*)^{-1}\mathsf{U}_0)\\
&\leq\operatorname{tr}(\mathsf{U}_0)\\
&=\operatorname{tr}(\widetilde{V}_0)\\
&\leq\rho\,\operatorname{tr}(\mathsf{F}_0^*\mathsf{F}_0)\\
&\leq\rho\min(K,N)\|F_0\|^2.
\end{aligned}
$$

Using the moment assumption (4), this yields

$$\mathcal{I}_\rho = \frac{1}{(M+1)N}\,\mathbb{E}\log\det(I+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*) + \mathcal{O}(1/M)$$

where $\mathcal{O}(1/M)$ is uniform in K, N. The same time of estimates yield that one can replace $\hat{\mathsf{H}}_{0,M}$ by $\mathring{\mathsf{H}}_{0,M}$ up to a $\mathcal{O}(1/M)$ correction, namely

$$\frac{1}{(M+1)N}\,\mathbb{E}\log\det(I+\rho\,\hat{\mathsf{H}}_{0,M}\hat{\mathsf{H}}_{0,M}^*) = \frac{1}{(M+1)N}\,\mathbb{E}\log\det(I+\rho\,\mathring{\mathsf{H}}_{0,M}\mathring{\mathsf{H}}_{0,M}^*) + \mathcal{O}(1/M)$$

with $\mathcal{O}(1/M)$ uniform in K, N, and the lemma is proven. □

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